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EDITORIAL

We have much pleasure in presenting this Centenary Number of the Journal of the Madras University, B (Science Series). In another few weeks this university will be celebrating its first centenary along with the sister universities of Calcutta and Bombay. Our minds naturally go back to those men of vision who launched these three universities in 1857, in the year of the Great Indian Mutiny. We salute them for their sagacity. In the comparative political quiet of 1957 we consider ourselves fortunate in being able to celebrate this occasion in Free India with this special Science number with invited articles from eminent scientists in India and abroad. When our invitations went out hardly five months ago, I had some misgivings about the response we were likely to have. Many distinguished alumni and well wishers of India and Indian science have responded generously to our invitation. It is my pleasant duty to offer on behalf of this university my most sincere thanks to all those whose articles are now before you. Many of our distinguished invitees have written articles in their own names, others have sent joint articles with their associates. In the latter cases, the names of the invitees in each case are indicated by an asterisk.

I shall be failing in my duty if I do not thank my colleagues on the editorial board and my own colleagues in this laboratory for the readiness with which they gave their time in many editorial matters. The Press deserves my special thanks for cordiality and co-operation and I offer it with pleasure.

University Botany Laboratory,
December 1956.

EDITOR

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J. Madras Univ., B, 27, No. 1. Centenary Number. 1957.

The Optical Behaviour of Polycrystalline Solids

BY

Sir C. V. Raman

(Memoir No. 89 from the Raman Research Institute, Bangalore 6)

ABSTRACT

The paper discusses the influence of the factors which determine the penetration of light into a polycrystalline solid, viz. the birefringence of the crystallites, their size and shape and their relative orientations. The results of the theoretical discussions are compared with the facts of observation. Five X-ray diagrams are also reproduced which exhibit the relation between the structure and optical behaviour of polycrystalline media in particular cases of interest.

1. *Introduction.*

As examples of polycrystalline solids which by reason of the desirable properties they possess have been extensively used for the fashioning of works of art may be mentioned especially the following minerals: marble, alabaster, chalcedony and jade. Visual observations supplemented, where necessary, by examination of thin slices under the polarisation microscope and the use of X-ray diffraction techniques reveal that individual specimens of these minerals may be widely different in regard to the size and shape of the crystallites of which they are composed and the manner in which these crystallites are disposed within the solid. Alongside of these differences and obviously related to them are noticed some striking differences in the optical characters of the specimens. As the external appearance of a finished work of art is influenced a good deal by such differences, it is not without interest to discuss in some detail the nature of the relationship between the structure and the optical behaviour of polycrystalline solids in general.

In all the cases mentioned, the basic substance is a transparent crystalline solid and accordingly the aggregate could also have been expected to be free from colour. Though this is so in some instances, it is not always the case. Indeed, the presence of some favoured colours often enhances the value of the material very greatly. The nature of the intrusive material res-

ponsible for such colour and the manner in which it is diffused through the substance and influences its optical behaviour are questions of considerable interest, but we shall not enter into them here.

It should be mentioned that the structure of polycrystalline solids and its relation to their optical behaviour has been the subject of both theoretical and experimental investigations by the present writer and his collaborators and the results have been reported in detail in some earlier memoirs of this Institute. In the present publication it is proposed to summarise broadly the results which have emerged from these studies and in addition to present some fresh observational material.

The five illustrations accompanying the present paper are X-ray-diffraction photographs of some interesting specimens recorded by Mr. A. Jayaraman at this Institute. In each case, the structure revealed by the X-ray diagram is found to be very clearly related to the optical characters exhibited by the specimen. We shall return to these diagrams and their significance later in the paper.

2. *Theoretical Considerations.*

If no extraneous material be present in a polycrystalline solid and its optical heterogeneity arises solely by reason of the varying orientations of the crystallites, the birefringences of the latter would give rise to a diffusion of the light in its passage through the material. At each incidence of the light on an intercrystalline boundary, it would be reflected and refracted and hence all trace of the original direction of propagation of the light would quickly disappear. After the passage of the light through a sufficient number of crystallites, a state of affairs would establish itself in which we have a diffusion in all directions, forwards and backwards as well as laterally. The greater the birefringence of the material, the fewer would be the number of intercrystalline boundaries needed for such a situation to establish itself. We may visualize the resulting phenomena by imagining a pencil of light to fall normally on a restricted area in the plane face of a block of the material. The observable results would then depend upon the various factors involved, namely the thickness of the block, the size of the crystallites and the magnitude of their birefringence.

In the absence of any absorbing material and if the thickness of the block be sufficiently great, all of the incident light must necessarily be turned back towards the original source of light and would re-emerge from the same face spread out from a much larger area than that on which it is originally incident. On the other hand, if the thickness of the block be not too great, part of the incident radiation would also emerge from the rear face of the block spread out over a large area. In each case, the observable effects would be profoundly influenced by the size of the crystallites. The smaller the crystallites are, the more quickly would the incident light be turned back towards the source and the less, therefore, would be the depth of the penetration of the light into the medium. The presence of any absorbing material in the solid would necessarily modify the observable results to a very great extent.

3. *Diffraction Phenomena.*

While the arguments set forth above based on geometrical optics may be adequate when the crystallites are of sufficiently large dimensions and the birefringence of the material is also fairly large, they would cease to be adequate and would fail to describe what is actually observed when the reverse is the case, i.e., when the crystallites are of small dimensions and the birefringence which they exhibit is also small. In the latter circumstances, the matter has necessarily to be considered from the standpoint of the wave theory of light, and the results indicated by such theory are very different indeed from those indicated by geometrical optics.

We may here broadly indicate the lines on which the problem can be handled on the basis of the wave theory. If the birefringence of the material be sufficiently small, the diffusion of light in its passage through it may be considered as arising from the varying retardations in phase of the light waves in traversing the differently oriented crystallites. Provided that the product of the linear dimension of each crystallite and of its birefringence is sufficiently small, we can regard the waves as moving undisturbed through the medium except for the changes of phase occurring over different areas of the wave-front; such changes may be regarded as being additive in depth and hence capable of being summed up over the whole of the path through a slab of the material, provided that the thickness of the latter is not too great. The wave-

front emerging at the rear surface of the block would thus exhibit fluctuations in phase over its area. The final observable effect may then be determined by the usual methods of diffraction theory.

The foregoing brief exposition ignores certain difficulties, namely, that inside each crystallite there would be two possible directions of vibration for which the refractive indices are different and these two directions would vary from crystallite to crystallite owing to their random orientations. This difficulty, however, may be circumvented by assuming that the light incident on the block is plane-polarised and that all the crystallites are so orientated that this original direction of vibration in the wave-front is preserved in its passage through the block, the operative refractive index for any one crystallite being one of its three principal values chosen at random. Such an assumption has the great advantage that it greatly simplifies the formulation and solution of the problem while retaining its essential features. Further, by assuming the probabilities of the three different orientations to be different, we can also determine the optical behaviour of polycrystalline solids in which the crystallites are not orientated at random but prefer certain directions. What the solution of the problem gives us is the intensity of the light regularly transmitted through the polycrystalline plate. This appears as an exponential function in which the argument has a negative sign and is directly proportional to the thickness of the plate and to the size of the crystallites and also to a quantity depending on the differences between the three refractive indices which is a measure of the birefringence; it is also inversely proportional to the square of the wave-length of the light. From the theory it follows that the light emerging from the block would be polarised in the same fashion as the light incident on the first face. The theory also shows that if the crystallites have no preferred orientation, the transmission coefficient is independent of the direction of vibration in the incident light. If, however, the crystallites are preferentially orientated, the intensity of the light emerging from the plate would vary with the circumstances, viz., with the relative probabilities of the three permitted orientations. In particular cases of this kind, e.g., when all the crystallites are orientated along a fibre axis which is also the optical vibration direction, the transmission may be actually complete.

We may summarise the foregoing by stating that on the basis of the assumptions made, the light incident on a plate of polycrystalline material would emerge from it greatly enfeebled in

intensity but retaining its primitive state of polarisation. The fraction of the light regularly transmitted for a given thickness of the material would increase with diminishing size of the crystallites and would reach the limiting value of unity if they are sufficiently small. This result is precisely the opposite of that indicated by the geometrical theory according to which diminishing particle size would result in greater opacity. These two contrasting results may be regarded as valid respectively in the limiting cases of a very weak and of a very strong birefringence.

The energy that disappears from the regularly transmitted light would appear as diffracted radiation. Its distribution of intensity in different directions would be determined by the factors entering into the problem, namely, the size of the crystallites, the thickness of the plate, the birefringence of the material and finally the wave-length of the light. It is a consequence of the special assumptions made regarding the orientation of the crystallites that the diffracted radiation would be polarised in the same manner as the incident and emerging pencils. It is not, however, to be expected that this particular result would represent the actual facts of the case. As a consequence of the unrestricted randomness of the orientation of the crystallites, the diffused or diffracted radiation would necessarily be depolarised to a greater or less extent. The greater the birefringence of the material or the thickness of the plate traversed, the more complete would be such depolarisation.

It is a particular consequence of the theory that if unpolarised light be incident on a plate of material having a fibrous structure and of sufficient thickness, the regularly transmitted beam would be completely polarised on emergence, while the rest of the light would appear as a fan of diffracted rays in the plane perpendicular to the fibre axis and polarised predominantly in the perpendicular direction. It has also to be remarked that a fibrous structure of the medium is favourable to the appearance of a regularly transmitted pencil. The same material with crystallites randomly orientated might not exhibit any transmission at all, but only a diffusion of light in various directions surrounding that of the incident pencil. Finally, we may remark that since these phenomena arise by reason of wave-optical considerations depending on the wave-length of the light, the transmitted light whenever it is observable should exhibit a reddening which increases progressively with the thickness of the material traversed.

4. *Some Illustrations of the Theory.*

Marble which in its purest forms consists of calcite furnishes excellent illustrations of the optical behaviour of a polycrystalline medium consisting of strongly birefringent crystallites. The influence of the grain size on the penetration of light into the material indicated by the theory is readily confirmed by observations with different specimens. A very interesting fact is that the light diffused by marble both backwards and forwards exhibits a surprisingly rich redness even when the marble itself is only of a faintly pinkish colour.

Alabaster which consists of randomly orientated crystallites of gypsum affords good illustrations of the behaviour of a polycrystalline medium in which the crystallites have only a feeble birefringence. The size of the crystallites can of course be readily determined by cutting a thin section and viewing it under a polarising microscope. The relationship between the size of the crystallites and the optical behaviour of the material can be readily confirmed by viewing a bright source of light through a slice of alabaster mounted between microscope cover-slips with a little Canada balsam. Only with the thinnest sections is a true optical transmission observable. This appears surrounded by a diffusion halo exhibiting streaky colours in white light. The depolarisation of the halo and the polarisation of the regularly transmitted light when the incident light is polarised are also readily confirmed by observation.

The X-ray diffraction diagram of alabaster depends very much on the size of the crystallites which it contains. Unless the alabaster is very fine-grained, a collection of spots only vaguely indicative of a ring pattern is obtained. Very different is the picture shown in Fig. 3 in Plate I which shows a readily recognizable ring system, though spotty in character. This was obtained with a very fine museum piece exhibiting a remarkably high degree of translucency.

Fig. 2 in Plate I is an X-ray diagram recorded with a dish of Chinese origin stated to be of jade, but not really so. The material was colourless and exhibited a high degree of translucency. The extremely small size of the crystallites of which it is composed is evident from the X-ray diagram.

Fig. 1 in the same Plate is the X-ray diagram obtained by passing a beam of X-rays through a plate of a white porcelain-like

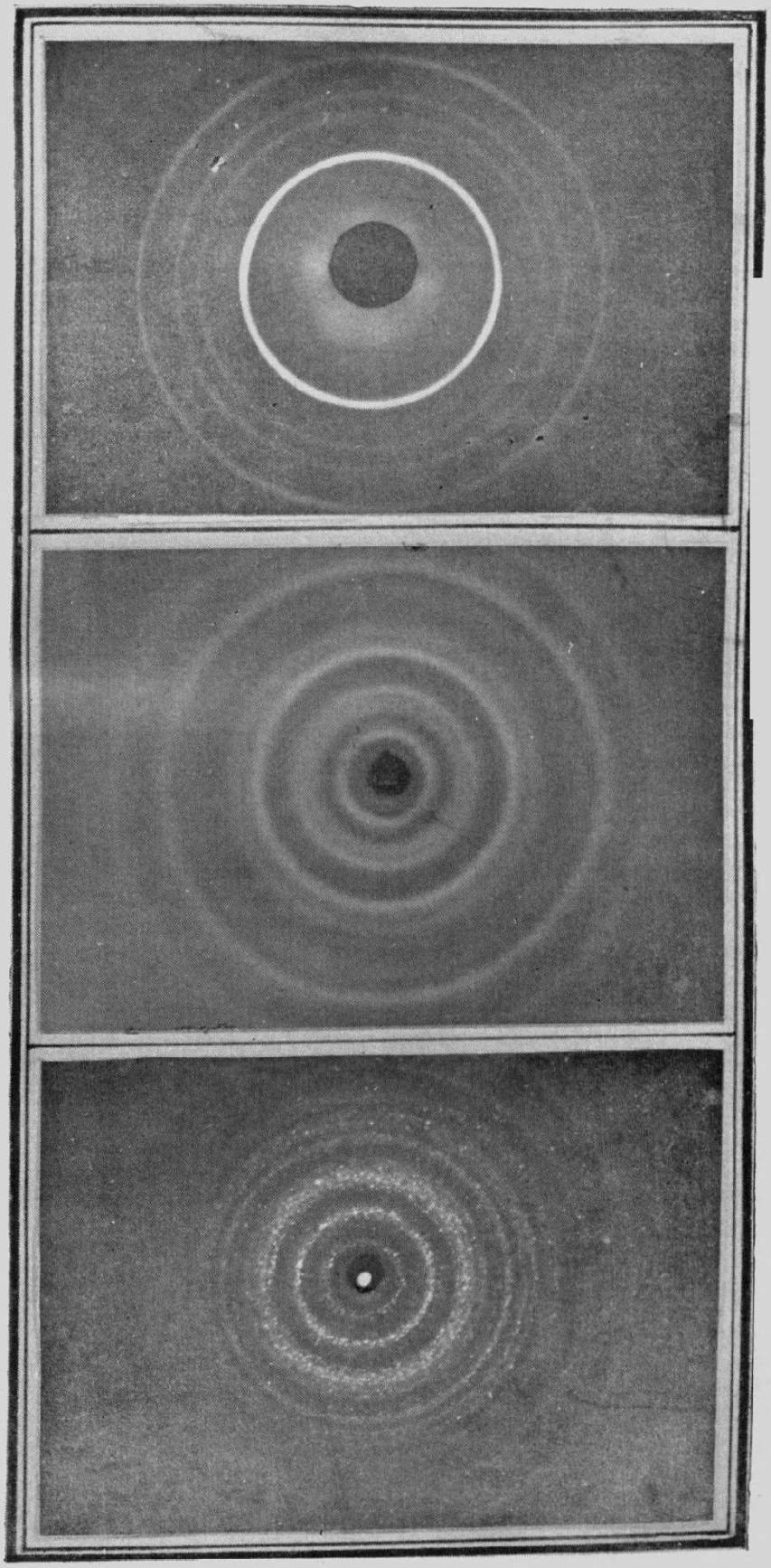

Fig. 1

Fig. 2

Fig. 3

C. V. RAMAN

material which was cut out from a block of agate and which was identified as α-cristobalite by its X-ray pattern. The extreme fineness of the particles of the material is evident from the X-ray diagram. A source of white light seen through the plate is visible, but appears of a deep red colour. The material thus furnishes an excellent example of a polycrystalline medium consisting of a birefringent material which nevertheless exhibits a true optical transmission by reason of the small size of the crystallites.

The phenomena arising from the propagation of light through a medium with a fibrous structure are beautifully illustrated by observing a source of light through a polished plate of chalcedony. They exhibit in a very conspicuous manner the various features indicated by the theory. It should be emphasized that chalcedony exhibits large variations in its structure and these correspond to striking differences in optical behaviour. Two extreme cases are illustrated respectively by Figs. 4 and 5 in Plate II. The first is an X-ray diagram recorded with a specimen of agate which exhibited only translucency but no real transmission of light. It will be seen from the X-ray diagram that this material consisted of extremely fine particles. Fig. 5, on the other hand, is an X-ray diagram of a specimen of chalcedony showing a coarsely fibrous structure. A source of light could easily be seen in perfect focus through the plate though much dimmed in intensity and reddened in colour. The transmitted light was found to be completely polarised, and this was evidently a consequence of the fibrous structure revealed by the X-ray pattern and also evidenced by the streams of diffracted light seen surrounding the image of the source.

J. Madras Univ., B, 27, No. 1. Centenary Number, 1957.

Some Experiments Aiming at an Isolation of Biochemical Mutations in *Pisum sativum* L.

BY

NILS FRIES

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ABSTRACT

The possibility of isolating biochemical mutations of the pea plant (*Pisum sativum* L.) after X-ray irradiation of the seeds was investigated with the aid of a root culture test method. Roots from seedlings representing the segregating X_2 generation were cultivated aseptically and their nutritional requirements tested. In 3 cases out of 75 the root clone required some additional factor present in the complete medium. After some time, however, all three clones changed back to wild type. In one of these cases it could be established that the necessary factor was a component of hydrolysed casein.

Introduction

During the last fifteen years the study of so-called biochemical mutations in lower heterotrophic plants have produced results of great interest to plant physiology, biochemistry and genetics. Efforts to isolate similar types of mutations in green plants, have, so far, been less successful. In seed plants, where such mutations, if present, would provide an extremely interesting material for various sorts of studies, only one clear-cut case has been reported, *viz.*, a thiamin-less mutant of *Arabis thaliana* (Langridge, 1955). In the present paper some experiments with the pea plant are reported which, although essentially negative, may be of some interest to other workers intending to follow the same track.

Material

The seeds for the experiments were obtained from Sveriges Utsädesförening, the Ultuna branch near Uppsala, and represented the "Torsdag III" strain of the ordinary pea plant, *Pisum sativum* L. They have been produced by X_1 plants derived from seeds treated with 6000 r and then sown in the field. Thus, the seedlings,

plants and excised roots of the experiments to be described represented the X_2 generation, in which homozygous mutant individuals, if they existed, ought to be found.

Methods

The procedure applied was based upon the following reasoning:

If a biochemical mutation has been induced in one of the X-ray irradiated seeds a certain number (at the most 25 per cent) of the X_2 plants should represent homozygous mutants. Each one of these plants should, therefore, be incapable of growing, unless supplied with a particular organic compound, a metabolite, the biosynthesis of which the mutant is unable to carry out. For obvious reasons this metabolite must be supplied under aseptic conditions. However, since the cultivation of whole pea plants aseptically presents great technical difficulties, the testing could be performed by means of root cultures. This seems justified if the metabolite in question is indispensable for every cell of the plant.

The rest of the plant could be planted in soil under ordinary green-house or field conditions, where it should be able to grow by means of adventitious roots and the remaining part of the root system. If it was a biochemical mutant its fate would probably depend on the cotyledons: the organic material stored there might be sufficient for a development of the plant to maturity, if the necessary metabolite was contained in this material, or it might not. In the former case homozygous mutant seeds would be obtained.

One seed from each X_1 plant was tested according to the principles outlined above. The seeds were chemically sterilized with ethyl alcohol (70 per cent) for 10 minutes, with mercury chloride (0·1 per cent water solution) for 15 minutes, and, finally, with calcium hypochlorite (saturated solution) for 30 minutes. They were washed with distilled water before and after the treatment with hypochlorite. After sterilization the seeds were soaked in distilled water for 18 hours and then placed for germination on plain water agar plates at $+20°C$ in darkness.

As soon as a 30 mm long root had been developed its end half was excised and aseptically transferred to a small cork float (Fig. 1) in a 100 ml Erlenmeyer flask with 20 ml of Robbins' (1946)

nutrient solution for root cultures, modified according to Fries (1954) and supplemented with the following additions per flask: 1 mg yeast extract (Difco), 2 mg hydrolyzed casein (NZ-case, Sheffield Farms Co., New York), and 2·5 mg lyophilized pea root exudate. Preliminary experiments had shown that these additions did not affect the growth of normal pea roots appreciably in any direction. The cultures were kept at +25°C in darkness.

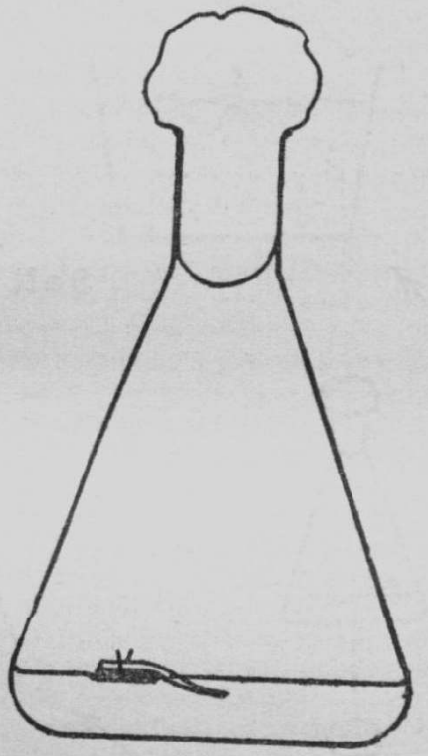

FIG. 1. The excised root placed on a float, consisting of a small cork disc covered with a thin layer of paraffin wax. Two thin chromenickel needles keep the root in position. The float is swimming on the surface of the nutrient solution.

The rest of the seedling was planted in a pot with garden soil and placed in a green-house.

After a few weeks the excised roots had increased in length and developed a number of lateral roots. Ten lateral roots, each attached to a small piece of the main axis, were cut out and transferred to new flasks with Robbins' solution: five without and five with the additions mentioned above. The former solution was called "minimal medium", the latter "complete medium". Only c. 10 mm long laterals were used, and one lateral was placed in each flask.

This treatment was performed with each of 75 seedlings obtained from just as many X_1 plants. The procedure is schematically outlined in Fig. 2.

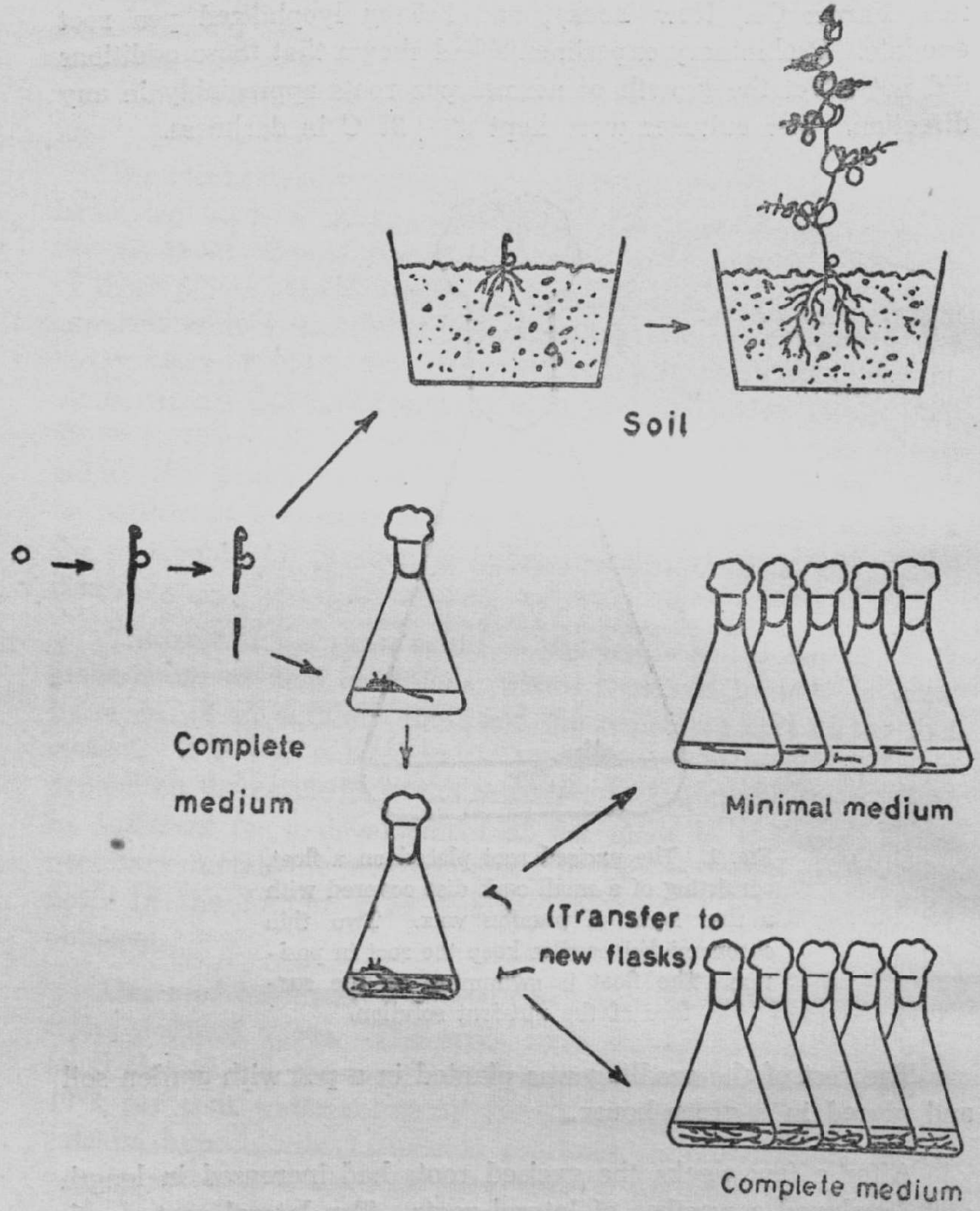

FIG. 2. Schematic illustration of the procedure used for testing the physiological properties of X_2 seedlings of *Pisum sativum*. The tested seed is assumed to be a biochemical mutant requiring a factor contained in the complete medium. The growth of the green part of the plant is limited by the amount of this factor present in the cotyledons, which in the illustrated case is assumed to enable a certain development of the plant. For further particulars see the text.

Results

Out of 75 tested root systems only three reacted in a way that differed from the normal (Table I). The laterals from these three root systems, *viz.*, Nos. 3, 6 and 33, grew only in the medium to which amino acids and other organic metabolites had been added. Since this behaviour could indicate a loss in the biosynthetic capacities of the roots in question, these root systems were further investigated.

TABLE I

Growth of root and shoot	Number of X_2 plants
Root growth in complete medium only; no shoot growth	.. 2 (= Nos. 6 and 33)
Root growth in complete medium only; normal shoot growth	.. 1 (= No. 3)
Root growth also in minimal medium; no shoot growth	.. 1
Root growth also in minimal medium; normal shoot growth	.. 71

From the five cultures on complete medium, laterals were transferred to four series of flasks, (a), (b), (c) and (d), each series comprising five flasks:

(a) Minimal medium;

(b) same as (a), plus NZ-case (2 mg per flask), to which certain L-amino acids had been added (serine, cysteine, alanine, tryptophane, γ—amino butyric acid, aspartic acid, asparagine, and proline);

(c) same as (a), plus a mixture of seven vitamins (riboflavine, pantothenic acid, *p*-aminobenzoic acid, biotin, inositol, ascorbic acid, and folic acid);

(d) complete medium.

In this experiment the growth of Nos. 6 and 33 was quite satisfactory in the minimal medium and not appreciably inferior to that

in the complete medium. The physiological properties of these two root clones had obviously changed and seemed to agree with those of the normal, wild type, as far as the nutritional requirements were concerned. No. 3 also produced a rather good growth in the control series (a), contrary to the result of the very first testing experiment, but the growth in series (d) was much better than in (a), especially in terms of lateral root production and dry weight values. The best growth, however, was obtained in series (b) (Table II).

TABLE II

The growth of roots from seedling No. 3 in liquid minimal medium with various additions

Time of incubation 3 weeks. The figures in the table represent average values from 5 or 6 root cultures.

Medium	Length of main root *mm*	Total length of lateral roots per root system *mm*	Dry weight *mg*
(a) Minimal (= control)	123	244	7·0
(b) Minimal + NZ-case + amino acids	147	520	15·2
(c) Minimal + vitamins	111	275	9·2
(d) Complete	135	404	11·9

Further experiments were performed only with the root clone of No. 3. Amino acids were tested in groups of four, added to the minimal medium, and the effect was compared with that of NZ-case. In no case, however, did the combination of amino acids improve growth, while the effect of NZ-case was considerable.

When NZ-case was compared with Difco hydrolyzed casein this latter preparation proved to be the more active of the two (Table III). Indole acetic acid produced a weak, not quite significant effect. The growth rate in this experiment was generally lower than earlier.

TABLE III

The growth of roots from seedling No. 3 in liquid minimal medium with additions of hydrolyzed caseins and indole acetic acid.

Time of incubation 3 weeks. The figures of the table represent average values from 10 root cultures.

Added substances µg per culture	Length of main root *mm*	Total length of lateral roots per root system *mm*	Dry weight *mg*
No additions	39	40	1·7
NZ-case hydrolysate, 200	64	155	3·7
„ „ , 2000	77	134	4·9
Difco hydrolysate, 200	52	116	4·1
„ „ , 2000	128	432	13·9
Indole acetic acid, 0·002	45	51	2·4
„ „ „ , 0·02	37	68	2·7

However, the investigation had then to be discontinued, since the next experiment (with graded dosages of Difco casein hydrolysate) showed that root clone No. 3 finally had reverted to wild type, like earlier Nos. 6 and 33. The growth in the control cultures was quite normal and not improved by casein hydrolysate, nor by yeast extract and pea root exudate.

The shoot belonging to root system No. 3 produced a few seeds in the soil culture, but it did not seem to be worth while to examine their nutritional requirements in aseptic culture. The green parts of Nos. 6 and 33 died at an early stage of development and did not produce any flowers.

Conclusions

Root systems developed from seeds of X_1 plants showed, as a rule, a quite normal behaviour in liquid culture. In a few cases only, the roots showed particular nutritional requirements during the first transfer or transfers. They needed, at least for good growth, some organic factor or factors included in the complete medium. In the case of root No. 3 the factor in question was a constituent of casein hydrolysate. A definitive identification of the nutritional requirements could not be carried out, since the roots after some time changed back to wild type.

The material is, of course, too small to permit the conclusion that mutations of a physiological type had been produced through the X-ray irradiation. By testing a larger number of X_2 seedlings and a corresponding number of seedlings derived from untreated ancestors it should, however, be possible to decide about this question.

It is probable that the three root clones, Nos. 3, 6 and 33, which originally proved unable to grow in the minimal medium, differed genotypically from the wild type. The great variation generally found in ordinary seed assortments of various plant species as regards the response to nutrient media is usually interpreted in this way. As a matter of fact, a successful cultivation of a species as a root clone very often starts with a sorting out and removal of those roots which show no or poor growth (Naylor and Rappaport, 1950). From this point of view, the material of the above experiments responded in a surprisingly uniform way despite the preceding X-ray treatment. This may be due to the fact that the "Torsdag III" strain used is a rather recently produced pure line.

If the three deviating root clones should be looked upon as mutants—spontaneous or induced—the sudden and deplorable reversions to wild type ought to be called back-mutations. However, in view of the well-known, as yet unexplained, variability of root cultures in the course of time (White, 1943) caution in interpreting these phenomena is warranted.

Acknowledgements

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Problematics of the Phenomenon of Phenocopy

BY

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An individual with an aberrant, non-hereditary type, identical phenotypically with that of a known mutant, is called a phenocopy. It has become a habit recently to call any non-hereditary modification, abnormality, morphose, monstrosity a phenocopy. This is wrong, as the original definition (and literal meaning) of the term restricts it to "copies" of a mutant. It is an old genetical experience that quantitative mutant characters may overlap with non-hereditary modifications of the same phenotype e.g., in Johannsen's classic experiments with pure lines of beans a plus individual of a small line may be as large or larger than a minus individual of a hereditary large line. A systematic study of this latter phenomenon was first made by Zeleny in a series of classical papers on the facet numbers of the series of Bar eye mutants of Drosophila. He showed that there is a temperature sensitive period during which different temperatures affect the number of facets in a Bar eye and that this number may be pushed by temperature action to the low level shown by the higher mutant ultrabar. Zeleny (1923) called this, a demonstration "of the fact that the gene ultrabar has the same type of reaction as a temperature difference. It is possible to state the effectiveness of particular germinal factors in terms of corresponding effects of temperature." This statement implies that mutant actions attack the same developmental processes in a quantitative way as temperature actions do. Actually I had drawn this conclusion already from different kinds of experiments (Goldschmidt, 1920), namely, the old temperature experiments with butterflies of Fisher (1895). Here a temperature action (a shock), applied during a definite sensitive period, could produce a wing pattern identical with that of a known geographic variety. Nowadays one could call this effect a phenocopy. I had drawn conclusions upon the identical actions in modification and mutation which were identical with those of Zeleny, and I had gone a step further in concluding that the reason of the identical effects is to be sought in the nature of genic action, namely by the control

of rates of some determining process. In a later summary of his work Zeleny (1933) joined me in these conclusions.

There is some other early work which went in the same direction. Kosminsky (1909) had shown that male pupae of Lymantria treated with heat shocks developed a double uncus (the hook in the genital armature) which is single normally, though concresced from a double anlage. Later when genetic intersexuality was discovered, I could show (1922) that this effect was identical with that of genetic intersexuality of a definite grade. Though this is not simply a mutant condition, it is genetically controlled and may be compared therefore with mutant action in the present context. It should not be difficult to find other comparable cases in which the same phenotype is known as a chance variation (though not experimentally induced), which is identical with that of a mutant. Such a parallel effect is expected, where a mutant acts embryologically by stopping prematurely a normal developmental process. Harelip, the lack of concrescence of the upper jaw anlagen, is one example, hypospadia penis is another one.

To a certain extent embryological work with vertebrates may be quoted in this connection, following the important work of Stockard (1910 ff). By treatment of embryos of different age with chemicals or abnormal temperatures, or X-rays (Little and Bagg, 1924) all kinds of monstrosities could be produced in a more or less regular way (full review by Ancel, 1950). In this work there was no emphasis on similarity with known mutants. But as many monstrosities in vertebrates are known to be hereditary, by necessity these induced monstrosities included also what are now called phenocopies.

All of this early work was of importance for problems of growth and embryonic determination and was discussed as such. But only the above quoted work of Zeleny (1923) and Goldschmidt (1920, 1922) attacked already the genetical problem, the latter (1920) already using Stockard's experiments for his analysis. This changed when the discovery was made that one and the same environmental action (heat shocks) could lead to the production of non-hereditary imitations of practically all known morphological mutants of Drosophila (Goldschmidt, 1929). A repetition on a larger scale and going into all the important phases of the problem (1935) showed beyond doubt the great importance of the phenomenon for many basic problems of genetics. The name "phenocopy" was proposed for the non-hereditary copies of the pheno-

type of a mutant. Since that time considerable literature has accumulated and new discoveries were made which not only enhanced the importance of the facts but also put new questions, which have not yet been answered. The most important of these findings is Rapoport's (1947) demonstration of the production of specific phenocopies in Drosophila by chemical treatment of the larvae during a sensitive period. The most elaborate repetition of the Drosophila work using heat shocks was done by Henke and students (1941) with essentially the same results as ours but adding many details and refinements.

The most important groups of facts which have to be taken into account when searching for a general explanation may be recounted, without quoting the individual contributions.

1. Almost all treatments of eggs and larvae of Drosophila which have been tried produce phenocopies e.g., temperature shocks, X-rays, neutrons, ultra-violet rays, many chemicals like ether, tryptophane, mustards, metal salts and many others. In chicken insulin (a protein) and a number of chemicals, including metal salts were effective and in mammals many types of lesions of the same type were successful. A much discussed phenocopy in man is the appearance of certain monstrosities in babies born from mothers who had German measles during pregnancy. At the other end of the phylogenetic tree are chemical phenocopies produced in Neurospora and bacteria.

2. In Drosophila all known types of morphological mutants of all organs have been copied in such experiments; but no reliable case of copying of biochemical mutants is known, though this has been produced in Neurospora. But quantity changes of pigments (darker or lighter) can be produced, which is theoretically significant.

3. Chemical treatment acts more specifically than radiations and temperature shocks do. In a few cases it seems that one specific mutant is copied after one specific treatment; but in view of recent work (see 8 and 9) this is not as simple as it appeared at first.

4. All treatments are most successful near the limit of lethal action and they may produce up to 100% phenocopies under proper conditions. Correlation of percentage of phenocopies with degree of lethality is observed, but does not occur in all cases.

5. All treatments act to some extent proportionally to intensity (dose, concentration, etc.).

6. All treatments are effective only during one or more than one sensitive periods of development. This is in most cases identical with the time of labile determination of the affected organ in development, as determined by effectiveness or failure in producing shifts of the phenotype of a wild type or a mutant in embryological experimentation.

7. In Drosophila the effect may be independent in either half of the body (no right-left correlation) and the same organ may form different phenocopies on the right and left side. One and the same imaginal disc may react as a whole or its parts may react separately. Thus an antenna may be duplicated (a phenocopy) and the two aristae may be different, one normal, and one a phenocopy. The effects always appear in different grades of expression which form a series like a multiple allelic mutant series, usually correlated to the intensity of the treatment. It might repeat a known multiple allelic series or contain more extreme members, which may be supposed to be lethal as mutants.

8. The quantity (penetrance and expressivity) of the phenocopic effect is dependent upon the genetic constitution of the treated line. In Drosophila wild type lines of greatly different sensitivity exist. The genetic basis is found in one or a few major loci and many modifiers and in mixed lines selection is successful. Comparable facts have been found also in chicken on a smaller scale.

9. The quality of the effect is also dependent upon the genetic constitution of the material. Each line has a specific pattern of phenocopies produced by the same treatment and some lines produce typically one definite phenocopy not found in others. Also the relative frequencies of the different phenocopies characterize each genetic line. Points 8 and 9 are best demonstrable for chemically induced phenocopies, and the most extensive data are derived from recent Drosophila work.

10. In chicken a correlation was found between the propensity of different stocks for production of spontaneous non-hereditary abnormalities of the same type as the mutants and induced phenocopies, and the sensitivity to phenocopic treatment of the line. But the experience in Drosophila is that the apparent non-hereditary variants of rare occurrence are not modifications but low penetrant mutants (see below).

11. It is frequently found that the phenocopic agents, when applied to mutants, either enhance their expression and also their

penetrance, if not completely penetrant, or they decrease their expression towards greater normality.

12. Cases have been found in which mutants which had lost their visible effect (by accumulation of minus modifiers) produced their typical effect after phenocopic treatment.

13. Phenocopic treatment of heterozygotes +/mutant brings out (if successful) the recessive character, or, more correctly, increases the effect of the treatment in regard to the phenocopy resembling the mutant type.

(The bibliography of the enumerated results is found in Goldschmidt (1938, 1955), Goldschmidt and Piternick (in press), Landauer (in press) and for vertebrate monstrosities in general without much attention to phenocopy in Ancel (1950)).

It is obvious that such facts pose many problems the solution of which may have far-reaching consequences. The first of these is to explain why in these experiments the mutant phenotype appears and not some irregular monstrosity. From the very beginning of this kind of work it was clear (Goldschmidt, 1920; Zeleny, 1923) that the simplest assumption would be that both the mutant locus and the shock (using this word for all phenocopic agents) affect normally occuring developmental reactions in a quantitative way by speeding or retarding them relative to the other not affected processes. The facts which have become known since require some such rather simple and generalized basic explanation. It would be hardly possible that such different agents used in producing phenocopies could have identical effects if they acted by producing specific substances or interfering with the quality of specific reactions or steps. But within this generalized rate shifting action many possibilities are given, actually demanded, if we want to account for the most disconcerting facts, the genetic control of quantity, quality and specificity of the effect (thus far studied only in chemically induced phenocopies). One could appeal in a general way to a modifier system for the details of processes of development which is different in different lines. Taking a single developmental process e.g., the differentiation of a part of the antennal disc into an arista (in Drosophila) all the processes of growth and differentiation must be properly integrated to produce a normal arista. The integration is under genic control and therefore liable to shifts if the combination of controlling loci varies in the usual way. In normal conditions the collaboration of this "modifier system" varies only within the limits for normal

development, though there might be variations which make one or the other part of the system come near to the point where normal development is not possible any more. Within such a system it is imaginable that such "weak points" are present for one or more definite developmental steps, differently in diverse genetic lines. A shock treatment will push the calibrated system out of gear at such points and abnormal development results for a specific process of differentiation. It is a different problem why, at least in Drosophila and also the fowl, this abnormality is identical with a mutant. One might say that integrated aberrant development is possible only in definite directions. We may use the simile of notches into which developmental processes must snap in order to succeed in the production of something orderly. These notches are the same whether the cause of the aberration is a shock or the activity of a mutant locus.

Landauer has tried to support an idea of this general type (though expressed in different language) by pointing out that in his experiments lines in which the phenocopy (e.g. rumplessness) appeared "spontaneously" reacted better to the phenocopic treatment with insulin or boric acid; and in addition in proportion to the incidence of the spontaneous variant. This could mean that these lines have already in their developmental system what I called above a "weak point," based upon a definite genetic constitution. But how can it be proved that the genetic basis of the "spontaneous phenocopies" is a definite constellation of modifiers and not rather the presence of a mutant of very low penetrance? We have studied in the course of time many mutants of low penetrance with monogenic or polygenic basis. Some of them can be selected for high penetrance, others cannot be shifted beyond a certain level which might be as low as less than one per cent. Without special inquiry they would be described as non-hereditary rare variations. As an example I might mention the concrescence of the first legs of Drosophila into a ventral, unpaired organ resembling an orthopteran labium. Some types of this variant were described long ago by Morgan *et al.* as a non-hereditary abnormality; but I found it later as a mutant of low penetrance, probably on a polygenic basis. Another example was found in *Lymantria dispar*, a case of prothetely, namely a pupal antenna in caterpillars, a rare anomaly, but once encountered as a low penetrant mutant. Thus I am inclined to consider all similar cases as mutants with low penetrance. If now, as in Landauer's case, a line producing such anomalies "spontaneously" is more liable to

produce also the phenocopy of the same phenotype; and if further a correlation exists between the percentage incidence of the "variant" and the sensitivity for producing it as a phenocopy, the suspicion arises that the phenocopy, in this case, is an enhancement of a low penetrant mutant.

We studied a number of cases in which a mutant of low penetrance can be enhanced by a phenocopic agent and such a case was also studied by Plaine and Glass (1955). In addition it is known that low penetrance may be so low that it amounts practically to absence. Further we found that a mutant which had lost completely its penetrance, probably by selection of minus modifiers may be restored to visibility by treatment with a shock which produces the same phenotype as a phenocopy (of course, this recovery is not hereditary). Taking all this together the hypothesis can hardly be avoided that all phenocopies are due to a bringing into light of already present non-penetrant, subthreshold or isoallelic mutants.

But a proof for such a hypothesis is very difficult. We have tried many ways (Goldschmidt and Piternick, in press); but in spite of a number of encouraging results thus far no fact has been established which could be considered a cogent proof. Actually such a proof would require the demonstration of a monogenic, subthreshold mutant for each successful and specific phenocopic action. But if the genetic basis were a polygenic condition the genetic element in the reaction to phenocopic treatment either would be called a set of subthreshold mutants, or, just as well, a condition of the modifier system controlling the set of different developmental conditions which permit a shift away from normal, as discussed above. Thus far we have not yet succeeded to overcome these experimental and logical difficulties.

It is obvious that a solution of such problems would greatly influence our ideas about genic action and the nature of mutation. Already at this point it becomes visible that the genic control of morphological development cannot be exercised by means of specific enzymatic actions but must require a system in which rates and timings of events play the preponderant role. This does not exclude the participation of enzymes in the setting of the proper order in the interwoven and interlocking developmental reactions e.g., by substrate affinities and competitions for substrates. Thus far I know only of one hint towards chemical identification of the actual processes involved, both in phenocopic and mutant action. Plaine and Glass found that L-tryptophane increased the

penetrance of the mutant erupt-suppressor and that previous oxygen treatment increases the penetrance still further. Now Hinton *et al.* (1951) found that L-tryptophane produces many kinds of phenocopies. If we add to these two groups of facts the well known influence of oxygen upon the production of mutants we are led to visualize an interrelation of these different genetic approaches which may lead one day to satisfactory solutions of the problems which have been touched upon.

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Three New Ascomycetes on *Spartium junceum* L.

BY

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Spartium junceum L. is a shrubby Leguminosa of the *Genisteae* sub-family in the Mediterranean region. According to Hegi (1923) it also occurs adventively in south and east Asia and in south America. The plant is the host of some peculiar Ascomycetes such as *Cainia incarcerata* (Desm.) Müller et v. Arx, *Arnaudiella coronae* (Pass.) Petr. and *Didymosphaeria spartii* (Cast.) Fabre. *Guignardia cytisi* (Fuck.) v. Arx et Müller with their imperfect stages and the *Mycosphaerella* described as *Sphaerella scopulorum* Sacc. et Cav. are always found on decayed branches. Besides these there occur a number of similar fungi imperfecti which resemble one another. They certainly belong for the greater part as conidial stages to the class Ascomycetes. We are, however, not as yet in a position definitely to determine their perfect stages. Research into Ascomycetes should in all these cases be based not only on finds but also on cultures so as to render inferences concerning the conidial stages possible.

In the last few years we have found on excursions in the Mediterranean region three apparently new Ascomycetes on *Spartium junceum* L. All three have also been isolated and cultivated on malt agar. After these experiments with cultures it is now possible to attribute to each of them a fungus imperfectus occurring on the substratum of the host. As these fungi belong to genera or subgenera which are not very well known they shall be described here in some detail.

1. *Plectosphaera spartii* nov. spec.

In several places in the south of France we have found a sphaerial Ascomycete with monocellular hyaline ascospores which cannot easily be referred to any described genera. Morphologically it almost fits into some species of the genus *Phyllachora* Nitschke, it differs, however, very distinctly in the formation of the conidial

stage and in the saprophytic growth from typical species of that genus. Another genus, which comprises quite a number of saprophytic species closely related to *Phyllachora* Nitschke is *Plectosphaera* Theiss. The fungus had best be classified with this genus, the natural connection being *Plectosphaera rosae* v. Arx and *Plectosphaera salicis* (Sacc.) v. Arx et Müller (v. Arx and Müller, 1954).

We could find the fungus in the following places:

France, Alpes maritimes, Fontan, July 31, 1955.

France, Vaucluse, Mt. Ventoux, southern side, 700 metres above sea level, June 18, 1956.

France, Vaucluse, Vaison-la-Romaine, June 18, 1956.

Of the first specimen we had a pure culture at our disposal. In this culture the fungus fairly soon formed a compact pink-coloured mycelium on malt agar, covered with a but slightly developed aerial mycelium. In this aerial mycelium there arise after about two weeks fairly large pycnidia in which spindle-like monocellular conidia are produced severally from filiform conidiophores 20 to 30 μ in length. This fungus also occurs on the natural substratum and can best be referred to the imperfect genus *Myxormia* Berk. et Broome.

We have compared the perfect and imperfect stages with fungi already described. According to original material of *Laestadia spartii* Pass. which agrees fairly exactly with our fungus in the mass of asci and spores, this fungus is nothing else than *Guignardia cytisi* (Fuck.) v. Arx et Müller. So there is no need to go into this further. There is, however, a possibility of our fungus, which is by no means rare, having been found before and determined as *Laestadia spartii* or *Physalospora euganea* Sacc. (= *Guignardia cytisi*).

The conidial stages could not be identified with any form already known. The fungus classified by v. Höhnel (1910) with *Phlyctaena* and before that time as *Sacidium juncea* Mont. has been carefully described by Petrak (1924) as the type of his new genus *Selenophomopsis*. Our own material of *Selenophomopsis* shows a fungus conforming to that description but differing sharply from the fungus in question. Also identical with *Selenophomopsis juncea* (Mont.) Petr. are *Phlyctaena spartii* Bub. (Jaap, 1916), *Rhabdospora Cocconii* Sacc. and *Cryptosporium lunulatum* Bäumler. As has only recently been proved (Müller, 1956) on

culture experiments, this fungus belongs to *Guignardia cytisi* (Fuck.) v. Arx et Müller.

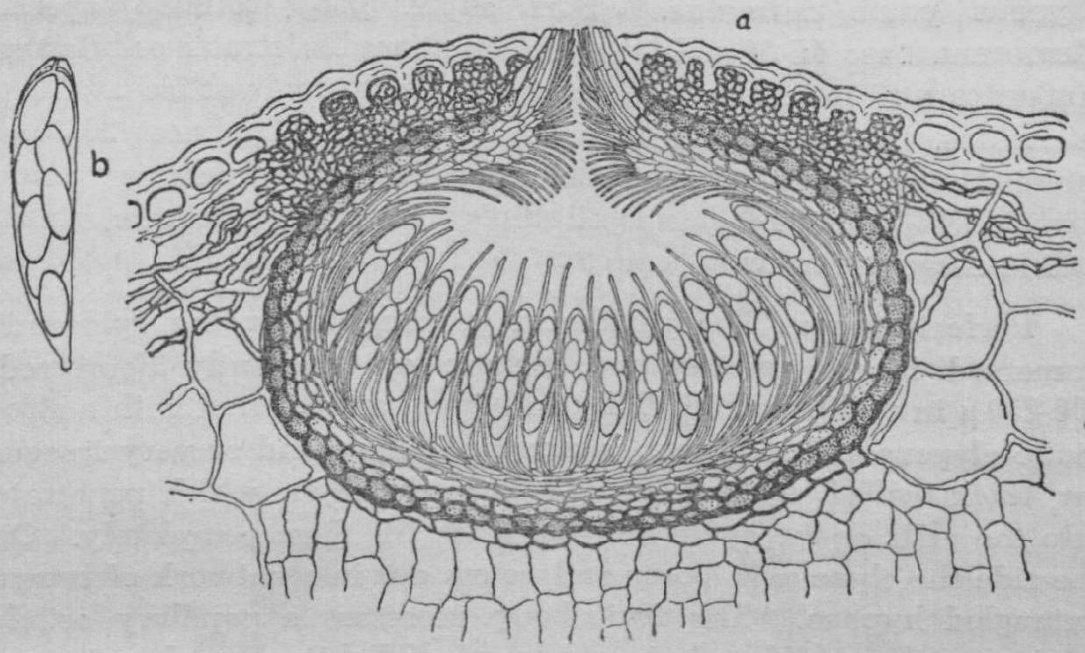

FIG. 1. *Plectosphaera spartii* (perfect stage).
(a) section of a fruit body (×250). (b) Ascus with spores (×1000)

Our fungus is, therefore, a new species which can be described as follows:

Plectosphaera spartii nov. spec.

Status conidiophorus: *Myxormia* Berk. et Broome.
Matrix: *Spartium junceum* L.
Loc.: Gallia part. Mediterr.

Status ascophorus: Perithecia subepidermalia, in substrato immersa, plerumque depressa vel globosa, 200-250 μ latitudinis, 180-220 μ altitudinis, clypeo stromatico tecta. Clypeus fuscis cellulis crasse tunicatis, 2-5 μ diam., epidermidem penetrantibus eamque complentibus. Perithecium apertum papilliforme ostiolo cellulis claris, applanatis, oblongis; centro ostioli periphyses numerosae, filiformes, hyalines. Membrana circiter 20 μ latitudinis, exteriore cellulis 8-12 μ diam., paulo applanatis, crasse tunicatis, nigris, interiore clarioribus et applanatioribus, quibusque ultimo paries fibrata oritur. Cellulae exteriores in cellulas hospitales penetrantes eiscum concrescunt. Asci numerosi, claviformes, octospori, unitunicati, 55-65 μ longitudinis, 10-12 μ latitudinis, breviter pedunculati ad marginem dispositi sunt. Sporae 12-17×5-5·5μ unicellulatae, hyalines, ellipsoides vel nonnumquam claviformes.

Status conidiophorus: Sori nigri et tecti, 350-550 μ latitudinis, curvata cuticula scissa circumdati, iuxta perithecia crescunt. Clypeus paulo formatus. Sori supra poro rotundo aperti. Membrana baso 15-20 μ latitudinis et nonnullis stratis cellularum applanatarum, exteriore crassissime tunicatarum, vertice ad 40 μ latitudinis. Conidiophori baso sori, filiformes, hyalines, 20-30 μ longitudinis. Conidia unicellulata, fusiformia vel recta vel paulo curvata, 17-21 × 4-6 μ magnitudinis, singulo hyalinia in multitudine rosae videntur.

Perfect stage: The perithecia growing under the epidermis immersed in the substratum, globose or frequently depressed, 200-250 μ in width and 180-220 μ in height, are covered with a stromatic clypeus. The clypeus is composed of rotund or very irregular, thick walled, fuscous cells, 2-5μ in diameter, which penetrate into the cells of the epidermis and often fill them completely. On the side the clypeus is looser and peters out in a network of brown entangled hyphae. The fruit body possesses a papillary ostiole rupturing the cuticle. It is composed of light-coloured, not very thick-walled, elongated and distinctly oblate cells which unravel inside at the centre into the numerous filiform, hyaline periphyses. The perithecial wall is c. 20μ in thickness and consists on the outside of opaque blackish-brown, thick-walled, but only slightly oblate cells, 8-12 μ. Inside, the colour grows lighter, the oblateness becomes more distinct and at last the cells form a fibrous fur or tunic of the perithecium. Outside, the perthecia grow together with the destroyed cells of the substratum and are sometimes intertwined or interspersed with them. The numerous claviform asci 55-65 × 10-12 μ in size with delicate simple perithecia sit on short stalks all along the perithecium. They have a distinct apical apparatus and are surrounded by not very numerous, filiform and muciferous paraphyses. The eight ascospores arranged in one or two rows are 12-17×5-5·5μ in size, unicellular, hyaline, ellipsoidal or sometimes slightly claviform in shape.

Imperfect stage: The covered black fruit bodies of the pycnidform appear mingled with the casings of the perfect stage. They are 350-550 μ in width, i.e., on an average twice as large as the perfect stage, and surrounded by the ruptured and concoidal cuticle. It is connected with the cuticle by a but faintly developed clypeus which penetrates into the epiderm. At the top the fruitbodies open in a rotund porus, sometimes they are rent wide open. The pycnidial wall is 15-20 μ at the base and consists of some layers of oblate cells, opaque, brown-black and thick-walled

on the outside, growing gradually lighter and thin-walled inside. Laterally they are just as thick, on the outside, however, they coalesce with the epiderm. In the apical area they reach a thickness up to 40 μ. The conidiphores are confined to the base. They are filiform, hyaline, 20-30 μ in length. At the base they pass into the hyaline interior cells of the fruit body, at the top they severally produce conidia, which are oblong, fusiform, straight or slightly curved, distinctly tapering at both ends, large, monocellular, 17-21 × 4-6 μ. These conidia fill the whole inside space. Alone they appear hyaline, in mass they are pink-coloured.

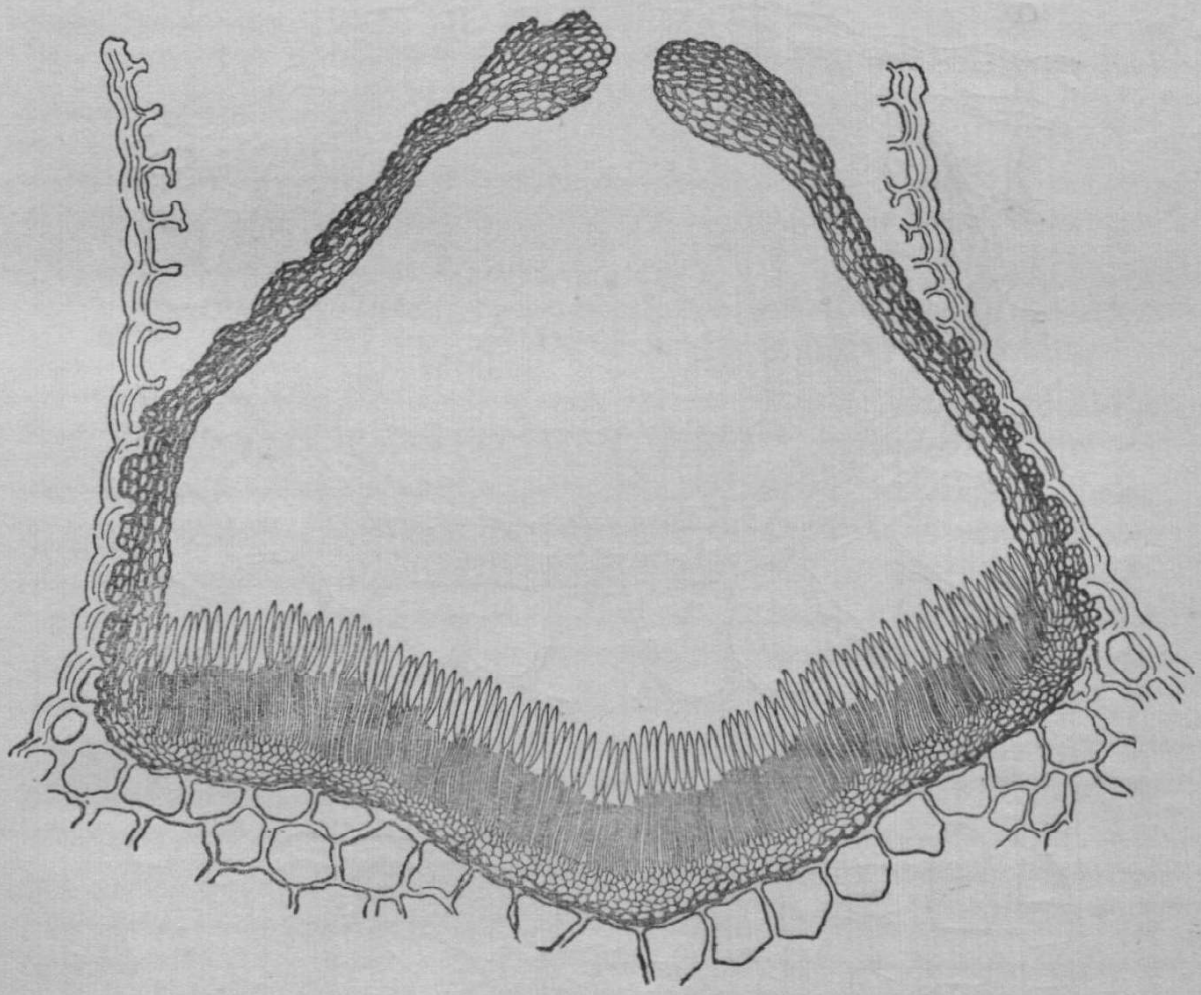

FIG. 2. *Plectosphaera spartii, Myxormia* conidial stage. Section of a fruit body, enlarged (×250).

2. *Trichometasphaeria gallica* nov. spec.

Another fungus found on *Spartium junceum* L. in several places of the Alpes maritimes has similar fruit bodies to *Plectosphaera spartii.* They, too, are covered by an epidermal clypeus. The ostiole, however, is shaped differently: it is flat or slightly

papillary and protruding. In the ostiolar canal there are dark bristles. The ascospores are multi-cellular, hyaline and enveloped in a thin mucous membrane. With all its characteristics the fungus belongs to the genus *Trichometasphaeria* as described by Munk (1953).

The fungus was found in the following places:

France, Alpes maritimes, Fontan, July 31, 1955.

France, Alpes maritimes, Gréolières, Le Foulon, June 22, 1956.

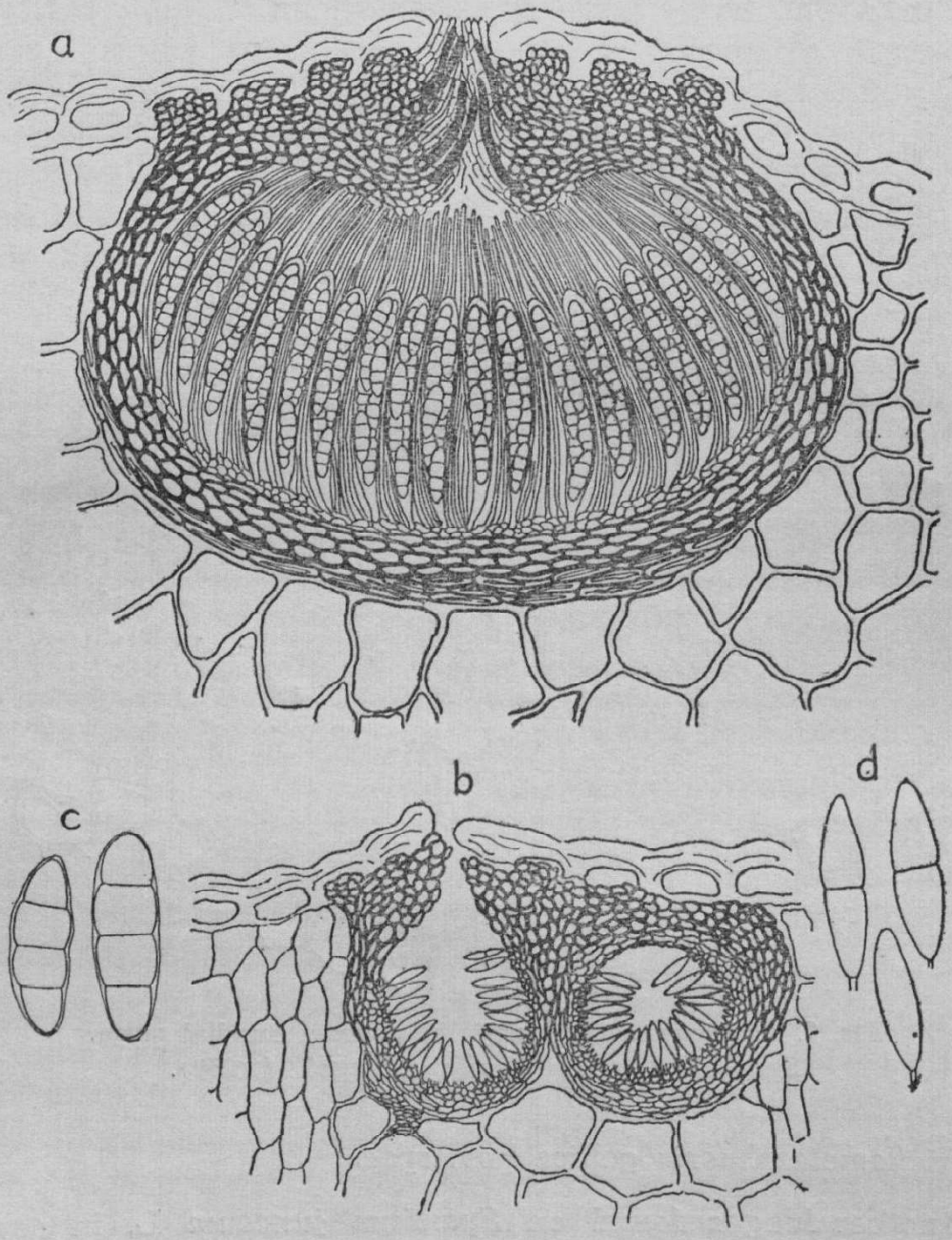

FIG. 3. *Trichometasphaeria gallica.*
(a) Section of a fruit body of the perfect stage. (b) Section of two pycnidia, (×250) each. (c) Ascospores. (d) Conidia, (× 1000) each

In both cases it has been isolated and examined in pure culture. The cultures are alike except for a few slight differences in colour.

In pure culture on malt agar the fungus forms a light grey, compact mycelium which only grows on the substratum. In this mycelium there arise pycnidia singly or mostly gathering in smaller or larger stromatic complexes. In the pycnidia fusiform ellipsoidal, monocellular or bicellular hyaline conidia are produced. We have also found this fungus imperfectus on thin branches of the host. The pycnidia growing on the natural substratum hardly differ from those growing on the agar, only the conidia are for the most part bicellular. This form of conidia fits best with the genus *Ascochyta* Lib. Presumably our fungus has already been described as *Aposphaeria mollis* (Lév.) Sacc.

Trichometasphaeria gallica nov. spec.

Status conidiophorus: *Ascochyta*

Matrix: *Spartium junceum* L.

Loc.: Gallia, part. Mediterr.

Status ascophorus: Perithecia in substrato immersa, 250-350 μ latitudinis, 180-250 μ altitudinus, globosa vel depressa, fusco clypeo cellulis 5-7 μ diam. rotundatis, crasse tunicatis, stromaticis, cellulas epidermales absolute complentibus tecta. Ostiolum planum vel cono depresso simile, perithecium per cuticulam aperiens, canale angusto perforatum, cornato fuscis saetis crasse tunicatis, hebetatis, 3-5 μ diam. Membrana circiter 15 μ latitudinis et nonnullis stratis depressarum cellularum crasse tunicatarum, fuscarum, centro clariorum, 8-10 μ diam. Numerosi asci claviformes, octospori 80-100 μ longitudinis 10-12 μ latitudinis, bitunicati crasseque, paralleli, ad marginem perithecii dispositi sunt. Asci nomerosis paraphysoidibus hyalinibus filiformibus circumdati sunt. Sporae fusiformes vel cylindricae quatercellulatae, hyalines, 16-22 × 5-6 μ, circumdatae strato mucoso.

Status conidiophorus: Parasitus caulibus hornotinis crescere videtur eosque interficiens. Pycnidia congregata, diversa forma et varia magnitudine, saepe secum coniuncta, in superiore parte clypeo simile augescens. Papiliforme ostiolum canale angusto perforatum. Membrana pycnidiarum superiore parte cellulis polyedricis, fuscis, 5-7 μ diam., crasse tunicatis; inferiore et latere clarioribus cellulis tenuius tunicatis et exteriore parte depressis. Conidiospori hyalines, 15-18 × 4-4·5 μ, fusiformes, uni vel bicellulati.

Perfect stage: The perithecia, 250-350 μ in width and 180-250 μ in height, depressed, are completely immersed in the substratum. They are covered by a fuscous clypeus consisting of rotund, rather thick-walled cells which are 5-7 μ in diameter. The clypeus fills the cells of the epiderm completely and dissolves into short, dark hyphae. The ostiole is flat or only scarcely conical and ruptures the epiderm from inside. It is pierced by a narrow canal which on the inside is spiked with dark, thick-walled, stout, blunt bristles and lined with highly flattened, hyaline, thin-walled cells. The perithecium is about 15 μ thick and consists of several layers of more or less distinctly flattened, rather thick-walled, fuscous cells which grow lighter in colour towards the centre, 8-10 μ in diameter. In the ostiolar area it grows up to 30 μ in thickness; it forms a light cavity and passes into the clypeus without a distinct borderline. The numerous claviform asci are 80-100 × 10-12 μ, thick-walled or even bi-tunicated. They are parallel all along the inside of the wall and enveloped in numerous filiform, hyaline paraphysoids. Each of them contains eight fusiform or almost cylindrical quadricellular, hyaline spores, 16-22 × 5-6 μ. The second cell is in most cases slightly thickened and all septa show a distinct constriction. They are enveloped in a thin transparent mucous membrane.

Imperfect stage: The conidium grows parasitically on the living twigs and kills them. The pycnidia are mostly immersed in the substratum in groups and very often stromatically grown together. They vary in size and shape, globose, upright, ellipsoidal or depressed, often very irregular owing to mutual pressure, 80-220 μ in diameter. At the top they are like the perfect stage thickened like the clypeus. A papillary ostiole pierced by a narrow canal breaks through to the outside. Here it consists of brown coloured rather thick-walled cells, 5-7 μ in size. On the sides and at the base these cells gradually grow lighter in colour, the wall is thinner and the cells are distinctly flatter. On the inside fusiform, hyaline, monocellular or bicellular conidia, 15-18×4-4·5 μ in size are produced by the conidiphores which stand up like short thorns.

3. *Pleospora spartium-juncei* nov. spec.

On *Spartium junceum* L. a species of *Pleospora* is often found which has been determined as *Pleospora spartii* (Sacc.) Berl. et Sacc. or as *Pleospora cytisi* Fuck. fa. *spartii* Sacc., but differs in nothing from the typical *Pleospora cytisi* Fuck. The fungus belongs

to the group of *Pleospora herbarum* (Pers.) Rabenh. It has depressed fruit bodies which vary in diameter from 8-300 µ at the beginning entirely closed, later opening in a rotund porus. The asci are broadly claviform or ellipsoidal, thick-walled or bitunicated, 100-140 × 20-25 µ, arranged in parallel order and enveloped in a cellular paraphysoid tissue. Each of them contains eight oblong ellipsoidal or slightly oval ascospores with a distinct hyaline mucous membrane clearly constricted in the middle, partitioned by 6-8 septa and 1-2 incomplete longitudinal septa, finely golden-brown coloured, 25-45 × 10-16 µ. A detailed description is to be found in Petrak (1921), who refers it to *Pyrenophora* Fr.

We have found *Pleospora cytisi* several times on *Spartium* and could isolate it in the following places:

France, Alpes maritimes, Fontan, July 31, 1955.
France, Alpes maritimes, Le Foulon Gréolières, June 22, 1956.

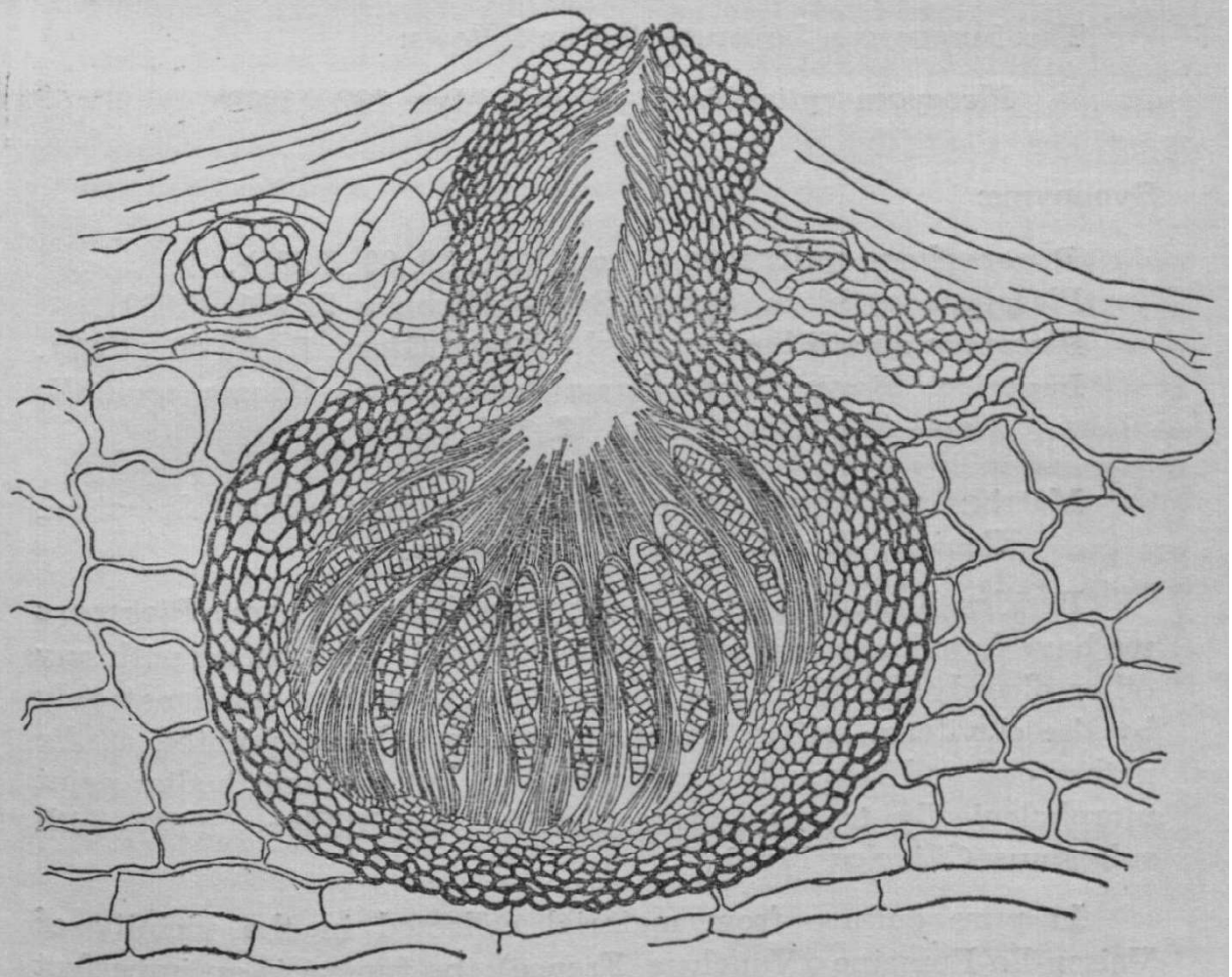

FIG. 4. *Pleospora spartium-juncei.*
Section of a fruit body (× 250).

In pure culture the two specimens behave somewhat differently. In the Fontan culture there arise, especially under intensive exposure to light, numerous pycnidia, in which byciform, monocellular, hyaline conidia 4-6 × 2 μ, are produced in filiform, hyaline conidiphores up to 10μ in length. The conidiophores dissolve soon so that they can no longer be recognized in the ripe stage. This fungus is certainly identical with *Phoma sagittalis* Jaap. Jaap (1904) expressly mentions that he found it on *Cytisus sagittalis* L. besides *Pleospora cytisi* Fuck.

But this conidial fungus is not a typical *Phoma* Fr. Morphologically it belongs to the neighbourhood of *Phoma acuta* Fr., which is the type of the genus *Leptophoma* v. Höhn and belongs as an imperfect stage to *Leptosphaeria acuta* (Moug. et Nestl.) Karst.

The Le Foulon specimen produced in pure culture only few pycnidia. On the other hand the development on malt agar is always closed by the formation of the perfect stage.

The fungus can be presented as follows:

Pleospora cytisi Fuck.—Symb. Myc. 132 (1869).

Synonyms:

Pyrenophora cytisi Petr.—Ann. Myc. 19, 99 (1921).
Pleospora cytisi fa. *spartii* Sacc.—Michelia 2, 251 (1882).
Pleospora spartii Sacc. et Berl.—Nuov. Giorn. It. 20, 86 (1883).
Imperfect Stage: *Leptophoma* described as *Phoma sagittalis* Jaap—Allg. Bot. Ztschr. 12, 124 (1904).

Matrix: shrubby Leguminosae such as *Cytisus, Spartium, Genista, Sarothamnus, Callicothome.*

Besides this striking and frequent species of the genus *Pleospora* we have found a second one which could not be identified with any other already described. It differs from *Pleospora cytisi* not only by the claviform, rather narrow spores which are pointed at the nether end and have a mucous appendix, but the fruit bodies point morphologically to the fungus classified by Müller (1951) with sub-genus *Cilioplea* Munk.

In pure culture from material collected in the vicinity of Vaison-La-Romaine (Vaucluse, France) the fungus is an imperfect form of the form genus *Coniothyrium* Corda and thus differs in its asexual phase distinctly from *Pleospora cytisi.* We could find this

imperfect form also on material of *Spartium junceum* L. The fungus can be described as follows:

Pleospora spartium-juncei nov. spec.

Status conidiophorus: *Coniothyrium.*

Matrix: *Spartium junceum* L.

Loc.: Gallia part. Mediterr.

Status ascophorus: Perithecia dispersa, sub-epidermide nidulatia, globosa vel depressa, 200-300 μ diam. Ostiolo plus minusve conico, canale hyalini periphysibus ornato. Pariete perithecii 20-40μ crasso, e stratis pluribus cellulis angulato-rotundatis, 10-15 μ diam. composito. Asci numerosi, clavati, bitunicati, 95-110 × 10-13 μ. Sporae clavatae media constrictae, base subacuminatae et appendiculis hyalinibus, mucosis ornatae, 5-septatae (rar. 4- vel 6- sept.), in loculis mediis praeterea longitudinaliter septatis.

Status conidiophorus: Pycnidia 100-250 μ diam. epidermide tecta, globosa; ostiolo conico erumpenti. Paries pycnidiorum fuscus, cellulis 4-7 μ diam. subcrassis compositus. Pycnidia prima interna composita cellulis hyalins, demum conidis ellipsoidibus, hyalinibus vel subfuscus 2-4 × 1-3 μ diam. formata.

Perfect stage: The fungus forms dark fruit bodies completely immersed in the substratum, each standing apart, globose or mostly depressed, 200-300 μ which break through in a conic ostiole up to 120 μ in height. The ostiole is perforated by a canal straightening at the upper end, with hyaline, filiform, paraphysoid hyphae projecting into it. The perithecium is 20-40 μ and consists of numerous layers of cells, thick walled, almost opaque, fuscous, regularly polygonal, 10-15 μ on the outside; on the inside, however, they are more thin-walled, slightly depressed and lighter, and at the centre smaller and thin-walled. The numerous asci are claviform, straightening into a stalk at the base, thick-walled and bitunicate, 95-110 × 10-13 μ. They are enveloped in numerous filiform paraphysoids and contain each of them eight claviform spores, constricted in the middle and distinctly tapering, with a hyaline mucous point, brownish-yellow, 20-25 × 7-8 μ, with 4-6, mostly 5 septa and often with a longitudinal septum in one of the middle parts.

Imperfect stage: The pycnidia are immersed in the substratum, upright ellipsoidal or depressed, 100-250 μ, but break through in a slightly conically projecting ostiole. The wall is brownish and

consists of not very thick-walled brown cells, 4-7μ. Inside they are almost completely filled with thin-walled, regularly polygonal hyaline cells. Later they sub-divide at the centre or in other part with monocellular, hyaline or faintly fawn coloured, ellipsoidal spores gradually arising from the part-cells. The spore-producing cells dissolve and the conidium formation proceeds from the inner to the outer cell layers. Later sterigmata are to be seen through which spores penetrate from the mother cells into the inside of the pycnidia and from there to the outside.

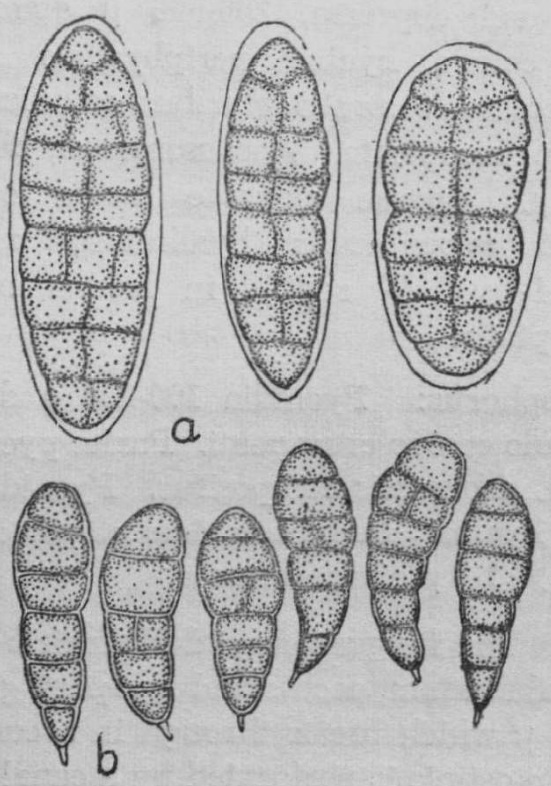

FIG. 5. Ascospores of (a) *Pleospora cytisi*, (b) *Pleospora spartium-juncei* (×1000).

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Temperature Relations of the Mineral Facies of Metamorphic Rocks

BY

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1. *Introduction*

The century old idea that all so-called igneous rocks had crystallized from a molten magma has, in the light of modern knowledge, met with severe criticism. Magmas are not at all abundant in the crust; on the contrary, they would seem to be very rare and local phenomena, perhaps to be compared to abnormal abscesses in the human skin. It is not possible any longer to maintain the idea that all the large batholithic bodies of granites and other coarse-crystalline rocks (anorthosites, etc.) grew by crystallization out of a magma.

In the crust of the continents the most important group of rocks, quantitatively, are those called metamorphic (gneisses, migmatites, granulites, crystalline schists, etc.). James Hutton (1726-97) of Edinburgh was the first to advance the opinion that some of these rocks were ordinary sedimentary strata subsequently altered by heat, and Sir Charles Lyell, who adopted the Huttonian theory, proposed the term metamorphic for the altered strata.

Metamorphic processes in rocks cannot be directly observed in nature, nor have they been adequately investigated by experiments. For this reason we are still very much in the dark as to the mechanism and thermodynamic relations of metamorphic processes. The study of transitional rock series from unmetamorphic to completely altered (recrystallized) rocks is the basis on which the principles of metamorphic geology must be built.

2. *Metamorphism and Metasomatism*

The alteration by recrystallization is generally accompanied by a change in the chemical composition, i.e., export and import of chemical matter resulting in the so-called metasomatic metamorphism, or simply metasomatism. One can often find statements to the effect that metamorphism without chemical change is the

normal thing; to be sure, supposed cases of isochemical metamorphism are still currently set down as "normal metamorphism". In harmony with this nomenclature it has become customary, and is perhaps expedient in an attempt to survey and classify the various petrographic types, to distinguish between the "normal metamorphism" (with but small changes in the total chemical composition) and the metasomatic metamorphism (with large changes in the chemical composition). But from a genetical point of view this distinction is illogical and obscures the broader relations. Metamorphic rocks *always* have suffered a change in composition, for transportation and transfer of chemical matter and heat are not only concomitant with, but actually essential parts of, the earth processes constituting metamorphism. In geologic discussions the fact that sediments at the very incipience of metamorphism lose volatile constituents (H_2O, CO_2, etc.), and consequently regularly change their chemical composition, has often been neglected. However, these changes are not to be neglected, and H_2O, CO_2, etc., are just as important as many other of the rock-making constituents, and serve to characterize the rocks in which they occur.

There should be no mistake about the fact that, whether one likes it or not, large scale chemical transport in and through large volumes of rock is the normal thing: gneiss-granites, migmatites, micaschists, etc., all exhibit a chemical composition different from that of the pre-existing rock complex. Similarly, the mineralogical composition changes according to the changing chemical composition and according to variations in temperature and pressure; thermodynamic equilibrium is approached during the course of the metamorphism.

What are the instruments of these changes? Chemical emanations, ichor, or watery solutions circulating in the poor spaces? The question is highly controversial, and I shall not go any further into the problem here, I shall only again emphasize the fact: large-scale chemical changes in solid rocks did take place; and it is no argument against "emanations" or "ichors" that the physicist is unable to explain them, as long as he is equally unable to offer any better solution to the problem.

For the sake of completeness: there is no doubt about the fact that the changes took place in essentially solid rocks — an abundance of evidence show that no melting occurred. Quite decisive is, for instance, the interesting account of Professor Bucher

(1953) on fossils in metamorphic rocks; so prudential were the metamorphic processes that in many cases the most delicate structures have been preserved.

To me it is a reasonable position to base the discussion of recrystallization and evolution of metamorphites on the postulate that free interchange of matter (atoms or ions) can take place over large volumes of rocks existing as essentially solid masses at temperatures well below 650° C. Above 650°, magmas may form, but below this temperature rock magmas do not exist. Thus there is a natural boundary between metamorphism and magmatism (Ramberg, 1952).

3. *The Facies*

It is of paramount importance to get better information about the temperature-pressure conditions of the recrystallization. A large-scale attempt in this direction was the development of the facies classification. The first adumbration exists in the pioneer work on the *étages* by Michel-Lévy (1878-79) and in the corresponding classification of the metamorphic minerals by Becke (1903) and Grubenmann (1904) according to the depth zones. But in the early work of G. Barrow (1893) on the classification of the metamorphic rocks of the Highlands of Scotland according to the so-called index minerals we find the true germ of the facies principle. V. M. Goldschmidt (1911), in his work on the hornfelses of the Oslo region, used the same ideas to contrive a well defined system built on a theoretical fundament. Finally Eskola (1915), in his work on the Orijärvi gneisses, perfected the ideas and shaped them into a tool of the highest value: the system of the mineral facies of rocks.

The concept of a *mineral facies* as defined by Eskola (1921) "comprises all the rocks that have originated under temperature and pressure conditions so similar that a definite chemical composition has resulted in the same set of minerals, quite regardless of their mode of crystallization, whether from magma or aqueous solution or gas, and whether by direct crystallization from solution, or by gradual change of earlier minerals".

In order to determine these associations, in other words, in order to find out which mineral associations are characteristic of high temperature, or of low temperature, each of them combined with high pressure and with low pressure, Eskola had to go into

the field and study naturally occurring mineral associations, i.e., rocks.

It has long been known that in an area of progressive metamorphism each successive stage, or each new zone of metamorphism, is reflected in the appearance of characteristic rock types. Rocks within the same zone may be called isofacial, or isograd as proposed by Tilley (1924) who, furthermore, proposed the term isograd for a line of similar degree of metamorphism.

In going from an area of unmetamorphosed sedimentary rocks into progressively more highly metamorphic rocks, new minerals appear in orderly succession. Thus, in a series of argillaceous rocks subjected to progressive metamorphism, the first index mineral to appear is usually *chlorite,* followed successively by *biotite, garnet* (almandite), and *sillimanite.* A line can be drawn on the map indicating where biotite first appears. This line is the biotite isograd. The less metamorphosed argillites on one side of this line lack biotite, whereas the more metamorphosed rocks on the other side contain biotite. An isograd can be drawn for each mineral. Actually the isograds are surfaces, and the lines we draw on the map are the intersections of these surfaces with the surface of the earth.

Further work along these lines resulted in the conclusion that it was possible to single out a well defined series of mineral facies. Sedimentary rocks of the lowest metamorphic grade recrystallized to give rocks of the *greenschist facies*—chlorite and epidote being characteristic minerals. Higher metamorphism produced the *epidote amphibolite facies,* still higher the true *amphibolite facies* in which both chlorite and epidote are forbidden minerals, hornblende and plagioclase mainly taking their place. Representative of the highest regional metamorphic grade is the *granulite facies* in which most of the stable minerals are water-free: pyroxenes, garnets, etc.

Any sedimentary unit will recrystallize according to the rules of the several mineral facies, the complete sequence of events being a progressive change of the sediment by deformation, recrystallisation and alteration in the successive stages:

Greenschist facies ⟶ epidote amphibolite facies ⟶ amphibolite facies ⟶ granulite facies.

During regional metamorphism a stationary temperature gradient is supposed to be established in the mountain masses. By

and large, the outer parts of a geosynclinal region are less affected, and in the ideal case the marginal parts contain unmetamorphosed sediments, clay, sand, limestone, etc., which gradually change into metamorphic rocks of successively higher facies as they extend into the central and deeper parts (depth zones).

This change, as demonstrated by the succession in mineral facies in the present surface rocks, would seem to correspond to the temperature gradient within the original mountain mass, a rapid change in facies indicating a previous steep temperature gradient and *vice versa*. This implied relation of mineral facies and temperature gradient of the metamorphism has been used to estimate the depth of erosion and, indeed, the original heights of eroded mountain ranges.

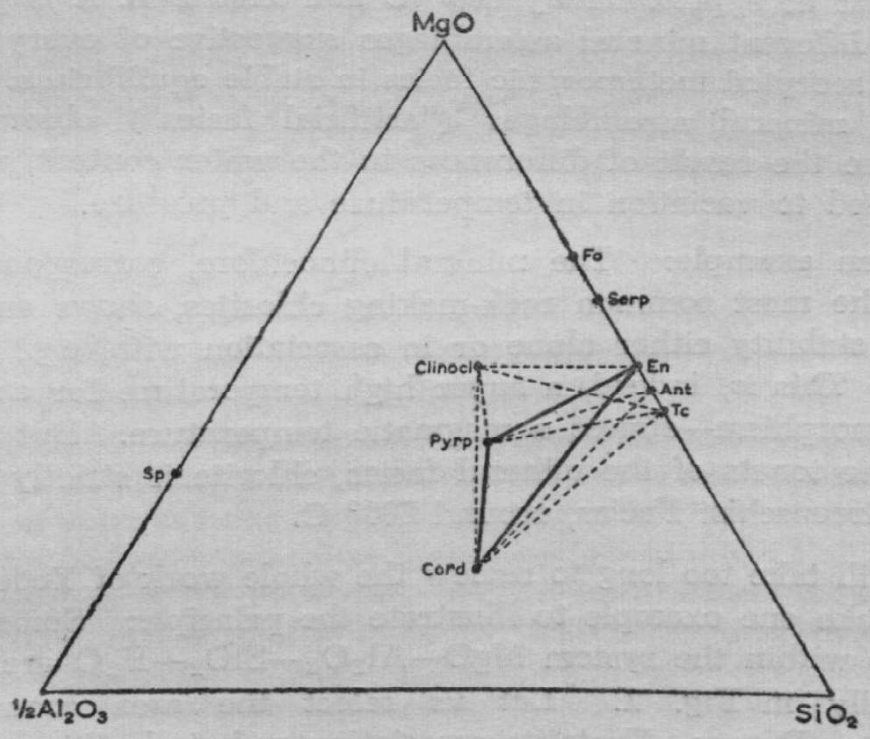

FIG. 1. A one-point projection of the MgO-Al_2O_3-SiO_2-H_2O system showing some of the constituent minerals. Ant = anthophyllite $Mg_7Si_8O_{22}(OH)_2$, Clinocl = clinochlore $Mg_5Al_2Si_3O_{10}(OH)_8$, Cord = cordierite $Mg_2Al_4Si_5O_{18}$, En = enstatite $MgSiO_3$, Fo = forsterite Mg_2SiO_4, Pyrp = pyrope $Mg_3Al_2Si_3O_{12}$, Serp = serpentine $Mg_6Si_4O_{11}(OH)_6 \cdot H_2O$, Sp = spinel $Mg\ Al_2O_4$, Tc = talc $Mg_3Si_4O_{11}(OH)_2$.

The assemblage shown in full lines refers to the anhydrous phases, cord-pyrp-en; the hydrous phases are joined by dashed tie-lines.

So strongly fixed in geological reasoning has become this direct relation between mineral facies and temperature (pressure) that without further reflection the various facies have been accepted as direct indicators of temperature (pressure). Consequently it

is possible, graphically, to correlate the facies with temperature and pressure. Using temperature and pressure as axes, each mineral facies can be assigned to a specific temperature-pressure region as shown in Fig. 3.

4. *The Experiments of Yoder*

It must be called almost a shock, then, that one carefully conducted series of experiments by Dr. Yoder (1952) at the Geophysical Laboratory in Washington has given compelling evidence that this relation is not at all so certain and as absolute as has been thought.

Work in the system $MgO—Al_2O_3—SiO_2—H_2O$ gave as one result that at approximately 600° C and 1200 atm. it is possible to have different mineral assemblages suggestive of every one of the now accepted metamorphic facies in stable equilibrium. These different mineral assemblages ("artificial facies") observed by Yoder are the result of differences in the water content, and are not related to variation in temperature and pressure.

As an example: The mineral clinochlore, corresponding to one of the most common rock-making chlorites, shows an upper limit of stability either alone or in association with, e.g., talc, of 680° C. This is, indeed, a super high temperature for any kind of metamorphism—almost a magmatic temperature. But according to the tenets of the mineral facies, chlorite is strictly limited to the Greenschist Facies, around 200° C.

It will take too long to review the whole work of Yoder. But I shall take one example to illustrate the principle: Some of the minerals within the system $MgO—Al_2O_3—SiO_2—H_2O$ are shown graphically in Fig. 1. Let us select the area around the Cordierite—Pyrope—Enstatite association and study this area at a constant temperature-pressure combination: 600° C and 1200 atm., but under varying concentrations of H_2O. The various mineral associations which then develop are shown graphically in Fig. 2.

Interpreted in terms of the mineral facies of rocks the associations are:

1. Enstatite—cordierite gneiss = granulite facies (~ 600°).
2. Anthophyllite—garnet gneiss=amphibolite facies (~ 500°).
3. Anthophyllite—talc schists = epidote — amphibolite facies (~ 400°).
4. Talc—chlorite schists = greenschist facies (~ 200°).

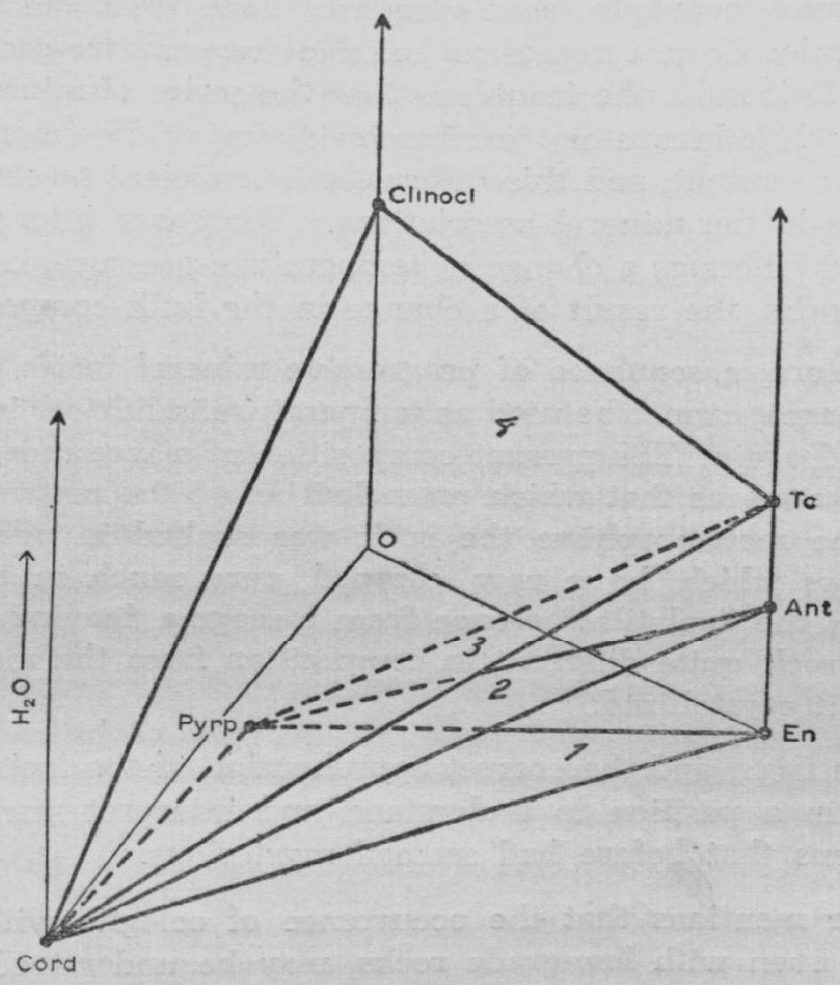

FIG. 2. A perspective projection of the system Cordierite-Pyrope-Enstatite-Water, constructed after observed and inferred data of Yoder (1952). Activity of water increases normal to the bottom face Cord-O-En.

(1) The water-free association is *cordierite-pyrope-enstatite* as shown in the bottom face of the illustration. With a small amount of water first enstatite yields part of itself to hydration (replaced by anthophyllite), with increasing concentration of water eventually all enstatite disappears and the association becomes (2) *cordierite-pyrope-anthophyllite.* With more water talc starts to form at the cost of anthophyllite and the association eventually becomes (3) *cordierite-pyrope-talc.* With still more water pyrope yields to hydration (replaced by chlorite) and the final association stable in the excess-water region is (4) *cordierite-talc-chlorite.*

According to the facies geology we go through a temperature range from 600° to 200° C, while the experiments prove that the temperature actually was constant; chemically we just go through a small range of composition from no water to 5-6% of water. The mineral association corresponding to the granulite facies is water-free. With increasing water content the minerals indicate facies of successively lower grade.

We must conclude, as Yoder has done, that the accepted mineral facies do not represent reliable temperature-pressure indicators. Evidently, the trouble is that the series of mineral facies chosen by Eskola are not isochemical: they differ markedly in their water content, and this difference is sufficient to cause great differences in the mineral associations. What was interpreted by Eskola and others as a change in temperature-pressure, is actually, in some rocks, the result of a change in the bulk composition.

Therefore, a sequence of progressive mineral facies established in one area, cannot be used as temperature indicators in another area. And again: The present composition of a metamorphic rock is not the same as that which prevailed before the metamorphism. During the metamorphism the rock was soaked in pore liquids and vapors which have now escaped, very much in the same manner as the "volatiles" escape from a magma, leaving behind a solidified rock quite different in composition from the magma out of which it crystallized.

These facts, and the correct apprehension of the facts, put the geologist in a position to understand and interpret many of the observations that before had seemed mysterious.

Yoder mentions that the occurrence of eclogite with amphibolite, or even with low-grade rocks, may be understood. To be sure, in an area of isofacial rocks the field geologist may frequently observe abnormal rock deposits indicating a change in mineral facies. As an example I may cite the description of Seitsaari (1954) of a chlorite schist occurring with pre-Cambrian rocks of the amphibolite facies near Tampere, Finland. In this case the feldspar of the chlorite schist is a bytownite, proving, that the chlorite was not formed at the temperatures of the greenschist facies ($\sim 200°$) but actually at higher temperatures ($\sim 500°$ (?)) corresponding to the amphibolite facies.

A similar explanation may, perhaps, be given to the anorthite-bearing chloritoid schist of the Hitachi District, Japan, described by Seki (1955).

5. *Conclusions as to the Relation between Temperature and Mineral Facies.*

I have mentioned that what has been described as a change from one mineral facies to another (= a change in the temperature-pressure conditions) is simply, in some rocks, the result of

a change in bulk composition. I have also mentioned a greenschist which does not belong to the greenschist facies. The question then arises: Are the established mineral facies "no good"? Is there no relation between temperature (pressure) and the several mineral facies of metamorphic rocks?

Although Yoder has proved that there is no absolute relation between temperature and facies, and although the greenschist from Tampere recrystallized outside the greenschist facies, I maintain that, to the field geologist, and to the laboratory man as well, the established facies classification will still remain as the best system of classification of metamorphic rocks; and in a majority of cases the facies will indicate the temperature-pressure conditions under which the several rocks recrystallized.

The point is that, generally speaking, there is a regular relation between the chemical activity of H_2O and the facies.

As shown by Yoder (1955) the role of water in metamorphism is determined by at least four variable parameters: rock pressure, temperature, water pressure, and the amount of water present. However, these four variables are geologically related: during a normal progressive regional metamorphism, rock

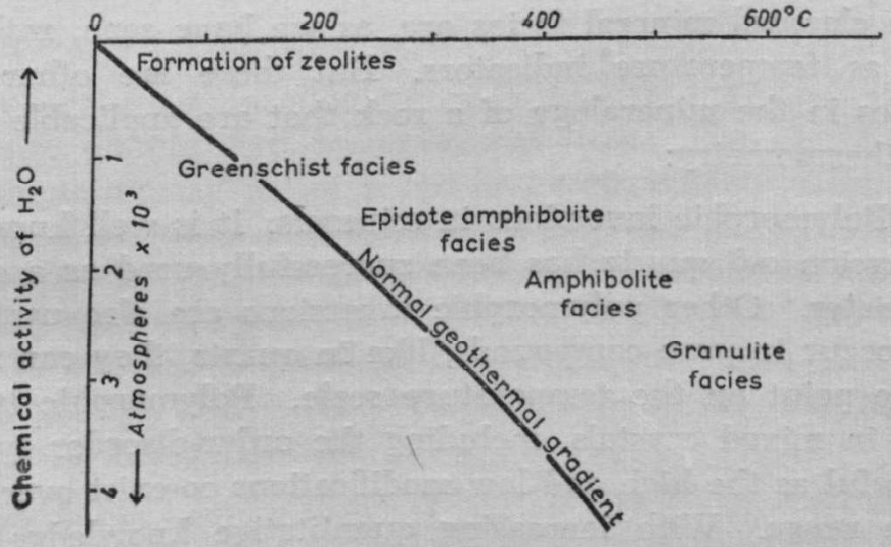

FIG. 3. The mineral facies of regional metamorphic rocks in their normal relation to temperature, pressure, and chemical activity of water.

pressure and temperature are interdependent, and the amount of water and the pressure of water are related to the encasing sediments and to the degree of metamorphism in such a way that, generally speaking, the low-grade metamorphic facies are characterized by the presence of an excess of water, the medium-grade by some deficiency in water, and the high-grade metamorphic facies by

virtual absence of water. In the usual diagrammatic illustration of the mineral facies of rocks, temperature and pressure (depth) are taken as coordinates; in regional-metamorphic rocks we may now add a third, dependent coordinate: the activity of H_2O running parallel to, but in opposite direction of, the pressure. See Fig. 3.

Thus, Fig. 3 represents my conclusion to the question of the relation between temperature and mineral facies of regional-metamorphic rocks. It illustrates the usual case but does not take into account all the exceptions that may occur. For the mineral facies of a rock is not a one-valued function of temperature (and pressure). So complex are the geological processes, that in addition to temperature and rock pressure a great number of additional variables enter into the picture: stress, rate of growth, position in the gravitational field, pressure of water, of CO_2, and of other volatiles, absolute amount of water, of CO_2, etc.—thus, at the present state of our knowledge, and for a long time to come, it will be impossible from the simple mineral facies of a rock to infer with certainty the temperature at which the metamorphism took place.

6. *Geological Temperature Indicators.*

The classical mineral facies are, as we have seen, rather uncertain as temperature indicators. But there are other useful properties in the mineralogy of a rock that are applicable in geological thermometry:

1. Polymorphic inversions in minerals: It is well known that the inversion α-β quartz has been successfully used as a geologic thermometer. Other polymorphic inversions are also useful, but if they occur in pure compounds, like in quartz, they can register only one point on the temperature scale. Polymorphic transformations in mixed crystals, including the order-disorder type, are more useful as the high and low modifications co-exist over a temperature range. With increasing quantitative knowledge of the mechanism of polymorphism in, e.g., pyroxenes and feldspars, these minerals will become useful temperature indicators.

2. The exsolution of mixed crystals and, generally, the transformation of a mineral into a mineral assemblage: Sanidine breaks up into perthite, basic plagioclase may break up into sodic plagioclase and either scapolite or epidote; there is always a relation between the break-down temperature and the composition of the produced mineral phases.

3. The ratio of distribution of an element between two solvent minerals: Thus soda (in terms of the albite molecule which is soluble both in orthoclase and in anorthite) will distribute itself between plagioclase and alkali feldspar, and the distribution factor is characteristic of the temperature (pressure). With better knowledge of the distribution factor the chemical composition of the individual feldspar phases would become a geologic thermometer of the highest value (Barth, 1956).

Still simpler temperature relations are obtained in the distribution of the trace elements between the various mineral phases.

4. The fractionation of isotopes between two minerals is probably the most accurate method. It has been used for minerals formed at relatively low temperatures, but recently also applied to quartz-calcite and to quartz-ironoxide associations at higher temperatures (Clayton and Epstein, 1956).

5. There are other, special methods for geologic thermometry, e.g., the observation of the temperature of the filling of inclusions in crystals, and others.

A complete general summary of geologic thermometry has been prepared by Ingerson (1955).

From these briefly mentioned facts, it is seen that prerequisite to a temperature determination come a long series of carefully determined quantitative mineralogical data. The determination of the simple mineral facies is the first step, it gives a first approximation. But today one can and should go much farther, for it is obvious that the ever increasing accumulation of knowledge of the chemical and physical properties of minerals, and of their relation to one another and to the temperature and their origin impose a greater task on modern petrologists than was borne by those of a former generation.*

Today is not enough to list by name and amount the minerals constituting a rock, but for each mineral its physical and chemical properties must be determined. Not until such complete data are available for various rocks can a refined system of mineral facies be worked out, and the mode of origin, temperature, and pressure of the rock be inferred.

* This sentence, slightly differently worded, was pronounced in 1902 in the conclusion of Cross, Iddings, Puison, and Washington's famous "A quantitative chemico-mineralogical classification and nomenclature of igneous rocks."

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On the Structure and Life-History of *Cylindrocapsopsis indica* gen. et sp. nov., A New Member of the Cylindrocapsaceae

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This alga came up in a pond in Madras in January 1938. It grew well in the pond for three months and then disappeared about the beginning of May. The next year, it came up again in the pond in February and disappeared in May as in the previous year. The alga was found in the pond only in a sterile condition.

Material and methods

The living material of the alga was taken to the laboratory and kept in large wide glass dishes with plenty of the pond water in a well lighted part of the room. The alga grew very well in the laboratory for some time in a sterile condition and finally showed several very interesting stages of sexual reproduction.

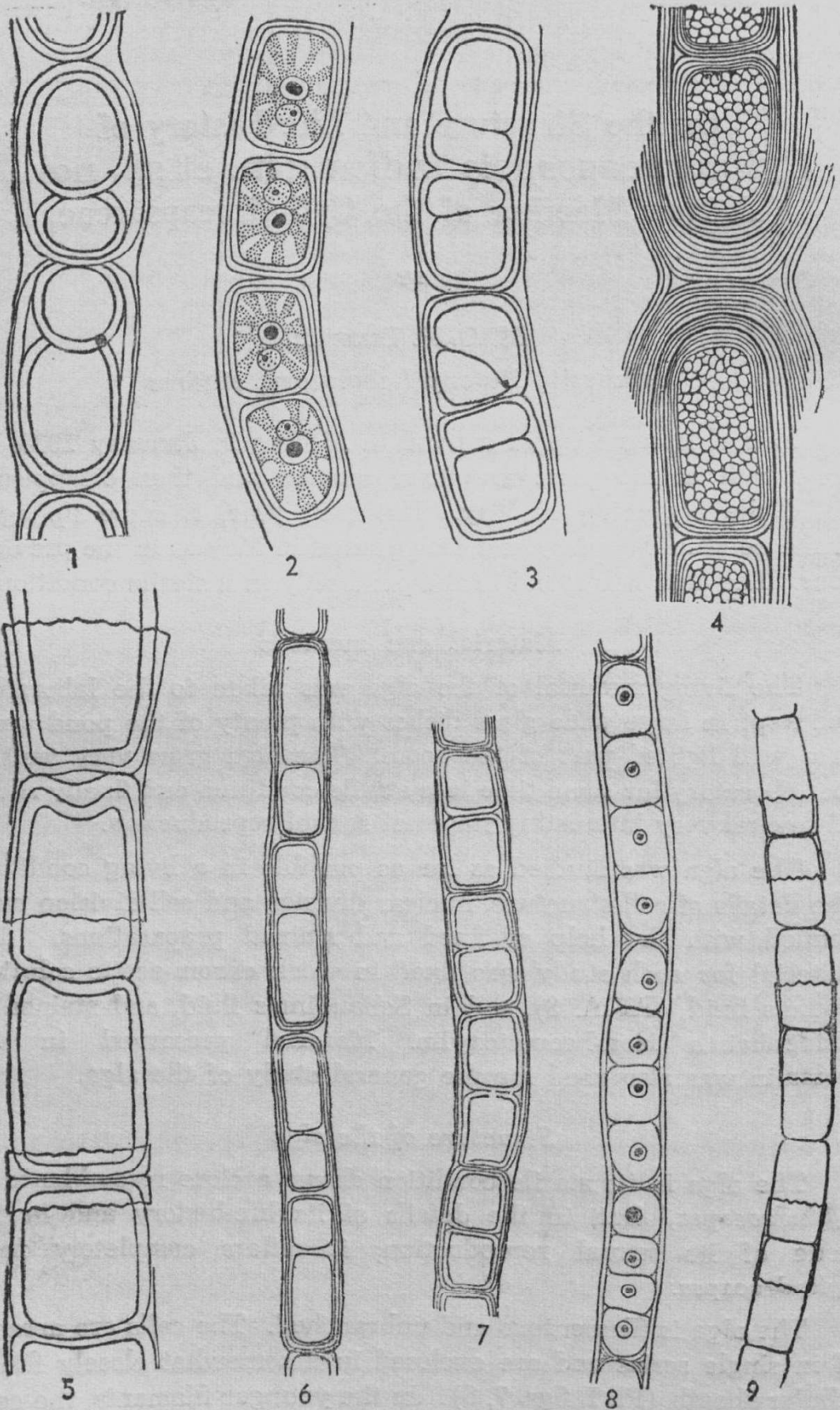
The alga was studied as far as possible in a living condition. The details of cell-structure, nuclear division and cell-division were studied with the help of fixed and stained preparations. The material for such study was fixed in weak chrom-acetic solution, Bouin's fluid (P.F.A. 3) and in Schaudinn's fluid, and stained in Heidenhain's iron-haematoxylin. Material preserved in 4% formalin was also used for the general study of the alga.

Structure of the alga

The alga in its sterile condition shows a close resemblance to *Cylindrocapsa*,[1] but, in the details of its life-history and in the mode of its sexual reproduction, it differs completely from *Cylindrocapsa*.

The alga is filamentous and unbranched. The cells are arranged in single series and are enclosed in a somewhat closely fitting tubular sheath (Pl. I, figs. 7, 8). In the younger filaments, the cells

1. A preliminary account of this alga was published under the name of *Cylindrocapsa geminella* Wolle (Iyengar, 1939).



TEXT-FIGS. 1-9. *Cylindrocapsopsis indica* sp. nov.

FIG. 1. Portion of a filament boiled in caustic potash to dissolve out the cell-contents, showing daughter-cells enclosed inside the mother walls. FIG. 2.

Cells with a single nucleus, and a stellate chloroplast with a pyrenoid. FIG. 3. Protoplasts of cells divided into two unequal daughter-protoplasts. FIG. 4. Old cells with walls made up of several layers. Note: H-shaped portion made up of remnants of mother-walls of several previous generations. FIGS. 5, 9. Cells showing ruptured mother-walls. FIGS. 6, 7. Cells enclosed in their mother-walls and grand-mother-walls. FIG. 8. Portion of a filament with three groups of four cells each, each group consisting of four grand-daughter-cells. FIGS. 1, 3, 5, × 750; FIG. 2, × 725; FIGS. 6, 7, × 370; FIGS. 8, 9, × 350.

are somewhat elongated and oblong with rounded corners. In the older filaments the cells are oblong to quadrate with rounded corners. The terminal cell of the filament is conical and elongated (Text-figs. 25, 27). The basal attaching cell also of the filament is somewhat elongated and roughly obconical (Text-figs. 32, 34, 38). In very old filaments, very often, in some parts of them, the cell-division, instead of being transverse, is oblique. The filament in these places becomes very much bent or contorted in various ways. Since active cell-division generally takes place in these, with the planes of division oblique or diagonal and occasionally also at right angles to the previous division, the filament in such places has a jumbled pseudoparenchymatous appearance, with the cells having all sorts of shapes and placed in a very irregular manner in more than one series (Pl. I, figs. 1, 2).

Cell-structure

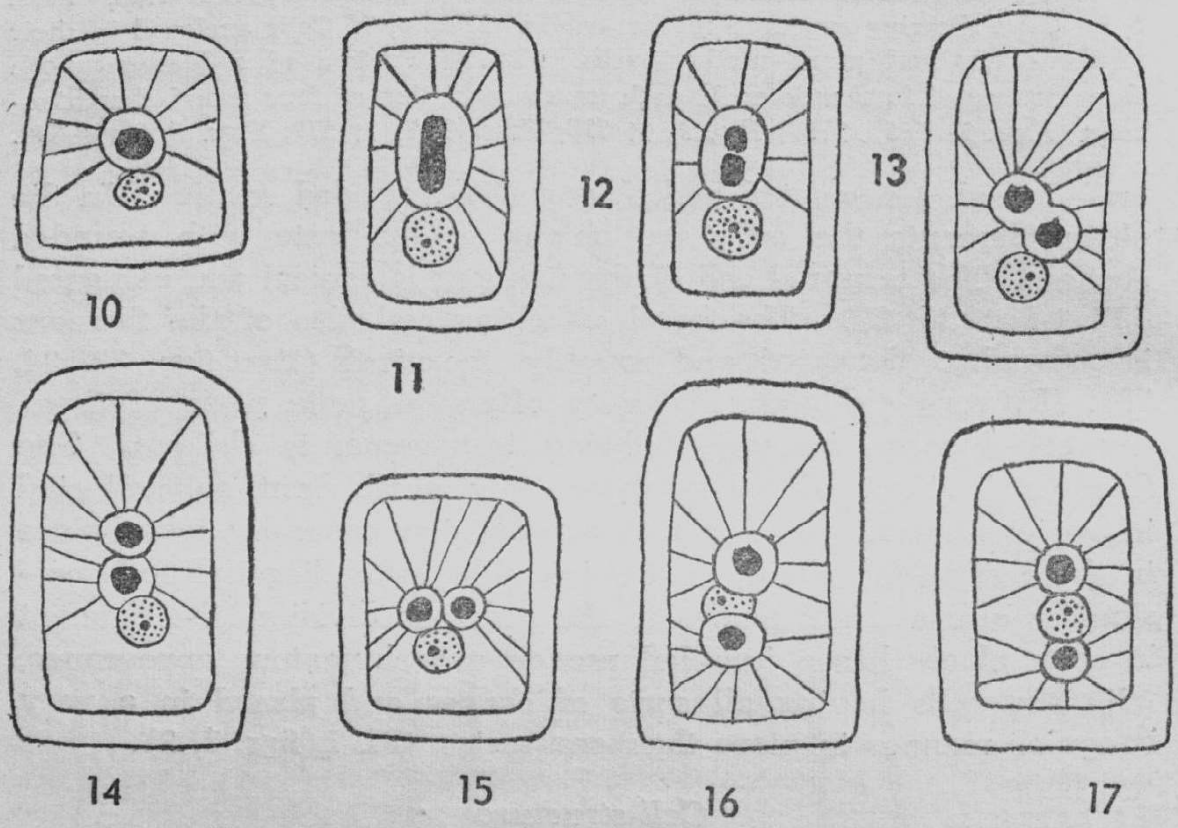
The cells are 22-28 μ thick, and about three-fourths to two and a half times as long as broad. The cell-wall is somewhat thick and lamellate, with up to four or occasionally more concentric layers, and has a bright refractive appearance. The cell-wall when treated with iodine and sulphuric acid becomes blue in all the layers, showing the presence of cellulose in it. When stained with ruthenium red, the cell-wall layers and the tubular sheath of the filament take a deep red stain, showing the presence of plenty of pectic substances also in them.

The cell contains a single nucleus and a stellate chloroplast in which is embedded a single pyrenoid (Text-figs. 2, 18, 30, 32). Owing to the cells being usually densely filled with starch grains, the nature of the chloroplast is not easily made out, but, if the alga is starved by being kept in darkness for a few days, most of the starch is used up and the stellate nature of the chloroplast is very clearly seen.

Cell-division

Just before cell-division, the pyrenoid divides into two. The details of the division of the pyrenoid and the movement of the

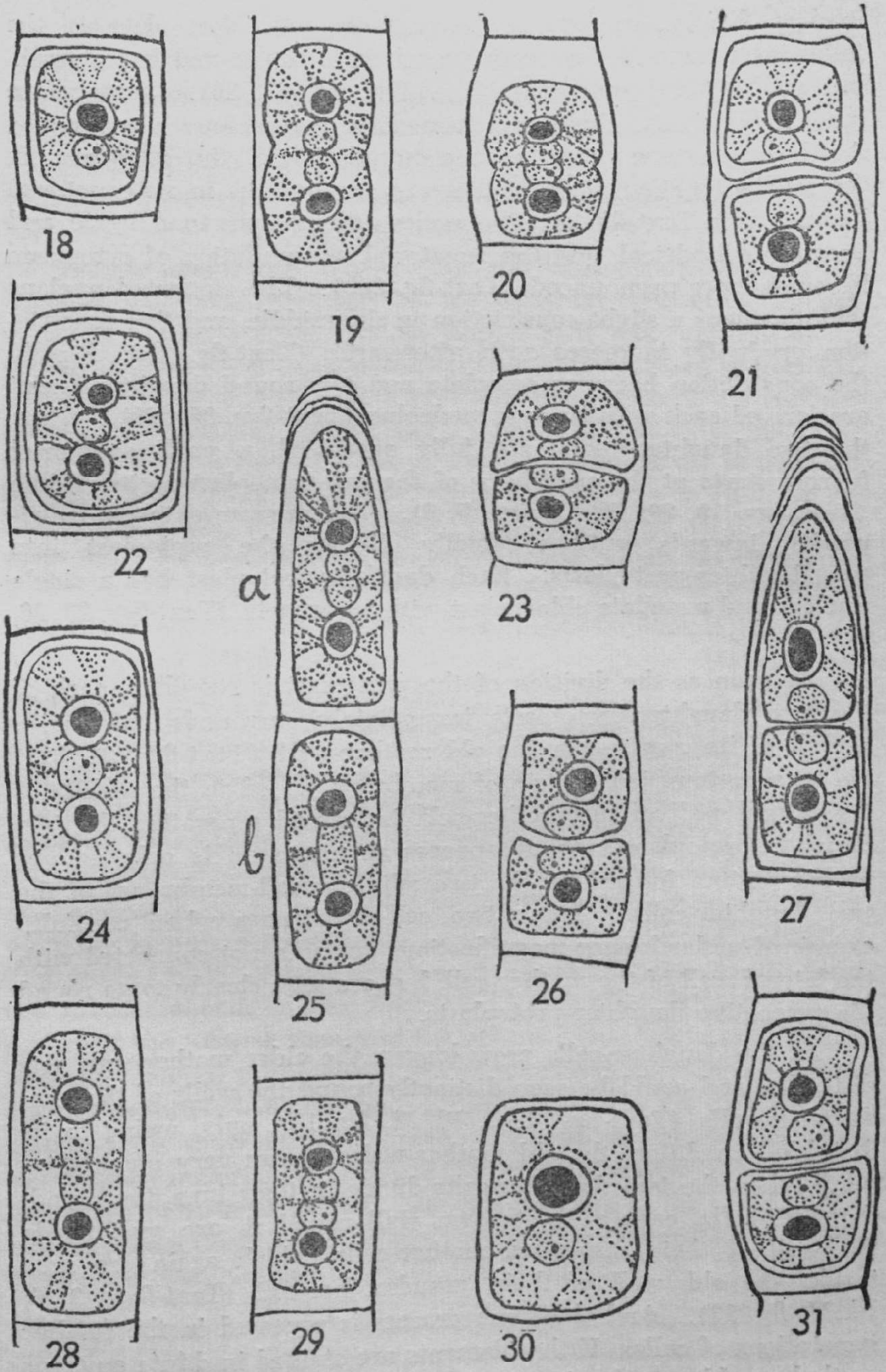
daughter-pyrenoids are shown in Text-figs. 10-17. The pyrenoid first of all becomes elongated along the longitudinal axis of the cell (Text-fig. 11) and then the pyreno-crystal divides into two



TEXT-FIGS. 10-17. *Cylindrocapsopsis indica* sp. nov.
Stages showing the division of the pyrenoids and the chloroplast, and their movement in the cell. Note: Nucleus and pyrenoids drawn with the camera lucida; the chloroplasts indicated diagrammatically by radiating lines. All FIGS. × 1000.

(Text-fig. 12). The pyrenoid then becomes constricted in the middle (Text-fig. 13) and divides into two daughter-pyrenoids. The two daughter-pyrenoids lie one behind the other as shown in Text-fig. 14. The upper daughter-pyrenoid then moves down so that the two daughter-pyrenoids remain at right angles to the longitudinal axis of the cell and in close contact with the nucleus (Text-fig. 15). After this the upper pyrenoid moves downwards round the nucleus (Text-fig. 16) and finally places itself close to the nucleus on its lower side (Text-fig. 17). As the pyrenoid divides, the chloroplast also divides (Text-figs. 14, 15) and the two daughter-chloroplasts with their respective pyrenoids finally place themselves at either end of the nucleus, as in Text-fig. 17. After this stage is reached the nucleus begins to divide.

During nuclear division in this alga no condensation of the chromosomes is seen, nor is there formed a nuclear spindle. The



TEXT-FIGS. 18-31. *Cylindrocapsopsis indica* sp. nov.
Stages of nuclear division and cell division. All FIGS. × 1125.

nuclear division appears to be amitotic and takes place in the following manner. The resting nucleus is round and has a small, but well defined nucleolus (Text-figs. 18, 22, 30). Just before nuclear division, a slight enlargement of the nucleus is noticeable. At the same time there is an accumulation of the cytoplasm in the median portion of the cell round the nucleus in an equatorial manner as in Text-fig. 24. The nucleus then begins to elongate, and becomes cylindrical, and the equatorial accumulation of cytoplasm becomes more pronounced (Text-fig. 25*b*). The elongated nucleus then develops a slight constriction in the middle, and this constriction gradually increases soon afterwards (Text-fig. 28). Finally the constriction becomes complete and two round daughter-nuclei are formed each with its own nucleolus (Text-figs. 25*a*, 29). When the two daughter-nuclei are fully organised, a small equatorial furrow starts at the periphery of the cell-protoplast at its middle (Text-figs. 19, 20; Pl. II, figs. 5, 6). This furrow extends centripetally inwards until it finally divides the protoplast into two daughter-protoplasts. Each daughter-protoplast has a single nucleus and a stellate chloroplast with a pyrenoid (Text-figs. 23, 26, 27).

As soon as the division of the protoplast is complete, each of the two daughter-protoplasts immediately surrounds itself with a wall of its own inside the old mother-wall (Text-figs. 21, 31). No true septum, i.e., a strip of membrane joined to the middle of the lateral wall of the mother-cell, is formed separating the two daughter-protoplasts. What appears as a septum is only an apparent septum formed by the two adjacent cell-membranes of the two daughter-cells. These two cell-membranes when carefully examined under higher magnifications are seen not connected with the lateral wall of the mother-cell, but are seen clearly to go round the respective daughter-protoplasts.

As the daughter-cells grow longer, the older mother-wall gets distended and could be seen distinctly round the walls of the two-daughter-cells (Text-figs. 1, 2; Pl. II, fig. 12), only, owing to its distended condition, the old mother-wall appears very much thinner. When the two daughter-cells divide in their turn, the granddaughter-cells, each with its own wall round it, are seen inside the wall of their respective mother-cells, which again are seen inside the old wall of their grandmother-cell (Text-figs. 3, 6; Pl. II, fig. 13). And the same process is repeated in the further generations of cells. If the filaments are stained in dilute aqueous methylene blue, safranin, gentian violet or ruthenium red, the walls of the several generations of cells are very clearly brought

out. If some material of the alga is boiled in strong caustic potash solution and thoroughly washed in water and mounted on a slide, the cell-contents are found completely dissolved, but the walls of the cells are seen quite intact. In such preparations, the walls of the daughter-cells are seen distinctly inside the wall of their mother-cell and quite unconnected with it (Text-fig. 1; Pl. II, figs. 7-9).

It was mentioned above that the wall of the mother-cell becomes distended and thin as the daughter-cells grow up and elongate. In some cases, the old mother-wall gets ruptured in the middle in a more or less circumscissile manner when it gets distended (Text-figs. 5, 9). In such cases the ruptured old mother-walls appear like H-pieces. After a few generations of cells, the end portions of these ruptured old mother-walls persist at the ends of the cells of the different generations (Text-figs. 4, 8; Pl. I, fig. 10, Pl. II, figs. 1, 13). Such cases show a close resemblance to *Ecballocystopsis,* a member of the Chlorodendraceae which has developed a filamentous habit (Iyengar, 1933). In fact the mode of cell-division and growth of the filament in the present alga is not at all unlike that of *Ecballocystopsis.*

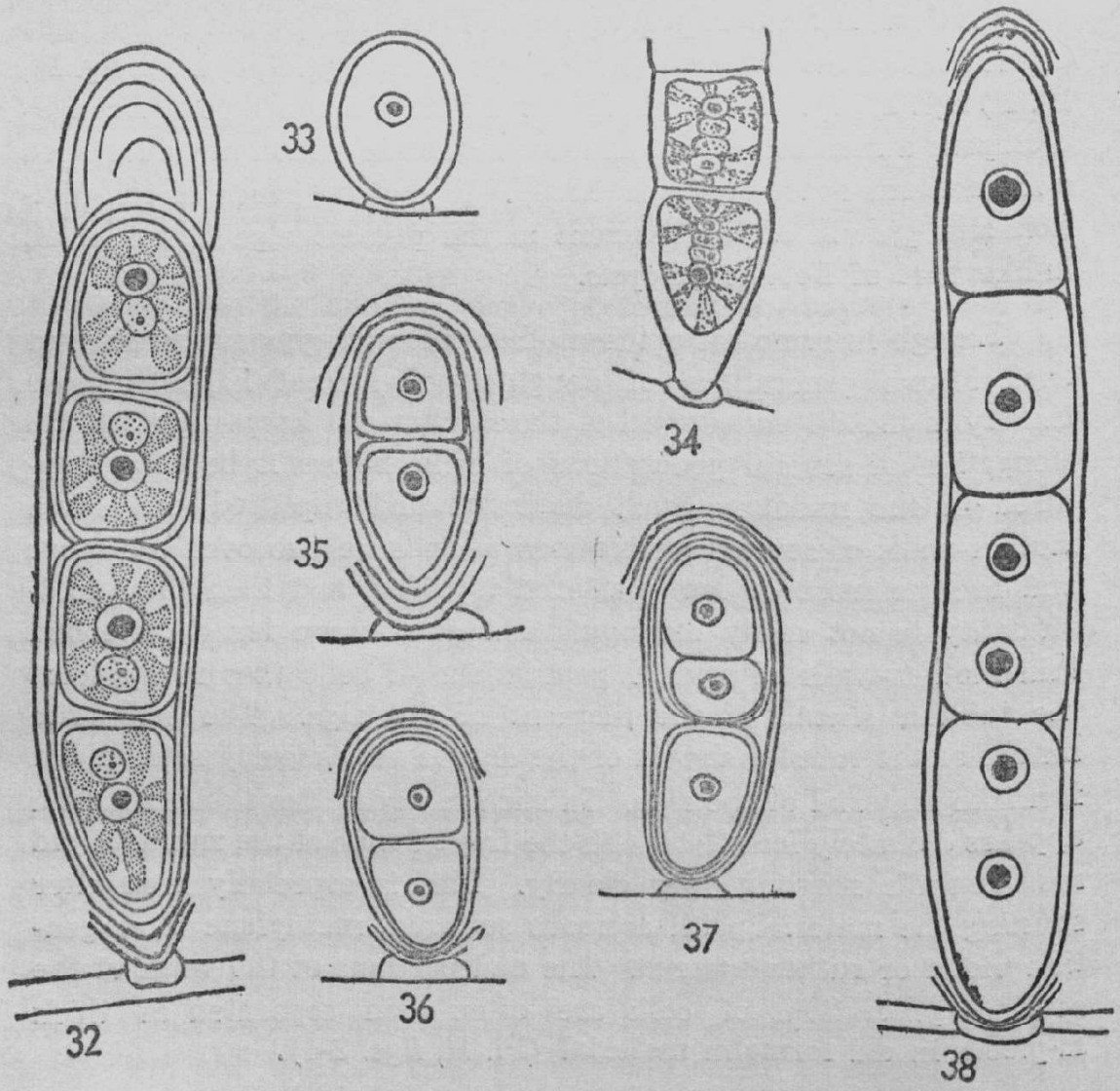
Though in some cases the mother-cell-wall gets ruptured after one or two generations of daughter-cells, what happens more usually is the distended wall of the mother-cell keeps up with the elongation of several generations of cells, before it becomes ruptured in the middle. Such distended and ruptured, very thin mother-walls of several generations of cells lie one over the other and become more or less plastered together, and then their individuality is not easily distinguished. It is these fused distended thin mother-walls of several generations of cells that really form the tubular sheath of the filament of the alga. The composite nature of the tubular sheath of the alga is very clearly seen in old formalin material of the alga. In such material, the tubular sheath appears definitely lamellate with the frayed ends of the ruptured old mother-walls showing very clearly. This is especially seen after staining the material with aqueous safranin, methylene blue, gentian violet or ruthenium red. The end portions of the walls of the several generations of cells are often seen separated as thick H-pieces in old formalin material (Pl. II, figs. 2-4).

During cell-division, the cells usually divide into two daughter-cells of equal size. But very often some of the cells divide into two daughter-cells of unequal size, one very large and the other very small (Text-figs. 3, 6). The significance of such an unequal division

is not quite clear. Quite frequently two adjacent cells divide unequally like this, and, when they do so, curiously enough, of the four daughter-cells, the two smaller daughter-cells are invariably side by side, while the two larger daughter-cells are away from each other (Text-figs. 3, 6).

Asexual reproduction

No zoospores were observed by the writer, even though he examined the living material of the alga both from the field and from the laboratory cultures for a long time on several occasions. But germlings of the alga in various stages of development were found growing on the older filaments of the alga and also on filaments of other algae, such as *Oedogonium* and *Pithophora,* in the



TEXT-FIGS. 32-38. *Cylindrocapsopsis indica* sp. nov.

FIGS. 32, 38. Four-celled germlings. FIG. 33. A single-celled germling. FIG. 34. The lowermost two cells of a fairly long attached filament. FIGS. 35, 36. Two celled germlings. FIG. 37. Three celled germlings. FIG. 32, × 1000; FIGS. 33-37, × 730; FIG. 38, × 965.

water. The writer evidently missed the time of their formation. The young germling is unicellular and obovate, and is attached to the host filament by a small mucilaginous pad at its base (Text-fig. 33). The germling is covered by a thin but firm wall. As the germling begins to grow and enlarge it secretes one or more new wall-layers below the first formed firm wall-layer. The new wall-layers are slightly thicker and more elastic. The unicellular germling soon grows in length and divides into two cells by a transverse division. Since the first formed wall-layer of the unicellular germling is firm and inelastic, it is unable to cope up with the growth in length of the germling, and so becomes ruptured at its base and remains as a cap at the top of the germling (Text-figs. 35, 36). The cap is seen persisting over the apical cell of even well grown filaments for quite a long time (Text-figs. 25, 27, 32, 38). The germling by further repeated division of its cells grows into a long filament. The young filament remains attached to the host alga for a good while and finally breaks away from it and grows in the water along with the other filaments.

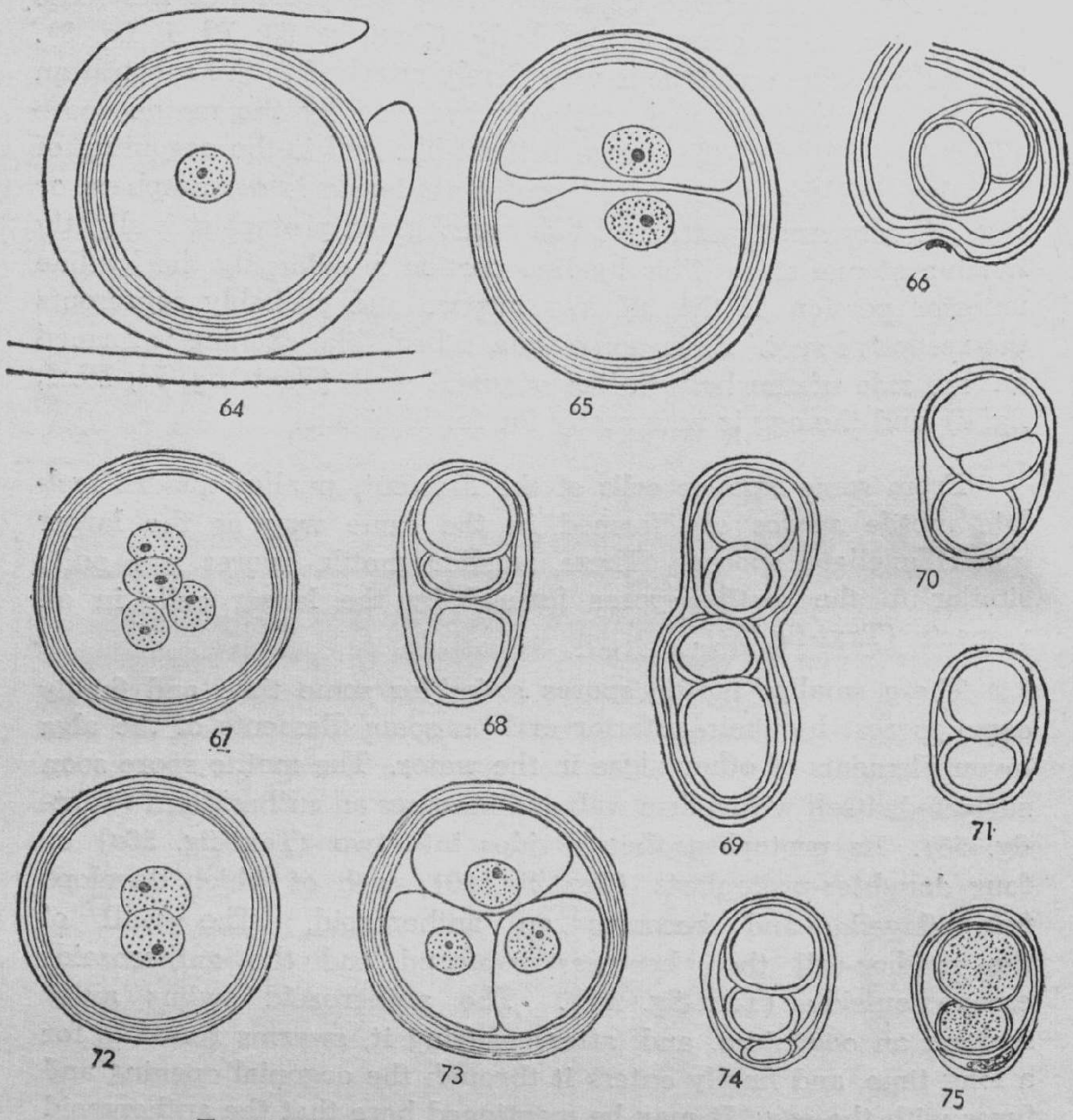
The alga also multiplies vegetatively by the fragmentation of its filaments into shorter pieces.

Sexual reproduction

Sexual reproduction was observed in the alga in the laboratory cultures during two successive years (1938 and 1939). Each year, sexual reproduction commenced only after the alga had lived in a sterile condition for about two months in the cultures. The details of the sexual reproduction are as follows. Some of the cells of the filament form large quadriflagellate motile spores (Text-figs. 41, 44). One motile spore is formed from each cell. The mode of formation of these motile spores in this alga is extremely peculiar and is quite unlike that seen in any other member of the Chlorophyceae, in fact, unlike that seen in any other known alga. The writer has, therefore, taken the liberty of giving it in some detail here below.

The cells which are to form the motile spores first of all become free and escape outside the filament. This escape of the cells from the filament appears like a passive process, but is evidently an active physiological one, being brought about by the gelatinization of the outer cell-wall layers, presumably through the active secretion of some enzymatic substance by the protoplasts of the cells concerned. The portion of

much broadened while its posterior end becomes somewhat conical (Text-fig. 58; Pl. I, fig. 5). It then pushes itself further against the egg when its protoplast becomes suddenly merged in that of the egg. When the two protoplasts fuse, there is a slight jelly-like vibration



TEXT-FIGS. 64-75. *Cylindrocapsopsis indica* sp. nov.
FIG. 64. A ripe zygospore with a thick lamellated wall inside an oogonium. FIGS. 65, 73. Protoplasts of zygospores divided into two and four daughter-protoplasts respectively. FIG. 66. An oogonium with a zygospore which had germinated and formed two daughter cells inside its old wall. FIGS. 67, 72. Zygospores whose nucleus had divided into four and two daughter-nuclei, respectively. FIGS. 68, 70, 74, 75. Three cells formed in a row through the germination of a zygospore. Note in FIG. 75 the degeneration of one of the cells of the first generation. FIG. 69. Four cells formed in a row from a zygospore. FIG. 71. Two cells formed from a zygospore. FIGS. 64, 65, 67, 72, 73 × 1400; the rest × 470.

of the fused protoplast for a second or so. The fertilized egg is spherical in shape. The actual fusion of the two protoplasts is very rapid and takes hardly a few seconds.

Sometimes after fertilization, the oospore covers itself with a wall. The oospore wall becomes very thick through the secretion of a number of wall-layers by the protoplast of the oospore (Text-fig. 62). These layers are quite firm and not gelatinous in consistency. The contents of the oospore after some days become bright red in colour.

The laboratory cultures of the alga were kept for a long time after the formation of the oospores. The oospores remained quite healthy throughout. In some of these oospores the nucleus had divided into two daughter-nuclei (Text-fig. 72) and in others the protoplast had divided into two daughter-protoplasts each with a single nucleus (Text-fig. 65). In some other oospores, the nucleus had divided into four daughter-nuclei (Text-fig. 67) and in some others again the protoplast had divided tetrahedrally into four daughter-protoplasts each with a single nucleus (Text-fig. 73). The further development of these daughter-protoplasts could not be followed. The division of the oospore nucleus into four nuclei suggests meiotic division, but no details of these nuclear divisions could be observed. But in a number of other cases, the oospores were observed to divide into two daughter-cells (Text-figs. 66, 71; Pl. I, figs. 6, 10), one or both of which divided again, so that there were three or four cells formed in a row (Text-figs. 68, 69, 70, 74, 75). Each of these cells possessed a thick wall and looked like resting cells. These cells were kept for a long time in the laboratory cultures, but they did not develop beyond these stages. The exact significance of these cells is not clear. Whether they represent the actual germination of the oospore or are merely cases of parthenogenetic development of unfertilized egg-cells could not be stated with certainty. Further careful cultural and cytological work is needed to elucidate these points.

General Discussion

The life-history of the present alga is very unique. From some large cells of the filament, large quadriflagellate motile spores are formed. These motile spores form one-celled female plants. Each of these one-celled female plants becomes bodily converted into an oogonium with a single egg inside. From some smaller cells of the filament, smaller quadriflagellate motile spores are formed. These motile spores form one-celled male plants. Each of these male

plants becomes bodily converted into an antheridium from which are formed later on two or four quadriflagellate antherozoids. The antherozoids seek the oogonia and fertilize the eggs inside.

The one-celled male plant of the alga may, in a way, be compared to the dwarf male plants of the Oedogoniales, in a number of species of which the dwarf males are unicellular. But the one-celled female plants are without any parallel in any other member of the Chlorophyceae. The motile spores which give rise to the male and the female plants cannot be considered as zoospores, since they do not reproduce the main plant in an asexual manner like all zoospores. They will have to be considered as special motile spores for the formation of one-celled sexual plants. The motile spore which forms the one-celled male plant may be called, following the terminology used for the Oedogoniales, an *androspore*. A new term, *gynospore,* may be created to designate the motile spore which forms the one-celled female plant.

Again, the mode of formation of these motile spores (androspores and gynospores) is very interesting. The cells which are to form these motile spores first of all escape from the filament with their cell-walls more or less quite intact, and then the contents of these loose cells escape outside as motile spores. Such an escape of walled-cells bodily from the thallus as a preliminary to the formation of motile spores is a rather unique feature and is quite unknown in any other member of the Chlorophyceae.

Another important feature of the alga is the amitotic division of its nuclei. Amitosis is considered by some authors as a degenerative or pathological phenomenon, and generally as occurring under abnormal or unfavourable conditions, such as are found in old cultures. But quite a number of authors state that amitosis could be normal and could be seen in cells which are quite healthy and show no signs of degeneration (Sharp, 1934, pp. 183-4). In the present alga, the amitotic division was found both in the material in the laboratory cultures and in the material freshly collected from the pond. The alga was in a very healthy condition and was growing actively in both the places. Amitosis is a very rare phenomenon in plants. Apart from the Characeae which is well-known for the amitotic division of the nuclei of its internodal cells, only a few cases of amitosis have been recorded so far in the Chlorophyceae, for instance, in *Valonia* by Fairchild (1894) and in *Spirogyra mirabilis* by Peterschilka (1923). The occurrence of amitosis in the present alga adds one more instance of amitosis

among the Chlorophyceae. But present day opinion is rather suspicious of all reported cases of amitotic division in algae, except in the internodal cells of *Chara*. The author, therefore, proposes to reinvestigate the nuclear division in the present alga using the Feulgen technique at the earliest opportunity, when suitable fresh material of the alga becomes available again.

The cell-division in the present alga is very interesting. In unicellular algae during cell-division the protoplast divides into two or more parts and the daughter-protoplasts cover themselves with new walls of their own inside the wall of the mother-cell. The daughter-cells after a shorter or longer period become free through the rupture or gelatinization of the mother-wall (West & Fritsch, 1927, p. 28; Printz, 1927, p. 6; Fritsch, 1929, p. 110; 1935, p. 17). In filamentous and other higher algae, cell-division is generally considered to take place in quite a different manner from that of the unicellular algae. The cell-division in these higher algae is considered to take place by a type of division known as *vegetative division* in which the parent-cell divides without rupture or gelatinization of its wall, the two daughter-protoplasts that are formed being merely separated by the development of a strip of membrane (septum) which is joined laterally to the membrane of the parent-cell. The two cells that are formed adopt the membrane of the parent-cell except for the intervening septum which is new and common to both (West & Fritsch, 1927, pp. 28, 142; Fritsch, 1929, p. 110; 1935, p. 17; Printz, 1927, p. 7). But according to Pascher (1924, p. 152), cell-divison in filamentous algae also takes place in the same manner as in unicellular algae and not by the formation of an intervening septum which is connected laterally to the parent-membrane. According to him, the protoplast of the cell divides into two, and the two daughter-protoplasts surround themselves with new walls of their own inside the parent-membrane, which consequently becomes suitably distended, so that the filament is composed of a linear series of such cells. The mode of cell-division is, in principle, in no way different from that seen in the unicellular algae except for the fact that only two daughter-cells are formed in each cell, and there is no marked contraction of the protoplast, so that the daughter-cells completely fill out the parent membrane. Cell-division in filamentous algae is nothing more than the formation of two autospores inside a cell. Printz (1927, p. 7) does not agree with Pascher's view. Vischer (1933) supported Pascher's view by giving a number of examples from the Chaetophorales. But Fritsch (1935, p. 18), while not com-

pletely rejecting Pascher's view, states that there is some difficulty in fully accepting Pascher's interpretation. He says, "While this interpretation appears to apply to a number of simple filamentous types (some Chrysophyceae; *Dinothrix,* fig. 237 A), it is difficult to harmonise it with the many accounts existing in the literature (cf. e.g. (173)) of the gradual ingrowth of a dividing septum during cell-division in green filamentous algae (figs. 69 D, E, I; 99 A). The data at present published by Pascher, as well as those afforded by Vischer (192), do not include a detailed consideration of such cases and one must await a fuller account before estimating their true value (cf. also (155) p. 6). Each cell of a filament, of course, develops its own appropriate thickening-layers subsequent to cell-division and in later stages the primary septum is often indistinguishable. The question to be settled is whether there is such a primary septum or not, and that can only be determined by a careful study of cell-division."

In the present alga the author made a very careful and detailed study of the cell-division. He found that the cell-division in the alga agreed fully in all details with Pascher's conception of cell-division in filamentous algae. During cell-division in the present alga, the protoplast divides into two and the two daughter-protoplasts then surround themselves with new walls of their own inside the parent-membrane. *No strip of membrane which is joined to the membrane of the parent-cell is formed at any stage in the present alga, nor could there be seen any indication of the formation of a primary septum connected laterally to the parent-membrane.* Whether Pascher's interpretation of cell-division in filamentous algae is true of all filamentous algae or not, it is certainly true in all details so far as the present alga is concerned.

Coming to the systematic position of the present alga, the alga in its vegetative condition shows a close resemblance to *Cylindrocapsa* in having cells with stratified walls, in having lamellated tubular sheath round the cells of the filament, and also in the frequent arrangement of the cells in some parts of the filament in more than one series through the oblique or longitudinal division of its cells. But it differs from *Cylindrocapsa* in its chloroplasts being stellate and not parietal as in *Cylindrocapsa.* The really main difference between the present alga and *Cylindrocapsa* is in its life-history and mode of sexual reproduction. The formation of quadriflagellate motile spores of two sizes, the larger forming one-celled female plants and the smaller forming one-celled male plants, the conversion bodily of these one-celled male and one-celled female plants

into antheridia and oogonia, respectively, and the fertilization of the oogonia outside the filament, are all features which are not found in *Cylindrocapsa*. Again, the motile spores in the present alga (androspores, gynospores and antherozoids) are all quadriflagellate, whereas, in *Cylindrocapsa*, the motile spores (zoospores and antherozoids) are biflagellate. So, except for a certain amount of superficial resemblance in the vegetative thallus, the present alga differs from *Cylindrocapsa* in all other respects.

The author had the privilege of discussing with the late Professor F. E. Fritsch, when the latter visited India in 1938, about the systematic position of this alga. Professor Fritsch, after considering all the different features of the alga, expressed the opinion that the alga could not be considered as a *Cylindrocapsa*, but might be placed in a new genus inside the Cylindrocapsaceae. The author takes this opportunity to express his indebtedness to Professor Fritsch for his great kindness in giving his opinion on the alga. The author agrees entirely with this opinion and accordingly places the alga in a new genus which he calls *Cylindrocapsopsis*. The genus is placed next to *Cylindrocapsa* inside the Cylindrocapsaceae. The alga itself may be called *Cylindrocapsopsis indica* sp. nov.[2]

Cylindrocapsopsis gen. nov.

Thallus, filamentous and unbranched, showing a close resemblance to *Cylindrocapsa;* attached when young, later free. Filaments composed of cells arranged in a single series; in older filaments, in places, cells arranged in more than one series through oblique or longitudinal divisions. Cells having a single nucleus and a stellate chloroplast with a single pyrenoid. Cell-membrane thick and stratified. Asexual reproduction very probably by zoospores. Sexual reproduction oogamous. Oogonia and antheridia unicellular and formed singly outside the plant by one-celled dwarf female and one-celled dwarf male plants, respectively, which again are formed by large quadriflagellate motile spores (gynospores) and smaller quadriflagellate motile spores (androspores), respectively. The large motile spores (gynospores) and

2. Randhawa (1941, p. 292) states that the present alga is the same as *Cylindrocapsa scytonemoides*, a new species described by him from Fyzabad (Randhawa, 1939, p. 275). But, unfortunately, *C. scytonemoides* is a very incompletely known species, and practically nothing is known regarding the structure of its chloroplast or the nature of its swarm-spores or of its sexual reproduction. Until these are clearly known, the question of the identity of these two algae cannot be profitably discussed.

small motile spores (androspores) formed singly from the cells of the filamentous plant. Two or four quadriflagellate antherozoids formed in each antheridium, and a single egg in the oogonium.

Cylindrocapsopsis indica sp. nov.

Characters same as for the genus. Cells elliptic when young and cylindric oblong to quadrate with rounded corners when older. Cells 22 to 28 μ broad and three fourths to two and a half times as long as broad. Antheridium 17 to 19 μ broad and 20-22 μ long. Oogonium 40-46 μ broad and 40-44 μ long. Oospores round, with thick stratified wall, and 23-29 μ in diam.

Hab. In a pond in Madras, growing on filaments of *Pithophora, Oedogonium,* etc., in the water.

Cylindrocapsopsis gen. nov.

Thallus filamentosus atque ramis carens, proxime accedens ad *Cylindrocapsam,* primo fixus, tandem liber. Filamenta constant ex cellulis dispositis in serium unicam; vetustiora filamenta aliquotiens monstrant cellulas dispositas in seriem multiplicem ob divisiones obliquas et longitudinales. Cellulae ornatae unico nucleo atque chloroplasto stellato, pyrenoideo unico. Cellulae membrana crassa atque stratificata. Reproductio asexualis probabilissime per zoosporas. Reproductio sexualis oogama. Oogonia atque antheridia unicellulata, atque evoluta singula extra plantam per unicellulatum minutam femineam plantam atque unicellulatam minutam masculam plantam respective, quae vicissim producuntur a sporis amplis quadriflagellatis sporis (androsporis) respective. Sporae amplae motiles (gynosporae) atque minutae motiles sporae (androsporae) efformatae singulae ex cellulis plantae filamentosae. In singulis antheridiis bina vel quaterna antherozoidea quadriflagellata efformantur, in singulis oogoniis vero ovulum unum efformatur.

Cylindrocapsopsis indica sp. nov.

Characteres ut in genere. Cellulae juniores ellipticae, seniores vero cylindrico-oblongae vel-quadratae angulis rotundatis. Cellulae 22-28 μ latae, et ¾-2·½ plo. longiores quam latae. Antheridia 17-19 μ lata, 20-22 μ longa; oogonia 40-46 μ lata, et 40-44 μ longa. Oosporae rotundae, ornatae parietibus crassis, levibus, stratificatis, dimetientes 23-29 μ in diam.

Habitat in stagnis in Madras, atque crescit affixa filamentis Pithophorae, Oedogonii, etc., in aqua.

Summary

The alga in its vegetative condition shows a close resemblance to *Cylindrocapsa*, except for its chloroplast which is stellate, unlike in *Cylindrocapsa* where it is parietal. It differs from *Cylindrocapsa* in its life-history and its mode of sexual reproduction. In this alga, quadriflagellate motile spores of two different sizes are formed from certain cells of its filament. The larger motile spore forms a single-celled dwarf female plant, which becomes bodily converted into an oogonium with a single egg inside. The smaller motile spore forms a single-celled dwarf male plant, which becomes bodily converted into an antheridium from which are formed later on two or four quadriflagellate antherozoids. The antherozoid seeks the oogonium and fertilizes the egg inside. On account of its unique life-history and its mode of sexual reproduction, it is placed in a new genus *Cylindrocapsopsis* and the alga itself is called *Cylindrocapsopsis indica* sp. nov. The new genus is placed in the Cylindrocapsaceae next to *Cylindrocapsa*.

During cell-division the two daughter protoplasts that are formed cover themselves with walls of their own inside the wall of the mother-cell. No septum is formed which is laterally connected to the parent-membrane. The mode of cell-division in this filamentous alga is very similar to that of the unicellular algae, and conforms to Pascher's interpretation of cell-division in filamentous algae. The nuclear division in the alga appears to be amitotic.

The author's sincere thanks are due to Rev. Father H. Santapau for rendering into Latin the diagnoses of the new genus and the new species.

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Explanation of Plates

Plate I

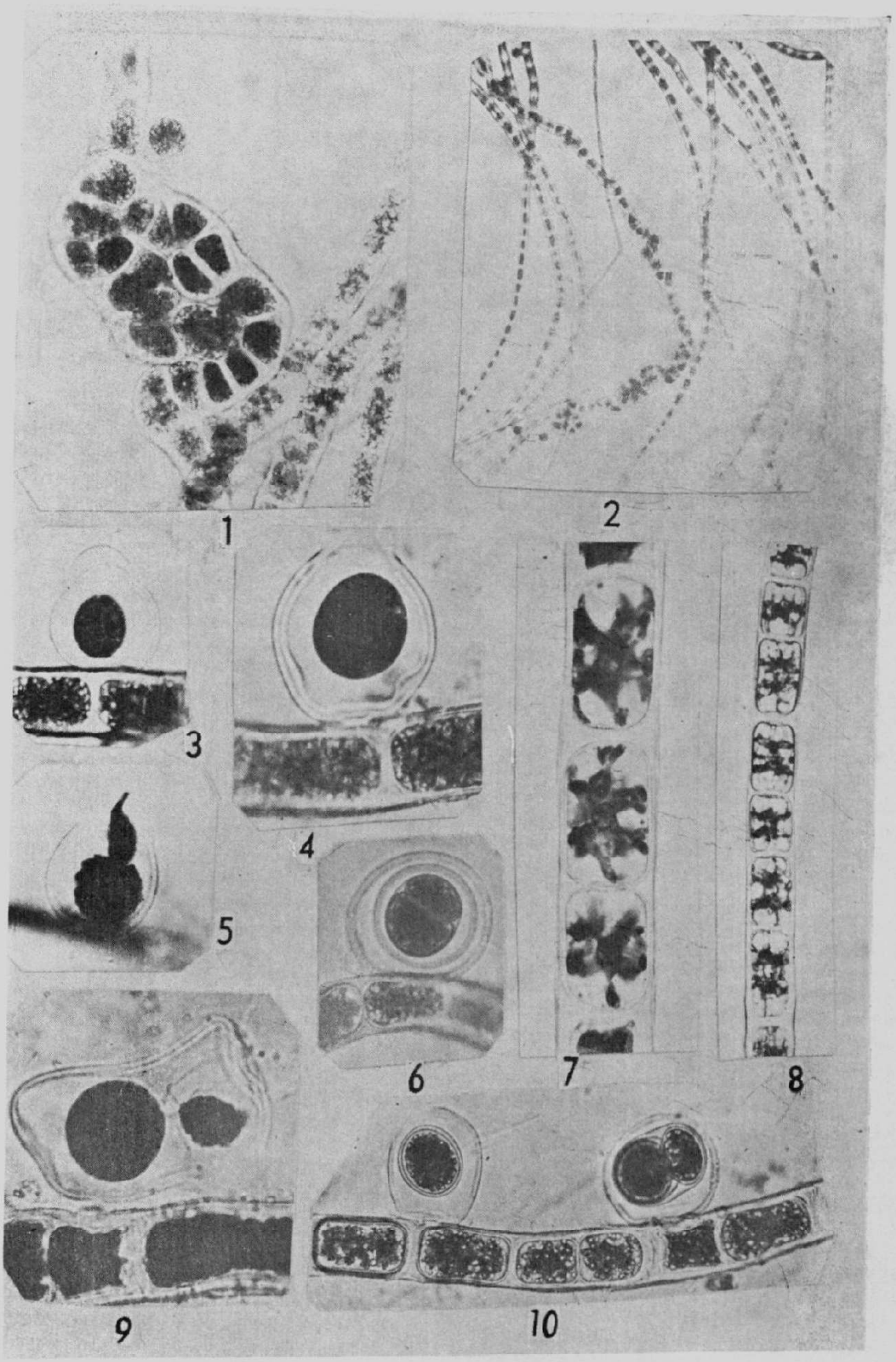
Cylindrocapsopsis indica sp. nov.

Fig. 1. Portion of a filament with pseudoparenchymatous growth. Fig. 2. A number of filaments of which one shows pseudoparenchymatous growth. Fig. 3. An oogonium with a single loose egg inside the enlarged oogonial wall. Fig. 4. An oogonium which has formed an opening in its wall. Figs. 5, 9. Fertilization of the egg by the antherozoid. Fig. 6. Contents of the zygospore dividing into two. Figs. 7, 8. Portions of the filament showing the cells arranged in a row inside a tubular sheath. Note the stellate structure of the chloroplast. Fig. 10. Filament of the alga with two oogonia settled on it; in one of which the oospore had divided into two daughter-cells inside its wall.

Plate II

Cylindrocapsopsis indica sp. nov.

Fig. 1. Portion of a filament showing an H-shaped thickening formed by the remnants of the mother-walls of the earlier generations. Figs. 2-4. Portion of filament from which the H-pieces have become loose. Figs. 5-6. Cells showing nuclear division and division of the protoplasts by furrowing. Figs. 7, 8, 9. Cells of filaments from which the contents have been dissolved out by boiling in caustic potash. Figs. 10-11. Stages in the escape of cells from the filament through the gelatinization of the tubular sheath of the filament. Fig. 12. A four celled germling growing on a filamentous alga. Fig. 13. Portion of a filament in which the cells of the first and second generations are enclosed in the mother-walls of the respective generations.



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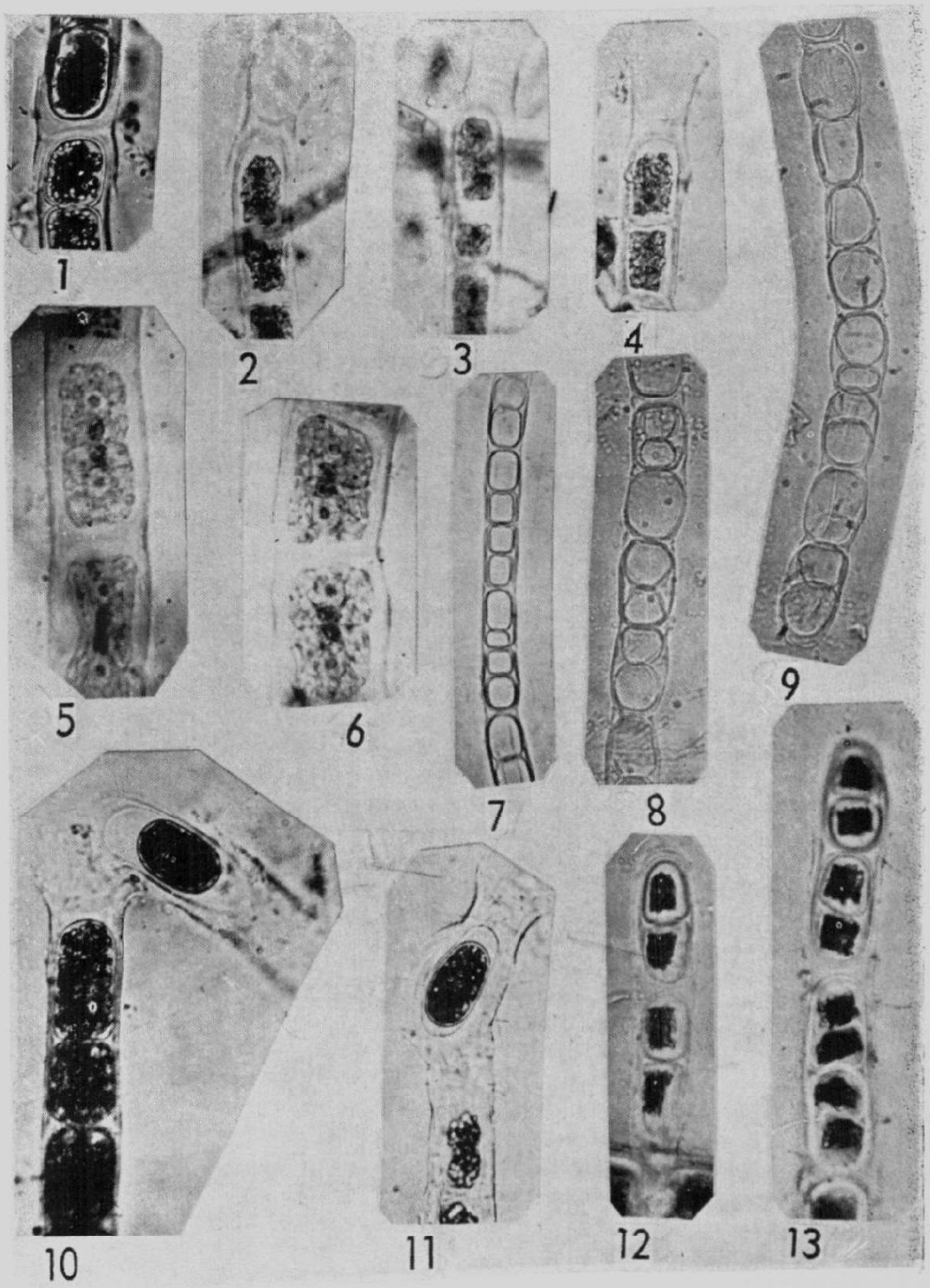

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Metamorphic Geology: Reflections on its Past, Present and Future

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The centenary of an institution devoted to all branches of sound learning may fittingly be made the occasion for taking stock of the position reached in the special disciplines. It may be worth while to pause and consider the present position of a science, for example, and to remember the painful and halting steps by which that position has been attained and to suggest the paths along which future progress may be made.

Such periodical assessments are particularly appropriate in the observational sciences where facts, or what pass as facts, accumulate with astounding speed. It must be remembered, however, that these accumulations can often be sorted out, interpreted or correlated only by personal predilection—but, nevertheless, such operations have their value. These conditions apply with special force to that branch of geology concerned with the metamorphic and granitic rocks—with the plutonic rocks, as I call them. In such studies, the personal factor is of profound importance since what has to be interpreted are field-facts and these, derived as they are from disconnected, partial and imperfect observations, constitute a very special class of facts indeed. In this essay, therefore, there must of necessity be a personal flavour—or bias, if you prefer it, about the discussion, but this flavour (or bias) is controlled by nearly half-a-century's fieldwork on the plutonic rocks and may accordingly be condoned.

I propose to deal with the past, present and future of metamorphic geology. Geology is the study of earth-history as revealed in the rocks making the crust—its essential aspect is the historical. I, therefore, have to survey the past, present and future of the study of the metamorphic rocks as historical records and not primarily as chemical, mineralogical or petrographical objects.

What is the past, or the present, or the future of a living science? For my purpose, I arbitrarily define the past as the time

before the First World War; I shall be especially concerned with what was put before students when I was one. The present lasts from the First World War to now; as a personal division it is my active period as a metamorphic geologist and is, I trust, not yet quite ended. The future is the active period of the potential metamorphic geologists now students. In what I have to say, I must select ruthlessly from the immense amount of available material and can mention only a few of the great names or the great ideas or the great pieces of work that embellish this branch of earth-history.

The Past

Half-a-century ago, the student of metamorphic geology found the subject simple and confined compared with the complexities and extensions that confront his successor of today. The main divisions had been established. There was a tendency, natural in view of the state of knowledge of the time, to oversimplify metamorphic processes and to explain what we now recognise as complex and diverse operations by a few distinct and separate mechanisms. In the orthodox view, there were two kinds of metamorphism, the one dynamic, the other thermal or contact. Dynamic metamorphism, the transformation of rocks through pressure and movement, had been developed from the early proposals of Lossen (1875) and Lehmann (1884) arising from their studies of German crystalline schists. Another classic of this time that we had before us was Teall's (1885) illustration of dynamic metamorphism by the transformation of diabase into hornblende-schist in the celebrated Scourie dyke in the North-west Highlands of Scotland. In the second kind of metamorphism, the thermal or contact, the modification of the rocks was due to their heating-up by invading igneous magma. The early study by Rosenbusch (1877) of Germany of the aureole of transformation around certain small granites in the Vosges had set the pattern for this class of metamorphic processes.

Rosenbusch, like most investigators of thermal metamorphism, held that no transfer of material had taken place across the contact of the magma with the invaded rocks, and this conclusion was applied to all contacts. But there was opposition from the French School of Michel-Lévy (1893-4), Lacroix (1898-9, 1899-1900) and others who had shown that at *their* granite contacts great transfer took place and that the processes of felspathisation and granitisation were widespread and were, indeed, essential parts of the mechanism

of granite emplacement. The ubiquitous gneisses, accordingly, might not all have been formed by the operations of dynamic metamorphism. The discerning geologist could have realised that metamorphic operations might be more diverse than the masters thought.

Early in this century, Sederholm (1907) of Finland had described the production of mixed rocks, the *migmatites,* by the admixture in various fashions and to varying degrees of granitic magma and country-rock. He considered that granitic magma could be made by ultrametamorphic processes, or anatexis, in the deeper levels of the crust. Adams and Barlow (1910) of Canada in their study of the Haliburton-Bancroft area of Ontario had proposed the formation of granitic magma by refusion. Another classic dealing with the migmatites was Fenner's (1914) interpretation of certain New Jersey gneisses as of mixed origin. In Scotland, Horne and Greenly (1896) had described the intimate penetration of country-rock by granitic material with an associated increase in metamorphic grade. Termier (1912) had introduced the notion of the 'oil-spot' mechanism in regional, widespread, metamorphism and in granitisation—fluids had spread unequally through a pile of strata just as oil does through a pile of cloths. Evidence of transfer and admixture was thus forthcoming from many countries, though Rosenbusch's ideas were dominant in academic circles. Nevertheless, if the student of that time was prepared to take a little trouble, he could indulge in as many heresies as can his successor of today.

In dealing with a broad spread of metamorphic rocks associated with what we would now call migmatites, Barrow (1892, 1912) in the South-east Highlands of Scotland had mapped out zones of progressive metamorphism marked by the appearance of certain index minerals in pelitic rocks, these minerals being, from low grade to high, chlorite, biotite, garnet, staurolite, kyanite and sillimanite. Barrow interpreted these metamorphic zones as something like a gigantic aureole of thermal metamorphism about a core of diffuse (migmatitic) granite. Another style of zoning had been worked out in a fundamental paper by Becke (1913) of Vienna. Becke showed that the mineral-associations and textures of metamorphic rocks depended on the pressure-temperature conditions of their formation and, on this principle, established the so-called depth-zones. These have since been elaborated by Grubenmann and Niggli (1924) into the highest or epizone controlled by low temperature and strong stress, the mesozone by

moderate temperature and stress, and the deep katazone by high temperature and low stress.

In the various fields of metamorphic geology sketched out in the foregoing, notable contributions were available from many countries. The student of my day could enlarge his knowledge of thermal metamorphism by study of Goldschmidt's (1911) magnificent account of contact metamorphism in the Oslo district of Norway, or of dynamic metamorphism by Niggli's (1912) description of the crystalline schists of the Gotthard massif in the Alps. A host of American workers had dealt with the phenomena of transfer shown at many of their limestone-granodiorite contacts. These samples of these contributions could be extended very greatly but, taking it by and large as the saying is, the student of that period found metamorphic geology reasonably simple. We had contact metamorphism (without transfer) and dynamic metamorphism as accepted and official, metamorphic zones and depth-zones were established notions, migmatisation was admitted but was to be connected with granitic magma, anatexis was considered rather an advanced idea whilst granitisation was scarcely to be mentioned in polite geological circles. Whatever modifications the present may have produced in these past ideas, the classic papers mentioned above, and many another not mentioned, may still be read with great profit. In them will be found the basis for many of the newest interpretations of metamorphic phenomena.

The Present

Developments in metamorphic geology since the First World War have been immense and varied though, as I have said, they were in many cases further growths of ideas already in existence. I propose to survey what I consider to be the major advances and to comment upon them where I find it appropriate to do so.

We may begin with Barrow's fundamental notion of *metamorphic zones*. In my opinion, Barrow's original Scottish zones have been applied correctly to the rocks of a few areas, but in general they appear not to be for export from their native home. For example, in Dutchess County, New York, they are found to be telescoped, as shown by the work of Balk and Barth (1936). Again, in the diverse metamorphic rocks encountered in Donegal, Ireland, none of Barrow's indexes are valid; garnet, staurolite, biotite, sillimanite and andalusite, for example, all occurring in the same rock. The reason for this is clear—the different minerals

record different episodes in the history of the metamorphic rock. As I emphasize later, time is just as essential a dimension in metamorphic geology as in any other branch of the science. Metamorphic zones can only be properly assessed in relation to the geological history of the associated rocks and dependence on a particular mineral is a hazardous proceeding. This is especially the case in dealing with rocks of different bulk-compositions, as the index minerals appear at different times in different rocks.

The danger of depending upon some key mineral to give the answer is exemplified to my mind in the use of the so-called *stress and anti-stress minerals* of Harker (1932). Since we now know that all the alleged stress minerals can occur in an environment in which no stress can have operated, and anti-stress minerals in a stress environment, history read by such aids is often wrong. Once again we find that the total geological history recorded in all components and all structures and textures is what has to be elucidated.

One of the most fertile of the modern developments in metamorphic geology has been the group of ideas clustered around the *metamorphic facies* principle of Eskola (1922). This principle has much in common with the extension by Grubenmann and Niggli (1924) of the Becke depth-zones. In the metamorphic rocks that have reached equilibrium, characteristic mineral assemblages in an isochemical series are controlled by the pressure: temperature conditions. For example, in Eskola's pioneer Karelian work (1925), he showed that green-schists, epidote-amphibolites and amphibolites formed an isochemical series developed from diabases by metamorphism with increasing temperature. As has been noted, the Grubenmann-Niggli zones are allied to metamorphic facies, as are also related developments by Roques (1941) and others. Properly applied, the facies principle and depth-zones are of great value in metamorphic geology—but they are not of infinite application and, in my opinion, care is needed. Again the time element in metamorphism must receive its due emphasis, and non-equilibrium assemblages may be of as great a significance as any others.

It is to the metamorphic rocks that the modern advances in *petrofabrics* are especially applicable. This branch of petrographical investigation is associated with the name of Sander (1930) of Innsbruck. Most metamorphic rocks show a preferred orientation of their constitutent minerals, either of form or of crystal

lattice. This fabric expresses the symmetry of the flow-movement to which the rocks have been subjected. A vast amount of laborious detailed observation is available, but its interpretation is often a matter for discussion. It appears to me that petrofabric studies may often be a valuable ancillary tool in the unravelling of metamorphic history, but they are not the main tool—this is the geological relationships determinable in the field. The proper sequence is, I feel, to use the geological history to interpret the fabric of a metamorphic rock, rather than the other way round.

The opinion just expressed applies with equal force to the interpretation of those structures of the metamorphic rocks that are grouped under the heading *lineation*. These arise by the linear orientation of crystals or crystal-aggregates, of fold or crumpling axes, of the intersection of planes of bedding, schistosity, cleavage, etc. and by other means, and have had the special attention of E. Cloos (1946) and G. Wilson (1946). Lineations in themselves, it appears to me, are rarely capable of supplying the correct interpretation of the geological history of an area; but when the geological history has been worked out from the totality of the phenomena, then the lineations can be correctly interpreted. The lineations mapped in an area, as for example parts of Donegal, Ireland, may as a whole give an appearance of a completely random arrangement, but this arrangement becomes orderly when each episode in the history of the area is found to be marked by the formation of a lineation with a special trend—once again, timing clearly becomes essential in the study of the metamorphic rocks. We have recently found in Donegal that a truly magnificent lineation developed over large expanses is not connected with tectonic movements of the normal kind but with the emplacement of a viscous granite body—the geological history of the region permits the lineation to be correctly interpreted.

Lineations are only one class of *microstructures* that are nowadays very much the concern of the metamorphic geologist. Small structures of all kinds—microfolds, crinkles, cleavages, drag-folds and the like—must all be mapped. Especially important here is the relative dating of the various structures observed, as this makes possible the formulation of the structural history of the area. Such a history can usually be correlated with the metamorphic history and a detailed geological history established. Thus, in parts of the Moine Series of the Northern Highlands of Scotland, Ramsay (not yet published) has related the small-scale

structures to two episodes of large-scale folding. Timing, as always, is an essential in earth-history, as in any other history.

Between the Wars, the notion of *metamorphic differentiation* had the attention of Eskola (1932), Read (1933, 1934B) and others. Material, especially in the juicy setting of the migmatitic complexes, moves around and is regrouped to provide profoundly different arrangements from the original ones. As an example, I may cite the formation of quartz-chlorite veins in chlorite-schists, of quartz-andalusite veins in andalusite-schists, of quartz-kyanite veins in kyanite-schists and so on. Exudation with increased mobility of material has clearly operated.

The regionally metamorphosed rocks are associated in their higher grades with areas of granitisation or migmatisation—granitisation here meaning the production of granitic rocks without their passing through a magmatic state. The relationship between such *regional metamorphism and granitisation* has naturally been the subject of much recent discussion. This is so because of the general debate on the origin of the granitic rocks as a group. I have proposed the Plutonic Series, composed of the metamorphic, the migmatitic and granitic rocks, as expressing a valuable genetic idea, generally valid. Regional metamorphism is to be connected in one way or another with a migmatisation or granitisation centre. If this is indeed the case, then the validity and extent of any proposed granitisation can only be judged by the geological setting in which it is observed. Large-scale granitisation must be accompanied by regional metamorphism and migmatisation. An apparent exception to this rule appears in *orogenic metamorphism*, the development over vast areas of lowly metamorphosed, highly folded, rocks—the 'universal phyllites'. But even this exception fits into the whole drama of orogeny and granitisation.

Time in Metamorphic Geology

We may recall that I have defined geology as earth-history, and in any history time is an essential dimension. In my opinion, the most important advance in metamorphic geology in the last 30 years has come from the increased concern with the timing of events on all scales. It is no longer sufficient to describe a metamorphic rock as, for example, a garnet-staurolite-mica-schist and leave it at that—a whole host of possible time-relationships must also be considered. We now look into this matter a little more closely—I have already made repeated mention of timing in the previous sections of this essay.

The style of metamorphic studies has been profoundly changed in many countries by the *establishment of stratigraphical successions* in completely metamorphosed rocks. This has been done chiefly by the use of original sedimentation structures such as current-bedding and graded bedding—the 'way-up' of the rocks has been determined. With a knowledge of the original stratigraphical order of the metamorphic rocks concerned, the geologist can now interpret with confidence the structure of the most complex areas. As samples of the methods and results, papers by Bailey (1930), Read (1936, 1955B) and Iyengar, Pitcher and Read (1954) may be consulted. The metamorphic geologist now has a weapon in his armoury as powerful as William Smith's two laws are to the stratigrapher.

A series of time-relations exists between the various episodes of crystallisation, deformation, migmatisation, granitisation, magmation and orogeny that may affect a metamorphic rock, and must nowadays be considered by the metamorphic geologist. We can begin with simple cases and lead up to the more complex.

The relation between the *times of crystallisation* of the mineral components of a metamorphic rock must be studied as, for example, in the researches of Cheng (1944) and Hietanen (Cloos and Hietanen, 1941). In different metamorphic rocks, too, *deformation and crystallisation* may bear different time-relations to one another. The familiar example of 'rolled garnets', with spirally arranged inclusions of the groundmass of the rock, indicates the contemporaneity of movement and crystallisation. The three great classes of time-relations of this kind have been established as pre-crystalline, para-crystalline and post-crystalline deformation, with obvious and important meanings.

By proceedings such as these, the *metamorphic history* of complex areas can often be deciphered. As an example, I may cite my own work in Unst, in the Shetland Islands, Scotland, (Read, 1934A, 1937) where it can be shown that a particular block of country has been subjected to three successive metamorphisms of different kinds: a first gave staurolite, kyanite, garnet and biotite in pelitic rocks; during a second this assemblage was unstable and one of sericite, chlorite and chloritoid took its place, to be replaced during a third metamorphism by one of chlorite and sericite. A long metamorphic history under varying physical controls has thus been read, and this history is obvious in the field and obvious under the microscope.

As a second example of this *polymetamorphism,* I may give the investigation of the Lewisian complex of the North-west Scottish Highlands by Sutton and Watson (1951). These workers have shown that an early complex of charnockitic facies has been elevated and cooled sufficiently so that diabase dykes intruded into it are chilled, and then the complex, with the dykes, has been subjected to a second metamorphism giving an amphibolite facies. This kind of investigation gives a new concept of the meaning of a geological era as applied to the metamorphic complexes which is as valid as that employed in more normal stratigraphy.

As the growth of a fold-belt continues, magma or migma becomes involved and the time-relations between *orogeny and the emplacement or manufacture of magmatic or migmatitic rocks* have to be considered. The dating of migmatisation or of the appearance of different kinds of granitic bodies is of great importance in metamorphic or plutonic studies. Intrusions have been classed as pre-tectonic, syntectonic or post-tectonic—terms with obvious implications. I have endeavoured to codify these relationships in what I call the Granite Series (Read, 1949), a series which relates the nature and form of different types of granitic bodies with their place in the fold-belt and the time of their final solidification. The *Granite Series* can be represented thus:—

TIME →

CRUSTAL LEVEL →

Autochthonous granitisation granites, migmatites and metamorphites.	Parautochthonous granites	Intrusive magmatic granites	Plutons

Deep in the fold-belt are formed, at an early stage of the orogeny, great complexes of granitisation granites associated with migmatites and widespread regional metamorphic rocks. As the orogeny continues, a part of these autochthonous granites becomes partly unstuck from its surroundings and moves higher in the fold structure. This process continues with the movement of true intrusive and magmatic portions late and high, and culminates in the emplacement of the granite plutons, highest and latest, pushing their way as almost solid bodies even into the post-orogenic sediments. It seems to me (though, as I foreshadowed in an earlier page of this essay, I may be biassed) that the Granite Series brings into orderly connection orogeny, granitisation, metamor-

phism and granite emplacement. I have suggested that we have here recorded related aspects of one operation—a down-buckling of the sima below the geosynclinal depression. Such a down-buckling, it is thought, would cause the liquefaction of the sima, to supply the ophiolites and related basic rocks in the geosynclinal piles; with the subsequent consolidation of this melted sima, fluids would be supplied to granitise the geosynclinal sediments, the Granite Series is initiated and, as the fold-belt develops, its members take their places in orderly succession (Read, 1955A).

As I have repeatedly emphasized in the foregoing pages, the timing of the diverse events that make up metamorphic geology is the most valuable aspect of present-day studies of the science. A metamorphic or granitic rock is now seen to record a certain precise stage in crustal evolution. Such rocks are not formed in any haphazard fashion, but arise at definite times in circumscribed places. When we know more about metamorphic geology we shall confirm a previously expressed opinion of mine, namely, that petrogenesis has its role in historical geology as much as has stratigraphy.

The Future

After a not inconsiderable experience of the metamorphic and granitic rocks, I may perhaps be permitted to set out the one main result of that experience and then to go on to suggest what I think would be desirable future developments in this study.

This one conviction is the simple one that no two metamorphic rocks are identical—the variables attending their formation are too many. There are pre-metamorphism variables, variables operating during one or more metamorphic episodes, and post-metamorphism variables. Each area of metamorphic rocks has, therefore, to be approached as if it were a piece dropped from some extra-terrestrial source. It has to be considered on its own merits and its rocks must be permitted to tell their own story, even though this story may be unlike any heard before. Generalisations in metamorphic geology should remain extremely general and the application of results attained in one field to another should be done with eyes exceedingly wide open. How then is the metamorphic geologist to proceed? The answer is that given to geologists in general by the illustrious Charles Lapworth—*map it, my boy, and it will all come out.* Metamorphic geology, like all geology, is based on field-work, and advances will come with more and better and, often, different maps.

But, first of all, a simple reminder is necessary; the value of a geological map depends in the first place on there being rocks to map. The importance of abundant exposure in field-studies of the crystalline rocks cannot be overemphasized. It would be advantageous to the progress of metamorphic geology for authors of papers to state clearly what the nature of visible outcrop was in the area under description and thereon to assess the validity of their results. The same remark applies of course to all enquiries of a statistical type; examination of one thin section or the chemical analysis of one gramme of rock cannot give automatically the composition of a large granite body of thousands of cubic kilometres in volume—the one section and the one analysis are valuable, it is true, but extrapolation from them, without controls, may be dangerous. Dare I venture to suggest that the routine mapping of poorly exposed ground should be left to Government Geological Surveys whose officers, after all, are paid to take the fat with the lean ?

Nevertheless, one quarry in metamorphic rocks properly mapped can yield results of great value. We need maps on all scales and especially maps on very large scales of very small areas. We need too a number of special maps, dealing with the results of one episode in what may be a long and complex history; one fold-system or one metamorphism or the distribution of different metamorphic minerals or of one set of small-scale structures can be dissected and given in separate maps. Such maps can then be assembled to give something like a film, though admittedly a jerky one, of this history.

If well-exposed and extensive areas of metamorphic and granitic rocks are to be investigated in such detail as to give the best possible results for the science, then it seems to me that team-work is essential. It is, for example, certain that one man, however able and energetic, working for a lifetime in Donegal, Ireland, could not have produced the body of knowledge that the team (including I am glad to say, Indian members) from my own College has assembled in half-a-dozen years. Mapping on a uniform basic scale and, with due allowance for personal fancies, in a uniform style, can thus be extended over considerable areas; such mapping all agrees with the geological climate of the time and can be uniformly adjusted to suit the geological climate at any instant in the future. Planning in geological research is essential to progress.

It is evident that an important class of maps of metamorphic areas should deal with the pre-metamorphic characters of the rocks that we now find metamorphosed. This is particularly desirable in

studies of the structure and correlation of metamorphic rocks over large regions. The original stratigraphy, the order of succession, the facies changes, the conditions of formation, the different places of deposition in the geosyncline, and similar topics are all essential to the full story of the metamorphic rocks. The present lithology of our rocks must be contrasted with the original lithology. Correlation can be done mainly with the help of this original lithology.

In the future it is desirable that the vast field of inquiry opened up by the development of rapid methods of chemical analysis of rocks and minerals should be extended in the metamorphic and granitic rocks. Such intense chemical investigation is especially necessary in coming to any conclusions concerning changes in composition, whether of rocks or of constituent minerals, during metamorphism or migmatisation. A dozen years ago, Lapadu-Hargues (1945) suggested that marked and significant changes in composition occurred in the series slate to migmatitic gneiss. Such studies can be extended and tested most expeditiously by the rapid methods of analysis now available. The compositional changes in aureoles and in metamorphic and migmatitic sequences, and how such changes are distributed among the mineral components, can now be established in a statistically satisfying manner. A beginning along these lines has been made, for example, by Pitcher, Sinha and Mercy in Donegal.

Another fruitful line of future work will be concerned with the determination of absolute ages of metamorphic and other plutonic rocks. Already valuable results have been achieved by Holmes (1951) and others in the dating of events in the complex histories of the ancient rocks of Africa and India. It must be remembered, nonetheless, that the extension of this work must be controlled by geological and mineralogical timing, but its possibilities appear to be very great indeed.

To my mind still, the fundamental requirement for advances in metamorphic geology remains more intensive, better co-ordinated and more properly assessed field-work. Geochemistry, geophysics, petrofabrics and the like provide invaluable ancillary weapons, but the main tool of the metamorphic geologist is still the 'silly little hammer'. Indian geologists, and especially those of Madras, the home of the charnockites, have magnificent opportunities for fundamental work in this branch of geology. On this auspicious occasion, I salute them and wish them success—they are indeed fortunate in that there is no lack of work ready to their hands.

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Raney Nickel Reductions

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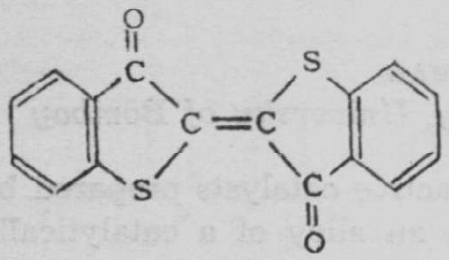
In 1925 and 1927 Raney[1] patented active catalysts prepared by dissolving out the inactive metal from an alloy of a catalytically active and an inactive metal. Raney nickel prepared in this manner by treatment of a nickel-aluminium alloy with caustic soda solution has been extensively used for catalytic reduction, desulphurization, and other purposes. The ability of Raney nickel to effect reductions in the absence of external hydrogen, utilizing the absorbed hydrogen, was demonstrated by Mozingo and his collaborators.[2] They thus removed sulphur from numerous sulphides, sulphoxides and sulphones, and from biotin methyl ester. Following the use of Raney (nickel-aluminium) alloy and aqueous caustic soda by Whitman and others[3] for the reduction of estrone to estradiols, Papa studied this reductive procedure, involving nascent hydrogen and nickel catalyst formed *in situ*, on a variety of organic compounds.[4]

Raney nickel desulphurization appeared to offer a useful approach to problems concerning the constitution of thioindigoids, sulphur dyes and sulphurized vat dyes, and dyes containing sulphonic groups. In a preliminary investigation[5] it was shown that under the Mozingo conditions thiodiphenylamine gave diphenylamine; benzidine-sulphone underwent desulphurization and *N*-alkylation, yielding *N* : *N'*-diethylbenzidine; carbazole gave tetrahydrocarbazole.

Treatment of thioindigo with Raney alloy and aqueous caustic soda at about 100° yielded diphenacyl (12% under the best conditions), 1 : 4-diphenylbutane and benzoic acid.[6]

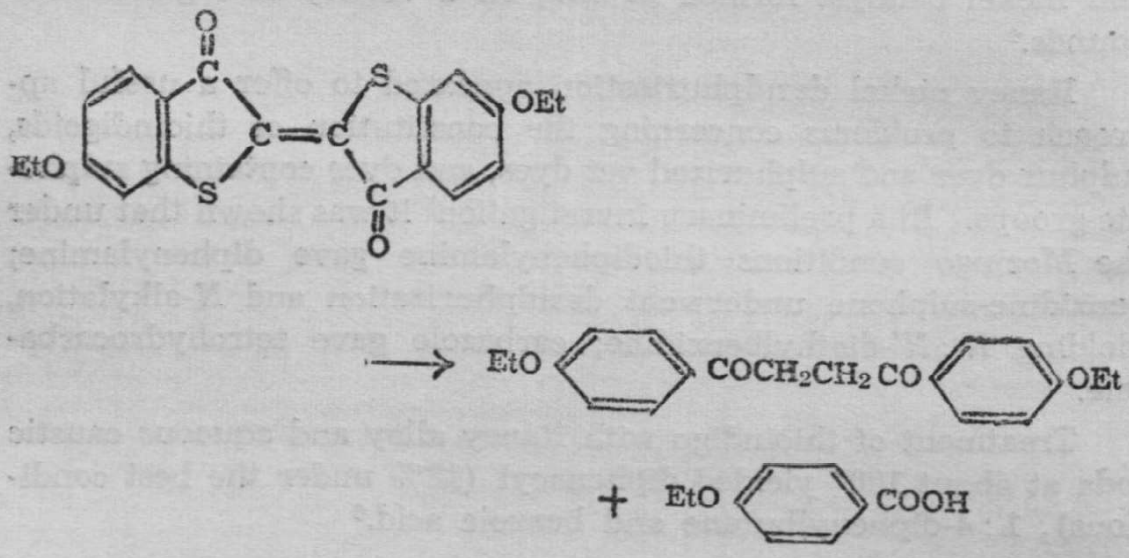
1. U.S.P. 1,563,787; 1,628,190.
2. Mozingo, *et al.*, *J. biol. Chem.*, 1942, 146: 475; *J. Amer. chem. Soc.*, 1943, 65: 1013; 1944, 66: 1859.
3. Whitman, *et al.*, *J. biol. Chem.*, 1937, 118: 792.
4. *J. org. Chem.*, 1942, 7: 587; *et sequa.*
5. Shah, Tilak and Venkataraman, *Proc. Indian Acad. Sci.*, A, 1948, 28: 142.
6. Kao, Tilak and Venkataraman, *ibid.*, 1950, 32: 162; 1953, 38: 244.

The isolation of compounds in which a carbonyl group attached to the benzene ring remains unaffected after a vigorous Papa reduction has not been recorded so far; Papa found that phenyl ketones,



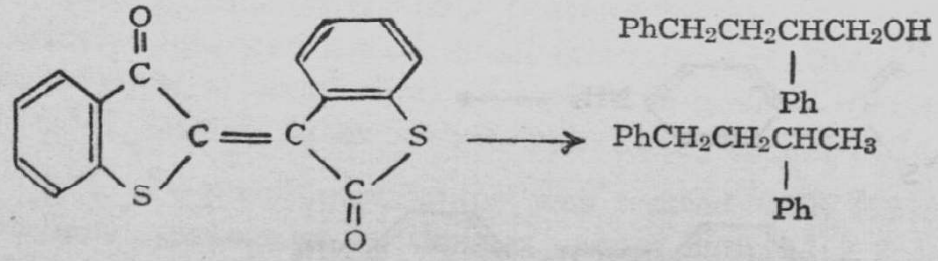
$$\longrightarrow \quad PhCOCH_2CH_2COPh + PhCH_2CH_2CH_2CH_2Ph + PhCOOH$$

PhCOR, were reduced to the corresponding hydrocarbons, $PhCH_2R$. Deactivation of the catalyst by poisoning with sulphur from thioindigo may account for the stabilization of part of the diphenacyl against further reduction. In connection with the isolation of benzoic acid, which may result from the hydrolysis of diphenacyl, it was noticed that acetophenone gave benzoic acid as one of the products of Papa reduction. 6:6′-Diethoxythioindigo (Durindone Orange R) led to 4 : 4′-diethoxydiphenacyl in 80% yield, indicating that the desulphurization of thioindigoid dyes is a practicable method for the preparation of diphenacyls, which are otherwise not readily accessible.

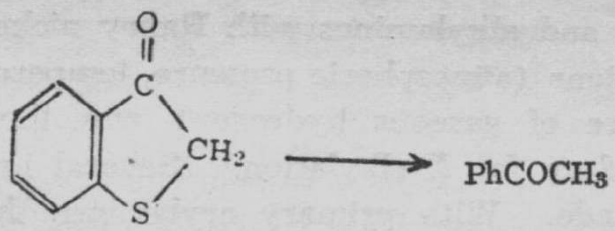


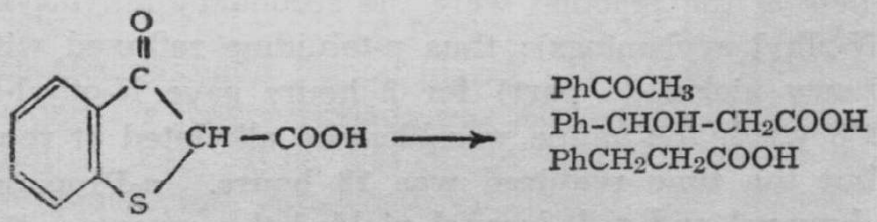
Papa reduction of the naphthalene analogue of thioindigo gave 1 : 2-di-α-naphthoylethane and a yellow non-fluorescent liquid which analysed for 1 : 4-di-α-tetralylbutane; a blue-fluorescent isomer of the latter, the two probably being the 1:2:3:4- and 5 : 6 : 7 : 8-tetrahydronaphthalene derivatives, was obtained by Mozingo reduction. The main product of the action of Raney

alloy and caustic soda solution on the unsymmetrical dye, 2:3′-*bis*-thionaphthene-indigo was 2:4-diphenyl-1-butanol; the corresponding hydrocarbon was obtained under Mozingo conditions.

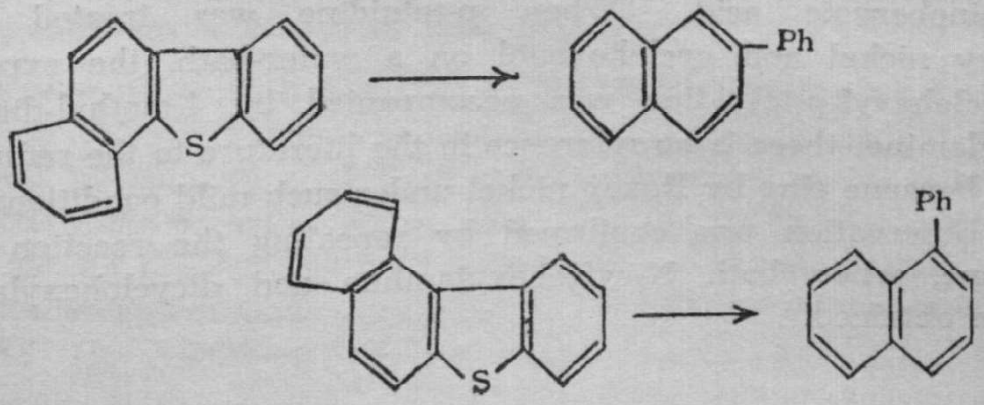


Reduction with Raney nickel and aqueous caustic soda at 0-5° gave very interesting results with thioindoxyl and thioindoxylic acid; the former gave acetophenone in 72% yield; the latter was completely desulphurized in five minutes with the formation of β-phenyl-β-hydroxypropionic acid as the main product, accompanied by acetophenone and β-phenylpropionic acid.



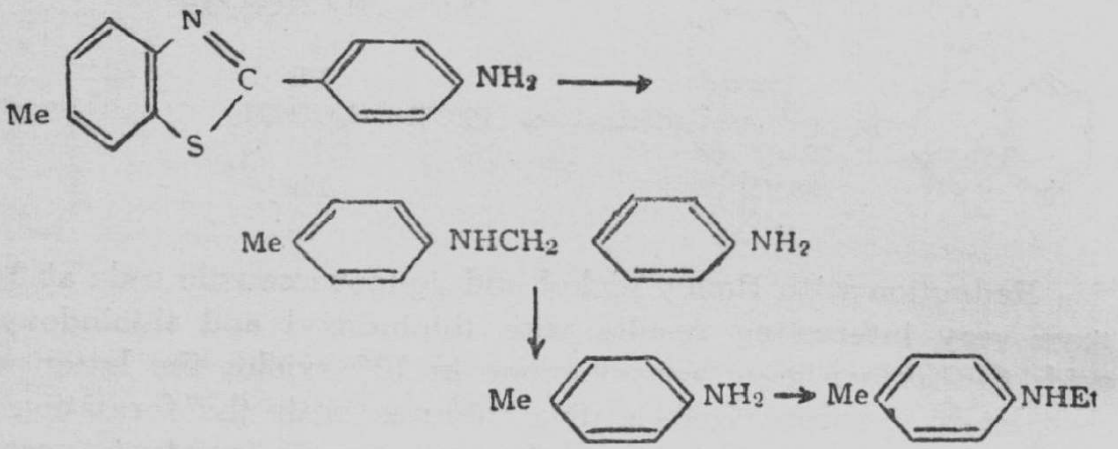


The isomeric 1:2- and 3:4-benzothiafluorenes were readily distinguished by Raney nickel in boiling alcohol, the respective products being 2-phenylnaphthalene and 1-phenylnaphthalene.



Reduction of dehydrothio-*p*-toluidine with Raney nickel in ethanol at room temperature gave *N*-ethyl-*p*-toluidine in good

yield; the reaction probably proceeded through the indicated mechanism, since benzylaniline under similar conditions gave *N*-ethylaniline.



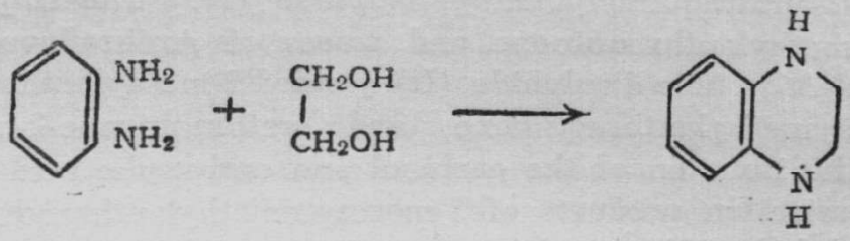
$PhNHCH_2Ph \rightarrow PhNHEt$

The interaction of arylamines and alkylamines with Raney nickel and alcohol under mild conditions (atmospheric pressure, temperature below 100°, and absence of gaseous hydrogen) was then examined as a general procedure for *N*-alkylation.[7] Several important observations were made. With primary arylamines the exclusive products of the reaction were the secondary alkylarylamines (mono-*N*-alkyl arylamines); thus *p*-toluidine refluxed with ethanol and Raney nickel (1 part) for 2 hours gave *N*-ethyl-*p*-toluidine in 88% yield; the amine was likewise alkylated at room temperature, but the time required was 18 hours. *n*-Propanol, isopropanol, *n*-hexanol and *n*-dodecanol yielded the corresponding *N*-alkylamines, but methanol did not react. *m*-2-Xylidine, presumably because of the steric effect of the *o*-methyl groups, did not react. The *N*-ethyl derivative was obtained in 70% yield from *p*-aminobenzoic acid. When *p*-toluidine was treated with Raney nickel and cyclohexanol on a water-bath, the expected *N*-cyclohexyl-*p*-toluidine was accompanied by 4-methyl-dicyclohexylamine; there is no reference in the literature to the reduction of a benzene ring by Raney nickel under such mild conditions, but the observation was confirmed by repeating the reaction with aniline, when both N-cyclohexylaniline and dicyclohexylamine were obtained.

7. Kao, Tilak and Venkataraman, *J. sci. industr. Res. India*, 1955, **14B**: 624. See this paper for references to relevant literature.

Alkylamines behaved towards Raney nickel and alcohols somewhat differently from the aromatic amines. Thus *n*-butylamine, Raney nickel and ethanol led not only to *N*-ethyl-*n*-butylamine, but also to di-*n*-butylamine (formed by the self-condensation of *n*-butylamine) and a very small quantity of diethylbutylamine. In the absence of an alcohol, *n*-butylamine gave di-*n*-butylamine and tri-*n*-butylamine on treatment with Raney nickel in boiling water.

When *o*-phenylenediamine was treated with Raney nickel in ethylene glycol on a boiling water-bath, 1:2:3:4-tetrahydroquinoxaline was isolable in low yield; if the conditions can be modified to improve the yield, a new route to quinoxaline derivatives may become available.



I II

After our work was completed, Rice and Kohn[8] and Ainsworth[8a] reported the *N*-alkylation of primary aromatic amines with alcohols, and Raney nickel. Rice and Kohn carried out the reaction for longer periods and Ainsworth used larger amounts of Raney nickel. Rice and Kohn failed to effect *N*-alkylation by isopropanol, but Ainsworth confirmed our observation that prolonged heating resulted in the formation of *N*-isopropylaniline in about 50% yield. Ainsworth found that the reaction of Raney nickel and alcohols with primary aliphatic amines was more complex; 2-phenylethylamine and ethanol gave di-(2-phenylethyl)-amine, and benzylamine gave dibenzylamine.

In connection with attempts to determine the constitution of Cibanone Yellow R and Cibanone Orange R, which are vat dyes prepared by thionation of 2-methylanthraquinone or its halogen derivatives, the action of Raney alloy and caustic soda solution at about 95° on the dyes and on anthraquinone, 2-methylanthraquinone, 2-mercaptomethylanthraquinone, *bis*-2-anthraquinonylmethyl sulphide (I; $AQCH_2\text{-}S\text{-}CH_2AQ$; AQ = 2-anthraquinonyl), and

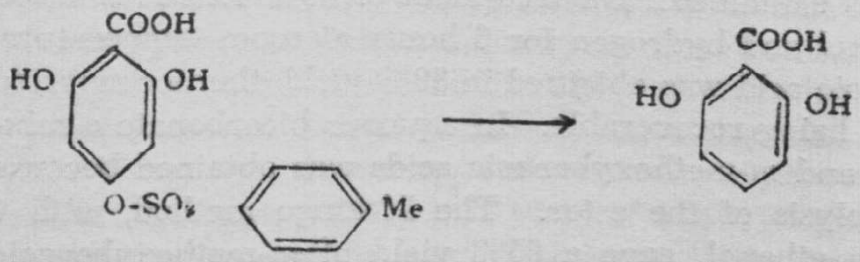
8. *J. Amer. chem. Soc.*, 1955, 77: 4052.
8a. *ibid.*, 1956, 78: 1635.

bis-2-anthraquinonylmethyl disulphide (II; $AQCH_2$-S-S-CH_2AQ) was studied.[9] Short heating for an hour was generally adequate. One of the objects was to remove sulphur from the dyes without hydrogenating the anthraquinone residues. Using 2·5 parts of Raney alloy, anthraquinone gave 1:2:3:4-tetrahydroanthroquinone, and 2-methylanthraquinone 1:2:3:4:-tetrahydro-6-methylanthraquinone, as the major products; the constitution of the latter was shown by oxidation with permanganate in acetone to 4-methylphthalic acid and adipic acid. For desulphurization it was convenient to use an equal part of Raney alloy since about 35% of 2-methylanthraquinone remained unchanged. Under these conditions 2-mercaptomethylanthraquinone gave 2-methylanthraquinone, 2-hydroxymethylanthraquinone, anthraquinone-2-carboxylic acid, and *bis*-2-anthraquinonylmethyl sulphide (I). The sulphide (I) gave 2-methylanthraquinone and traces of Anthraflavone (AQ-CH=CH-AQ). The disulphide (II) yielded 2-methylanthraquinone, 2-hydroxymethylanthraquinone and anthraquinone-2-carboxylic acid. The isolation of the carbinol and carboxylic acid from the desulphurization products of 2-mercaptomethylanthraquinone and the disulphide (II) is of interest. From Anthraflavone a small quantity of 1:2:3:4-tetrahydro-6-methylanthraquinone, formed as a result of fission at the ethylene bridge, and a product analysing for decahydroanthraflavone were isolated.

Kenner and Murray[10] found that, when an alcoholic solution of the *O*-tosyl (*p*-toluenesulphonyl) derivative of a phenol, was shaken with Raney nickel and hydrogen at atmospheric pressure, the products were the aromatic hydrocarbon and nickel tosylate ($2Ar\text{-}O\text{-}SO_2C_7H_7 + Ni + H_2 \rightarrow 2ArH + Ni(OSO_2C_7H_7)_2$). Thus *m*-acetamidophenol could be converted to acetanilide (90%), and methyl salicylate to methyl benzoate (80%), via the tosylate. The naphthalene nucleus also underwent hydrogenation under these conditions, β-naphthyl tosylate yielding tetralin. Hydrogenolysis of the tosyl derivatives of phenols is useful for several synthetic purposes. It provides for instance a new and convenient route to γ-resorcylic acid,[11] the sodium salt of which has been stated to be more active than sodium salicylate as a therapeutic agent in rheumatic fever. γ-Resorcylic acid (2:6-dihydroxybenzoic acid) can be readily prepared from phloroglucinol in an overall yield of

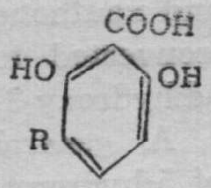
9. Ramanathan, Tilak and Venkataraman, *Proc. Indian Acad. Sci.*, A, 1953, 38: 161.
10. *J. chem. Soc.*, 1949: 178.
11. Ramanathan and Venkataraman, *Curr. Sci.*, 1952, 21: 283.

about 75% by carboxylation, mono-*O*-tosylation to form (III), and treatment of an aqueous solution of the acid (III) and potassium bicarbonate with Raney nickel at room temperature in a current of hydrogen.

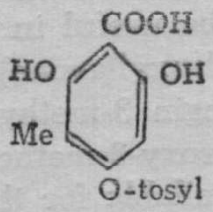


III

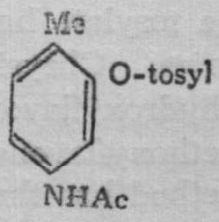
The preparation of 4-methylresorcinol-2-carboxylic acid (IV; R = Me) and 4-*n*-hexylresorcinol-2-carboxylic acid (IV; R = *n*-C_6H_{13}) was next attempted. Unlike (III), the tosylate (V) was not reducible to (IV; R = Me) by Raney nickel and hydrogen; the product was a mixture of the unchanged compound (V) and methylphloroglucinol resulting from hydrolysis of the tosyl group and decarboxylation. Considering the possibility that the steric effect of the methyl group *ortho* to the tosyloxy group may be responsible for the failure of the hydrogenolysis of (V), the reduction of 4-acetamido-*o*-cresol tosylate (VI) was studied. Under Mozingo conditions, the tosyloxy group was split off giving acet-*p*-toluidide. Hexylphloroglucinol was carboxylated by aqueous potassium carbonate, but the monotosylate could not be prepared, since the acid was unstable and decarboxylated readily.



IV; R = Me or n-C_6H_{13}



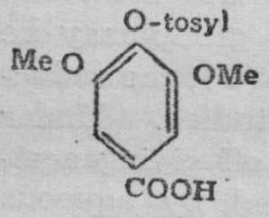
V



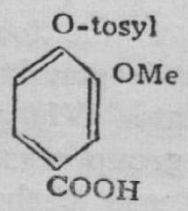
VI

A new route for the preparation of 3:5-dimethoxybenzoic acid, which is normally prepared from benzoic acid by sulphonation, alkali fusion and methylation of the α-resorcylic acid thus obtained, was explored. The tosyl derivative (VII) of syringic acid was prepared in the usual manner in aqueous solution by means of caustic potash and tosyl chloride. This compound on treatment with Raney nickel catalyst and hydrogen did not undergo reduction; hydrolysis took place when aqueous bicarbonate was used as the solvent, giving syringic acid, and in alcoholic solution the unreacted compound was recovered. The tosyl derivative (VII) also

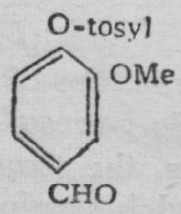
remained substantially unaffected under the Mozingo conditions of reduction, a small part being hydrolysed to syringic acid. With a view to study the effect of one *o*-methoxyl group instead of two, the action of Raney nickel on the tosyl derivative (VIII) of vanillic acid was examined. On treatment of an alcoholic solution with Raney nickel and hydrogen for 6 hours at room temperature, *m*-methoxybenzoic acid was obtained in 30% yield, the unconverted tosyl derivative being recoverable. In aqueous bicarbonate a mixture of vanillic and *m*-methoxybenzoic acids was obtained because of partial hydrolysis of the ester. The Mozingo method, with 4 hours boiling in ethanol, gave a 65% yield of *m*-methoxybenzoic acid, together with 15% of unreacted material. The recovery of some quantity of unreacted material in these reactions shows the slight effect of one *o*-methoxyl group. The reduction of vanillin tosylate (IX) by the Mozingo method led to *m*-cresol methyl ether (60%) and a trace of a sulphur-containing oil. Under the Kenner-Murray conditions vanillin tosylate (IX) gave *m*-methoxybenzyl alcohol (X) in 80% yield.



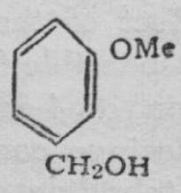
VII



VIII



IX

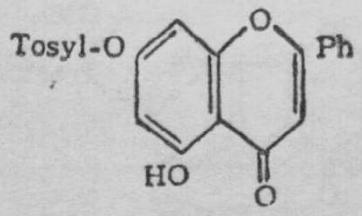


X

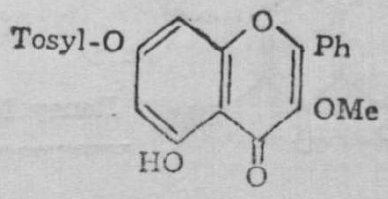
Removal of a phenolic hydroxyl group by hydrogenolysis of the tosylate has proved to be useful in the flavone and xanthone series. Chrysin (5:7-dihydroxyflavone) has been converted to 5-hydroxyflavone, and galangin 3-methyl ether (5:7-dihydroxy-3-methoxyflavone) to 5-hydroxy-3-methoxyflavone.[12] A general method has thus become available for the synthesis of 5-hydroxyflavones, which are of interest because of their possible occurrence in nature and also as intermediates for the synthesis of naturally occurring 5:8-dihydroxyflavones by the Elbs-Seshadri persulphate oxidation. Chrysin and galangin 3-methyl ether readily gave the 7-tosyl ester (XI) and (XII) on treatment with a molar proportion of tosyl chloride and excess of anhydrous potassium carbonate in boiling acetone. It was necessary to carry out the hydrogenolysis of the tosyl derivatives under controlled and mild conditions, leaving a part of the tosyl derivative unreacted to poison the nickel

12. Ramanathan and Venkataraman, *Proc. Indian Acad. Sci.*, A, 1953, 18: 40.

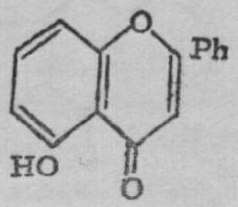
catalyst in order to avoid undesirable side reactions. The crude reaction mixture was treated with alcoholic caustic potash to hydrolyse the unreacted tosyl derivative and the mixture of hydroxyflavones separated by chromatography. Thus by the hydrogenolysis of the 7-tosyl ester (XI) of chrysin, a mixture of 5-hydroxyflavone (XIII) and chrysin was obtained; when the mixed flavones were dissolved in benzene and chromatographed on Florex xxx (an extruded fuller's earth supplied by the Floridin Co., Inc.), development and elution with benzene yielded 5-hydroxyflavone (XIII) in a yield of about 25 per cent calculated on the quantity of the tosyl derivative (XI) consumed in the reaction. Elution with ethyl acetate then led to the recovery of the more strongly adsorbed chrysin. In the case of the 7-tosyl ester (XII) of galangin 3-methyl ether Raney nickel treatment and hydrolysis of the product yielded galangin 3-methyl ether and 5-hydroxy-3-methoxyflavone (XIV), readily separable by chromatography. It was found, however, that galangin 3-methyl ether and 5-hydroxy-3-methoxyflavone can be separated in a simpler manner by taking advantage of the sparing solubility of the former in benzene.



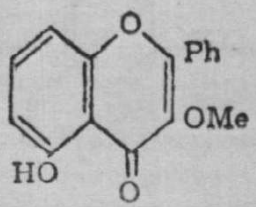
XI



XII



XIII



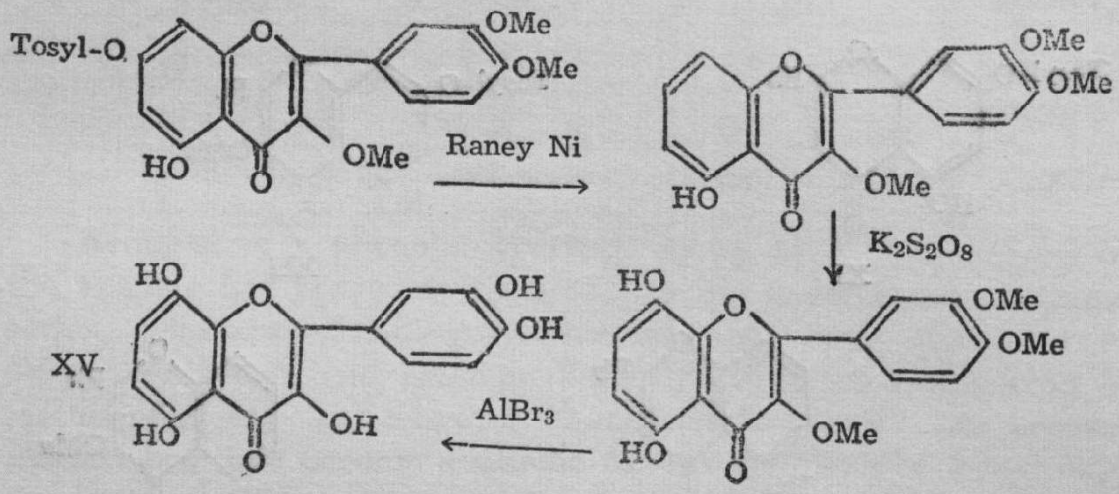
XIV

After the present work was published, Seshadri and co-workers have reported experiments on similar reductions.[13] For the purification of the mixture of flavones obtained after alkaline hydrolysis of the hydrogenolysis product, they first precipitated the coloured impurities from an ethyl acetate solution of the mixed flavones by the addition of petroleum ether. Then the two flavones were

13. Jain and Seshadri, *J. sci. industr. Res. India*, 1953, 12B: 503.

separated by the difference in solubility in aqueous sodium carbonate in which chrysin was soluble, while 5-hydroxyflavone was not. Though this method appears to be simpler, it has been found in the present work that purification through chromatography on Florex is more satisfactory and affords a quantitative separation of the flavones. Florex xxx has been found to be a very useful adsorbent for the chromatographic separation of hydroxyflavones and other coloured phenolic compounds such as the hydroxyanthraquinones.

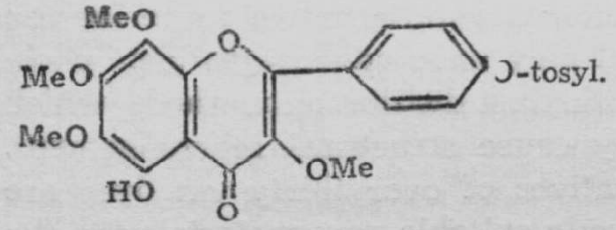
Kurth and Hubbard[14] isolated from *Ponderosa* pine bark a bright yellow colouring matter, to which they assigned the structure 3:5:8:3′:4′-pentahydroxyflavone (XV). The flavone (XV) was synthesized by the indicated route, and found to be different in its m.p. and other properties from the Ponderosa colouring matter.[15] Subsequent work has shown that the latter is a complex mixture, from which quercetin, taxifolin (2:3-dihydroquercetin), and two new colouring matters, 6-methylquercetin (pinoquercetin) and 6-methylmyricetin (pinomyricetin) have so far been isolated.[16]



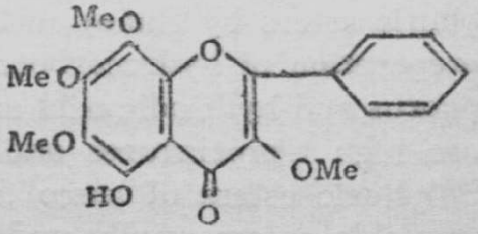
With the object of reducing hydroxyl groups in the 2-phenyl nucleus, the 4′-tosyl ester (XVI) of calycopterin was prepared; Raney nickel treatment and hydrolysis of the product gave calycopterin and 5-hydroxy-3:6:7:8-tetramethoxyflavone (XVII).

14. *Industr. Engng. Chem.*, 1951, 43: 896.
15. Ramanathan and Venkataraman, *Proc. Indian Acad. Sci.*, A, 1954, 39: 90
16. Kurth, Ramanathan and Venkataraman, *Curr. Sci.*, 1955, 24: 157; *J. sci. industr. Res. India*, 1956, 15B: 139. Contrast Gupta, Kurth and Seshadri, *ibid.*, 1954, 13B: 886

Extending this procedure to the xanthone series 1:3-dihydroxyxanthone has been converted into 1-hydroxyxanthone. The tosyl ester (XVIII) on hydrogenolysis and hydrolysis gave a product which on direct crystallization gave 1-hydroxyxanthone (XIX) in 80% yield. From these experiments it is clear that the

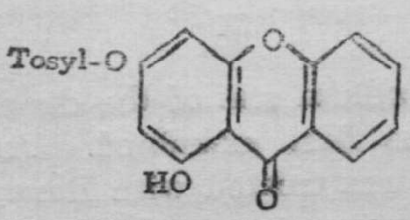


XVI

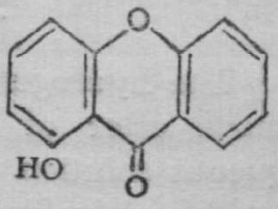


XVII

stability of the pyrone ring to reduction plays a vital part in the yield and purity of the desoxy product. In the case of chrysin, a flavone, the yield of the product is about 25% and it increases through 45% with galangin 3-methyl ether, a flavonol, to 80% in the case of 1:3-dihydroxyxanthone in accordance with the increasing order of stability of the systems.



XVIII



XIX

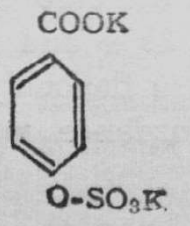
After the present work was completed, Seshadri and co-workers reported the reduction of 1-hydroxy-3-tosyloxyxanthone to 1-hydroxyxanthone, 1-tosyloxyxanthone to xanthone, and 3-tosylgentisein to euxanthone.[17]

The action of Raney nickel on the sulphuric esters of phenols has been studied to determine if sulphuric esters can be used as intermediates in place of tosyl esters for the removal of phenolic hydroxyl groups. *p*-Hydroxybenzoic acid was sulphated using pyridine and chlorosulphonic acid. The sulphuric ester was not isolated, but converted into the dipotassium salt (XX), which was

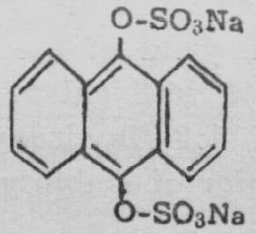
17. Jain, Mittal and Seshadri, *J. sci. industr. Res. India*, 1953, 12B: 647.

treated with Raney nickel and hydrogen. From the reaction mixture benzoic acid (65%) and a little *p*-hydroxybenzoic acid were isolated.

A new general method for the conversion of quinones to the corresponding hydrocarbons through the disulphuric esters of the leuco derivatives thus becomes available.[18] The cleavage of sulphuric esters by Raney nickel affords an alternative route for the preparation of hydrocarbons and may have advantages over phosphorus and hydriodic acid and zinc dust distillation methods, which use high temperatures and may cause structural rearrangement. Sulphuric esters of leuco derivatives of over thirty vat dyes are available commercially and provide suitable raw materials for the preparation of various polycyclic hydrocarbons, azahydrocarbons, etc., which are of interest in connection with cancer studies.



XX

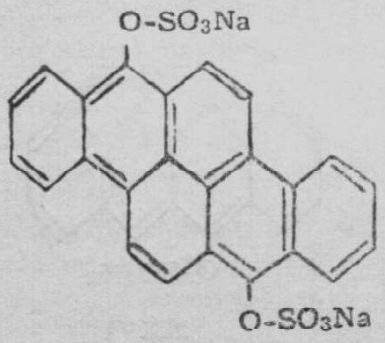


XXI

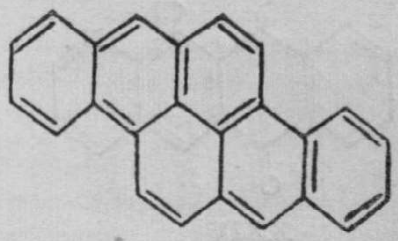
The hydrogenolysis of (XXI), the disodium salt of the disulphuric ester of anthrahydroquinone, was first examined. An aqueous solution of the disulphuric ester was treated with Raney nickel and hydrogen for 4 hr. at room temperature and filtered. Benzene extraction of the nickel residue after deactivation gave a mixture of 1:2:3:4-tetrahydroanthraquinone and hydrogented anthracene; the latter was then dehydrogenated by selenium to anthracene in an overall yield of 30%. From the aqueous filtrate, after oxidation with sodium nitrite and sulphuric acid, a further quantity of 1:2:3:4-tetrahydroanthraquinone was recovered, indicating that hydrogenation of anthrahydroquinone disulphuric ester to 1:2:3:4-tetrahydroanthrahydroquinone disulphuric ester also takes place as a side reaction. The formation of a quinonoid derivative, because of the incomplete hydrogenolysis of the ester even when a large amount of Raney nickel was used, was practically eliminated by carrying out the reaction using Raney alloy

18. Desai, Ramanathan and Venkataraman, *J. sci. industr. Res. India*, 1955, 14B: 330.

and aqueous alkali instead of Raney nickel catalyst. This modification gave an excellent yield of hydrogenated anthracene (85%) which could be readily dehydrogenated to anthracene.



XXII



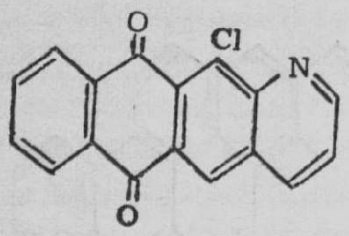
XXIII

Reduction of Indigosol Golden Yellow IGK (XXII), the disodium salt of the disulphuric ester of Indanthrene Golden Yellow GK (2:3:7:8-dibenzpyrene-1:6-quinone), by means of Raney alloy and aqueous alkali afforded an excellent yield (80%) of a mixture of hydrogenated hydrocarbons, which were separated by chromatography into hexahydrodibenzpyrene (m.p. 240-41°) and decahydrodibenzpyrene (m.p. 235-36°). The mixture of hydrogenated hydrocarbons was dehydrogenated to dibenzpyrene (XXIII) by treatment with iodine and sodium acetate in boiling nitrobenzene. 2:3:7:8-Dibenzpyrene (XXIII) has been prepared earlier by other workers from 2:3:7:8-dibenzpyrene-1:6-quinone by zinc dust distillation, Clar reduction, and reduction with phosphorus and hydriodic acid.

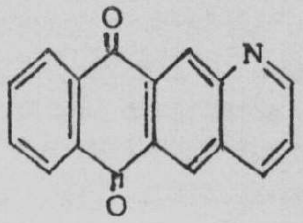
A further application of the Raney nickel method for the reduction of quinones to the corresponding hydrocarbon derivatives is the synthesis of the hitherto unknown 1-azanaphthacene (XXVIII).[19] The Skraup reaction on 2-amino-1-chloroanthraquinone, adding arsenic pentoxide to suppress the benzanthrone reaction, gave a good yield of 12-chloro-1-azanaphthacene-6:11-dione (XXIV). Dechlorination by means of caustic soda and sodium hydrosulphite and air oxidation of the vat yielded 1-azanaphthacene-6:11-dione (XXV). Treatment of the disodium salt (XXVI) of the disulphuric ester of (XXV) with Raney alloy and caustic soda at 100° gave an octahydroazanaphthacene, the absorption spectrum of which showed that it had the structure (XXVII).

19. Desai, Ramanathan and Venkataraman, *ibid.*, 1956, 15B: 279.

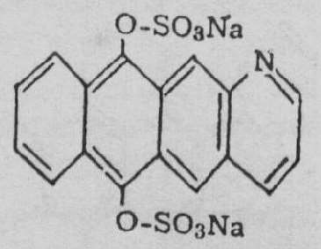
After attempting other procedures which effected only partial dehydrogenation, (XXVII) was ultimately dehydrogenated to 1-azanaphthacene (XXVIII) by prolonged heating with palladized carbon in boiling Dowtherm.



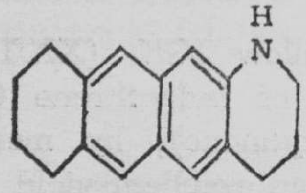
XXIV



XXV



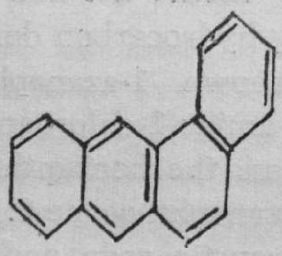
XXVI



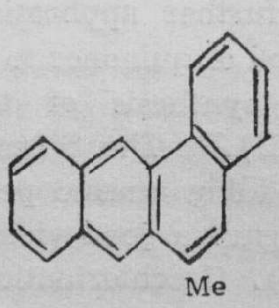
XXVII

XXVIII

The hydrogenolysis of leuco sulphuric esters of quinones by Raney alloy and aqueous alkali to the corresponding hydrocarbon derivatives has been extended to the leuco sulphuric esters of 1:2-benzanthraquinone and its derivatives for the preparation of benzanthracene and methylbenzanthracenes, which are of interest in connection with their carcinogenic activity. Thus the hydrogenolysis of the sulphuric ester of benzanthrahydroquinone gave a



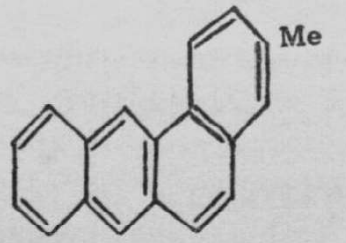
XXIX



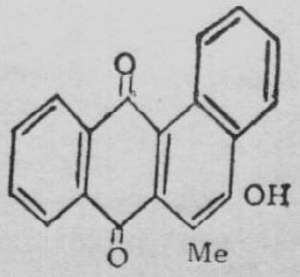
XXX

hydrogenated benzanthracene, which was dehydrogenated to 1:2-benzanthracene (XXIX) in an overall yield of 70%. With the object of preparing 4-methylbenzanthracene (XXX) the hydrogenolysis of the leuco sulphuric ester of the methylchlorobenzanthraquinone, obtained by the Friedel-Crafts condensation of

1-chloro-2-methylnaphthalene and phthalic anhydride and cyclisation of the resulting keto acid, was carried out; the product, however, was 3′-methylbenzanthracene (XXXI), indicating that the phthalic anhydride condensation had taken place in the unsubstituted benzene ring (the 5-position of 1-chloro-2-methylnaphthalene).



XXXI



XXXII

The synthesis of 4-methylbenzanthracene (XXX) through the trisulphuric ester of 3-hydroxy-4-methylbenzanthrahydroquinone is being studied. The desired anthraquinone derivative (XXXII) was obtained by condensing 3-hydroxybenzanthraquinone with aqueous caustic soda, sodium hydrosulphite and formaldehyde (the Marschalk reaction). As a pilot experiment, the trisulphuric ester of the leuco derivative of 2-hydroxyanthraquinone was reduced by Raney alloy and caustic soda to a hydrogenated anthracene, which was then dehydrogenated to anthracene.

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On the Mineral Facies of Charnockites

BY

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Introduction

The charnockites, thought to be rocks specific for Peninsular India by their first investigator, Sir Thomas Holland (1900), have later on been found to be universally distributed and have attracted general interest among petrologists all over the world. Holland had concluded that the charnockites are magmatic rocks. Later Vredenburg (1918), Groves (1935), Ghosh (1941), J. A. W. Bugge (1943), Ramberg (1948, 1949, 1952, 1956) *et al.* presented theories of their genesis, different one from the other, but agreeing therein that the charnockites are metamorphic in one way or other and not primary magmatic rocks.

Since the Mineral Facies classification (Eskola, 1921) was expanded to include the Granulite Facies, to which all charnockites belong (Eskola, 1939), the charnockite problem took partly a new aspect, as it could be understood that the charnockites like other rocks formed at rather high temperatures may be either magmatic, metamorphic, or metasomatic. More important, in many respects, for the nature of a rock than its mode of crystallization, is the temperature and pressure under which it has crystallized. Therefore, the question whether the charnockites are magmatic or metamorphic is, to a certain degree, irrelevant as compared with the character of their mineral facies.

This article is written in order to make the mineral facies principle familiar to the students of charnockites. This might appear unnecessary, as, after 1939, excellent expositions of the mineral facies classification have been published by Turner (1949), Turner and Verhoogen (1951), Barth (1952), Ramberg (1949, 1952), Rosenqvist (1952) and others. However, it seems to me that discussion on the basis of this principle still can afford elucidation to the charnockite problem. In this short paper I cannot go deep into the genesis of charnockites but must refer to the works cited and especially to Quensel (1951) who gives a fair review of recent literature. I hope later, in another connection, to discuss the genetic problems in the light of the mineral facies principle.

The mineral facies conception was based upon the physico-chemical thesis that, in a rock which has reached chemical equilibrium at constant temperature and pressure, the assemblage of phases, *i.e.*, mineral constituents, will depend solely upon the chemical bulk composition. A certain mineral facies, then, is a group of rocks which, at a given bulk composition, have the same assemblage of minerals but, at varying bulk composition, show mineral compositions varying according to definite rules.

The chief mineral facies classes of rocks separated in the Sial Crust, supposed to represent different temperature and pressure ranges are as follows.

Decreasing temperature ⟶

Increasing pressure ⟶

Metamorphic facies :	Sanidinite	Zeoliths
Magmatic „	Diabase	
Metamorphic „	Hornfels, Amphibolite, Epidote amphibolite, Greenschist	
Magmatic „	Gabbro, Hornblende gabbro	
Metamorphic „	Granulite	
Magmatic „	Mangerite	
Metamorphic „	Eclogite	
Magmatic „	Griquaite (see Eskola 1921, p. 185)	

Recent experimental research has proved that the phase equilibria of rocks are far less sensitive for pressure than for temperature. Nevertheless, everybody will agree that, in the vertical column of the table, the sequence of rocks is that of in-

creasing depth and pressure. The influence of pressure is partly indirect but, as will be shown later, its direct effect probably is not quite negligible either.

The mineral facies, excepting those of the lowest temperature ranges, occur in magmatic as well as in metamorphic crystallizations. This is also true of the granulite facies. V. M. Goldschmidt (1922) distinguished in the Norwegian Caledonides magmatic charnockites, called mangerites. Recently Rosenqvist (1956) pointed out that this is an appropriate name for the magmatic granulite facies. At present, after the research of Bowen and Tuttle has proved that many granites and charnockites have almost the minimum-melting composition of the haplogranitic system, the existence of magmatic charnockites cannot be doubted.

In a remarkable paper Yoder (1952) published his first results of experimental research on the thermal properties of metamorphic minerals, some of which are important constituents of the charnockites. It is by such investigations that the value of the mineral facies theory can be tested. The theorist has to adjust his conclusions after them. These results, as a matter of course, will be briefly discussed in this paper.

The methods of investigation at high temperatures and pressures have advanced far enough to enable the establishment of phase equilibria under the temperature and pressure conditions prevailing in all parts of the lithosphere where the existing magmatic and metamorphic rocks have originated (Roy and Tuttle, 1956). Thus, it is probable that in the near future the stability conditions of all rock-making mineral systems can be made clear. Then it will be possible to lay an exact basis of the mineral facies classification instead of the empiric methods practicable up to date.

Characteristics of the Granulite Facies

The granulite facies comprises the crystalline rocks which, at a normal granitic composition, are composed of quartz, potash feldspar, sodic plagioclase, and almandine-pyrope garnet. The last mentioned mineral can be formed only when the rock contains alumina in excess over the amount that can be contained in the feldspars, thus substituting the mica of the granite. Almandine-pyrope equals hypersthene + corundum :

$$3\,(Mg,\,Fe)\,SiO_3 + Al_2O_3 = (Mg,\,Fe)_3\,Al_2Si_3O_{12}$$

In acid granulites the alumina in excess over the garnet ratio forms sillimanite. In many occurrences, however, the excessive alumina first gives rise to cordierite :

$$2\,(Mg,\ Fe)_3Al_2Si_3O_{12} + 4Al_2O_3 + 9SiO_2 = 3\,(Mg,\ Fe)_2Al_4Si_5O_{18}.$$

Cordierite as well as sillimanite in its respective areas occurs in the manner of a stable constituent, so as to suggest distinction of two subfacies. They show, also, some textural and tectonic differences, the sillimanite granulite being foliated and its quartz of the platy quartz type, the cordierite granulite more coarse-grained and massive. Both may represent typical compositions of argillaceous sediments.

When the amount of ferro-magnesium silicate exceeds the garnet ratio, the granulites contain hypersthene. If, at the same time, the rock contains calcium in excess over the ratio required by anorthite with the aluminium present, no garnet is formed.

More extreme types of charnockites are the khondalites of India and Ceylon. They are quartz-sillimanite-garnet rocks of argillaceous composition grading over into quartzites or, at the opposite end, into corundum-bearing rocks equivalent to bauxite. Peculiar to many Indian charnockite provinces are magnetite-quartz rocks.

In lime-rich rocks calcite, dolomite, and meionitic scapolite occur in association with diopside or forsterite. Examples were described from the Aldan Plateau in Siberia by Korzhinsky (1936) who also discussed their thermodynamical significance stressing the idea that they have formed under extremely high pressures. From western Uusimaa, Finland, scapolite-bearing sedimento-geneous charnockitic rocks have been described by Parras (1941). Numerous and most remarkable examples of calcium-rich rocks metamorphosed in the granulite facies are known from India (Ghosh, 1941; Rama Rao, 1945, p. 149; Muthuswami, 1951) and Ceylon (Adams, 1929). Paulose (1956) describes diopside-scapolite granulites and diopside-wollastonite rock from Cape Comorin. Subramaniam (1956) has published a masterful treatise of the Sittampundi complex, Salem district, Madras, where the main rocks are anorthosites, partly with exceptionally pure An, accompanied by eclogites and chromitites. The complex is interpreted as a layered intrusion, while a bytownite-cordierite-corundum rock in it is said to represent a metasediment. The complex, situated

in one of Fermor's charnockite provinces, has suffered a retrograde metamorphism in the amphibolite and epidote-amphibolite facies connected with the emplacement of younger granite, but its preserved relict constituents clearly indicate its original belonging to the granulite and eclogite facies.

It is to be remembered, moreover, that in the charnockite provinces the charnockites frequently alternate with layers of crystalline limestone and quartzite which thus have been metamorphosed in the granulite facies although this does not show in their minerals, as calcite, dolomite, and quartz are "Durchläufer" (Fr. Becke), stable in most facies.

The minerals of the granulite facies have several persistent peculiarities which distinguish this facies from all others. Some of these give important indications of the conditions that were prevailing during the crystallization of the rocks. An almost invariable characteristic of the charnockites is a dirty brownish or bluish hue of their feldspars, giving the rock a darker colour than it would have otherwise on account of its small amount of mafic minerals. This colour is due to pigment-like inclusions, probably of iron and aluminium oxides which have been originally present in the mineral lattices after crystallization at high temperatures, possibly as an iron feldspar (Rosenqvist, 1951). The quartz also is often dark bluish and is seen under microscope to contain slender prisms of rutile. Pichamuthu (1953) stresses the occurrence of "clouded feldspars" in the Indian charnockites.

Another general characteristic is a wider range of diadochy in the series of isomorphic minerals than in other facies excepting the eclogite facies. Thus, the garnets sometimes contain as much as 55 per cent pyrope and up to 20 per cent grossularite, or appreciably greater proportions than the garnets of ordinary amphibolites or granites of the amphibolite facies (Eskola, 1952, p. 150, 153).

The hypersthenes of charnockites are often highly aluminous. In the charnockites of Uganda, Groves (1935) found hypersthene containing as much as 9·39 per cent Al_2O_3. In the granulites of Lapland (Eskola, 1952) the maximum percentage of Al_2O_3 in hypersthene found was 8·26. About a half of the aluminium here replaces silicon in the SiO_4-tetrahedra. Thus, we see here an extended diadochy of cations just as in the granulite garnets. From a charnockite at Bidaloti, India, Rama Rao (1945, pp. 19-21)

describes a lilac-coloured mineral called Bidalotite, which otherwise resembles hypersthene in characters and composition, but contains 10·55 per cent Al_2O_3.[1]

The wider diadochy in the orthopyroxene and garnet might be caused by high temperature of crystallization, as is known to be the case in many pairs of isotypic compounds. Thus, solid solutions of diopside and (Mg,Fe)-clinopyroxene, or pigeonite, exist in rocks of the diabase facies. But this kind of widened diadochy does not occur at all in the granulite facies. Naidu (1955), in his monograph of the charnockite minerals, states that pigeonite is present in the non-metamorphic dolerites intersecting the charnockites but not in the latter, and Leelananda Rao (1954) describes pigeonite dolerite dikes at Pallavaram. Therefore, we cannot assume that the extended diadochy in the granulite facies minerals would be effected by higher temperature, especially as the latter also differ from the high-temperature magmatic minerals in their colour and pleochroism. Therefore, the conclusion appears to be well founded that this feature in the charnockites is due to higher pressure (cf. Rosenqvist, 1952).

In many charnockites the hypersthene is strongly pleochroic: α purple violet, β pale reddish yellow, γ pale green. The intensity of pleochroism varies greatly being apparently independent of the Mg/Fe ratio, the rate of oxidation of the iron, and the amount of total iron. The problem of the colours of hypersthene in the charnockites is still unsolved. Possibly the intensity of the colours depends upon the degree of unmixing of iron oxides just as the darkness of feldspar and quartz, though in the hypersthene the pigment remains submicroscopic (cf. Quensel, 1951, p. 10).

The potash feldspar in the Indian charnockites, according to Holland, is invariably microcline, but later on orthoclase has been also reported. In the granulites of Finnish Lapland the potash feldspar according to Eskola (1952) appears to be primarily orthoclase. In many thin sections of coarse-grained cordierite-granulites and garnet-granulites of granite-like appearance the potash

1. Dr. P. R. J. Naidu of the Madras University kindly called my attention to a fact which, unfortunately, I had forgotten, viz., that Dr. J. C. Rabbitt and Professor C. E. Tilley have identified this mineral as anthophyllite (Rabbitt, Am. Min. 33, p. 310). The strong colour and high titanium content of this anthophyllite (or gedrite) seems to place it in a type comparable with the hornblende or hypersthene of charnockites in contrast to the same minerals in rocks of the amphibolite facies.

feldspar shows straight extinction on (001), is clear and free from perthite. Similar orthoclases were also encountered in foliated, more fine-grained granulites with platy quartz. In some specimens the clear orthoclase shows undulating extinction from which gradual passage into typical cross-hatched microcline may develop. Independent of the change to microcline is the appearance of minute sparse perthite spindles. When the perthite contains nearly equal amounts of Or and Ab, as the so-called mesoperthite in Greenlandic charnockites (Ramberg, 1956), it is surely a high-temperature product.

Goldsmith and Laves (1954) have investigated the potash feldspar of the Lappish granulites and stated by means of X-ray studies gradual change of orthoclase into microcline characterized by an increase of triclinity. In some cases the whole range from monoclinic to 100 per cent triclinic variety is represented in a single thin section.

The potash feldspar of the charnockites is mostly developed as extremely fine exsolution perthite called hair-perthite or string-perthite with hair-like threads of plagioclase. Rosenqvist (1952, pp. 81-87) has carried out experiments to determine the diffusion ellipsoid for radium ions in feldspars. He found its longest axis, or the direction of the greatest velocity of diffusion, to be $\perp$ (801) and the anisotropy to increase with temperature. Assuming that the Na-ions behave like the Ra-ions he found herein the cause why the albite in the perthites has formed hairs or strings or veins parallel to (801). The hair perthite, therefore, indicates the highest temperatures whereas the vein and patch perthites, found, *e.g.*, in granite pegmatites, have formed at lower temperatures.

Not infrequently hornblende occurs in association with the typomorphic granulite facies minerals in all charnockite provinces of the world. It is a special variety, greenish brown in colour, rich in titanium and abnormally deficient in hydroxyl, as found by Groves in the Uganda charnockites and as apparently is the case in other charnockites as well. The mode of occurrence of this brownish hornblende in association with hypersthene and garnet is perfectly that of a typomorphic mineral unlike the common green hornblende found as a product of alteration of pyroxene and garnet and representing change to some lower temperature facies. Biotite also occurs in two different manners, as an apparently typomorphic constituent and as a product of alteration. The typomorphic biotite, like the brownish hornblende, also seems to be low

in hydroxyl. The stable-looking biotite is common in the rather coarse-grained granite-like charnockites and in the cordierite granulites, but absent in the foliated granulites, especially those containing sillimanite.

Muthuswami (1953) in his meritorious investigation of the mineral facies of hornblende- and biotite-bearing charnockites from Pallavaram near Madras places these to the almandine-diopside-hornblende subfacies of the amphibolite facies. This may be quite correct, but similar rocks from Lapland, Saxony, and Lower Austria known to me and containing that typical brownish hornblende, in my opinion are better grouped as a subfacies of the granulite facies. Al-rich varieties of the same areas usually contain cordierite, so these may possibly belong to the same subfacies.

Epidote is, on the whole, rather an uncommon product of alteration in the charnockites, and so is also sphene, which is typomorphic in the amphibolite facies. The typomorphic titanium minerals in the granulite facies are rutile and ilmenite, the former occurring in the acid charnockites and the latter in the basic charnockites.

Yoder's investigation

Yoder (1952) discusses the metamorphism of the mineral assemblages in the system MgO- Al_2O_3- SiO_2- H_2O on the basis of experimental investigations. He concludes (p. 615) that "at approximately 600° temperature and 1500 psi pressure of water vapour, it is possible to have assemblages suggestive of every one of the now accepted metamorphic facies in stable equilibrium." Yoder continues: "The differences between these facies are the result of variation in bulk composition and need not represent variations of pressure and temperature. This conclusion, based on experimental fact, appears to be at variance with field observations.....The sequence of progressive metamorphism established in one area, presumably in homogeneous layers, cannot be directly applied to another area as a temperature scale unless the bulk compositions have been established as identical. For example, the temperature of appearance of enstatite in a bed deficient in water will be different from that in a bed containing excess water. It is necessary, therefore, to construct a new facies classification based on a fixed composition, or on several independent compositions, particularly sensitive to temperature and pressure. This project may have to await further experimental work."

"In no way should these remarks be constructed as detracting from the value of the concept behind the Facies Classification. Eventually sufficient data will be accumulated from which a critical group of assemblages can be selected as an indicative of grades of metamorphism. Until these data have been collected, one should be wary of assigning temperatures and pressures of formation for metamorphic rocks of different bulk compositions which are not continuous in the same pressure and temperature gradients."

Yoder's results were surprising to those who like me had been accustomed to regard the minerals of the different facies to be stable at different temperature ranges. The result, in fact, as he says himself, appeared to be at variance with the field observations.

As a criterion of chemical equilibrium in mineral assemblages I had considered the circumstances that it can be reached both from below and from above. From below at rising temperature, for instance in contact metamorphism, and from above at falling temperature, as in magmatic rocks after consolidation. All crystalline rocks have after their final magmatic or metamorphic crystallization passed through levels representing successive decreasing temperatures. It is to be expected, therefore, that any crystalline rock should, besides its stable main constituents, contain products of alteration formed at successively lower temperatures. This, in fact, is generally the case.

Let us take, as an example, olivine dolerite, a hypabyssal rock, crystallized from a water-poor magma. Its chief minerals, as a rule, constitute an equilibrium assemblage, but in most specimens one will also find a series of products of alteration whose sequence of crystallization can be established with certainty. Yoder found experimentally (1952, p. 608) that the metamorphism of argillaceous sediments in the presence of excess water would give at increasing temperature successively kaolinite, serpentine, forsterite, enstatite (and sillimanite), cordierite, tridymite. Now, this sequence is just the opposite to that observed in the products of alteration, as is to be expected from an equilibrium series at decreasing temperature. Thus there is, in Yoder's own results, clear evidence of the progressive metamorphism in the series of mineral facies from the lowest to the highest.

But, on the other hand, Yoder's experimental result implies that the same sequence of mineral facies could also be reached at a constant temperature by change of bulk composition, in this

case the amount of water present. In other words, the products of alteration in the olivine dolerite would be the same, if the temperature was constant and the water concentration increased due to decrease of residual liquid at progressive crystallization. Everybody will agree that in this case fall of temperature has taken place at the same time. To a certain degree this must generally be the case in normal metamorphism.

Charnockite, in contrast to dolerite and most other rocks, mostly contains very small amounts of alteration products, although the rock certainly has cooled. An original deficiency in water must have been an important, sometimes perhaps the only cause to development of the granulite facies.

Occasionally charnockites and granulites, however, may be quite strongly altered to lower facies containing green hornblende, epidote, *etc.* In Lapland, Eskola (1952, p. 166) found in places a peculiar mode of occurrence of amphibolite alternating with granulite, described as follows: "Such rocks in many cases show no relict features from the granulite facies. Their granoblastic texture is very similar to that of the granulites, but the minerals are strikingly different: common green hornblende, sphene instead of rutile or ilmenite, and flame perthite instead of hair perthite."

It is hardly possible to assume different temperatures in such alternating layers, but after Yoder's work it is clear that the amphibolite layers have been more permeable and soaked with water vapor. Tectonized zones are frequently amphibolitized in charnockites. A good example is furnished by the so-called trapschotten granulites in India (Pichamuthu, 1953).

When writing his last exposition of the mineral facies the present author was well conscious of the difficulties resulting from the variation of bulk composition (Eskola, 1939, pp. 343-344). The belonging of a rock to a certain facies is ascertained from the so-called critical minerals or assemblages, stable in one facies only. It is particularly the amount of water and carbon dioxide that causes complications, and this is especially true of the granulite facies. The mineral facies classification, as Yoder says, must be reformed on an exact basis to meet these complications. It seems to me, however, that the experience gained in the application of the facies principle by various workers in many countries has proved its usefulness even in its present, only qualitative state, and that the cases where it has lead to erroneous results are neither numerous nor very serious.

Role of water-deficiency in the genesis of charnockites

One of Yoder's conclusions is that pyrope is stable in water-deficient systems only.[2] He continues the discussion in another paper (Yoder, 1955) objecting to the opinion of Ramberg (1952) that water is universally present and "always in adequate supply". Yoder concludes that often the genesis of hydroxyl-bearing minerals will be prohibited due to lack of water. Argillaceous sediments have "more than enough water to transfer the rock into the most hydrous assemblage of metamorphic minerals", whereas a sediment that "initially consisted of an assemblage of minerals that are completely dehydrated on compaction, as may be the case in some arcoses" have not, "the water-content would then depend on the porosity and permeability of the sediment. In the extreme case of a nonporous, impermeable sediment, only anhydrous minerals could be produced by metamorphism. Granulites, for example, may have formed in the absence of water at low temperatures, or at high temperatures in the presence of water above the stability range of hydrous minerals ..... There are objections, therefore, to the concept of the ubiquitous occurrence and adequate supply of water as a free phase in metamorphism" (Yoder, 1955, p. 518).

To this may be remarked that in many cases the initial rock has been just of the kind that Yoder considers to have been water-rich originally. Thus, the Khondalites have been argillaceous sediments. The same is true of the cordierite granulites of Finnish Lapland which, moreover, are rich in graphite, a probable evidence of their origin as sapropelic sediments (Eskola, 1952, p. 158). Similarly, charnockite pegmatites in India (*e.g.* Rama Rao, 1945), Greenland (Ramberg, 1948, 1956) and Lapland (Eskola, 1952, p. 169), being of an origin like ordinary granite pegmatites, either primary magmatic, as I believe, or metasomatic, as believed by Rama Rao and Ramberg, have probably been hydrated originally, but are now composed of typical anhydrous granulite facies minerals including hypersthene and garnet. The absence of water most probably was a consequence of high temperature which, therefore, appears to be the *prima causa*. In this point I agree with Ramberg.

2. This result suggests, in my mind, the question whether natural pyrope which, in the eclogites and olivine rocks as well as in granulites, invariably contains appreciable amounts of almandine, behaves like pure pyrope in regard to water-deficiency. In Yoder's papers I have found no comment to this question.

From my field experience in the Archaean of Finland I can afford some evidence that in fact water must have been universally present wherever needed in metamorphism. The rocks mostly contain hydrous minerals of the amphibolite or some still lower facies, but many of them have most probably been anhydrous originally, such as basalt or arcose. Yet surrounded by such rocks is the charnockite district of Western Uusimaa where one will find rocks of any primary origin crystallized in the granulite facies with solely anhydrous minerals.

The hydroxyl as well as the oxygen ions tend to migrate upwards in the lithosphere. Szadeczky-Kardoss, in several important papers (1955 a and b), has demonstrated the relations of element migrations to the potentials of ions, atoms and compounds, as well as their specific gravities or "wichtes". For this reason the hydroxyl tends to move upwards until, at the fall of temperature, it is eventually arrested in the lattices of minerals.

Many facts, on the other hand, speak for Yoder's conclusions. Rosenqvist (1952), who like me believes that the granulites and eclogites generally owe their facies to temperatures and pressures higher than those of the amphibolite facies, admits that water-deficiency at low temperature may lead to assemblages of the granulite and eclogite facies and calls this process dipsenic metamorphism (from Greek διψην = thirsty). Such rocks as anorthosites, closely related to the granulite facies (*cf.* Goldschmidt's "Anorthosit-Charnockit-Stamm"), indeed behave as though thirsty rocks, being often thoroughly saussuritized and otherwise hydrated.

Goldschmidt (1920) probably was the first to seek the cause of water-deficiency of magmatic charnockites (mangerites) in the dryness of the magma and the surrounding older rocks, whereas, according to him, the trondhjemites have differentiated from magmas intruded into folded water-rich argillaceous sediments. In the Turku district, S.W. Finland (Hietanen, 1947), zones of trondhjemite and charnockite alternate. As both rocks occur near each other and both are now similarly injected with younger granite and migmatized, they cannot be assumed to have crystallized at different temperatures, and the water deficiency of the charnockites can in this case only be sought in the dryness of their past materials.

After all, we find that in some cases water-deficiency and in others high temperature causes the genesis of the granulite facies. The two causes do not exclude each other, for crystallization in

dry materials, as in dry magmas, takes place at much higher temperatures than in such soaked with water vapor, while, on the other hand, high temperature promotes drying. Scheumann (1954) also finds that the genesis of the eclogite and granulite facies requires strong pressures and high temperatures. He regards the Saxonian granulites as metamorphic rocks, although largely soaked with magmatic solutions. An important moment in their metamorphism is what he calls phyrasis, or dry grinding of the minerals, apparent in the marginal zones of the granulite massif as the typical lamination with platy quartz. Scheumann, like Eskola (1921), sees the best evidence for the great pressure at the genesis of eclogites and granulites in the great density of many of their minerals, like garnet, jadeite, omphacite, kyanite, rutile, diamond. In later years it has been found that in many cases (jadeite, glaucophane schist) the supposed high-pressure minerals cannot have formed under exceptional hydrostatic pressures. It seems, therefore, now rather that low temperature may act in the same direction as high pressure provided stress or mineralizers, acting as catalyzers, can bring forth reactions at low temperatures. It cannot be doubted, however, that high pressure, in accordance with the equation of Clausius-Clapeyron, has been a general promoter at the genesis of the granulite facies.

Rosenqvist (1952) proposes that phase transitions of the feldspar minerals and the characters of perthites should be used to define the exact boundaries between different facies. As recently much progress has been achieved in the experimental studies on feldspars, this proposal seems worthy of consideration. New geological thermometers and barometers, however, have been found, and may still be found, in other mineral groups as well.

Mode of origin of charnockites

India is the true home of the charnockites. It is quite true, as Pichamuthu says, that the charnockite problem cannot be solved without consideration of the Indian occurrences. The writer has not been in that country and the literature dealing with the Indian charnockites is too comprehensive and partly inaccessible here. Fortunately we now possess the exhaustive, critical review by Pichamuthu (1953). This work for me personally is the more congenial, as his general conceptions closely agree with mine. His book is particularly valuable, because he reviews at length, with numerous quotations, the classical memoirs on charnockites by Holland.

Holland studied the charnockites primarily at St. Thomas Mount and at Pallavaram near Madras. The charnockites occur as great masses, one of them, the Nilgiri Plateau, being 700 square miles. Sometimes they are quite massive and have a uniformity of composition throughout. Holland says there is no equivalent among sedimentary formations or the old metamorphic rocks like the Dharwars for such homogeneous masses of rock. In detail they are, however, schlieric and may contain angular fragments of basic rocks as well as darker schlieren with unsharp contacts. Almost every mass also affords examples of vein-shaped schlieren. According to Holland these represent the consolidation of the more acid residual portions of the magma, being comparable to venitic pegmatites. Except the general schlieric banding there are veins of a much more perfect kind, which, in some cases at least, are due to *lit-par-lit* injection.

The importance of the chemical composition as a criterion of origin was well understood by Holland. He was the first to distinguish acid, intermediate, basic, and ultrabasic charnockites. Each of these types has its counterpart among the common magmatic rocks: granites, diorites, gabbros, and peridotites. This fact was regarded as a strong evidence that the charnockites also are magmatic rocks.

Today, after Bowen and Tuttle, in a work unpublished when this is being written (summarized by Abelson, 1955), have definitely established the crystallization of granite as the end phase of crystallization differentiation from the residual magma, we know the composition of the lowest melting "haplogranitic" magma and, on the other hand, that a great majority of the granites closely approach this composition. A comparison with analyses of charnockites shows that great many acid members of this group also agree with these rocks in their cation ratios, while intermediate, basic and ultrabasic charnockites are like diorites and so forth. These facts afford a firm basis for the assumption presented by Holland that the charnockites in their differentiation behave like the calcalkaline suite of magmatic rocks, and that the charnockites can crystallize directly from magmas.

To this conclusion, however, we must make one important reservation: it is not sure that the charnockites have attained their present mineral composition, or their mineral facies, directly by magmatic crystallization and not by later metamorphism with unchanged bulk composition.

One of the earliest petrologists in India to suggest a metamorphic mode of origin for Holland's charnockite series was Vredenburg (1918). He started from the observation that, in some epidioritic Dharwars, the augite was formed by a recrystallization of uralite. Vredenburg considered this as an example of "a schist caught, so to speak, in the very act of transformation into a granulite". The association of charnockites and khondalites recalls the occurrence of metamorphosed igneous and sedimentary rocks of the Dharwars and would have nothing to surprise us, as it would simply represent an alternation of volcanic and sedimentary beds. In a summary, the charnockites according to Vredenburg are metamorphosed magmatic rocks, primarily like basalts and andesites, and hypabyssals. The khondalites and other clearly sedimentogeneous rocks were not regarded by him as charnockites.

A more extreme transformistic view was taken by Ghosh (1941) in his study of the charnockites of Bastar and W. Jeypore, interpreted as impure calcareous sediments originally. Hypersthene was regarded as secondary after diopside and the change of basic rocks to acid charnockites as a result of the influx of alkalies from granitic magma.

Rama Rao (1945) found in the charnockites of Mysore much more variation in texture, composition and mode of origin than Holland had found in neighbouring areas. He assumed transformations of minerals in connection with their origin, effected by reactions from influx of acidic solutions. For instance, in the Biligirirangan Hill range charnockite has been formed from sedimentogeneous sillimanite quartzites by feldspathization of quartz, and transformation of ilmenite to mafic minerals from the injection of alkaline fluids connected with the intrusion of the younger Closepet granites which, in many cases, have been held responsible for charnockitization of various rocks. Basic charnockites of noritic composition were explained as crystallizations from contaminated gabbroic magmas that had assimilated aluminous sediments, an idea first suggested by H. H. Read.

Pichamuthu criticizes Rama Rao's conclusion concerning the role of the Closepet granite: "Where such basic charnockites have come within the influence of the younger granites, the effect has been for the progressive transformation of minerals in accordance with the reaction series: Hypersthene-augite-hornblende-biotite. By the intrusion of this granite, therefore, far from charnockites being formed, the existing members of this series became degraded

or 'decharnockitized'." Indeed, from the mineral facies principle, a granite which itself belongs to a lower facies, cannot cause the appearance of a higher facies. The water-content, normally present in the granite magma, alone prohibits charnockitization.

Holland had described intrusive tongues or apophyses from the charnockite into a biotite gneiss in a quarry near the town of Salem and regarded this as an evidence for the intrusive nature of charnockite. I wish to put on record some of his words in this connection, as they, in my opinion, clearly show that this pioneer in Indian petrology was far ahead his time in objectivity of observation. He has written (quoted by Pichamuthu, 1953, pp. 132-134): "Whilst the tongues of charnockite are easily, at a glance, distinguished from the gneiss they protrude into, close examination of the junctions shows that, instead of there being a sharp line between the two rocks, there is a very rapid, though gradual, transition from one to the other. Besides this, dark patches occur in the charnockite parallel to the dark patches of micaceous mineral which in the biotite-gneiss indicates its foliation. The charnockite veins look, therefore, as if they were pseudomorphs very imperfectly retaining the structures of the rock they have corroded and replaced". After a long discussion he left the problem undecided and reverted to it the following year but could not reach any other conclusion except that "the absence of chilled selvages and the presence in the charnockite of indistinct dark patches parallel to the dark patches in the gneiss show: that the charnockite tongues have not been intruded into fissures in the gneiss after the fashion of a simple igneous injection; but rather that the 'trespass' was more of a nature of corrosion, imperfect pseudomorphs of the old structure being preserved by this corroding charnockite magma".

Pichamuthu (1953, p. 134) expresses Holland's conclusion more in the terminology used at present time as follows: "These facts, especially the last one regarding the presence of gneissic relics in charnockites, are in favour of replacement origin. Charnockitization appears to be the best term for explaining this conversion of biotite gneiss into charnockite".

Pichamuthu's own vision of the genesis of the Indian charnockites seems to be approximately as follows, though I am not quite sure whether I have not therein mixed my own ideas from the study of charnockites and granulites in Finland: in India, ancient Archaean formations, the Dharwars, are mainly meta-

morphosed in the Amphibolite facies, just as the Svecofennides in Fennoscandia. In South India erosion in Archaean times has laid bare formations metamorphosed at deeper levels and in the granulite facies. These are the Charnockite Provinces of Sir Leigh Fermor. This old complex comprises plutonic, vulcanic, and sedimentary rocks, including hypabyssal dyke-rocks, all disguised as charnockites. Later, but still under the granulite facies conditions, younger magmas were intruded in the charnockites giving rise to the transgressive dykes. They may be palingenic dykes, but whether they are juvenile or palingenic is of no concern for the problem of their genesis. There are "charnockites and charnockites", as Pichamuthu has put it (1953, pp. 126-129).

Among the granulites and charnockites (enderbites) of Finnish Lapland, likewise, the primary character of the rocks is most variegated (Eskola, 1952). The enderbites and many of their pegmatites were interpreted as primary magmatic. Many noritic granulites have originally been hypabyssal dyke-rocks, and there is an abundance of primary argillaceous sediments, now cordierite and sillimanite granulites, and quartzitic granulites. Cordierite-quartz rocks or laanilites were possibly products of magnesium metasomatism originally, while many of the highly tectonized acid granulites by their high and variable silica and chemical abnormities suggest some kinds of metamorphic differentiation.

The charnockite province of western Uusimaa in Finland is another area where rocks of the granulite facies are exceedingly variegated in structure, composition and probable origin. Parras (1941) has published a preliminary report of this district. As his exhaustive investigation is expected to be soon available, I do not wish to anticipate his results.

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The Floral Morphology and Embryology of the Loranthoideae (Loranthaceae)

BY

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Danser (1929, 1933a, b) published several monographs on the taxonomy of the Loranthoideae and revised the nomenclature. He classified this subfamily into 51 genera and erected an additional genus, *Thaumasianthus*. Engler & Krause (1935) followed him in many respects but recognized only 25 genera.

Ernst (1942) pointed out that the divisions of the Loranthoideae into tribes and subtribes by Danser (1929, 1933a, b) and Engler & Krause (1935) should be checked on the basis of embryological evidence.

The first reasonably satisfactory account of the development of the reproductive organs of the Loranthoideae was given by Treub (1881). Working on *Loranthus sphaerocarpus* (= *Macrosolen cochinchinensis*), he reported that the tip of the floral axis grows into the ovarian cavity forming a three- to four-lobed mammiliform structure, the "mamelon", which is fused with the ovary at as many points as the number of carpels. Treub regarded the lobes as rudimentary ovules and the central part as the placenta. Concerning the development of the female gametophyte, he stated that the megaspore mother cell divides to form a row of three cells of which the uppermost[3] gave rise to the embryo sac. Treub also observed the formation of a composite endosperm arising by the fusion of the endosperm of several embryo sacs and the extremely long suspensors carrying the proembryo downwards into the ovary. In *Loranthus pentandrus* (= *Dendrophthoe pentandra*), Treub (1883) found the elongation of the sporogenous cells to result in a slight elevation at the base of the ovarian cavity which remained free from the ovary wall. He interpreted this as a "rudimentary placenta". Six to seven embryo sacs developed in the

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3. Treub regarded the basal megaspore (the one away from the epidermis of the mamelon) as the uppermost.

same ovary and grew up into the style. The zygote divided vertically and further divisions were transverse leading to the formation of a long biseriate proembryo.

Subsequent to Treub's classic work, Rauch and Schaeppi & Steindl, working in Ernst's Laboratory, greatly extended our knowledge of the embryology of the Loranthoideae. Rauch (1936) gave a detailed account of *Scurrula atropurpurea* and *Dendrophthoe pentandra,* while Schaeppi & Steindl (1942) published some notes and observations on *Amyema gravis, Barathranthus axanthus, Dendrophthoe pentandra, Elytranthe parasitica, Helixanthera hookeriana, Lepeostegeres gemmiflorus, Macrosolen cochinchinensis, M. formosus, M. pseudoperfoliatus, Scurrula atropurpurea, S. lepidota, S. montana* and *S. parasitica.* The embryo sac was found to conform to the Polygonum type; its lower end extended down to the collenchymatous pad or tube, but the tip attained different heights in different genera sometimes extending close to the stigma.

Several members of the Loranthoideae have been studied at the University of Delhi: *Helixanthera ligustrina* (Maheshwari & Johri, 1950); *Macrosolen cochinchinensis* (Maheshwari & Singh, 1952); *Dendrophthoe falcata* (Singh, 1952); *Scurrula cordifolia, S. ferruginea, S. parasitica* and *S. pulverulenta* (Agrawal, 1954); *Helicanthes elastica* (Johri & Agrawal, 1954); *Amyema congener, A. miquelii, A. pendula* and *A. preissii* (Dixit, 1954); *Lysiana exocarpi, Dendrophthoe neelgherrensis* and *D. falcata* (Narayana, 1954a, b); *Nuytsia floribunda* (Narayana, 1955a); and *Lepeostegeres gemmiflorus, Tolypanthus lagenifer* and *T. involucratus* (Dixit, 1955, 1956a). Investigations on *Elytranthe capitellata, Taxillus cuneatus, T. sclerophyllus* and *Tapinanthus uhehensis* have already been completed and are awaiting publication. Studies on five other genera (*Atkinsonia, Barathranthus, Phyllodesmis, Struthanthus* and *Tapinostemma*) are in progress.

Of the above plants, *Dendrophthoe falcata* was collected from Agra, Bareilly and Lucknow; D. *neelgherrensis* from Kodaikanal (Madras) and Ceylon; *Elytranthe capitellata* from Bombay; *Helicanthes elastica* from Courtallam (Madras), Changanacherry (Kerala) and Mahabaleshwar (Bombay); *Helixanthera ligustrina* from Nepal; *Macrosolen cochinchinensis* from Dacca; *Scurrula cordifolia* from Kakinada (Madras), Courtallam and Hyderabad (Andhra); *S. parasitica* from Dacca; S. *pulverulenta* from Dehradun, Mussoorie, Courtallam and Mahabaleshwar; *Tolypanthus*

involucratus from Gauhati (Assam); *T. lagenifer* from Khandala and Bombay; *Taxillus cuneatus* and *T. sclerophyllus* from Ceylon; and *Barathranthus lagenifer* and *Lepeostegeres gemmiflorus* from Indonesia.

In addition, material of *Amyema congener, A. miquelii, A. pendula, A. preissii, Atkinsonia ligustrina, Lysiana exocarpi* and *Nuytsia floribunda* was obtained from Australia; *Phyllodesmis kaempferi* from Japan; *Scurrula ferruginea* from Singapore; *Struthanthus androstylus, S. marginatus* and *S. stphylinus* from Rio de Janeiro (Brazil), *Tapinanthus uhehensis* from Bakoba (Africa) and *Tapinostemma acaciae* from Sudan (Africa).

Some of the more important results obtained during the last several years are discussed below in the light of earlier interpretations.

Calyculus—The morphological nature of the calyculus has been an object of much controversy. Roxburgh (1874) and Bentham & Hooker (1883) considered it to be a true calyx. Danser (1931) too uses the term calyx for the outer floral envelope, but Eichler (1878), Engler (1889), and Schaeppi & Steindl (1942) regard it as a special organ.

Eichler (1878) considered the calyculus to be an axial structure and this view was accepted by Engler (1888-1889) and Engler & Krause (1935). Schaeppi & Steindl (1942), who agree with the above interpretation, emphasize that the calyculus is never calyx-like, and the lobing of its margin, wherever present, is always irregular. Besides, the absence of any vasculature in the calyculus led them to conclude that there is no justification for regarding the calyculus as the calyx.

Working at Delhi, Narayana (1955a) has, however, reported the presence of vascular bundles in the five- to seven-toothed calyculus of *Nuytsia floribunda*. There is an outer whorl of six traces in the ovary and five of these enter the lobes of the calyculus whereas the sixth fades away. While further work is necessary, it now appears likely that the calyculus may in fact represent the calyx.

Mamelon—The ovary does not contain any normal ovule, but several species show a special mammiliform structure, called the mamelon, which arises from the base of the ovarian cavity (Schaeppi & Steindl, 1942; Maheshwari & Singh, 1952; Johri & Agrawal, 1954).

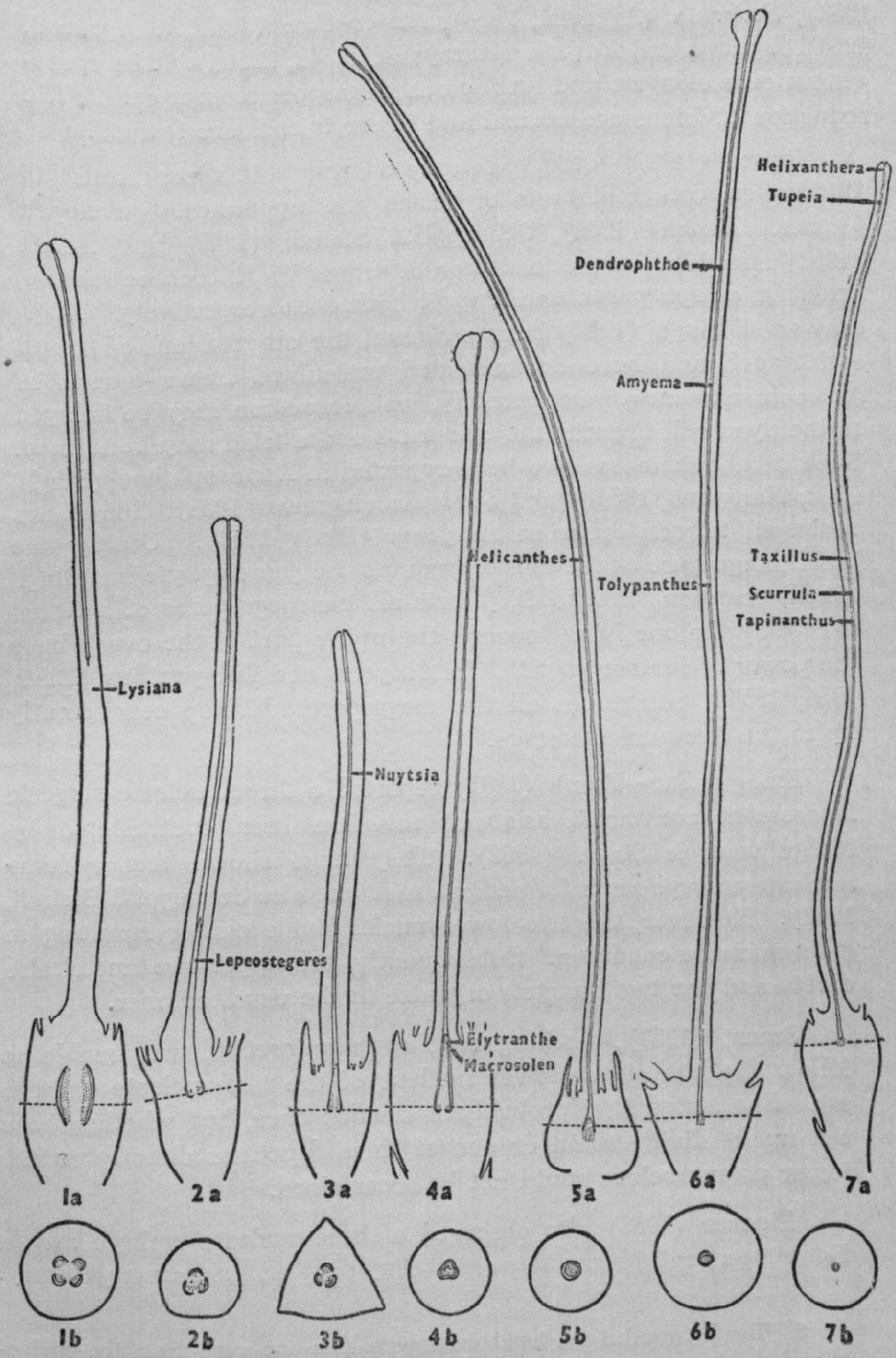
Griffith (1836, 1844) called the basal swelling in the ovary of *"Loranthus"* as the free central placenta, and in *Hyphear europaeum,* Hofmeister (1859) considered it equivalent to a single erect ovule. Treub (1881) interpreted the mamelon of *Macrosolen cochinchinensis* as a free-central placenta and the three to four basal lobes as equivalent to rudimentary ovules. Schaeppi & Steindl (1942), and Maheshwari & Singh (1952) agree with this view.

A feature of considerable significance is the presence of vascular elements in the mamelon. Isolated tracheids were first reported by Dixit (1954) in *Amyema miquelii* and since then they have also been recorded in *Dendrophthoe neelgherrensis* (Narayana, 1954b) and *Nuytsia floribunda* (Narayana, 1955a). It would be desirable to check up this feature in other species of the Loranthoideae, but on present evidence the vascular tisssue in the mamelon seems to represent the vestiges of the ventral bundles of a free central placenta.

Starting from a hypothetical inferior ovary with axile placentation and a single ovule in each locule, one may visualize the disappearance of the integuments and the reduction of the ovules to result in a condition similar to that of *Lysiana* (Figs. 1a, b). The next step in reduction is the retention of the partitions only in the lower region of the ovary, as in *Lepeostegeres* (Figs. 2a, b) and *Nuytsia* (Figs. 3a, b). In *Elytranthe* and *Macrosolen* the partitions have disappeared completely, but the mamelon retains a lobed appearance (Figs. 4a, b). In *Helicanthes* the lobes are already lost (Figs. 5a, b). In *Amyema, Dendrophthoe neelgherrensis, D. pentandra, Scurrula montana* and *Tolypanthus* (Figs. 6a, b) there is only a transitory swelling at the base of the ovarian cavity formed by the enlargement of the sporogenous tissue, and in *Helixanthera, Tapinanthus, Taxillus, Tupeia* and several species of *Dendrophthoe* and *Scurrula* there is a virtual disappearance of the mamelon (Figs. 7a, b).

Megasporogenesis—One to four archesporial cells differentiate in *Elytranthe* (Schaeppi & Steindl, 1942), *Macrosolen* (Maheshwari & Singh, 1952), *Lepeostegeres* and *Tolypanthus* (Dixit, 1955, 1956a), while there is a much larger number in *Dendrophthoe* (Singh,

Lysiana; 2a, b. *Lepeostegeres;* 3a, b. *Nuytsia;* 4a, b. *Elytranthe* and *Macrosolen;* 5a, b. *Helicanthes;* 6a, b. *Tolypanthus, Amyema* and *Dendrophthoe;* and 7a, b. *Tapinanthus, Scurrula, Taxillus, Tupeia* and *Helixanthera.* The levels to which embryo sacs rise in the style are indicated by the names of the genera.



FIGS. 1a-7a. Diagrammatic representation of gynoecia (l.s.); note the mamelon in FIGS. 1a-6a. FIGS. 1b-7b. Cross sections of the ovary at levels indicated in longisections. The dotted regions represent sporogenous tissue. FIGS. 1a, b.

1952), *Scurrula* (Agrawal, 1954), *Helicanthes* (Johri & Agrawal, 1954), *Lysiana* and *Nuytsia* (Narayana, 1954a, 1955a). The sporogenous cells elongate and function as megaspore mother cells. The reduction divisions are normal and linear tetrads are formed.

Treub (1881) observed only a row of three cells in *Macrosolen cochinchinensis* of which the uppermost[3] developed further. Rauch (1936), Schaeppi & Steindl (1942) and Smart (1952) also describe the formation of 'triads' in *M. cochinchinensis, Scurrula atropurpurea* and *Tupeia antarctica* respectively. Maheshwari & Singh (1952) have criticised the observations of Treub and Rauch and have demonstrated undoubted linear tetrads in *M. cochinchinensis.* Occasionally, wall formation may be delayed in the dyad cell towards the epidermis and a triad may be formed. Such a behaviour has also been reported in *Dendrophthoe* (Singh, 1952; Narayana, 1954b) and *Scurrula* (Agrawal, 1954). Sometimes ⊥-shaped tetrads occur in *Amyema* (Dixit, 1954). Usually the basal[4] megaspore of the tetrad produces the gametophyte but Singh (1952) mentions that in *D. falcata* all the megaspores of a tetrad excepting the one lying towards the upper part of the ovary have the power of further growth. In *Amyema* and *Tolypanthus* (Dixit, 1954, 1956a) sometimes all the megaspores develop concurrently up to the two-nucleate stage.

Embryo Sac—Treub (1881) reported a three-nucleate stage in *Macrosolen cochinchinensis* but this has been contradicted by Maheshwari & Singh (1952). They have shown that a four-nucleate embryo sac is formed as usual (*see* also Schaeppi & Steindl, 1942). However, the two lower nuclei divide earlier resulting in a six-nucleate condition. Subsequently, the upper two nuclei also divide and the two quartets organize in the usual manner.

Smart (1952), working on *Tupeia antarctica,* was unable to follow the fate of the lower nucleus of the two-nucleate embryo sac. She stated that probably it divides once, but whether both the nuclei divide again or one of them functions directly as the lower polar nucleus could not be ascertained.

Pienaar (1952) also observed a three-nucleate embryo sac in *Loranthus rubromarginatus* (= *Tapinanthus rubromarginatus*).

3. Treub regarded the basal megaspore (the one away from the epidermis of the mamelon) as the uppermost.

4. The megaspore facing the epidermis of the mamelon should be regarded as the uppermost.

The two nuclei at the upper end divide to produce the egg apparatus and the upper polar nucleus. However, Pienaar was unable to follow the further fate of the nucleus at the lower end of the embryo sac.

Several embryo sacs may develop simultaneously in the same ovary (Figs. 8, 10); during their growth they digest the contents of the adjacent cells which are rich in starch.

The embryo sacs extend down up to the collenchymatous pad. Upward they go up to the base of the style or still higher (Figs. 1a-7a). In *Macrosolen cochinchinensis* (Fig. 4a) (Maheshwari & Singh, 1952) the tips of the embryo sacs extend only to the base of the style; up to one-fifth the length of the style in *Lepeostegeres gemmiflorus* (Figs. 2a, 8, 9a-c) (Dixit, 1955); half the style in *Helicanthes elastica* (Fig. 5a) (Johri & Agrawal, 1954), *Lysiana exocarpi* (Fig. 1a) (Narayana, 1954a), *Taxillus cuneatus, T. tomentosus* (Fig. 7a) (Narayana, 1955b), *Tolypanthus lagenifer* (Fig. 6a) (Dixit, 1956a), and *Tapinanthus uhehensis* (Fig. 7a) (Dixit, 1956b); and to two-thirds the height of the style in *Taxillus sclerophyllus* (Schaeppi & Steindl, 1942), *Dendrophthoe falcata* (Singh, 1952), *D. neelgherrensis* (Fig. 6a) (Narayana, 1954b), *Amyema congener, A. miquelii, A. pendula, A. preissii* (Fig. 6a) (Dixit, 1954), *Nuytsia floribunda* (Fig. 3a) (Narayana, 1955a) and *Tolypanthus involucratus* (Dixit, 1956a). In *Helixanthera hookeriana* (Fig. 7a) (Schaeppi & Steindl, 1942) and *Tupeia antarctica* (Fig. 7a) (Smart, 1952) the apex of the embryo sac reaches up to the stigma and in *Helixanthera ligustrina* (Figs. 10, 11) (Maheshwari & Johri, 1950) it grows as far as the stigmatic epidermis.

The antipodal cells appear to be ephemeral and have been missed in a number of species. Occasionally, only two cells are organized, one being binucleate, e.g. *Lepeostegeres gemmiflorus* (Fig. 9c) (Dixit, 1955). Persistent antipodal cells have been reported in *Macrosolen cochinchinensis* (Maheshwari & Singh, 1952), *Helicanthes elastica* (Johri & Agrawal, 1954) and *Lepeostegeres gemmiflorus* (Dixit, 1955). Schaeppi & Steindl (1942) misinterpreted the antipodal cells of *M. cochinchinensis* as the three non-functioning megaspores. Those of *L. gemmiflorus* were mistaken for a part of the proembryo (*see* Dixit, 1955).

Polar fusion has so far been observed only in *Macrosolen cochinchinensis* (Maheshwari & Singh, 1952) and *Lysiana exocarpi* (Narayana, 1954a) but may be presumed to occur in other species also. It is the lower polar nucleus which migrates to the upper part of the embryo sac and fuses with the upper polar.

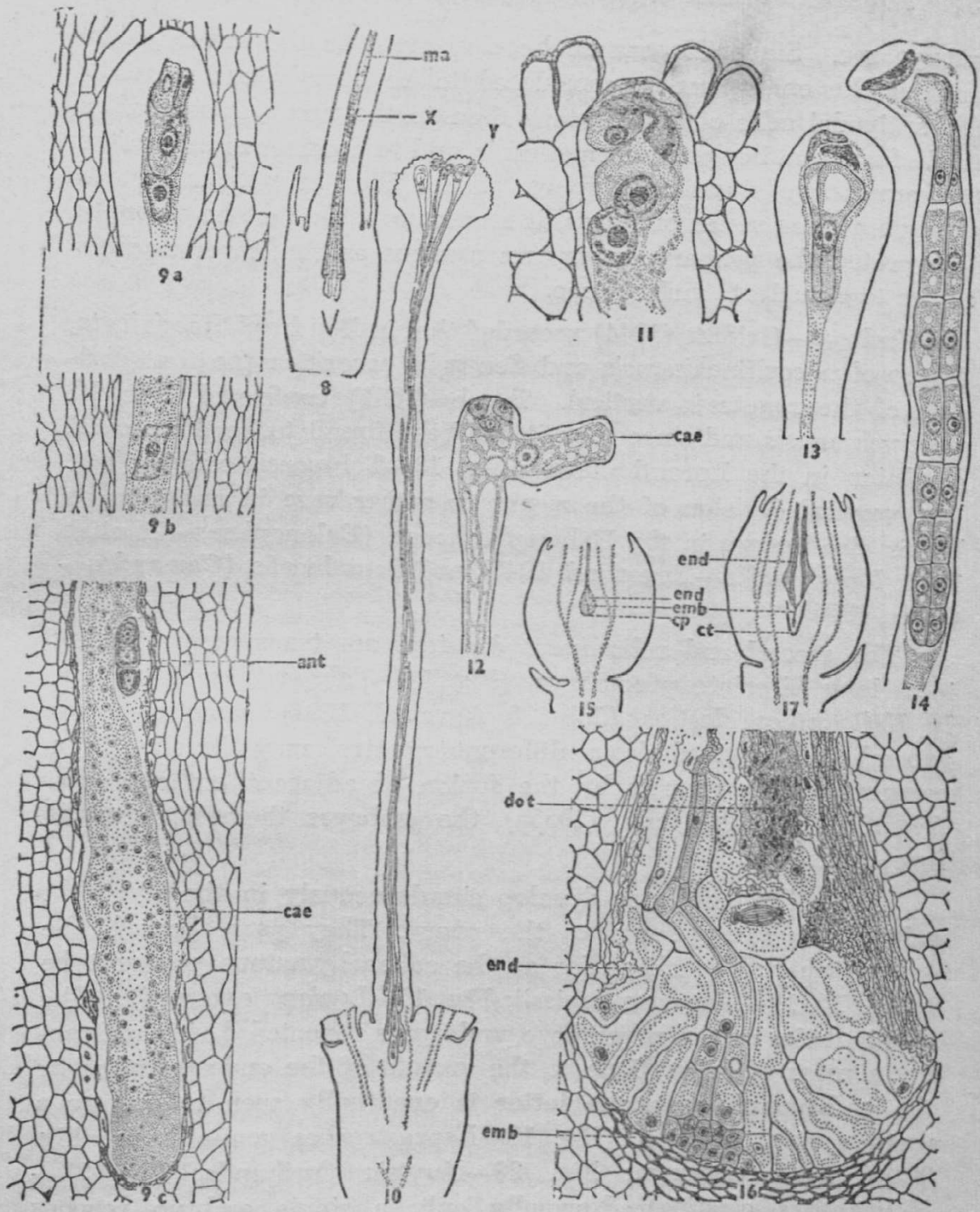
Maheshwari & Singh (1952) showed that in *Macrosolen cochinchinensis* a lateral pouch arises slightly above the level of the antipodals and progresses downwards as far as the collenchymatous tube forming a caecum. This has since been confirmed also in *Lysiana* (Narayana, 1954a) and *Lepeostegeres* (Dixit, 1955) (Fig. 9c). In *Nuytsia* (*Narayana,* 1955a) a lateral process also develops from the upper end of the embryo sac and is directed towards the wall of the style (Fig. 12). This feature has not been observed in any other member of the Loranthoideae and is comparable to the condition met with in *Comandra* (Santalaceae) (Ghosh, 1955).

Endosperm—The earliest observations on the development of the endosperm are by Griffith (1844) on *Dendrophthoe falcata* and *Macrosolen cochinchinensis,* and Hofmeister (1859) on *Hyphear europaeum.* A more detailed account has been given by Maheshwari & Singh (1952) in *Macrosolen* and Singh (1952) in *Dendrophthoe* (*see* also Rauch, 1936; Schaeppi & Steindl, 1942).

The initial stages in endosperm formation are known only in *Macrosolen cochinchinensis* (Maheshwari & Singh, 1952). The primary endosperm nucleus travels to the lower end of the embryo sac and the first division is followed by a transverse wall. Further divisions produce a four-seriate arrangement followed by additional divisions. Endosperm formation occurs in several embryo sacs in the same ovary and finally all of them fuse to form a composite structure (Figs. 15, 16—*Amyema preissii;* Fig. 17—*Lysiana exocarpi*). The mature endosperm is vase-shaped and its appearance largely depends on the vascular skeleton of the ovary (Figs. 18-31). During these changes the collenchymatous pad may often be shifted to a lateral position (Figs. 30, 31—*Tolypanthus lagenifer*).

Pienaar's (1952) description of the endosperm of *Tapinanthus rubromarginatus* leaves many doubtful points. According to him the upper polar nucleus degenerates soon after the differentiation of the egg. The large nucleus situated at the lower end of the embryo sac is regarded as the lower nucleus of the two-nucleate

Nuytsia floribunda (Labill.) R.Br. Upper end of the embryo sac with lateral caecum. (After Narayana, 1955a). × 194. Figs. 13, 14. *Tolypanthus lagenifer* (Wight) Van Tiegh. (After Dixit, 1956a). Fig. 13. Two-celled proembryo. × 305. Fig. 14. Biseriate proembryo. × 305. Figs. 15, 16. *Amyema preissii* (Miq.) Van Tiegh. (After Dixit, 1954). Fig. 15. Outline diagram of l.s. ovary showing endosperm and proembryo. × 10. Fig. 16. Same, enlarged. × 117. Fig. 17. *Lysiana exocarpi* (Behr.) Van Tiegh. L.s. ovary showing endosperm and embryo. (After Narayana, 1954a). × 6.



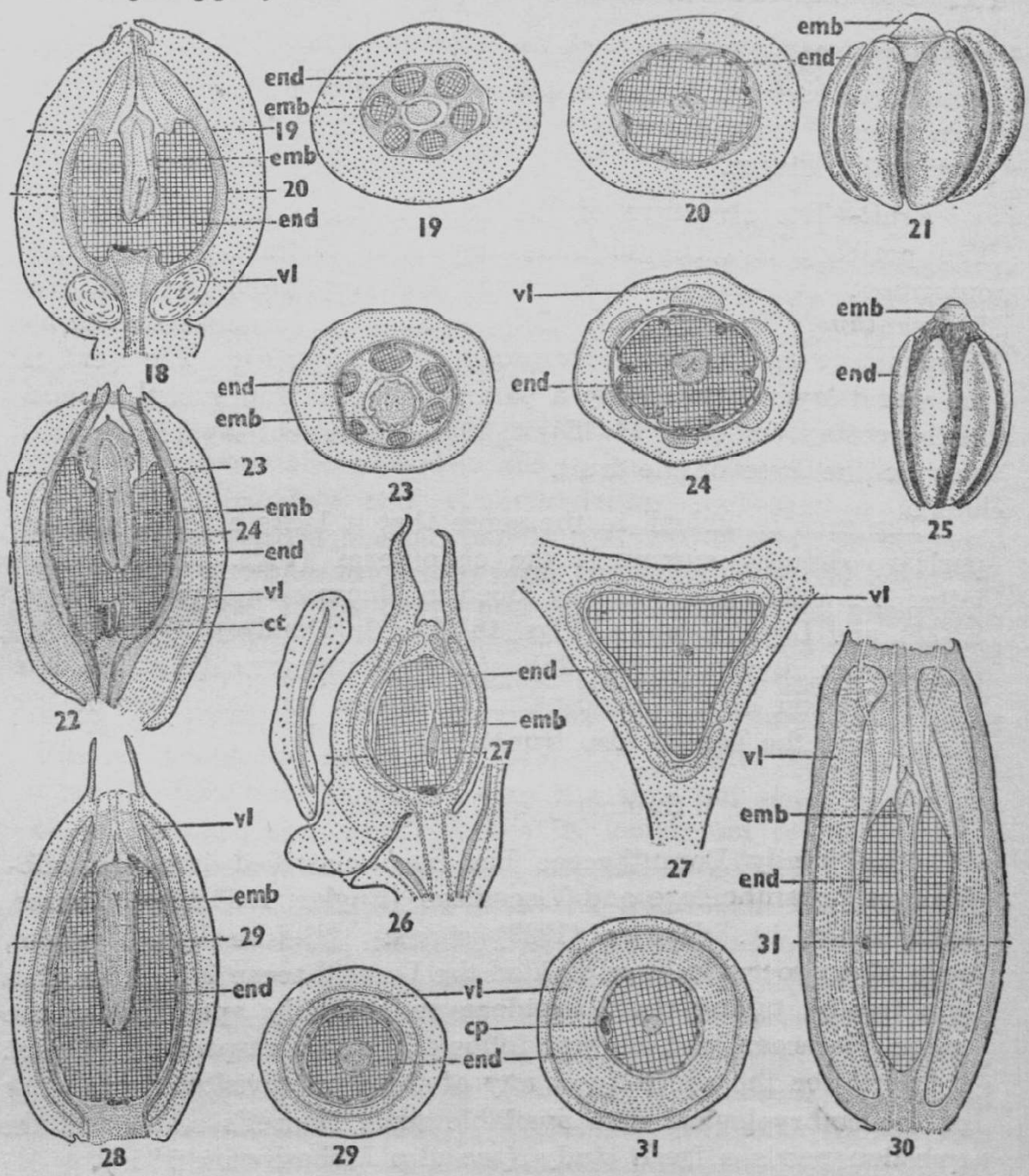
FIGS. 8-17. (*ant*, antipodal; *cae*, caecum; *cp*, collenchymatous pad; *ct*, collenchymatous tube; *dot*, degenerated ovarian tissue; *emb*, embryo; *end*, endosperm; *ma*, mamelon). FIGS. 8, 9. *Lepeostegeres gemmiflorus* Bl. (After Dixit, 1955). FIG. 8, L.s. ovary and lower portion of the style showing embryo sacs. × 10. FIGS. 9a, b, c. Upper, middle and lower parts of the embryo sac marked X in FIG. 8; note the caecum in FIG. 9c. × 384. FIGS. 10, 11. *Helixanthera ligustrina* (Wall.) Dans. (After Maheshwari & Johri, 1950). FIG. 10. L.s. carpel, tips of five embryo sacs are seen in the stigma. × 9. FIG. 11. Embryo sac (marked Y in FIG. 10) in contact with stigmatic epidermis. × 273. FIG. 12.

embryo sac. Since this cannot be equivalent to the lower polar nucleus the endosperm is considered to be homologous more to the antipodal cells of angiosperms than to the tissue arising after triple fusion, although its formation is said to commence only after the proembryo starts developing. Since Pienaar was unable to observe a close series of stages, it seems probable that he wrongly interpreted the primary endosperm nucleus as the lower nucleus of the two-nucleate embryo sac.

Embryo—Griffith (1844) recorded that in *Dendrophthoe falcata, Macrosolen cochinchinensis* and *Scurrula parasitica* the first division of the zygote is vertical. Treub (1881) confirmed it in *M. cochinchinensis* and since then it has been found to be the normal condition in the Loranthoideae (Fig. 13—*Tolypanthus lagenifer*). The vertical division of the zygote is rather rare in angiosperms but is also known in the Balanophoraceae (*Balanophora*—Zweifel, 1939; *Langsdorffia*—Fagerlind, 1945) and Santalaceae (*Comandra*—Gosh, 1955).

The second and subsequent divisions are transverse resulting in a long biseriate proembryo (Fig. 14—*Tolypanthus lagenifer*). In *Helicanthes elastica* (Johri & Agrawal, 1954), sometimes two proembryos, derived from different zygotes, may develop in a common cavity formed by the fusion of adjacent embryo sacs. According to Narayana (1955a) the embryos themselves might fuse in *Nuytsia.*

Several proembryos develop simultaneously in the style but only a few grow down to the ovary. They pass through the endosperm and reach as far as the collenchymatous pad or tube (Figs. 15, 16—*Amyema preissii;* Fig. 17—*Lysiana exocarpi*). The terminal tier of the proembryo undergoes repeated divisions forming a mass of cells. Due to the growth of the endosperm below the embryonal mass, the latter is eventually seen to occupy a more central position (Fig. 18—*Lepeostegeres gemmiflorus;* Fig. 22—*Lysiana exocarpi;* Fig. 26—*Nuytsia floribunda;* Fig. 28—*Helicanthes elastica*). Normally only a single embryo reaches maturity but two well developed embryos have been seen sometimes in *Dendrophthoe* (Singh, 1952; Narayana, 1954b) and *Lepeostegeres* (Dixit, 1955). The embryo is dicotyledonous but the cotyledons may fuse for part of the distance resulting in a pseudomonocotyledonous condition - Examples: *Helixanthera, Dendrophthoe, Scurrula, Helicanthes* (Figs. 28, 29), *Amyema* and *Tolypanthus* (Figs. 30, 31). A true radicle is absent and the radicular end has endarch collateral bundles (Singh, 1954). In *Lysiana*



FIGS. 18-31. (*cp*, collenchymatous pad; *ct*, collenchymatous tube; *emb*, embryo; *end*, endosperm; *vl*, viscid layer). FIGS. 18-21. *Lepeostegeres gemmiflorus* Bl. (After Dixit, 1955). FIG. 18. L.s. fruit. × 4. FIGS. 19, 20. T.s. at levels marked in FIG. 18. × 4. FIG. 21. Mature endosperm and embryo from whole mount. × 5. FIGS. 22-25. *Lysiana exocarpi* (Behr.) Van Tiegh. (After Narayana, 1954a). FIG. 22. L.s. fruit. × 4. FIGS. 23, 24. T.s. fruit at levels marked in FIG. 22. × 4. FIG. 25. Mature endosperm and embryo from whole mount. × 4. FIGS. 26, 27. *Nuytsia floribunda* (Labill.) R.Br. (After Narayana, 1955a). FIG. 26. L.s. fruit × 3·5. FIG. 27. T.s. at level marked in FIG. 26. × 4. FIGS. 28, 29. *Helicanthes elastica* (Desr.) Dans. (After Johri & Agrawal, 1954). FIG. 28. L.s. fruit. × 4. FIG. 29. T.s. fruit at level marked in FIG. 28. × 4. FIGS. 30, 31. *Tolypanthus lagenifer* (Wight) Van Tiegh. (After Dixit, 1956a). FIG. 30. L.s. fruit. × 4. FIG. 31. T.s. at level marked in FIG. 30. × 4.

(Figs. 22, 23) and *Helicanthes* (Fig. 28) it shows papillate outgrowths whose significance is not quite clear. Perhaps they may form lateral roots but this has to be confirmed by a study of germination stages.

Fruit—The structure of the fruit is more or less uniform. The pericarp is distinguishable into four distinct zones. The outermost is leathery (*Amyema, Dendrophthoe, Helicanthes, Tapinanthus, Taxillus* and *Tolypanthus*) or sometimes fleshy (*Lepeostegeres, Lysiana, Macrosolen* and *Nuytsia*). The next is the viscid layer followed by a parenchymatous and vascular zone. In *Lepeostegeres* (Fig. 18—Dixit, 1955) the viscid layer is limited only to the base of the fruit.

The seed is 'naked' in the sense that it lacks a testa and the pericarp directly surrounds the endosperm (Figs. 18-31). The latter is massive and is deeply lobed in *Macrosolen, Lysiana* (Figs. 22-25) and *Lepeostegeres* (Figs. 18-21). In *Dendrophthoe falcata, Macrosolen, Lysiana* (Fig. 25) and *Lepeostegeres* (Fig. 21) the apical portion of the endosperm is drawn out into five or six prongs and the base is also truncated.

Summary and Conclusion

The family Loranthaceae has been classified into the subfamilies Loranthoideae and Viscoideae (Engler, 1888-1889; Danser, 1929, 1933a, b). Danser (1929) states: "Before Van Tieghem, only Miers seems to have divided the Loranthaceae into more than one family, taking the Viscoideae as a family apart under the name Viscaceae, and he was followed in this respect by Miquel. I do not see the necessity of any of these subdivisions." In view of the embryological data available since Danser's statement, the question needs a fresh study (*see* also Maheshwari, 1954).

In the Loranthoideae the flower may be unisexual or bisexual but a calyculus is invariably present. The ovary may contain a lobed mamelon which may be fused with the ovary wall in between the lobes, e.g. *Lepeostegeres; Lysiana* and *Nuytsia;* or the mamelon may be free as in *Elytranthe* and *Macrosolen.* A free mamelon but without any lobes occurs in *Helicanthes;* it is extremely short in *Amyema, Dendrophthoe neelgherrensis* and *Tolypanthus;* and is absent in *Barathranthus, Dendrophthoe falcata, Helixanthera, Scurrula, Tapinanthus, Taxillus* and *Tupeia.* The presence of a collenchymatous pad or tube in the ovary is also a characteristic feature of the Loranthoideae. The embryo sac is of the Poly-

gonum type. Many embryo sacs may develop simultaneously in the same ovary and their tips grow up into the style. The lower end of the sac sometimes produces a caecum, e.g. *Amyema, Lepeostegeres, Macrosolen* and *Tolypanthus*. The endosperm is a composite structure formed by the fusion of all the endosperms developing in the same ovary. The first division of the zygote is always vertical and is followed by transverse divisions resulting in a long biseriate proembryo. The fruit shows four distinct layers and the viscid layer is situated outside the vascular bundles of the 'corolla' (*see* Danser, 1931).

In the Viscoideae, the flowers are unisexual (Danser, 1931), and the calyculus is generally absent but is known to be present in *Viscum orientale* and *V. articulatum* (Schaeppi & Steindl, 1945). The central ovarian papilla has two or more archesporial cells, each of which divides to form two dyad cells. Of these, the upper dyad cell is the larger and functions, while the lower soon degenerates. Thus the development conforms to the Allium type and has been described in *Ginalloa* (Rutishauser, 1937), *Korthalsella* (Rutishauser, 1935, 1937; Schaeppi & Steindl, 1945) and *Viscum* (Schaeppi & Steindl, 1945). A peculiar feature is that after the four-nucleate stage there is a slow but steady curvature of the embryo sac, which causes its lower end to bend out of the papilla and proceed upwards into the carpellary tissue. The egg apparatus differentiates from the quartet in the originally lower pole, which is, however, now situated at a higher level than the upper — Examples: *Dendrophthora* (York, 1913), *Phoradendron* (Billings, 1933) and *Korthalsella* (Rutishauser, 1935, 1937). The embryo sac is straight in *Viscum* (Schaeppi & Steindl, 1945), *Arceuthobium* (Thoday & Johnson, 1930) and *Ginalloa* (Rutishauser, 1937). Generally only one embryo sac develops which is said to show inverted polarity. Maheshwari (1948) draws attention to the fact that the embryo sacs "although apparently inverted, are not really so." The endosperm is derived from the primary endosperm nucleus of a single embryo sac. The first division of the zygote is usually transverse but sometimes vertical as in *Korthalsella* (Rutishauser, 1935) and the embryo lacks a suspensor (Thoday & Johnson, 1930; Schaeppi & Steindl, 1945). The viscous part of the fruit is situated within the vascular bundles of the perigone (Danser, 1931).

Thus, the Loranthoideae and the Viscoideae show important differences in floral structure, mode of development of the embryo sac, endosperm and embryo, and in the location of the viscid zone

of the fruit. These differences adequately warrant the raising of the two subfamilies to the rank of families. As proposed by Miers (*see* Danser, 1929), the name Loranthaceae may be retained for the Loranthoideae and the Viscoideae may be designated as Viscaceae.

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Some Early British Contributions to Essential Oil and Terpene Chemistry

BY

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One hundred years ago the University of Madras was founded. It is an interesting coincidence that in this year the centenary of Perkin's discovery of the first synthetic dyestuff, mauveine, is also being celebrated.[1] The opening of a factory by Perkin to manufacture this dyestuff inaugurated the first organic chemical industry thus stimulating an intensive study of the chemistry of organic compounds. It appears, therefore, not to be without interest to place on record the early contributions of British chemists to essential oil and terpene chemistry.

From very early times essential oils have provided the main source of materials for the manufacture of perfumes and cosmetics, whilst they have also found an extensive use in pharmacy and as ingredients of spices. In mediaeval literature India is frequently mentioned as the source of such well-known perfumes as attar of roses and sandalwood. It is, however, only during the last hundred years that any detailed knowledge of the constituents of essential oils has been obtained. This is perhaps not surprising in view of their complexity and the difficulties involved in the separation and the purification of their constituents since, apart from camphor and borneol, they are liquids.

A perusal of the articles in the standard text-books would lead to the conclusion that our knowledge in this important field of chemistry has been due in the main to the researches of foreign chemists and that British chemists made only minor contributions. Whilst not desiring to minimise the important investigations associated with the names of Aschan, Baeyer, Barbier, Berthelot, Komppa, Semmler and Wallach it is the purpose of this brief article to direct attention to the pioneering researches of British chemists to whom are due many of the experimental methods so

1. Robinson, *Endeavour*, 1956, 15: 92; Simonsen, *Research*, 1956, 9: 166.

successfully applied in this difficult field. To these chemists are undoubtedly due (i) the first method for the characterisation of essential oil constituents, (ii) the use of dehydrogenating reagents for their conversion into aromatic hydrocarbons, later identified by oxidative degradation and (iii) the introduction of the use of physical constants as a means of identification. In considering these early investigations it must be borne in mind that at the time they were undertaken knowledge of structure, as we now understand it, was extremely limited.

The first important investigation to which attention may be directed was by Greville Williams,[2] who, by the pyrolysis of rubber and guttapercha, obtained a hydrocarbon, cautouchcene, $C_{10}H_{16}$. He recognised that this hydrocarbon was isomeric with the hydrocarbon present in oil of turpentine and he proceeded to investigate the action of bromine on the two hydrocarbons. He found that they both gave bromo derivatives which, after treatment with sodium, furnished the same hydrocarbon, "cymol", yielding on oxidation with potassium dichromate and sulphuric acid an acid, "isolonic acid". In this early communication we find the first use of methods later applied with such success, namely, dehydrogenation to an aromatic hydrocarbon followed by oxidative degradation. The method employed was essentially the same as that generally associated with the names of Baeyer and Villiger.[3] They were probably unaware of the work of Greville Williams since they do not refer to it. As is well-known, the use of sulphur for dehydrogenation is attributed to Vesterberg[4] but as will be shown later (p. 140) Alder Wright[5] treated an alcohol, absinthol, $C_{10}H_{18}O$, from wormseed oil with potassium pentasulphide obtaining thereby a mixture of cymene and thiocymene. The first use of this reagent, phosphorus pentasulphide, as a dehydrogenating agent was apparently due to Pott,[6] whilst Kelbe[7] studied the action of sulphur on colophony (abietic acid).

At this time an increasing interest was being shown in the chemistry of essential oils and under the aegis of the British Association for the Advancement of Science a committee was formed

2. *Proc. roy. Soc.*, 1860, 10: 517.
3. *Ber.*, 1892, 25: 1401; 2067.
4. *ibid.*, 1903, 36: 4200.
5. *J. chem. Soc.*, 1874, 27: 1.
6. *Ber.*, 1870, 3: 422.
7. *ibid.*, 1878, 11: 2174.

with Dr. J. H. Gladstone as the Chairman and Dr. Alder Wright as Secretary. Stimulated by this committee important results rapidly accrued. In his first major memoir Gladstone[8] described the fractionation of a number of oils and the isolation of their main constituents. From the wide range of oils which he examined he isolated three hydrocarbons (i) $C_{10}H_{16}$, b.p. 160-176°, (ii) $C_{15}H_{24}$, b.p. 242-265° and (iii) $C_{20}H_{32}$, whilst from the oil of *Matricaria Chamomilla* he separated a blue hydrocarbon, "coecula", which he subsequently found to occur in the high boiling fractions of many other oils. This is probably the first mention of the hydrocarbon now known as azulene. Gladstone introduced a novel method for the characterisation of the $C_{10}H_{16}$ hydrocarbons. He determined their densities, refractive indices for the lines, A, D and H and their rotatory powers. He noted that the densities of the lower boiling hydrocarbons did not differ greatly from 0·9, the refractive indices lying between 1·46 and 1·5, there was, however, a marked variation in their rotatory powers. In attempting to differentiate the hydrocarbons he introduced the constant "specific refractive energy", which was the product of the division of the value for the refractive index by the density. From this constant, Gladstone deduced that the $C_{10}H_{16}$ hydrocarbons present in the oils from orange, citrus and lime were identical but that they differed from those present in the oils from wormseed, anise, thuja, mace and laurel, the latter hydrocarbons being identical with that occurring in oil of turpentine. Gladstone was therefore apparently the first chemist to make use of a physical constant to show the identity of hydrocarbons, a method which was later to find such wide application. Some years later[9] he extended the use of physical constants by the determination of vapour densities and he compared also their solubility in ethanol and their reaction with hydrogen chloride. From some of the lower boiling hydrocarbons he prepared a crystalline product, $C_{10}H_{16}$, 2HCl, and from the higher boiling hydrocarbons a similar product, $C_{15}H_{24}$, 2HCl, the yields in the latter case being very poor. In this second memoir he discussed for the first time oxygenated derivatives. He isolated from citronella oil, citronellol, b.p. 202-205°, incorrectly formulated as $C_{10}H_{16}O$. He also obtained an oxygen derivative, cajupetol, from the oils of wormseed, absinthe and cajeput. He was able to confirm from their physical

8. *J. chem. Soc.*, 1864, 17: 1; *Brit. Assoc. Rept.*, 1862, Pt. 2; 47.
9. *ibid.*, 1872, 25: 1.

constants the identity of "carvol" and "carotol", whilst from spearmint oil he obtained menthol and from oil of nutmeg myristol. He further recognised that the carvols from dill oil and from caraway seed oil were identical by the preparation of their derivatives with hydrogen sulphide and that this carvol differed from menthol which did not react with this reagent. In considering Gladstone's work we must bear in mind that he did not distinguish between an alcohol, an aldehyde and a ketone.

As the Secretary of the British Association Committee Alder Wright carried out a long series of experiments[10] which would appear to have been somewhat overlooked although he obtained results of considerable interest. Unlike Gladstone, Alder Wright was not specifically interested in the physical constants of the various products which he described and he studied mainly their chemical reactions, more especially their oxidation products. Thus, for example the hydrocarbon, hesperidine b.p. 178°, from orange oil yielded on oxidation with potassium dichromate and sulphuric acid a substance, $C_{10}H_{18}O$, b.p. *ca* 215°, together with carbon dioxide, acetic and formic acids. Since, as we now know, hesperidine is identical with limonene, this oxygen product must have been α-terpineol, a hydration and not an oxidation product. Results of much greater interest, however, accrued when, following Greville Williams, he treated the hydrocarbon with bromine when he found that the bromo derivative gave, on distillation, cymene. He also obtained cymene from the hydrocarbon present in oil of nutmeg but of much greater interest was his conversion of the alcohol, myristol, b.p. 212-218°, from the same oil, into cymene. For this purpose he employed a somewhat different method. He converted the alcohol into the chloride, $C_{10}H_{17}Cl$, by the action of phosphorus pentachloride, the chloride yielding on distillation cymene. Alder Wright considered myristol to be a secondary alcohol, isomeric with camphor which at that time was thought to be in aldehyde. [Myristol is now known to be a mixture of α-terpineol, borneol and terpinen-4-ol.]

Whilst in their later extensive experiments Alder Wright and his colleagues were able to correct the formulae given by previous investigators to some of the essential oil constituents, they were mainly concerned in establishing the fact that they all gave on

10. *Brit. Assoc. Rept.*, 1871, Pt. 2; 83; 1872; 311, 313, *J. chem. Soc.*, 1872, 25: 549; 1873, 26: 686; 1874, 27: 1, 317, 619; 1876, 29: 5; 1877, 32: 548; *Pharm. J.*, 1873; 2: 135; 1874, 4: 317; 5: 233.

dehydrogenation the same aromatic hydrocarbon, cymene, from which they obtained on oxidation p-toluic acid. They were thus able to prove that cymene could not be a diethylbenzene[11] but that it must be methyl *iso*propyl (or propyl) benzene.

It has already been emphasized that one of the main difficulties encountered in the study of essential oil constituents was that in the majority of cases they were liquids from which no crystalline derivatives had been prepared suitable for their identification and differentiation. It is true that Gladstone had used their physical constants but obviously the preparation of crystalline derivatives would be far more satisfactory. A solution to this problem was provided by the science master at Clifton College, W. A. Tilden. In 1874 he prepared nitrosyl chloride.[12] He did not undertake the preparation of this substance with the object of contributing to our knowledge of inorganic chemistry but in order to investigate its reactions with organic substances. Thus, he found that it reacted with phenol to yield chloroanil. In the following year[13] he made the very important observation that if nitrosyl chloride was allowed to react at $-10°$ with oil of turpentine, b.p. 158-161°, $\alpha_D - 15°$, a crystalline product was obtained. For this new substance he recorded no melting point but he showed that on treatment with ethanolic potassium hydroxide it furnished "nitrosoterpene", m.p. 120-130°. He regarded this new substance to be of sufficient importance to have its crystalline form described by Maskelyne.

Tilden in collaboration with Shenstone[14] now proceeded to examine the action of this new reagent on the hydrocarbons present in a great variety of essential oils. It is beyond the scope of this brief article to discuss in detail their important results. It will suffice as an indication of their importance to mention that they were able to show the identity of the hydrocarbons present in the French and American oils of turpentine since they gave the same "nitrosoterpene"; they differed, however, from the Swedish oil of turpentine from *Pinus sylvestris* which did not give a crystalline derivative with nitrosyl chloride. The hydrocarbon, hesperidine, from caraway and berganot oils was different from that in oil of turpentine yielding a nitrosochloride, m.p. *c.* 103° and a

11. *c.f.* Oppenheim, *Ber.*, 1874, 7: 628.
12. *J. chem. Soc.*, 1874, 27: 670, 851.
13. *ibid.*, 1875, 28: 514.
14. *ibid.*, 1877, 31: 554.

"nitrosohesperidine", m.p. 71°. The value of this new reagent having thus been established Tilden in subsequent communications[15] proceeded to examine the hydrocarbons in a great variety of essential oils. Based mainly on the fact that they all gave on dehydrogenation cymene, he suggested formulae for them. These suggestions are now only of historical interest.

It is not possible to over-emphasise the importance of the introduction by Tilden of nitrosyl chloride as a reagent in the study of the chemistry of the terpenes. It ranks in importance with the application by E. Fischer of phenylhydrazine in his investigations on the chemistry of the carbohydrates. Following its introduction the chemistry of the terpenes was studied by many investigators and our knowledge is now very complete. Tilden himself in a series of lectures has given an excellent summary of his early work.[16] Many years later a British chemist was to provide the final rigid proof of the structure of a simple terpene when Perkin in 1904 synthesized dipentene and the related alcohol, α-terpineol.[17]

We have now dealt with the past, what of the future of organic chemistry? We seem to be at the beginning of a new epoch. Whilst there is no doubt that from the apparently limitless natural sources new substances will be isolated and their structures elucidated, this task, with the techniques now available, is comparatively simple. Surely the future now lies in applying the great skill and originality of the organic chemist to a study of the problems of the living organism. Here, there lies before him an almost virgin field of undoubted interest, if of great complexity. Can we doubt that in the future he will provide a solution to these problems as he has in the past solved the problems of the structures of so many natural products? Is it too much to hope that when the bicentenary of the University of Madras is celebrated the names of her alumni will be famous for their contributions to this field of scientific research?

15. *Pharm. J.*, 1877 [iii], 8: 191; *J. chem. Soc.*, 1878, 33: 80, 247.
16. *Perf. Ess. Oil Rec.*, 1912.
17. *J. chem. Soc.*, 1904, 85: 416.

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An Essay on Computation

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Mathematicians generally have little fondness for computation, and are inclined to regard the need for lengthy routine calculations as an unpleasant but inescapable concomitant of more interesting features of mathematics and its applications. I am sure that no good mathematician has ever shirked for long the calculations essential to his researches, but I am equally sure that until recently very few mathematicians have been inspired to devote much time or effort to exploring the art of computation. The invention of the high-speed electronic computer has, however, begun to alter the mathematician's attitude toward this art and has already been responsible for a good deal of professional mathematical interest in its modern aspects, both practical and theoretical. The possibilities being opened up by the development of the modern computing machine, even though it is too soon for us to begin to delineate them with any clarity, are indeed an astonishing challenge to the professional interest and skill of mathematicians as well as to the curiosity of the general public. That is why I venture to write this short essay, in an attempt to describe some of these new vistas from a mathematician's point of view.

The two characteristics of the modern electronic computer which mark it as a truly revolutionary invention are speed and versatility. Speed gives the machine its power to complete in a few days or weeks calculations so massive that in a lifetime even a large staff of mathematicians and computers working with ordinary manual aids and mechanical devices would be unable to finish the task. Versatility renders the machine adequate for handling an almost unlimited variety of problems, not excluding those of a very high degree of complexity. The opportunity for building machines with these two important characteristics was created by the fortunate confluence of two quite disparate and otherwise unrelated scientific developments—the perfection of the vacuum tube, and the logical analysis of the structure of mathematics which is associated above all with the names of Russell and Whitehead. The attain-

ment of the extremely high operating speeds necessary in handling voluminous or exceptionally complex calculations within acceptable limits of time depends upon the ingenious application of electronic techniques. On the other hand, the achievement of versatility results from the implementation, by means of a few simple electronic devices, of our modern knowledge of how to reduce the complexity and variety of mathematics to a small number of elementary logical operations. Looking backward, we can see how the invention of the high-speed multi-purpose computer might have sprung from this opportunity, by virtue of a conscious effort to join these two scientific components. However, the fact appears to be that sporadic and more or less unrelated advances, of a practical rather than a theoretical inspiration, provided the actual basis for modern computer design. Indeed, some of these advances were undoubtedly prompted by the realization that the development of mathematics and its applications had reached a point where the associated computational programs could no longer be handled with the available human and mechanical resources. Thus need as much as opportunity inspired invention in this particular instance.

If the origins of the modern developments in the field of computation are to be examined more closely, it will be necessary to assemble historical materials from a good many different sources, some of them perhaps obscure or difficult to get at. Some twenty five years ago I had occasion, in lecturing on Fourier series, to discuss the principles involved in the design of harmonic analysers; and I well remember how meagre was the material I could bring together about the history of computing machines in the nineteenth and twentieth centuries up to about 1930. Of course, a considerable variety of mathematical instruments had been invented over that period for carrying out arithmetical calculations and for performing other more elaborate mathematical tasks. Many of these machines, such as integraphs, differential analysers, and harmonic analysers, were based on special applications of kinematical principles; and all were mechanical in nature, consisting of more or less numerous moving parts which in the interests of computational accuracy had to be designed and fabricated with care. Of these early mathematical machines, the one which marks the highest achievement is perhaps the harmonic analyser, so useful in analysing and predicting the tides. A development which seems, in a historical sense, to constitute a transitional link between the early period and the modern was the construction by Dr. Vannevar Bush

and his associates at the Massachusetts Institute of Technology of a series of differential analysers, machines designed to solve differential equations of a fairly broad type. While these machines involved no sharp break with traditional techniques, their power and usefulness were impressive enough to kindle a wide interest in the potentialities of machine computation. Because of this stimulating influence of Bush's work it may be considered as inaugurating the modern phase of the computing art, despite the fact that it did not point out the paths later followed in order to achieve speed and versatility. A very different approach to the problem of mathematical computation was made in the early thirties by Dr. W. J. Eckert of Columbia University, who showed that extremely elaborate and voluminous astronomical calculations could be carried out rather rapidly be a skilful use of Hollerith machines alone. The origins of this approach can be found in the theoretical contributions of the mathematician Charles Babbage (1792-1871), some of whose ideas found their embodiment in the machine actually constructed at a later time by Hollerith. The Hollerith machine is essentially a simple device for sorting punched cards into specified classes—in other words, it is a machine which mechanically performs the basic logical operation of classification and which, therefore, should in principle serve to carry out all kinds of mathematical calculations. In three or four years' time Dr. Eckert was able to run through all the definitive calculations made upon the motions of the moon by the celebrated mathematician and astronomer E. W. Brown in the course of his entire life. This was a conclusive practical demonstration of the versatility and mathematical power of a logical machine. To what extent this demonstration may have influenced subsequent developments, I am simply unable to say. So far as I am aware it did not enter into the thinking of Dr. Howard Aiken of Harvard University, to whom should go much of the credit for the initial advances in modern high-speed computation. Dr. Aiken conceived the idea of applying the complex electrical techniques developed by the telephone engineers to the design of a large and elaborate multi-purpose computing machine which could be set up to carry out at high speed and in a fully automatic manner any desired program of mathematical calculations. He had to overcome many difficulties and some fortuitous obstacles before he won an opportunity to build the first model of his machine, but today the Computation Laboratory he created at Harvard University has been functioning for more than a decade and has earned its world-wide renown,

To tell the truth, I am not certain that Dr. Aiken's first machine was completed earlier than the ENIAC, a machine constructed at the University of Pennsylvania during World War II. It should be emphasized that the earliest types of high-speed machine were built and operated under conditions of secrecy as part of the war effort, with the consequence that the historian eager to pin down the facts may be faced with a fairly complicated task of documentation. Be that as it may, the ENIAC embodied some important technical advances over Dr. Aiken's original conception, as it achieved much greater operating speeds through the use of electronic rather than mechanical switching. The first pioneer models have been followed by others in fairly rapid succession, with many technical improvements designed to increase speed, capacity, and reliability of operation. The distinguished mathematician Professor John von Neumann of the Institute for Advanced Study (Princeton) played the leading part in the design and construction of one of the later machines. Although the design and operation of modern computers are now sufficiently well understood that various models are currently in commercial production, it is almost certain that further technical and theoretical advances of a fundamental kind lie ahead. Indeed, the replacement of tubes by transistors is already in view, opening up the prospect that future machines will be more compact and more economical in respect to energy-consumption.

The new machines have been seized upon as extremely valuable instruments by research laboratories and industrial establishments making use of applied mathematics. As a result there has been a rapid and large-scale expansion of computing facilities, particularly in the United States. This expansion, contrary to what the public may have been led to believe by newspaper accounts of these new scientific wonders, has opened up a whole new field of employment for mathematicians trained at intermediate and higher levels. The fact is that the machine can do nothing until it has received instructions in the form of a carefully worked out program drawn up in terms suited to the structure of the machine itself. Thus the preparation of mathematical material for machine computation requires a thorough knowledge of the underlying mathematical theories and a specialized skill in casting them into efficient programs for computation. The large-scale use of computers has, therefore, added heavily to the demand for mathematical services already in short supply. A few years ago I had occasion to discuss this point with Dr. John Curtiss, who was then directing

the mathematical activities of the United States Bureau of Standards. My recollection is that his estimate of the staff requirements for governmental computing programs alone indicated an immediate need for about four hundred mathematicians trained up to the master's degree level or above. In the future there will very likely be a reduction and modification in the demand per machine, but the certainty of further rapid increase in the number of machines in use guarantees a rising total demand for a rather long time to come.

The introduction of high-speed electronic computers has augmented man's mathematical powers in so spectacular a fashion that it is necessary to remind ourselves of some of the limitations affecting present and future uses of the machine for mathematical purposes. Some of these limitations can perhaps be overcome by technical developments, others may have to be accepted as permanent bars to progress in certain directions. There seems to me to be a fundamental difference between computing machines and the machines which constitute our present industrial apparatus, a difference which is reflected in an essential difference in the limitations to which they are respectively subject. To make the point somewhat too sharply, it is that the computing machine is devised to utilize time more efficiently while most industrial machines are designed to tap and utilize the sources of energy known to us. To be sure, many of our modern technological aims involve speeding up the use of some of our natural stores of energy and accelerating other forms of human activity, so that more can be accomplished in a given time. However, it is now widely recognized that the principal achievement of the industrial revolution has been to increase the per capita utilization of energy in human affairs and that in the long run the only sources of energy on which we can count for the maintenance and development of a technological society are those, like the sun and the nucleus of the atom, which supply us at rates beyond our power to affect or control. Thus the ultimate limitation upon the industrial use of machines is set by considerations of energy. Beyond the foreseeable time when our coal and our oil will have been exhausted and our water power fully exploited, our hopes rest upon the discovery of new ways to employ nuclear and solar energy. On the other hand, the essential function of the modern computer is to enable us to do in a relatively short time tasks so complex that their completion by ordinary means would be possible but only by dint of

coordinated efforts scheduled over inacceptably long periods of time. The ultimate limitations in the field of computation must, therefore, include the limitations upon speed of machine performance. The latter arise from two factors, one of a logical nature, the other physical. When any reasonably difficult mathematical calculation is analysed into its elements, it appears as a complicated pattern made up of interrelated sequences of simple operations. The number of steps in the longest such sequence occurring in the pattern obviously is one of the determinants of the time which the machine will need to carry out the calculation in question. The other major determinant is the speed with which the machine can perform the individual steps into which all computations are to be analysed. For a given calculation logic must evidently set a limit below which we cannot force the number of steps needed to carry out the calculation. For a given basic machine operation, Nature, so far as we now know, sets a limit below which we cannot hope to reduce the time-span requisite for effecting the operation. Of course we can hope to apply our knowledge of logic and of physics so as to choose the best possible calculation patterns and the best possible combinations of machine elements within the ultimate limitations to which we have alluded; but we cannot expect to overcome them altogether. So far as the machine-components are concerned, the fact that to attain the present operating speeds we have already had to utilize electronic devices suggests that future improvements may have to be sought in some sort of nuclear engineering. From some of the reports which have been published, like the one stating that the ENIAC could compute the trajectory of a bullet in less time than the bullet would take to complete its motion, it might be thought that the limitations which we have been considering are theoretical rather than practical from a computer's point of view. However, this is by no means the case, as the problems of short-term weather forecasting clearly show. To predict tomorrow's weather from today's observations over an area as large as that of the continental United States poses the practical problem of collecting a huge volume of data, carrying out upon them the necessary calculations, and disseminating the results to potential users, all within twenty-four hours' time. Evidently the computing machines used must be capable of completing their share of the work in only a fraction of a day. Existing machines can handle the numerical data currently supplied in an hour or so, it is said. While this speed is entirely acceptable, it seems quite possible that the machines could prove to be

inadequate if more computations were required of them in this connection—either because of an increase in the number of weather stations, or because of the introduction of forecasting methods requiring more calculations with the data made available. Thus the fundamental limitations which we have been discussing here are already felt at the practical level.

The complexity of the calculation-patterns arising in the applications of mathematics offers a serious challenge to the machine-designer. In principle this challenge might be met by enlarging the machine, by enabling it to handle segments of the computation seriatim, or by combining both expedients. As the present machines are not very bulky and may soon be replaceable by similar but less bulky machines based on the transistor, there does not seem to be any immediate practical obstacle to building much larger and more complicated machines than any now in operation. However, there is an important technical difficulty which becomes apparent when an attempt is made to arrange the segments of a computation in series, and which proves to be inherent in the nature of mathematical calculation. The point is that in complicated calculation-patterns, results obtained at one stage may have to be stored for use at various later stages. The machine has, therefore, to be provided with a memory, capable of retaining information conveyed to it and of making this information readily available as required. Thus mere enlargement is not enough to produce machines capable of handling more complex computations; it must be accompanied by an enlargement of the memory with which the machine is equipped. Here the technical obstacles have been discovered to be quite serious indeed. They constitute a genuine limitation upon progress in the art of machine computation at the present time, though one which we can hope eventually to overcome in large measure. Of course, even if there were no technical barriers to the indefinite enlargement of the machine or of some particular part of it, such as the memory, the construction and maintenance of very large machines would surely be impractical. Apart from the obvious inconveniences of sheer size, it seems likely that increases of size and complexity may be subject to a law of diminishing returns in the case of computing machines as well as elsewhere. In conclusion, it seems that some limitation upon size must be accepted, although this appears to be a less serious and a less fundamental one than the limitation on speed which has been discussed earlier in this essay.

Before turning to some aspects of the modern art of computation which hold a special interest for the mathematician, I must mention, if only briefly, one of its ramifications which bids fair to develop into an independent sister art under the name of "Automation". Throughout this discussion emphasis has been laid on the versatility of the modern types of computing machine. Their ability to handle the complex patterns of arithmetical operations involved in computing numerical answers to mathematical problems extends to patterns composed of other operations, mathematical or non-mathematical. In particular, the electronic techniques applied in building a high-speed computer are now finding application in the construction of automatic controls of manufacturing and managerial operations. The new industrial revolution which such applications portend has been vividly described in Professor Norbert Wiener's book "Cybernetics". The fully automatic factory and the fully automatic counting-house are no longer unrealizable dreams; pioneering realities today, they will be tomorrow's commonplaces. For a long time we have known how to mechanize individual operations, even highly complex ones, such as the machining of a metal part, the production of a light-bulb, or the writing and mailing of a cheque. What we can now do is to mechanize the coordination and control of the logical pattern in which the component operations of a complete manufacturing or managerial process are to be performed. An essential feature of this mechanization is that it provides for the automatic correction or suspension of component operations which may drift from the prescribed operating norms. This aspect of control is important in the case of computing machines, but very much more so in the case of an automatic factory. It requires that the control mechanism be capable of recording certain information about the execution of the process while under way and acting upon this information in an appropriate manner. At the very least it must be able to stop the whole process when something goes badly wrong and to signal an attendant who can then put matters right and start up the process again. In the case of a computing machine, for instance, the failure of a single tube may throw the entire computation off, and so should be detected and remedied at once. Some of the modern machines have, therefore, been so designed as to check their own operations and to suspend computing whenever the check shows something out of line. Long ago the invention of the governor for steam-engines established the basic principles of automatic control. It is the subtle elaboration of these principles through

modern electronic techniques which has given us a new vision of man in his industrial world.

While the mathematician may be intrigued and duly impressed by modern developments in the field of computation, it is natural that he should take a somewhat detached attitude towards them until he can relate them to the problems on which he is personally engaged. For the majority of mathematicians this means raising the question "Can modern machine methods be of genuine help in carrying out fundamental mathematical investigations, not directly related to finding numerical solutions to problems of applied mathematics?" The answer is certainly in the positive, though our experience is so far very limited. Perhaps I may first cite an occasion many years ago when in my own work I noted how convenient it would have been to possess a small electrical device for checking the properties of certain finite algebras. I was studying Boolean algebras at the time and wished to discuss the logical independence of certain of their properties. This involved setting up various finite algebras *ad hoc* and laboriously checking the validity or failure of the various properties in question. It seemed evident to me that under pressure of necessity the task could be mechanized and that to this end electrical means would be preferable because they would greatly facilitate the change-over from one algebra to another. There are certainly many situations, of which this one is typical, where the mathematician will find modern machine methods a great boon to him in his work, and some will shortly be cited as instances where such methods have actually been applied to research problems of pure mathematics. Logically, however, it may be more fitting to begin with some interesting results in the theory of numbers which depend only on the increased arithmetic power of the newer machines. Like most of the results which we shall mention here, the computations were performed on the SWAC, a high-speed machine which was built at the University of California in Los Angeles by the Bureau of Standards and which was made readily available to mathematicians interested in using its services. Professor D. H. Lehmer used SWAC to test the primality of various Mersenne numbers, $2^p - 1$ where p is a prime, and found that $2^{2281} - 1$ is indeed a prime. I believe that this number, with nearly seven hundred digits in its decimal development, is still the largest known prime. Professor Lehmer also tested a hypothesis based on a remark of Fermat, finding that the huge number $2^{2^{16}} + 1$ is not prime as surmised. In collaboration with Professor H. S. Vandiver of the University

of Texas, he also verified the truth of a very large number of special cases of Fermat's Last Theorem not yet amenable to theoretical treatment. More recently, Professor Marshall Hall of the Ohio State University, also using SWAC, has carried out the computations involved in verifying the Desarguesian character of the finite projective planes of all orders through 9. This problem is essentially an algebraic problem analogous to the one referred to above. Another program, reported on by Dr. George E. Forsythe, determined all semi-groups of order 4, finding that there are 3492, of which 126 are distinct under isomorphisms and anti-isomorphisms. These results coincide with those obtained independently by Professors Takayuki Tamura and M. Yamamura in Japan, working by hand. Calculations like these are interesting in themselves and they throw a certain amount of light on important problems of number theory, geometry, and algebra; but they are obviously incapable of settling any but finite questions. For example, the machine cannot give a definitive treatment of Fermat's Last Theorem simply by calculating a number of special cases, however large, and so long as the machine does not turn up a failure of the Theorem the most it can do is to strengthen the feeling that the Theorem is perhaps true. It might appear, therefore, that the role of the machine in mathematics may be positive for the case of finite systems but is condemned to be negative in the case of non-finite questions about infinite systems. Logically this is perhaps true, but in the case of one branch of mathematics, analysis, we have a great deal of empirical and mathematical evidence that the structure of infinite systems can often be satisfactorily approximated by that of well-chosen finite systems. In such instances, therefore, we are justified in drawing some tentative conclusions, either qualitative or quantitative within known limits of error, about the infinite system as soon as we have calculated enough information about the approximating finite system. It has been pointed out by Professor von Neumann that we may hope to learn a great deal about the theory of partial differential equations by using modern computing machines to provide ourselves with this kind of approximate empirical information. Certainly, one of our difficulties in studying partial differential equations from a general point of view is that we have very meagre knowledge of the behavior of their solutions, even in a numerical sense, and thus lack guidance in formulating fruitful questions. I am unable to cite any case where the machine has helped us increase our knowledge to the point that it could be used as the base for deeper

theoretical investigations, but I can well imagine that such cases will become part of our mathematical experience as time goes on.

When a mathematician is engaged in making a theoretical study of the properties of a mathematical system, finite or infinite, he aims at establishing general propositions of interesting content by applying in a systematic and appropriate way the basic principles of logic. Thus a machine designed to imitate what the mathematician does would have to be capable not only of carrying out various numerical or algebraical calculations but also of executing logical manipulations. As we have already pointed out, the work of formalizing logic was carried so far that since the fundamental contributions of Russell and Whitehead we have known how to reduce these manipulations to a routine and hence how to mechanize them. The modern high-speed electronic computers could certainly be set the task of carrying out such routines and thereby proving theorems. Thus in principle the only parts of logic and of mathematics which we cannot turn over to machines are those, if any, which somehow escape the reduction to a routine. Since in the present state of the art we can confidently equate the phrase "reducible to a routine" with "mechanizable", all that we have said here emerges as a tautology. We see, therefore, that the problem we are trying to discuss is, when purely practical limitations are put to one side, a problem of logic which does not become precise until we have defined more precisely what we mean by a routine. To go into this matter more fully would involve us in subtleties which would be out of place in an essay of this kind. They are of such a nature that it has taken logicians much time and effort to clarify them and to explore their implications. The work of logicians, particularly the monumental contribution of Gödel, has made it clear that there are indeed aspects of logic and of mathematics which cannot be routinized. In the present context it will be most fruitful for us to examine briefly some results established independently by Professor Alonzo Church of Princeton University and A. M. Turing of Cambridge University. In order to objectify the particular logical problems he wished to treat, Turing described a conceptual machine and determined what it could or could not compute; but Church proceeded without making such an appeal to the concrete. It is of interest to note that Turing's researches were pursued in the thirties, before the modern computational developments were well launched. At that time Turing very likely had no thought that anything resembling his machine would ever be built. As a matter of fact, I believe

that no Turing machine has yet been built, though the marked current interest in it suggests that sooner or later one will be. For a description of a Turing machine, reference may be made to Turing's original paper in volume 42 of the Proceedings of the London Mathematical Society (with a theoretical correction in the following volume). It should be noted that the machine works with an infinite tape, on which it prints symbols and carries out operations of other kinds, thus being equipped in effect with a memory of unlimited capacity. Needless to say, such a conceptual machine transcends the practical limitations we have already noted, and only practical modifications of it can ever be realized. Although Turing described a machine which in this and other respects achieved a maximum of flexibility and versatility, he was able to show that there are real numbers definable in terms of a specified logic which the machine cannot compute; and also that in pure logic, as formalized by Hilbert and his school, there is no standard procedure which could be carried out by the machine and which would suffice to decide the truth or falsity of the formal statements expressible in that logic. Thus it appears that neither the Turing machine nor any modification of it which anyone has yet been able suggest could be expected to replace a skilful mathematician in finding proofs for theorems. In short, a close examination of formal logic brings us back to the thought from which we started—namely, that finite methods are in general insufficient for handling infinite systems—with the additional knowledge that they are in particular insufficient for handling the infinite systems of classical logic. This outcome may serve to remind us that a good many years ago the great mathematician Poincaré advocated the point of view that mathematics cannot be fully formalized. I believe that the intuitionists, led by Brouwer, entertain a similar point of view towards both mathematics and logic.

If these considerations do indeed point to the conclusion that there is a gulf between the mind and the machine, we are nevertheless well aware that in many respects the mind does function mechanically. In consequence we see that we cannot leave these questions without trying to square whatever we have said about the nature of mind with whatever is known about the physiology of the nervous system. The illuminating studies of McCulloch and Pitts have shown how certain known features of the nervous system enable it to function as a logical machine, which could be simulated by an electronic device. A very careful discussion of the logical power of such a machine has been given by Professor

S. C. Kleene of the University of Wisconsin in a paper published in "Automata Studies" (Princeton University Press, 1956; C. E. Shannon and J. McCarthy, editors). He finds that the machine is subject to certain explicitly determined limitations, and remarks that it is necessarily less versatile than a Turing machine because the latter has the advantage of using an infinite tape, as mentioned above. In their theory of nerve nets, McCulloch and Pitts suggested a possible mechanism to perform the function of a memory, but the best evidence shows that human memory cannot be explained in terms of it. As a matter of fact the nature of memory remains essentially a mystery. The same may be said, of course, about other functions of the mind. In an effort to understand some of these, such as learning, much thought has been devoted in recent times to the design of mechanisms which might simulate various features of human behavior. Several such conceptual machines are described in the volume "Automata Studies" mentioned above. While these descriptions demonstrate that a great many different aspects of the mind may well be explainable in mechanistic terms, and suggest that some, at least, will eventually be so explained, they clearly leave unaltered the position which we have reached in this discussion—namely, that in its logical faculties, and probably in others as well, the mind transcends the machine.

Our exploration of the subject of computation, practical and commonplace though it may have seemed at the outset, has thus brought us to a point where technology, mathematics, logic, physiology, psychology, and philosophy can be seen to meet and to interact. We have seen how success in solving the simpler and more immediate problems raised in the art of computation has opened up deeper and more intriguing problems in these various disciplines, how the dissipation of the more superficial difficulties and mysteries has disclosed others of a more fundamental nature. What we have thus observed in the case of computation seems to me to be somehow characteristic of the stage of knowledge we have now reached. Perhaps this is an illusion created by my own personal experience of science and learning, but I do not think it is. Wherever we may look in the natural sciences today we see the merging and interpenetration of disciplines which twenty-five or fifty years ago either did not exist at all or did not touch one another; and we see that the problems we want to solve become more complex and profound as we master more aspects of reality. The nineteenth century dream of a closed system of knowledge has had to be

abandoned for the adventure of working in what appears instead to be an open system with infinite possibilities. Nowhere is this clearer than in the discipline we have essayed to examine here, or than in the modern physical theories of the constitution of matter, but our point can also be illustrated in the case of biology by reference to recent work on the nature of viruses. In physics the study of the elementary particles—electrons, positrons, protons, mesons, neutrons, and so on—has reached a critical stage where a serious maladjustment between physical facts, mathematical tools, and working metaphysical principles prevents the evolution of a satisfactory theory of microscopic events. In biology the pioneering work of Professor W. M. Stanley and his associates on the viruses has brought us to the mysterious borderline which separates living from inert matter. In his laboratories at the University of California in Berkeley the recent synthesis of a virus from apparently inert components is but the latest in a series of remarkable discoveries showing how finely Nature has drawn this line. On these widely separated frontiers of science we thus find that our deepening knowledge invests with new meaning and new relevance the ancient questions of philosophy—"What is matter?", "What is energy?", "What is life?", "What is mind?" Far from being alien to the proper concerns of science, these questions are inseparable from its essential fabric, and are not without significance even in its practical applications.

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Fifty Years of Chemistry in Indian Universities

BY

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Fifty years of chemistry in Indian Universities is a theme of surpassing interest for a study in retrospect.

My study of chemistry was initiated in the Calcutta University 52 years ago and I have now had the privilege of teaching the subject in two of the premier universities in India, viz., the Universities of Calcutta and Madras. It is an unique experience and it seems appropriate that on the occasion of the centenary celebrations of the old Universities, I should relate some of it and try to present a picture of the conditions of scientific education which prevailed in our universities fifty years ago.

Science was mostly neglected in our schools in the early years of this century. The subjects which were taught compulsorily for all in the Entrance Examination of the Calcutta University, i.e., the examination at the end of the school final course, were English, Sanskrit or French or Latin or any other foreign language, Mathematics, including arithmetic, algebra and geometry, History of England and of India, and Geography. In the paper in Geography one or two questions in Elementary Science were set from a small text book prescribed for study in the final year of the course, but the subject, as far as I remember, was seldom taught seriously in the class, and there was no provision in the schools of a laboratory for experimental work by the students. Chemistry as a distinct subject first appeared in the curriculum of studies for the so-called First Examination in Arts equivalent to the Intermediate Examination of the present day. The course included English and Sanskrit or a foreign language, History of Rome and Greece, Logic, Physics, Chemistry and Mathematics which included Trigonometry and Conics. All these subjects were compulsory except in the case of Logic which could be substituted for Roman and Greek History. There was no practical examination in the sciences at the end of the course and no laboratory was provided for the

students to do any experimental work. The lectures too were illustrated mostly with chalk and blackboard and it was only on rare occasions that some simple experiments were performed in the class, such as the determination of the composition of air by burning phosphorus in a closed volume of air, or the separation of sulphides of metals from the solutions of their salts by passing sulphuretted hydrogen gas from a Kipp's apparatus. These were red-letter days for the students who eagerly awaited the experiments and watched their performance with keen interest and excitement. The B.A. degree course in the Calcutta University was grouped into "A" course comprising humanities and "B" course consisting of science and mathematics. Even in the latter course, laboratory work was not compulsory, and the candidates were not subjected to a practical examination at the end. It was in the B.Sc. degree course which was opened by the Calcutta University in the year 1902, that practical work in science was first introduced. The only college affiliated for this course at that time was the Presidency College, Calcutta, as the other private colleges in the city and outside were unable to satisfy the conditions of having the requisite well-equipped laboratories and competent staff for the purpose.

The B.Sc. degree course comprised Physics, Chemistry, Mathematics (Statics, Dynamics, Hydrostatics and Astronomy) and either Botany and Physiology or Geology and Mineralogy and there was a good practical examination in each science at the end. The private colleges teaching science even up to the standard of the B.A. degree had no laboratories worth speaking of, and most of their scientific apparatus were stored in a few almirahs kept in the lecture theatre. It was only after some twelve years, i.e., towards the middle of the second decade, that some of these colleges in the city began to equip themselves with the necessary laboratories and staff and seek affiliation to the B.Sc. degree course.

The teaching of Chemistry even in the highest standards was confined in those days almost entirely to the Inorganic branch of the science and little attention was paid either to Organic or Physical Chemistry. The practical course, too, consisted of only inorganic analyses, both qualitative and quantitative, the acme of perfection being supposed to have been reached by a student who could detect the presence of as many as twelve different radicals in a mixture of salts, and estimate accurately the quantities of two or three different metals present in an ore or an alloy. Organic preparations and analyses were unknown, and there was no proper syllabus

drawn up for experiments in Physical Chemistry. The little Physical Chemistry that was taught was in connection with the theoretical portions of Inorganic Chemistry.

The M.Sc. degree course was first instituted in the Calcutta University in 1908. It was a very comprehensive course in which one of the branches of Chemistry, either Inorganic or Organic or Physical, had to be offered by the candidate as his special subject and the other two branches as subsidiary. A heavy syllabus both in theory and practical was drawn up and the list of text books prescribed for study included almost all the advanced text books that had been published at the time. The two volumes of Advanced Organic Chemistry by Cohen and the series of books in Physical Chemistry edited by Sir William Ramsay as well as Soddy's Radioactivity and Lehfeld's Electrochemistry were to be found in this list. While the standard aimed at for the Master's Degree in Science was reasonably high, little provision was made for imparting effective instructions to the students on the high standard prescribed. Indeed, until comparatively recently, the universities in India did not take any active part in teaching, and were content merely with laying down curricula and conducting examinations. The teaching was provided exclusively by the colleges affiliated to the universities. In this respect, the Presidency Colleges in Calcutta and Madras, as Government institutions, held special positions in those days. They had well equipped laboratories and highly qualified staff, enabling them to provide instructions in the different sciences up to the highest standards.

The two members of the higher staff in science in the Presidency Colleges in Calcutta and Madras respectively, who were brought out from England as early as 1875, were Professor (Sir) Alexander Pedler and Doctor W. H. Wilson. The former was a student of the old Royal College of Science in London and Prof. Wilson appears to have had his training under Bunsen in Heidelberg.

Both these professors belonged to the old school and were brought up in the traditions of teaching only inorganic or mineral chemistry. They had both been known to be excellent teachers in their way and to have taken great pains in preparing their experiments for purposes of lecture demonstrations. Many of the apparatus set up by them are still preserved in the chemical laboratories of the Presidency Colleges of Calcutta and Madras, and bear testimony to their excellence as teachers and their skill

in experimentation. Professor Pedler's tradition was continued by Acharya (Sir) Prafulla Chandra Ray who had studied chemistry in Edinburgh under Professor Crum Brown. He confined his teaching and research also mostly to the inorganic branch of the science. In Madras, Prof. W. Erlam Smith who succeeded Dr. Wilson in 1907, also carried on the same tradition of teaching inorganic chemistry. As a consequence, there was a great dearth fifty years ago of teachers who could sponsor the exposition of the other two equally important branches of organic and physical chemistry. The writer had the distinction to belong to the first batch of students who prepared for the then new M.Sc. degree examination of the Calcutta University. The students numbering about six were largely left to themselves to learn these subjects both in theory and practice, but the way in which they combined to a well-knit team for the purpose of teaching themselves and overcame their many deficiencies and difficulties particularly in the guidance of their experimental work, remains to this day a vivid memory and an unforgettable experience. The knowledge gained by the students in this way left an abiding impression which was of invaluable help to them in later life. Notwithstanding the defects of teaching pointed out, the knowledge of theoretical science acquired by the Indian student was not inferior to that of a good honours graduate of a British university; his real weakness lay in his acquaintance with practical science in which he found himself at a disadvantage. This fact impressed itself on the mind of the writer when he went for higher studies to London in 1911 after taking his Master's Degree in science and doing post-graduate research in the Calcutta University for a year. While he acquitted himself with credit in theory, the confidence and ease with which his British fellow-workers did their experimental work filled him with envy. It was not perhaps without reason, therefore, that the old universities in Great Britain did not consider the standard of the degree examination of an Indian university to be any higher than that of their Matriculation examination, and refused to grant the Indian graduate, even though the latter might possess the Master's degree, exemption from their Intermediate examinations. It was only as a special case that the writer, by virtue of his published researches and an exacting preliminary test, was permitted for the first time to register himself as a candidate for the higher degree in science (D.Sc.) of the University of London without passing the examinations for the lower degree. The reasons for the real superiority of the English student in his practical equipment can be readily understood when it is

remembered that the majority of the students of science there had begun their careers in school where they had well equipped laboratories where experiments formed the basis of their scientific training, whereas in India no scientific education worth the name existed in the schools and no facilities were available for the students to engage themselves in experiments. It is very necessary that science teaching in our schools is raised to a healthy level without which our efforts to stimulate and encourage the development of science in the universities and technological institutes may be largely frustrated.

I have tried to give a general idea of the way in which chemistry was taught in our college days early in this century. Although the phenomena of radioactivity had been studied and understood, the Daltonian conception of the atom as the smallest indivisible particle of an element still continued to all intents and purposes to be the basis of teaching chemistry, and Mendeleeff's Periodical Law of classification of elements based on atomic weights as their unalterable physical constants, continued to serve as the sheet-anchor of the chemist. The new conception of the structure of the atom and its consequences on atomic numbers and isotopes and the electronic theories of valency, etc., did not seriously invade chemistry in our undergraduate days. The old traditions of teaching began slowly to crumble only in the second decade of this century and gave way to the new ideas making chemistry more and more of an exact science and less of empiricism. In our universities, too, the old order changed slowly and the subjects of Organic and Physical chemistry were assigned their legitimate places of importance in the scheme of general education. The improvements which occurred mostly during the second decade of this century were helped by several circumstances which may be worth recording. In the first place, the University of Calcutta, through the energy and inspiration of its great Vice-Chancellor, Sir Asutosh Mookherjee, changed from an examining to a teaching University. This was made possible primarily through the munificent benefaction of two of India's great lawyers, Sir Rashbehari Ghosh and Sir T. Palit, who gave their life's savings for the purpose of establishing a college of science and technology for the University. A post-graduate council in science was established next, and teaching and research for the Master's and other higher degrees in science was undertaken by the University in co-operation with the Presidency College and some of the other bigger colleges in the city. This marked the beginning of teaching and research

in science by an Indian university, and other universities were not slow to follow the example of the University of Calcutta. Secondly, some of the British professors who came out to India at this period, of whom the names of Dr. J. L. Simonsen (Madras), Dr. E. R. Watson (Dacca) and Dr. A. N. Meldrum (Bombay) may be mentioned as most prominent, were all specialists in Organic Chemistry, and vigorously pursued teaching and research in this branch in the universities to which they were attached. As regards Physical Chemistry, a number of young men who had offered this as their special subject of study for their Master's degree in science acquired a real taste and aptitude for post-graduate research in this field. Several of them went abroad for higher studies in Physical Chemistry and took up, on their return, teaching and research in our universities. Among those may be mentioned the names of Professor Nil Ratan Dhar (Allahabad), Dr. J. C. Ghosh (Dacca and Bangalore), Dr. J. N. Mukherjee (Calcutta) and the late Dr. S. S. Bhatnagar (Banaras and Lahore). There were two other events which must be considered here, as they had an important bearing on the development of chemical research in India. The first All India Institute for research was established in Bangalore about 1912, thanks to the great vision of the late Mr. J. N. Tata. Sir William Ramsay was brought out to India to choose a site for the Institute and draw up a programme of research and he selected Bangalore as the most suitable, and deputed later his co-worker, Dr. Morris William Travers as its first Director. With Travers came several other distinguished scientists such as Dr. J. J. Sudborough, Dr. Gilbert J. Fowler, Dr. H. E. Watson, Dr. Alfred Hayes and others. The presence of these men in the country was undoubtedly responsible for ushering in an era of great change for the better in the attitude of our universities to advanced studies and research in science. The other important factor was the inauguration of an annual Indian Science Congress, the first session being held in Calcutta in 1914. The annual meetings have been held up to the present day in various different university centres where men engaged in scientific and technological research in different parts of India have opportunities of meeting and discussing the results of their work in special fields and comparing notes.

The writer had opportunities of attending several of these meetings since 1916, and can testify to the great stimulation of thought and endeavour which these meetings created among the young scientists of our country.

I joined the Presidency College, Madras, in the summer of 1920, after serving the Calcutta University for nearly four years as Professor of Chemistry in the Presidency College, Calcutta, and participating in the activities of the newly started post-graduate council in science of the University. The conditions of teaching and research prevailing in Madras at the time were found to be not dissimilar to those in Calcutta during my college days (1906-1910). The only course in science offered by the University of Madras was that of the B.A. Degree. This was divided into two groups which were called the II-A and the II-B, respectively. Physics was the main and chemistry was the subsidiary subject of study in the II-A, and vice versa for II-B. The standard of main chemistry in the II-B group was comparable to that of the old B.Sc. degree of the Calcutta University, but the standard of subsidiary chemistry in II-A was only a little higher than that of the Intermediate Examination. Honours courses in chemistry were not introduced till as late as 1911, and the Presidency College remained for thirty years and more the only college in the whole province which was permitted to teach this course in chemistry. The usual practice in those days was to admit about four students in the Honours classes in a year, and as it was a three-year course, the maximum number of students at any time in this course rarely exceeded a dozen. The best students of the year competed for admission to the Honours classes in chemistry, and it thus became possible to maintain a high standard for the Honours degree in Madras which approached that of the M.Sc. degree of the Calcutta University. The B.Sc. degree course (Pass and Honours) which did not differ materially from the old B.A. —II-B course except in the enlarged syllabus for practical work, was not begun till 1932, and the Presidency College as usual was the pioneer in being affiliated to the new course, while the other colleges followed suit gradually. The M.Sc. degree course was introduced simultaneously and was awarded entirely by research on the merit of a thesis submitted by the candidate. As a rule, only Honours graduates were accepted for this course and there were not more than one or two candidates in a year for this degree.

In Madras no tradition had been built up for research either in physics or in chemistry, until Dr. Simonsen joined the Presidency College in 1910. The Calcutta University was more fortunate in this respect. The two professors, Sir J. C. Bose and Sir

P. C. Ray, who adorned the chairs in Physics and Chemistry respectively in the Presidency College, Calcutta, during the years 1890 to 1915, developed lines of research of their own and trained students for research. Professor W. H. Wilson who taught chemistry in the Presidency College in Madras from 1889 to 1907 and Professor W. Erlam Smith who succeeded him do not appear to have seriously engaged themselves in research or encouraged students to do so. Dr. Simonsen who acted as additional professor of chemistry during the years 1910 to 1918 engaged himself in researches in organic chemistry and made valuable contributions in the chemistry of terpenes which he made his special field for study. He also appears to have created enthusiasm for research among his students, some of whom, like the late Mr. M. Gopala Rao, who died unfortunately very early in his career, gave brilliant promise for research. Researches in organic chemistry have been continued to the present day and nearly a hundred publications of original memoirs have resulted from the researches in the course of the past thirty years. The problems ranged from synthetic chemistry to those of natural products and enzymes and about a score of students at different times have taken part in the investigations. During the past decade, the Council of Scientific & Industrial Research in Delhi had subsidised two distinct lines of research in the chemical laboratory of the Presidency College, Madras, viz., (a) the chemistry of gland products for which the raw materials were supplied by the slaughter houses, and (b) the electro-synthesis of dye-intermediates. Two teams of researchers had been engaged in these investigations and their results, published in many scientific journals, have attracted the attention of well known workers in these fields outside India. Many of these students have now made their mark in the world of research and some have attained positions of eminence in academic life. The present Chief Professor of Chemistry in the Presidency College is himself a product of this school and is now vigorously pursuing researches in organic chemistry.

From the rather gloomy retrospect with which I began this article I come now to a period when India is emerging into an era of glorious prospect in matters of research and advanced studies in pure and applied science. Where there was sometimes not even a single scholarship available for chemical research in a university or a province, we have now a multiplicity of scholarships and fellowships provided for the purpose. For these, not only are the funds of the university available, but liberal grants are also being

made by the Ministry of Education at the Centre, by the Council of Scientific & Industrial Research, by the National Institute of Sciences of India and even by private commercial firms like the I.C.I and the Burmah Shell Company. There is almost a glut of these research fellowships and scholarships, and the problem is not so much one of lack of funds as of their utilization to the greatest advantage to the country. People are beginning to view with anxiety not only the increase at times in the duplication of research, but also of spurious work masquerading in the name of research. There is great need to-day for the creation of healthy public opinion in matters of science and research and of vigilance for the prevention of waste. There should be better and more intensive co-ordination of efforts made in the various research establishments such as the universities, the research institutes and the dozen National laboratories of India of which no less than seven are being devoted to problems of different aspects of chemical research.

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Biological Study of Mechanism of Carcinogenesis—I

BY

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The answers to the questions of *why* and *how* things happen are the keys that unlock the doors of all scientific mysteries, for they enable one to understand *the processes* instead of being satisfied with *the facts*—Berenblum (2).

Cancer is essentially a problem of abnormal growth; abnormal in the sense that it appears purposeless and devoid of a self limiting propensity. One of the usual characteristics of cancer is the transformation of some cells from a relatively static to a rapidly proliferating condition. Almost all cells of the body, with few exceptions, are endowed with a capacity for division and multiplication and are, therefore, possessed of an inherent potentiality for persistent or malignant growth. The mechanism which initiates and sustains this unusual property has intrigued many investigators and has interested us for some time. Much research in other countries had suggested that the causation of cancer in experimental animals depended on an interplay of cancer producing stimuli and a susceptibility of a particular tissue to react to them. The susceptibility was mainly determined by the genetic constitution of the individual. The stimuli for any particular type of cancer could either be elaborated within the body (intrinsic) or could operate from sources outside it (extrinsic). The problem before us, therefore, was to unravel the tangled skein of fact and theory which had been wound round during the last quarter of a century of cancer research.

A study of the mechanism of carcinogenesis in experimental animals was, therefore, started in our laboratories by Ranadive and has been continued by her, with a batch of young colleagues for the last ten years. It is now intended to review briefly the work which has been carried out in our laboratories without entering

into details of microbiological techniques or commenting upon the statistical significance of experimental data. We shall limit ourselves to the general object, design and sequence of experiments and refer occasionally to an interpretation of some of the salient observations.

Two types of cancer were selected for our study.

1. Cancer of the breast (spontaneous and induced).
2. Cancer of the skin (induced by chemical substances).

This review is confined to the work on the cancer of the breast in mice.

For the type of work we wished to undertake it was essential to determine certain objective criteria to commence with and later to assess the role of the various intrinsic and extrinsic factors in the production of this particular type of cancer in mice. The problem was, therefore, attacked along three different lines of approach:

(a) Evaluation of the role of intrinsic and extrinsic factors in the production of tumours.

(b) Behaviour of the tumour in a new host on transplantation.

(c) Behaviour of tumour cells in tissue culture.

In planning our experiments it was decided to work with genetically pure strains of mice, with known incidence of spontaneous breast cancer, to study the morphology of mammary glands at stated periods of their normal life and to investigate the naked eye and microscopic structure of the glands, the vaginal mucous lining and the endocrine glands, particularly the ovaries, adrenals and the pituitary. For the purposes of our experiments four pure strains of mice obtained from the U.S.A. and bred by brother-sister mating over a period of several years, in our laboratories, were mainly used. One other strain LC (Paris) obtained from Europe and later maintained as a separate line L (P) was also employed for certain special studies and will be referred to later in the text. The following table summarises some relevant information concerning the strains.

TABLE 1.

Strain	Abbreviation used in the text.	*Etiological Factors* Heredity	Hormonal Factor	Milk Factor	Reported present tumour incidence from the laboratories of origin. Virgins	Breeders
1. C_3H (Bar)	C_3H	P	P	P	95-100	95-100
2. dba (Bar)	dba	P	P	P	55	55
3. Strong A	A	P	A(V) P(B)	P	4-5	83
4. C_{57} (Bl)	C_{57}	A	A	A	—	0·5
5. L. (Paris)	L(P)	P	A(V) A(B)?	A	—	—

P — Present A — Absent V — Virgin B — Breeders

Cancer of the breast happens to be the commonest type of spontaneous tumour in laboratory mice and for that reason is probably the most thoroughly investigated subject in experimental studies on cancer. As a result of much work, it had been established that three intrinsic factors are mainly responsible for the development of breast cancer in mice. They comprise of a genetic susceptibility, hormonal stimulation and a milk-borne tumour agent (MTA), an ultra-microscopic particulate material having the chemical properties of a nucleoprotein and possessing characteristics suggestive of a "virus". It needs to be stated that the genetic factor "cannot be considered as the cause of the development of the tumour, but rather as the determinant of what the tissue could or could not do when submitted to the cancer producing agent" (2).

The first series of our experiments was designed to study the morphology of mammary glands by whole mount preparations in males and females at different age periods in the four strains of animals, and to identify differences in the architecture of their mammary glands. These studies formed the basis of reference in all our later investigations. The developmental differences in the breast glands of strains C_3H, A and C_{57} revealed a positive cor-

relation between the structure of the mammary ducts and acini and the degree of the susceptibility of mice to cancer. Precancerous lesions in the form of focal acinar clusters were seen only in the highly susceptible strain C_3H and were absent in the mammary glands of virgins in the other two strains (18).

The complement of the three intrinsic factors in the three strains of mice and the incidence of breast cancer in them differ markedly. It was, therefore, considered advisable to study the action of these factors individually. In this next series of experiments, about six hundred mice belonging to these three strains were used.

For the purposes of studying the comparative response of the mammary glands to sex hormones, a large number of mice of the cancer susceptible and cancer resistant strains were gonadectomised soon after birth and their mammary glands examined when they were 5 to 7 months old, after an identical treatment with sex hormones. With such treatment the mammary gland morphology as well as the urogenital reaction presented significant differences. The pattern was determined by the amount of the gonadal hormone which was administered as well as, by the reactivity of the mammary tissue in the particular strain of mice. The tissue susceptibility appeared to be strain specific and was mainly influenced by its genetic constitution (21).

The action of the milk-borne tumour agent (MTA) in mammary carcinogenesis was studied by foster-nursing new born mice of each of the three strains on mothers of the other two strains. The litters were exchanged within 17 hours after birth and the effect of cross suckling was traced in adults when they were 7-8 months old. These experiments showed that foster-nursing distinctly altered the morphology of the mammary structure in the direction which was characteristic of the foster mother (15).

The part played by these three factors in the production of cancer was studied in the highly susceptible strain C_3H and the one resistant to it, C_{57}. The new born mice were ovariectomised soon after birth and foster-nursed on the mothers of the opposite strain. When they reached maturity they were divided into two groups. A pellet of estradiol, approximately weighing 0·9 mg. was introduced subcutaneously in mice of one group whereas the other received a balanced dose of 0·2 mg. estradiol and one mg.

testosterone. At the age of five to seven months the mammary pattern and the urogenital response was compared in these experiments, where the hormonal factor and the MTA were kept under control. It was found that the genetic susceptibility had little or no effect on the mammary pattern. The breast glands of C_{57} mice nursed on C_3H milk showed the same extent of acinar proliferation as the C_3H mice nursed by their own mothers. On the other hand, C_3H and C_{57} mice which were reared on C_{57} milk presented only bare mammary ducts. It became, therefore, evident that the ductal pattern was determined by the hormonal influence and the acinar proliferation (indicative of abnormal growth) by the milk-borne agent (MTA). Our observations indicated the dominant role of the MTA over the other intrinsic factors in the development of mammary cancer in mice and confirmed the opinion that the nodular overgrowths which had been commented upon by other workers were definitely precancerous in nature. It could also be concluded that although the hormonal stimulation was necessary for an orderly development of the breast tissue, the agent transmitted through the milk in the early stages of life acted invariably as the initiating, as well as, the promoting factor for this particular type of cancer in mice (22).

Many workers have repeatedly considered heredity, hormones and the MTA as separate factors working along three parallel lines, each equally important and designated as 'mammary tumour preparatory influence or inciter' (3). The genetic constitution of the animal occupies a more fundamental position, at the base of the three pathways involving the milk factor, the hormonal stimulation and the susceptibility of mammary tissue. It controls the potency of the milk factor and the degree of hormonal stimulation (11). However, our experiments suggested that the milk factor exerts a definite influence on the general metabolism of experimental animals.

The influence of the MTA on the hormonal metabolism of the animal was first indicated in our preliminary observations on a group of 'intact' foster-nursed virgins. During the course of these experiments, it was often observed that the female mice of the cancer susceptible strain C_3H when foster-nursed on the milk of the resistant strain continued in a state of long diestrus, whereas, the resistant strain females fostered on the milk of the susceptible strain remained in a long estrus. These findings invited a systematic study of the ovaries of the foster-nursed mice. It was noticed

that foster-nursing significantly altered the number of corpora lutea in the ovaries (20).

TABLE 2.

Strain	Total number of females studied.	Number of corpora lutea per largest section available (average)
Group I:		
C_3H/C_3H	10	$10{\cdot}0 \pm 1{\cdot}24$
C_3H/C_{57}	12	$4{\cdot}5 \pm 0{\cdot}47$
Group II:		
C_{57}/C_{57}	12	$4{\cdot}4 \pm 0{\cdot}48$
C_{57}/C_3H	13	$9{\cdot}6 \pm 0{\cdot}66$

The letters following the / indicate the strain of foster mother and those preceding it the parentage of offspring, thus C_3H/C_{57} indicates a C_3H new born foster-nursed on C_{57} mouse.

The striking differences in the number of corpora lutea of the cancer susceptible and the resistant strains had been reported by other workers. The average number of corpora lutea in C_3H decreased from 10 to 4·5 after foster-nursing on the C_{57}—while the average of 4·4 in the C_{57} increased to 9·9 when nursed on the C_3H milk. The number of corpora lutea is indicative of the degree of hormonal stimulation in an animal. Our observations demonstrated for the first time a definite change in the maturation of ovaries—indicating a change in the hormonal stimulation—after foster-nursing. This, to our mind, was a finding of some considerable importance and invited further experimentation for confirming our hypothesis. A third series of experiments was, therefore, designed to study the mode of action of the milk-borne agent. Four hundred mice of four strains were utilized for these experiments. The milk-hormone relationship was studied both in 'intact' virgins and in ovariectomised mice. Bittner, during the course of his extensive studies on 'milk factor', had suggested that a comparison of endocrine organs of two lines of a strain, one possessing the milk factor and the other devoid of it, would constitute the simplest method for testing the influence of the milk borne agent on

the endocrines. Three lines of strain dba (Bar), dba (new) possessing the milk factor, dba (—MTI) a line free of the milk factor raised from full term foetuses removed from the uteri of young breeders of dba (Bar) and fostered on C_{57} Milk, and dba (old), another old line of dba (Bar) presumed to have lost the milk factor during years of intense inbreeding, were used for these studies. A comparative study of the maturation of ovaries in these three lines presented significant differences in the average number of mature corpora lutea. Under the conditions of our experiment such differences could only be attributed to the presence or absence of the milk-borne agent (MTA) in the three lines. The mammary pattern also presented corresponding differences, so that hyperplastic nodules were present only in the first line and absent in the other two. Such observations, specially the gradation of differences in the three lines of dba, confirmed the view that although the hormonal factor is primarily conditioned by the genetic constitution of the animal, the virus received through the milk exerts a secondary influence on the hormonal metabolism. We were, therefore, led to the conclusion that the milk-borne agent probably acts on the mammary tissue in two ways, directly on the glandular epithelium and indirectly through its action on the ovaries leading to a more intense luteal stimulation (26).

If the milk factor acted through the agency of the gonads, it was felt that it may be worth investigating the mode of its action in the absence of ovaries. The first step in this direction was to follow the effect of gonadectomy in the three strains of mice C_3H, C_{57} and A. New born mice of these strains were castrated within 36 hours of birth. It was observed that such mice from C_3H strain at the age of 8 months showed continuous keratinisation of the vaginal mucosa, a profuse multiplication of mammary ducts and a marked adrenal hyperplasia. This manifestation of hyperactivity of the adrenal cortex and an estrogen-like stimulation of secondary sex organs was completely absent in the gonadectomised mice of the other two strains, viz., C_{57} and A. It was, therefore, assumed that in the absence of the gonads a well regulated hormonal brake was not being applied to the relevant activity of the pituitary gland and culminated in an excessive secretion of pituitary gonadotrophins. If the tissue is genetically susceptible, the increased gonadotrophins act on the adrenal cortex and induce cortical hyperplasia. An "inherited hormonal influence" is probably associated in some manner with the adrenal tissue susceptibility to pituitary gonadotrophins (4, 5, 12). A susceptibility to

spontaneous development of mammary cancer does not necessarily induce a response in the adrenal cortex, as the mice of A strain which are genetically susceptible to breast tumour formation are lacking in the "inherited hormonal influence" and do not react by adreno-cortical hyperplasia after gonadectomy (25).

The next step was to study effect of the milk-borne agent (MTA) on the endocrine system after gonadectomy. The new born mice of C_3H and C_{57} after castration were foster-nursed on the milk of the opposite strain. The next table shows the results of the experiment.

TABLE 3.

Strain	Total Number of Mice.	*Number of Mice with* Adreno-cortical Hyperplasia.	Normal Ductal Development.	Calcium Deposits.
C_3H/C_3H	16	14	11	7
C_3H/C_{57}	14	6	13	6
C_{57}/C_{57}	12	—	4	1
C_{57}/C_3H	12	6	6	9

One other striking observation deserves particular mention. The adreno-cortical hyperplasia in C_3H strain of mice is invariably associated, not only with a profuse burgeoning of mammary ducts, but also with a heavy deposition of calcium salts in the stroma, in which the ducts lie embedded. The phenomenon of calcium deposition in stroma of C_3H mice was observed and reported for the first time by Ranadive in 1949. We have been interested in this finding and the influence of gonadectomy and foster-nursing has been investigated in some detail in our laboratories since that time (25). Adreno-cortical hormones were known to have a depressant action on bone formation. Adrenal hyperplasia or an excessive stimulation of adrenal cortex in gonadectomised animals probably sets up a steroid imbalance and may lead to a decalcification of bones followed by a deposition of calcium salts in the vascular stroma of the breast and the interstitial tissue of the kidneys.

The adrenal response to gonadectomy appears to be strain specific. There was no appreciable change in the adrenal response, breast development or calcium metabolism in the castrates of the C_3H mice after foster-nursing on the C_{57}. However, C_{57} castrates foster-nursed on the C_3H milk, presented adreno-cortical hyperplasia and calcium deposition in more than half the number of animals (17). The agent received through the milk might have some effect on the adrenals through excessive gonadotrophic stimulation. The direct effect of the agent on the pituitary is at present being investigated. We are of the opinion that these experiments have thrown some light on the mode of action of the milk-borne agent. The agent appears to have a two-fold action—partly an 'initiating' action directly on the breast tissue and partly a growth promoting action by stimulation of the endocrines.

Several spontaneous breast tumours in mice are known to have originated in different laboratories, in lines of mice which were free of the milk factor, particularly after forced breeding. This would not militate against the assumption that although the milk factor is probably responsible for 'initiating' the formation of these tumours, the "promoting stage" of the disease is mainly maintained by an excessive hormonal secretion which was partly induced by the agent itself.

As a continuation of our studies on the 'intrinsic' factors of mammary cancer it was essential to investigate the role of 'extrinsic' factors, such as chemical carcinogens in the production of mouse mammary carcinoma. Although American and European workers (1, 16, 6, 13, 14, 7) have studied in great detail the different aspects of chemical carcinogenesis in breast tumours, the manner in which the chemical substance induces the development of cancer, is still not very clearly understood. A systematic study was, therefore, planned to investigate the action of methylcholanthrene (MCA) in different strains of mice, varying in their content of known 'intrinsic' factors. The main problem was to ascertain if a chemical carcinogen like methylcholanthrene could replace any or all the intrinsic factors, in the production of breast cancer and later to study the pathways along which it acts (10).

Two lines of dba (Bar) besides the two strains A and C_{57} were selected for these experiments. The skin of one hundred female mice was painted twice a week with a 0·25 per cent methylcholanthrene dissolved in thiophene free benzene, at an age of 2½ months. The paintings were continued till the animals looked too

weak, either because of the development of breast tumours or skin lesions. The females were kept either as normal or forced breeders. After autopsy, whole mounts of mammary glands, and histological sections of breast tumours, ovaries and adrenals were prepared for examination. The following table summarises the results in different groups of experimental mice.

TABLE 4.

Strain	Type of Breeder	Number of Animals.	Animals showing palpable breast tumours and microscopic cysts in the breast.	*Ovaries.* Average No. of Corpora Lutea.	Early granulosa cell lesions.
Dba (new)	Painted Breeders	16	10/16	5·34 ± 0·28	7/16
	Control Breeders	10	2/10	10·31 ± 0·44	0/8
Dba (-MTI)	Painted Breeders	20	13/20	9·9 ± 0·30	7/20
	Control Breeders	6	0/6	6·24 ± 0·08	0/6
A	Painted Breeders	37	1/37	2·51 ± 0·15	7/37
	Control Breeders	6	1/6	2·16 ± 0·66	0/6
C_{57}	Painted Breeders	10	0/10	5·72 ± 0·49	3/10
	Control Breeders	4	0/4	3·12 ± 0·62	4/4

Similar studies were also carried out on another new strain of white mice L (P) (24) which was imported from Paris in the year 1948. Even on intense breeding this strain had never developed a single spontaneous breast tumour in our laboratories and had not evinced the presence of the milk-borne agent on biological testing. The inherited hormonal influence is lacking in the virgins of the strain and hormonal stimulation is not adequate to induce hyperplastic nodules in the mammary glands of the breeders. A complete biological study of the mammary glands and endocrine organs at different age periods presented a picture closely comparable to that of the cancer resistant strain C_{57}. The mammary glands developed bare mammary ducts without any acinar proliferation in young and old virgins and breeders. The strain presented marked deficiency of the 'luteal factor'. The ovaries exhibited a typical picture of an organ very deficient in luteal stimulation, as the average number of corpora lutea was particularly small.

The virgins as well as breeders of strain L (P) were given bi-weekly paintings of methylcholanthrene from maturity till death. The following table summarises the results of such treatment.

TABLE 5.

Strain L(P)	Number of Animals.	Number of animals showing a palpable breast tumour and microscopic cysts.	*Ovaries.* Average No. of Corpora Lutea.	*Ovaries.* Early granulosa cell lesions.
Painted Virgins	8	0/8	0·43 ± 0·18	0/8
Control Virgins	10	0/10	1·20 ± 0·20	1/10
Painted Breeders	26	17/26	4·28 ± 0·09	14/26
Control Breeders	22	0/22	2·04 ± 0·18	2/22

On comparing the results on four strains of mice the following facts stand out—(i) In the dba possessing all the three intrinsic factors, methylcholanthrene only accelerated the tumour production. (ii) In the dba (-MTI), methylcholanthrene induced the same proportion of tumours as in the last group, dba with many multiple tumours. It was noticed that the tumours appeared slightly later. It could, therefore, be concluded that the milk-borne agent was not required for chemical carcinogenesis. The MCA breeder group developed the largest number of multiple tumours and showed a significant increase in the number of corpora lutea as well as granulosa cell lesions. (iii) The genetically susceptible strain A after being subjected to excessive hormonal stimulation of forced breeding failed to develop palpable tumours in any of the treated animals except one. Although the strain possessed the genetic susceptibility for tumour production, the luteal stimulation was probably insufficient for developing tumours after treatment with MCA. This is also suggested by the appearance of the ovaries. (iv) In the C_{57} with an absence of all the three intrinsic factors, hormonal stimulation by forced breeding was insufficient to induce cancer even after MCA painting.

The observation on the dba and A confirm the view that although genetic constitution determines the tissue susceptibililty to carcinogen treatment, the susceptibility to spontaneous breast cancer and the susceptibility to carcinogen induced tumours are two different conditions. The milk-borne agent has no essential

role in chemical carcinogenesis but there is an indication in the strain dba (-MTI) that it is the increased 'luteal stimulation' in the breeders which acts as a developing factor. Bonser (6) and Jull (14) have reported the special susceptibility of strain 'IF' virgins to methylcholanthrene treatment, which could also be attributed to particularly heavy luteal stimulation as indicated by the appearance of its mammary ducts which were studded with acinar buds closely comparable to a gland during pregnancy and giving the appearance of a 'tree in flower'.

The importance of the 'luteal factor' in chemical carcinogenesis was further confirmed by our observations on the new strain L (P). The mammary glands of strain L (P) as well as, the ovaries suggested a marked inadequacy of luteal reaction at different age periods, in both virgins and breeders. In the animals treated with methylcholanthrene the tumours developed (17/26) grew rapidly and occurred in all the breast glands of the affected animals. The ovaries also gave some indication of increased luteal stimulation. The nature of this action is still obscure and is not cleared up from these experiments. It may, however, be surmised that MCA possesses growth stimulating properties in view of its chemical structure which closely simulates some steroid hormones and its hormono-mimetic action (13). In our experiments methylcholanthrene appears to act first as an initiating and on continued treatment as a developing factor in the production of mammary tumours. The hormonal stimulation during pregnancy, particularly by the progesterone appears to be one of the essential 'developing' factors. Investigations on the role of the pituitary gland might elucidate further the mode and mechanism of action of methylcholanthrene.

Methylcholanthrene has an additional effect on the reproductive function of the ovary besides its effects on its endocrine functions. Ova during their process of maturation or probably later during fertilisation are affected by MCA. The effect of MCA is reflected in the abnormalities observed in the progeny of animals painted with the chemical. Interesting strain differences have been observed in the progeny and are briefly tabulated (Table 6).

These studies on the mechanism of carcinogenesis in the spontaneous and chemically induced breast cancer reveal that although the etiological factors differ, the fundamental process of carcinogenesis is essentially the same. It consists of three stages 'initiation', 'promotion' and 'development' of the disease.

TABLE 6.

Observations on the Progeny of Painted Breeders.

Strain	Number of Breeders Painted with MCA	Total No. of young ones born.	Total No. survived.	Number observed till death.	Breeding behaviour of progeny.
Dba	16	168	7	4 ♀s, 3 ♂s	Females did not mate.
Dba (-MTI)	20	193	0	0	
A	13	260	221	10 ♀s, 6 ♂s	Normal healthy Breeders.
C_{57}	4	38	26	8 ♀s, 4 ♂s	Normal healthy Breeders.
L(P)	6	62	42	10 ♀s, 4 ♂s	Normal healthy Breeders.

In spontaneous carcinoma, the virus like agent (MTA) which is conveyed through the mother's milk is the 'initiating' factor. It acts on the breast parenchyma cells and on the hormonal system. The hormonal stimulation specially the estrin stimulation—acts as the 'promoting' and the 'developing' factor. A high dose of estradiol can promote cancerous growth in the breasts of the susceptible castrates possessing the milk-borne tumour agent. In chemical carcinogenesis, the chemical substance instead of the virus in the milk acts as an 'initiating' and to a certain extent as a 'developing' agent when the treatment with it is prolonged. Whatever may be the 'initiating' factor of the disease (MTA or MCA) the promotion and development of the mammary tumour once it has started is essentially controlled by adequate hormonal stimulation. The voluminous literature on this subject has repeatedly stressed the importance of estrin stimulation in mammary carcinogenesis. In the chemical carcinogenesis, however, effective progesterone stimulation in female appears to function as an important developing factor.

Tumour Metastases:

An action and interaction of intrinsic and extrinsic factors initiates the process of carcinogenesis which finally results in the development of a carcinoma. Once the cancer has been launched on its career there follows a spread of the tumour in the body. The process of metastatisation and the factors affecting it need careful investigation. At attempt is being made to study the role of the reticuloendothelial system in this process of metastasis. Some work in our laboratories (9) suggests that after blocking the reticulo-endothelial system with agents like trypan blue and cortisone, it has been possible to induce secondary growths of the tumour in many organs like liver, kidney, lung, etc., as seen in the table below.

TABLE 7.

	Total Number of animals.	Metastases in			
		Liver only	Lung only	Liver & Kidney	Lung, Liver & Peri-nephric region
Trypan blue treated	29	3	4	1	1
Control	28	0	0	0	0

It appears that the R.E.S. plays a dominant role in preventing a metastatic scattering of the tumour in mice.

Tumour Transplantation:

What has been stated above refers to the behaviour of the mammary tumour in the animal in which it originates, grows, spreads and scatters. The nature and behaviour of the tumour could be further studied by transplanting it in other hosts. Fifteen spontaneous and methylcholanthrene induced breast tumours from lines of mice with and without (MTA) have been studied for more than 30 generations of transplantation. The milk factor adenocarinoma is not noticeably altered through transplantation but the agent free carcinoma simplex has developed into highly anaplastic type in its 25th generation and the methylcholanthrene induced tubular adenocarcinoma has shown a sarcomatous transformation of its

stroma very early during repeated transplantation. It may be mentioned here that in these studies, besides recording the local growth rate, site of metastases and microscopic structure of the tumour at each transplantation, detailed "in vitro" observations have been carried out. Growth rate, growth pattern, cell morphology and cytology of different tumours have been compared in 'hanging drop cultures' and significant differences have been observed in the behaviour of different tumour types (23). Twelve adenocarcinomas, one carcinoma simplex, one adenoacanthoma and a carcinosarcoma have been thus standardised as a valuable material for later use in experimental chemotherapy.

Experimental Cytology:

Studies on the biological mechanism of carcinogensis would be incomplete unless it is explained at cell level. An evaluation of the direct action of cancer inducing substances on normal cells, viz., the effect on their morphology, cytology and intracellular enzymatic activities may shed further light on this phenomenon of a transformation of healthy into malignant cells. It is possible to undertake these studies on living tissue cells of different origin, grown in tissue culture. The possibility of growing successfully and maintaining indefinitely different cell-types in appropriate nutritive media of known chemical composition (tissue farming) constitutes the main obstacle in these studies. Different normal and cancer cell lines of mouse and human origin are being maintained in our laboratories for investigations in experimental cytology. There are many facets to these studies (biological and biochemical). For instance, it may be fruitful to study the nutritional requirements and the metabolism of different cell types under controlled conditions. Several mouse mammary tumours besides mammary glandular tissue from adult virgins and pregnant female mice have been grown and maintained by different techniques in tissue culture in our laboratories. It is intended to use this material for studying the direct effect of various hormones on the normal and malignant breast tissue.

An attempt has been made in this review to give a bird's eye view of the several investigations which have been completed or are at present in progress in one department of our laboratories. The work reveals a persistent effort to unmask one of the unsolved riddles in cancer research and the different directions in which it has led us so far.

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Note: No illustrations and photomicrographs are included in this review. They may be consulted in the publications cited above.

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Cultivars of the Fungus, *Fusarium*

BY

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The use of variety names for plants arose in ancient times when man began to recognize the advantages of one planting stock over another. When he found that a certain grape tasted better, yielded more heavily, or made a choicer wine than another grape, he needed a distinguishing name for it. Such a name is our horticultural variety or, preferably, cultivar, of today. It preceded formal botanical classification, which is a product of modern time, and has survived. The cultivar provides a means of informally naming plants. It has nothing to do with taxonomy, or classification, and, therefore, is entirely independent of the botanical variety which implies relationship and position in a scheme of plant classification. These two systems of naming serve different purposes, and may supplement one another, but neither takes the place of the other.

The role of the simple, convenient name of the cultivar is as important today as it was centuries ago. Actually it is more important now because of the development and use of more and more cultural varieties.

The more we learn about plants and the more we use them, the more variation we recognize in them. And the more variation we see, the more opportunity we find for the selection and perpetuation of new cultural varieties, and thus grows the need for more variety names. As it is with the grape, so is it with all plants. And so it becomes with fungi. Indeed, so must it be with all kinds of plants and animals the cultivation or husbandry of which serves man.

The cultivation of fungi today is increasing phenomenally. We are learning rapidly, now, how better to grow fungi, and we understand more fully their variability. This enables us to utilize them to greater advantage, to select superior variants for the production of vital medicines, chemicals, foods; and to approach more intelligently the control of those which destroy our crops and which in other ways are harmful to man. We need names for the subspecific entities or variants within the species, for the clone or

group of clones the cultivation of which is in some way important to man. The situation now arising in fungi is analogous to that which arose in crop plants a long time ago, and the solution seems the same, i.e., the employment of cultivar names. This paper is submitted in support of the wider use of cultivar names in fungi.

Need of cultivar names in fungi

The better we know an organism the broader becomes the species concept applied to it, and the greater the need for names for intra-specific categories. Man, perhaps, knows more about his own species, *Homo sapiens*, than any other. He accepts without question a wide range of stature, skin color, behavior, and racial characteristics in this species. He provides names not only for all categories from the species to the individual but even gives one or more names to every individual. In no other species on earth can every individual boast a name. The dog, *Cannis familiaris*, probably comes nearest to man in this respect. This species, too, has been known long and well, and here again a broad species concept has evolved. Some of the variations in this species have been selected and perpetuated under such names as Poodle, Great Dane, and Mexican Hairless, just as Perfection, Green Giant, and Marvel have been selected for culture out of the pea species, *Pisum sativum*. All of these names of dogs, and of peas, are in effect "horticultural" varieties or cultivars, that is, they are the cultivated intra-specific variants of species.

Another organism man has known long and well is wheat. He has cultivated it for at least 3000 years. When taxonomists began to classify wheat botanically they tended to regard the different forms of wheat as distinct species, and so named them. They separated, taxonomically, awned from awnless, winter from spring, and red from white wheats. It has been only since the knowledge of the inheritance of these features in wheat that many such species were merged together. Now taxonomists agree that most common varieties of wheat belong to one species, and the former species, botanical varieties, etc., have given way to one or more cultivar names. Now wheat, *Triticum aestivum* L., is a large species composed of a wide range of variants, the more important of which have been designated by cultural variety names.

Less well-known species of organisms are apt to be narrow and small. This is because the first step in classification, after collection, is analysis, where the different collections are searched

especially for differences from the then described species. At this stage any differences are apt to be seized upon to make a new species, and a sounder classification must wait the accumulation of knowledge which finally permits the recognition or synthesis of fewer but better species.

The science of mycology is growing up. Knowledge acquired through genetical and other experimental disciplines is beginning to reflect this advance in the fungi. Already there has developed a marked trend toward classifying fungi into fewer and larger species. And it is only natural that following upon this trend there should be an increasing need to provide cultural names to the ever-growing number of variants within these species to which man is now turning his interest.

Examples of cultivars in fungi

When man selects from nature a particular edible mushroom for cultivation he has started a cultivar of it, whether it is so named or not. Through mutation or other variation superior strains or clones are selected again, and these become additional cultivars. Similarly, clones in *Aspergillus* and *Penicillium* species are selected for their superior production of citric acid in one case and penicillin in the other. These, again, may be cultivars by definition. Although not generally named as such, it is clear that cultivars already are in wide use in industrial mycology.

But there are other fields of economic mycology where the use of cultivars is just as soundly indicated. One of these fields is plant pathology.

If one wanted, for some reason, to produce sclerotia of *Sclerotinia sclerotiorum,* using agricultural methods, he might grow such crops as the crucifers, forage legumes, and lettuce on infested land. Were he to do this he probably would find that he was obtaining a variety of sclerotial sizes following parasitism of these crops. If the uses for the sclerotia called for small sclerotia, he might select lettuce for the propagating crop, if large sclerotia, cauliflower. Such procedures would be recognizing variation in *Sclerotinia sclerotiorum* and mean the selection of cultivars of the fungus by controlled propagating methods. That such horticultural varieties in *Sclerotinia sclerotiorum* do exist, and are of economic importance, has been recognized already by Purdy (3) who has named the small, intermediate, and large sclerotial forms of

Sclerotinia sclerotiorum as cultivars Minor, Intermedia, and Major, respectively.

In the genus *Fusarium* similar examples of cultivars exist, and only remain to be named, for the same reasons. In the species *F. oxysporum* both Gordon (1) and Snyder and Hansen (4) agree on the taxonomy and nomenclature of its component variants except for the former species *F. redolens.* Gordon considers this species to be sufficiently distinct from the range of morphology in *F. oxysporum* to warrant making it a botanical variety, *F. oxysporum* var. *redolens,* rather than to make it a synonym of *F. oxysporum* as Snyder and Hansen have done. The present writers acknowledge that there are features in *F. redolens* which are on the level of the cultivar and propose to recognize these features by naming the eariler species, *F. redolens,* as a cultivar: *F. oxysporum* "Redolens" with description as given for his botanical variety.

Similarly, Gordon (1) is in essential agreement on the taxonomy and nomenclature of the species *F. solani,* except that he retains separately from this species *F. coeruleum,* which Snyder and Hansen (5) merge with *F. solani.* Because of the economic importance of the fungus formerly known as *F. coeruleum,* but at the same time recognizing that its distinguishing features are unsuited for nomenclatural distinction on a species or even a botanical variety level, it is proposed here to give this fungus a cultivar name, *F. solani* "Coeruleum," with description as provided for *F. coeruleum.*

The most striking cultivars are, perhaps, in the species *F. roseum.* Into this species Snyder and Hansen (6) merged such former species as *F. graminearum, F. culmorum, F. avenaceum, F. acuminatum, F. equiseti,* and *F. sambucinum.* The basis for this consolidation of species was provided by Oswald (2) and on the grounds of the wide variation obtained from crosses made by Gordon and Snyder and Hansen with component clones of *F. roseum.* However, in spite of similar experimental evidence in the hands of the proponents of both taxonomic schemes, Gordon prefers to keep as acceptable species several of those which Snyder and Hansen prefer to combine into *F. roseum.* It seems to us that these differences of judgment may be resolved by the use of cultivar names for those differences in morphology which all of us recognize, but which are given different weight taxonomically. It is proposed here, then, that the following cultivars be recognized in

the species *F. roseum*: Culmorum, Graminearum, Avenaceum, Acuminatum, Equiseti, and Sambucinum. These cultivars are recognized chiefly on appearances, and are independent of such biological characteristics as are indicated in forma and race names. (The forma, and its subdivision, race, represent distinctions in behavior rather than primarily in appearance).

These cultivars may be written, therefore, in the following fashion, when no pathogenesis is indicated:

F. roseum "Graminearum"
" " "Culmorum"
" " "Avenaceum"
" " "Acuminatum"
" " "Equiseti"
" " "Sambucinum"

and as follows when pathogenesis is indicated:

F. roseum f. cerealis "Graminearum"
" " " " "Culmorum"
" " " " "Avenaceum"

It seems to us that the designation of the transitory or minor variations in conidial morphology or appearances, which are too unstable to find a place in botanical classification, are best labeled in the above fashion as cultivars.

Other cultivars in *Fusarium* which we believe at this time should be named are as follows.

F. episphaeria "Dimerum" for the small-spored forms
" " "Gigas" for the large-spored forms
F. moniliforme "Subglutinans" for those forms of the species which usually do not produce microconidia in chains.

These cultivars in *Fusarium* may not be cultured directly by man, nor intentionally. Nevertheless, they are a direct result of man's selective crop culture. If he intended to produce the cultivars of *F. roseum* he could do no better than he is already doing unintentionally. Graminearum, Culmorum, and Avenaceum represent variants in *F. roseum* which man, inadvertently or not, selects out and propagates in the course of selecting out and cultivating agronomic varieties of *Triticum aestivum.*

Are these differences too great for cultivars?

Some may object that the morphological differences between the cultivars Graminearum, Culmorum, and Avenaceum, in the species *F. roseum,* are too large, that it is better to recognize these differences on a species level.

To explore this situation, let us look at cultivars in, for example, their host plant, wheat, as shown in Plate I A.

The differences here appear far greater than those shown between the pathogenic and non-pathogenic cultivars of *F. roseum,* often associated with wheat culture, illustrated in Plate I B and C, respectively.

Certainly the differences between cultivars in *F. roseum* are a great deal less than those commonly accepted in onion Plate II A, in squash Plate II B, or in radish Plate III.

Some of the cultivars of the higher plants shown here are based on vegetative structures, as in *F. roseum,* others on the fruit.

And we should remember that some of these cultivars of wheat were at one time species, as were the cultivars of *F. roseum.*

Some may feel we are not justified in drawing a parallel between fungi and higher plants. Why not? Why not draw comparisons in the taxonomy and nomenclature of all living things, whether plant or animal, whether they are higher or lower forms of life? Might not the classification of one group of living things benefit from the knowledge acquired in classifying and naming another group of living things? Wheat, *Triticum aestivum,* has been studied longer, more intensively, by more individuals than the fungus, *Fusarium roseum.* Is it not desirable for one who is studying *Fusarium* to look to those who have studied *Triticum* for guidance and ideas? More than this, actually no comparison is being made between fungi and higher plants, the comparison is between the man-made procedures or systems of classifying and naming organisms. Is there anything wrong with man being consistent about such procedures? Rather, how can he afford to be inconsistent in applying all his knowledge and his most proven concepts in the classifying and naming of *all* organisms? And finally, were we to compare the morphological differences of cultivars of wheat, and of *F. roseum,* we should have to remember that the variability in the diploid sexual fruit of a plant is less than that of a haploid vegetative stage.

Discussion

The cultivar is an informal device for naming subspecific clones, or group of clones of living things which have no taxonomic standing and the use of which conveniently serves man's everyday need. Examples of established cultivars are Poodle, in the dog, *Cannis familiaris;* Perfection, in the pea, *Pisum sativum;* and Minor, in the fungus, *Sclerotinia sclerotiorum.*

The need for cultivars increases with the consolidation of many species into one, such as in wheat. The cultivars of today are often the species of yesterday. Unrecognized variability in living things brings into existence innumerable man-made species. Recognition of variability, on the other hand, eventually brings down sharply the number of species. The younger or less studied the taxonomy, the greater the number of species and the fewer cultivars to be found in it, and conversely. Most fungi are little known and are vastly over-speciated. The trend among the most intensively studied genera of fungi is towards the recognition of fewer species. This trend brings about the need for cultivar names by which to indicate certain former species and any other subspecific forms man has had reason to learn to recognize, as a plant breeder has reason to learn to distinguish between common wheat varieties, whether directly cultivated or not.

Some pathologists have reason to distinguish between cultivars Culmorum and Graminearum, even though both may cause the same diseases of wheat and taxonomically are in the same species, *F. roseum,* because the latter produces perithecia in the field and, therefore, having ascospores, is readily wind-borne. Other pathologists may not have need for this specialized knowledge, but are more concerned with distinguishing between the foot rots caused by *Helminthosporium sativum* and *Fusarium roseum.* For them the use of species names in place of cultivar names would delay species diagnosis.

The adoption of cultivar names in fungi is a step towards the simplification of fungus identification. An example of great need in this respect is in the Actinomycetes. Progress has been so rapid in the industrialization of these fungi used in the manufacture of antibiotics that taxonomy and nomenclature have been outstripped and confusion is the result. It seems inevitable that in *Streptomyces,* as in most economic genera of fungi burdened with large numbers of species, the solution will lie in the delimitation

of very few, broad species, with their biologically distinct formae and races, and with sufficient cultivars to take care of the every day needs of the economic mycologist.

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Explanation of Plates

PLATE I

Cultivars of wheat, and of the *Fusarium roseum* often associated with its culture.

A. Wheat cultivars, left to right: Sonora, Onas 53, Ramona, Poso and Kanred.
B. Cultivars of *F. roseum* pathogenic to wheat, left to right: Graminearum, Culmorum, and Avenaceum.
C. Non-pathogenic cultivars of *F. roseum*, left to right: Sambucinum, Equiseti, and Acuminatum.

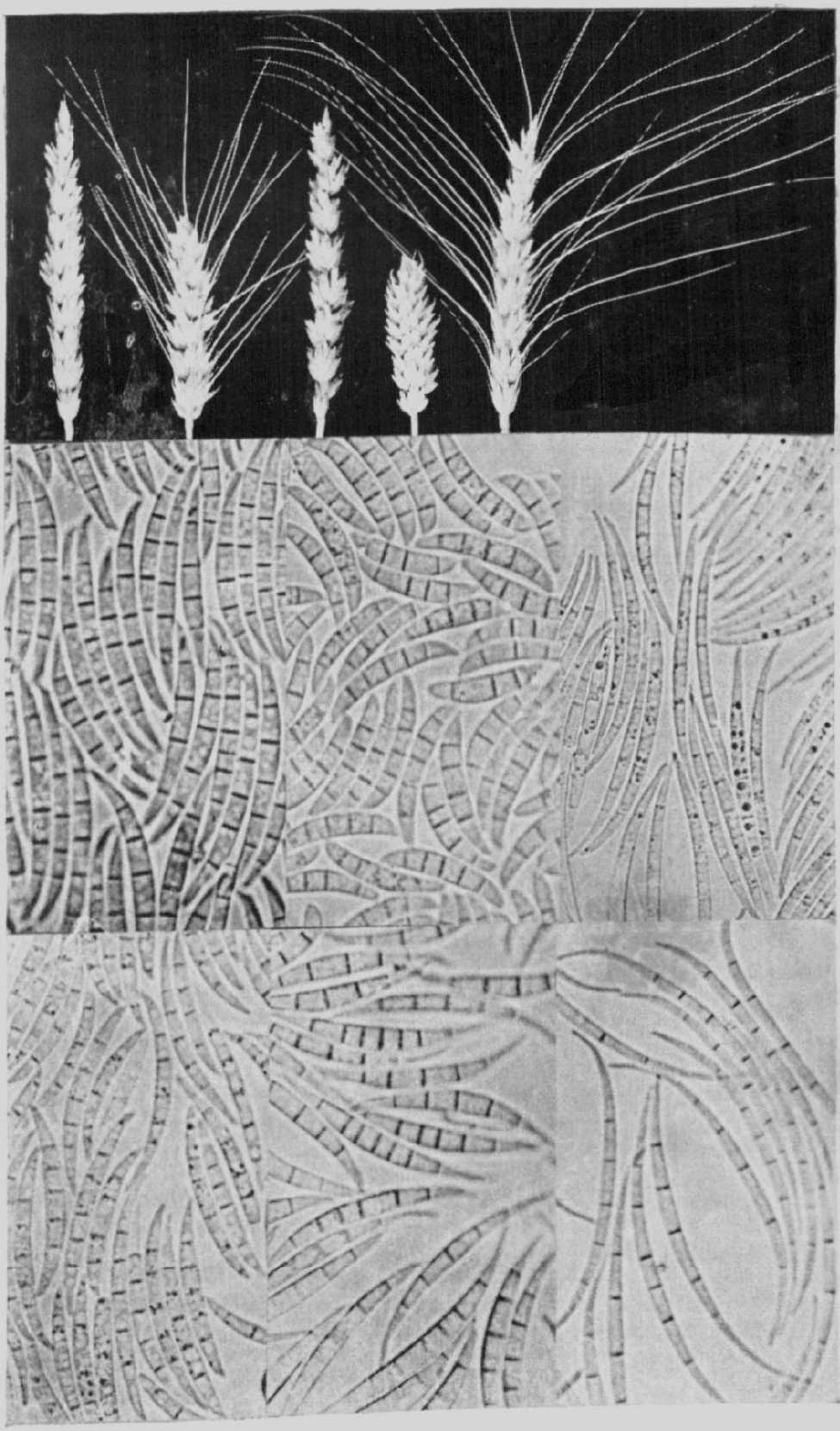
PLATE II

Cultivars of onion and of squash.

A. Left to right: California Early Yellow, Italian Red, and Extra Early Flat White onions.
B. Left to right: White Bush Scallop, Yellow Summer Crookneck, and Dark Green Zucchini squashes.

PLATE III

Radish cultivars. On the left, White Icicle; on the right, Cherry Belle.



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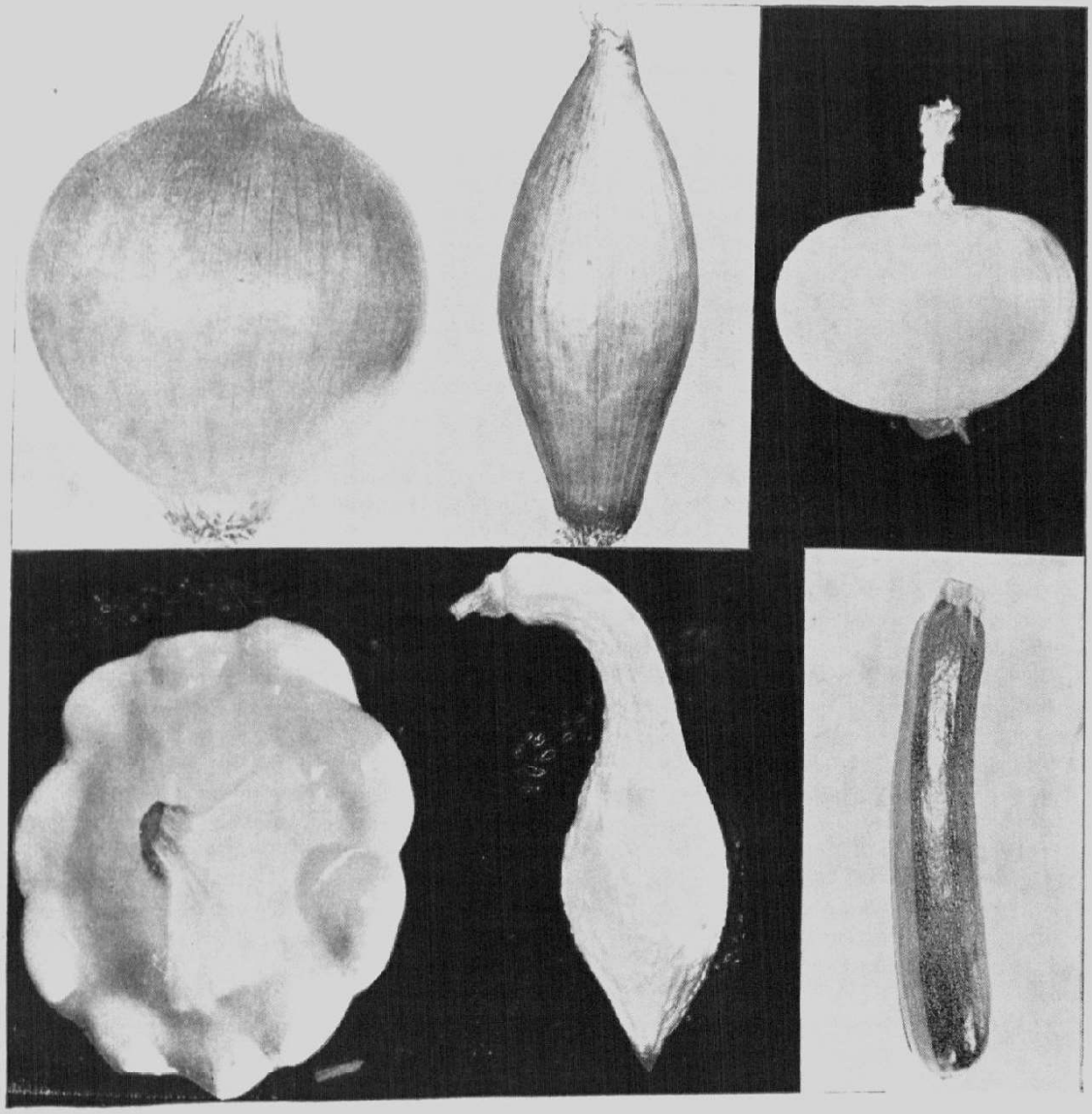

A

B

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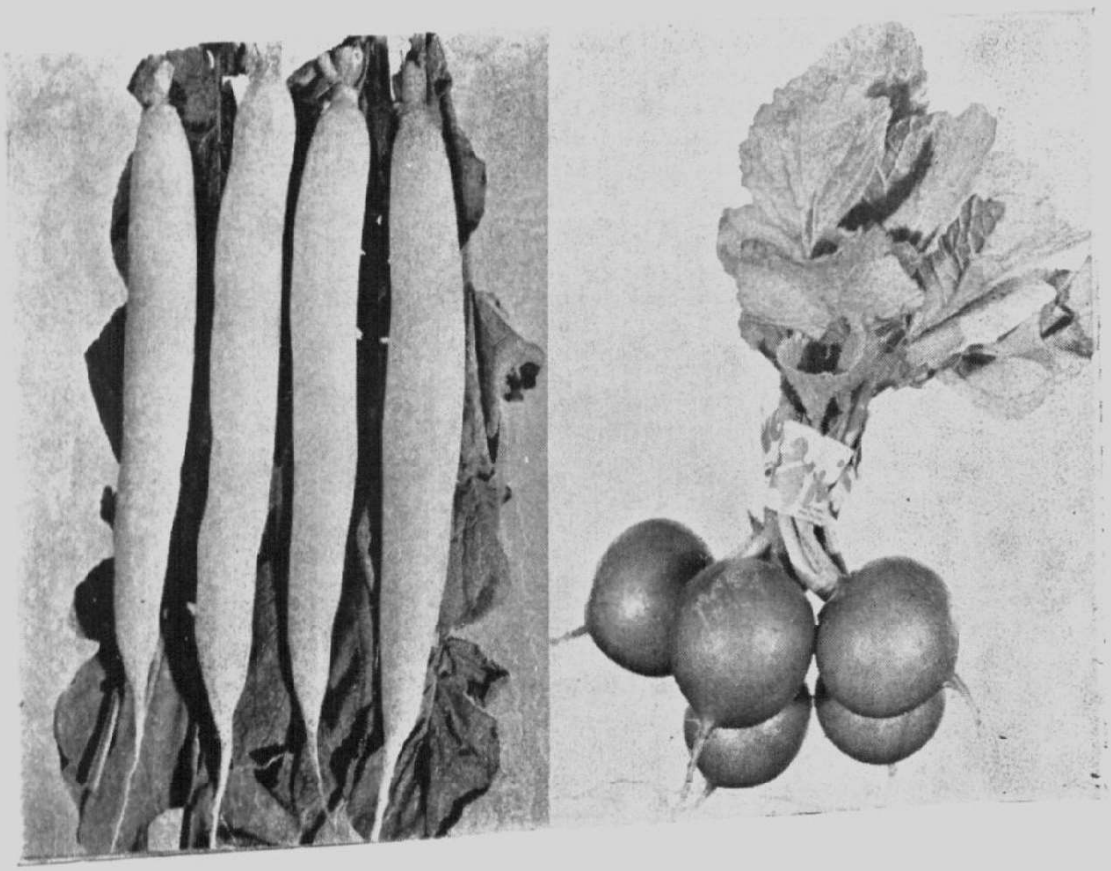

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Volcanic Episodes in Indian Geology

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It is now generally agreed that the earth passed through a stage when its surface temperature was high enough to melt the rocky material of which it is made, or at least to soften it to a viscous state. The heat of the original earth was gradually radiated out and a solid crust began to form. During a period of many million years, until its surface became cool enough to collect large pools of water, the crust was composed entirely of partly crystallised molten rock whose constituents were probably much the same as the rock-forming minerals we now find in the igneous rocks exposed at the surface.

At the present time, we cannot be sure about the constitution of the primordial crust. It had probably the mineral composition of a gabbro or basalt. Among the shield areas of the world, i.e., those containing the oldest rocks, confined to a few limited regions, the more important rock types are granitoid gneisses containing mixtures of acidic and basic materials in varying proportions. From what we know of the phenomenon of radioactivity, the earliest formed minerals which have been preserved in the shield areas—and therefore the rocks containing such minerals—are about 2300 to 2500 million years old. As the radioactive minerals occur generally in granitic and pegmatitic rocks which are later in age than the gneissic and schistose host rocks it is likely that the host rocks are many million years older than the age indicated by the radioactive minerals contained in the pegmatites.

The stratification of the earth into various shells inferred from seismological studies leads us to believe that the core, which occupies more than half of the volume, is of the nature of nickel-iron. This idea is indirectly supported by iron meteorites which fall to the earth from outer space and which represent the original material out of which the planets of the solar system were made. Surrounding the core is a shell of metallic sulphides and ferro-magne-

sian silicates, perhaps resembling the material of the stony meteorites which also come from inter-planetary space. The stony meteorites are related to such rocks as peridotites, pyroxenites and eclogites. Still nearer the surface is another shell of the composition of the basic igneous rocks (dolerite or basalt) which make up the bulk of the great lava flows found in many parts of the world. The outermost shell is apparently of rather mixed character but predominantly acidic and granodioritic in composition. The crust itself is only a few kilometers thick and composed of sedimentary rocks containing large patches of metamorphic and igneous rocks. The bulk composition of the earth would therefore approximate to that of basic igneous rocks.

The vast span of time covering the Archaean and pre-Cambrian era must have witnessed repeated igneous action. As the cooling of the earth continued and sedimentary basins became established, there would have been a slowing down of the phenomenon of intrusion and extrusion until stable conditions were established when this phenomenon assumed a periodic character.

The existence of radioactive material in the earth and its comparative concentration in the outer shell is an important factor in the temperature status of the earth and in the preservation of heat in spite of ceaseless radiation from the surface into space. Calculations based upon various assumptions about the quantity of heat generated by radioactive materials in the earth's crust and its interior have enabled geologists and cosmologists to infer that enough heat accumulation occurs to produce igneous action, in some part of the earth or other, every 20 to 30 million years.

Stratigraphical data indicate that three or four major mountain building movements took place since the last movement before the beginning of the Cambrian. Several earlier movements are known, but only in an imperfect way. The four movements are:—

(1) The post-pre-Cambrian movement after which the Cambrians with their rich fauna were deposited. Evidences of this movement, which is about 550 to 600 million years old, are seen in several parts of the world.

(2) The next movement occurred in the Silurian and is called the *Caledonian.* Effects of this movement, which is about 350 million years old, are seen in Central and Western Europe as well as in North America.

(3) The succeeding important movement is called the *Variscan* or *Hercynian,* of Carboniferous age. Evidences of this movement which took place about 250 million years ago resulting in the Karakoram—Hindukush mountain ranges to the north of India, and several other ranges in Northern Asia, are seen in other parts of the world also.

(4) The last great mountain building activity which is called the *Alpine-Himalayan* movement, commenced in upper Cretaceous about 60 or 70 million years ago and lasted through the Tertiary era right down to the Pleistocene Ice Age.

Very little work of high quality has been done on age determinations of radioactive minerals in Indian rocks. The earliest rocks so far known are in the Mysore State where some magnetites indicate ages of the order of 2300-2500 million years. The minerals from the pegmatites of Nellore district of Andhra State indicate an age of around 1650 million years, while some radioactive minerals from the mica pegmatites of Gaya and Hazaribagh in Bihar give an age of 950 million years. Minerals from some post-Delhi pegmatites of Rajasthan indicate an age of 730 million years. No further data are available but it should be possible to determine the ages of the various sedimentary systems in India by careful investigation of radioactive minerals present in igneous rocks associated with them. It will be noticed that there are large gaps in the ages which will have to be filled up by data to be obtained by further investigations.

Periods of mountain building are intimately connected with igneous activity. It would be of interest to state here that the precursor of the birth of a mountain range in geological history is a *geosyncline.* A geosyncline is a depression in the earth's crust of fairly large dimensions occupied by sea in which sediments are washed down from the neighbouring land masses. Geosynclines may also form parts of large oceans. The accumulation of sediments is accompanied by the gradual sinking of the basin. The blanketing effect of the sediments at the bottom leads to the heating of the basal portions which soften and give access to molten igneous materials from the basaltic shell of the earth, these forming sills and dykes. Contemporaneous volcanism found in geosynclinal sediments is probably to be explained in this manner. When the sedimentary basin has been in existence for several million years and has received considerable thickness of sediments, there

comes a stage at which the basin is subjected to horizontal compression which affects the segment of the crust containing the geosyncline. The cause of such compression is not known but has been explained as due to the contraction of the earth or to the effect of convection currents acting beneath the crust. As a result of the compression, the strata formed in the geosyncline are pushed down deeper and deeper, breaking the crust and inducing even rocks from the ultramafic or peridotitic shell to intrude into the sediments near the surface. Such ultramafic intrusions are found generally in the root zone of mountains which were raised up in the region formerly occupied by the geosyncline. At an early stage of the compression, a part of the outer zone of the geosyncline on the ocean side (i.e. away from the land mass) is raised up as a wrinkle to form an *island arc* which may partly be submarine and partly above the sea waters. Violent volcanic activity is manifested in such island arcs as is now seen in the East Indian Archipelago and in the strings of islands in the Pacific ocean bordering Asia.

H. H. Hess (1938) states that all major occurrences of ultramafic rocks in the world are confined to the root zones of recent or ancient folded mountain ranges. Basaltic rocks are also extruded, and occur as sills and contemporaneous flows in the geosyncline during the operation of the compression, for this might last for as much as 30 to 50 million years. During the later stages of compression the sediments at the base of the geosyncline, which become depressed to depths of several scores of kilometres, are more or less completely melted or reduced to a viscous stage and are extruded as granite batholiths in the root zone. The intrusive activity may persist throughout the period of compression, for we find that the granites are followed, at a somewhat later date, by aplites, pegmatites and mineralising solutions which often produce metalliferous mineral deposits containing workable amounts of gold, tin, tungsten, copper, lead, zinc, mercury, etc.

Volcanic action may also be manifested in areas in which the crust is subject to tension, producing cracks and fissures. Such tensional cracks developed in parts of the Gondwanaland during the Jurassic period and ultimately led to its disruption and separation into the various continental units which are now found scattered in the Southern hemisphere. This type of activity is exemplified in the Stormberg lavas of South and Southeast Africa and the Rajmahal volcanics of Bengal which are of Jurassic

age, and the Deccan Traps which welled out from fissures in western parts of Peninsular India in Upper Cretaceous times.

This introduction will give some idea of the mode of formation of volcanic rocks, including both hypabyssal and extrusive phases. A brief account of the various episodes of volcanic activity, deciphered in the different periods of the geological history of India, follows.

PRE-CAMBRIAN

More than half of Peninsular India is occupied by gneisses amidst which there are bands of dominantly schistose rocks, usually referred to the Dharwar System. It is now recognised that the schistose rocks may include materials of different ages. As our knowledge of individual areas is still sketchy, our remarks will apply broadly to regions rather than to limited districts.

Mysore: The Dharwar System consists of various types of schists, quartzites, ferruginous quartzites and subordinate crystalline limestones. The schists are dominantly chloritic in the north and hornblendic in the south. Granulitic texture is prominent where the hornblendic schists are in contact with or included in granitic and gneissic rocks. A stage of alteration intermediate between the two is probably represented by epidiorites. In the Shimoga schist belt, B. Rama Rao (1940) has divided the rocks into three divisions each of which contains effusive igneous rocks. Post-Dharwar volcanism has also been deciphered, as shown in the accompanying table.

Post-Dharwar	Felsites and porphyries, generally unaltered, occasionally sheared. Amphibolite and norite dykes.
Upper Dharwar	Some basic sills and other igneous materials.
Middle Dharwar	Rhyolites, tuffs, felsites. Basic intrusives, mainly sills and chloritic schists.
Lower Dharwar	Basic flows and sills. Amphibolites, Hornblende schists, chlorite schists, greenstones; Keratophyres, rhyolites and tuffs often sheared and schistose.

The lower Dharwars are considered to be mainly of igneous origin, made up of amphibole and chlorite schists, amphibolites

etc., and flows and tuffs of more acidic character. Where highly metamorphosed, the basic lavas have been converted into granulites. The felspars are of generally medium plagioclase showing ophitic relation to the amphiboles; while biotite, titaniferous magnetite and ilmenite are also present. Occasionally pyroxene schists occur, as at Javanhalli. The fine-grained schistose types may represent tuffs or volcanic flows but some of the chlorite schists may be sedimentary in origin also.

In the middle Dharwars of the Chitaldrug schist belt, there are basic lavas some of which are spilites and pillow lavas, which show vesicles filled with secondary calcite, silica etc. The rocks have not been altered to schistose or granulitic types. Some of the norite, porphyry and felsite dykes appear to be of post-Dharwar age as they do not show much alteration.

Cuddapah Basin: In the Cuddapah basin of Andhra, the Vempalle limestone belt as well as some younger zones contain thick sills, dykes and sheets of basic rocks which are doleritic or basaltic, and generally somewhat altered. One sill traversing the Vempalle dolomitic limestone has produced workable deposits of chrysotile at the contact zone. These dolerites carry besides augite and plagioclase (andesine-labradorite), hornblende, biotite, magnetite, apatite and pyrite. They are similar to the newer dolerites of Singhbhum and Keonjhar and differ from the Deccan Traps in not containing pigeonite and are perhaps differentiates of a tholeitic magma. They are of Upper Cuddapah age, contemporaneous with the Nallamalai or Kistna series (Coulson, 1933).

Rajasthan and Gujarat: The pre-Cambrian formations known here are the pre-Aravalli gneisses, the Aravalli System, the Banded Gneissic Complex, the Delhi System and the Gwalior System. The Bijawar Series of Central India is also considered to be equivalent of the Delhi System. In Western Rajasthan, fairly large areas are occupied by acid igneous extrusives (rhyolites, felsites) which are considered to be of Lower Vindhyan age. In the pre-Aravalli schists there are some amphibolites and hornblende schists which may represent altered or reconstituted ultramafic rocks.

In the Aravalli System, which is generally correlated with the Dharwars, there are intercalated bands of amphibolite and dolerite. These amphibolites are schistose and are composed largely of green pleochroic amphibole, plagioclase, felspar and some titaniferous magnetite.

The banded gneissic complex is considered by A. M. Heron (1917) as older than Aravalli while N. L. Sharma (1953) and H. Crookshank think that they are later in age and intrusive into the Aravallis; the basic inclusions and streaks in them being derived from the Aravallis. The rocks vary from foliated amphibolites to hornblende-schists and epidiorites. The constituent minerals are actinolite, granular quartz, magnetite, chlorite and epidote, the last two being alteration products of the earlier amphibole. Microcline and calcite are found in some varieties.

There are some dolerite dykes traversing the Aravallis which are only partially altered. They contain amphibole, labradorite, a little quartz, occasional hypersthene and biotite, with apatite, zircon and magnetite as accessory minerals. There are also some dykes in this region, the rocks in them containing some amphibole and a little partly altered augite, in addition to quartz, microcline and oligoclase.

In the succeeding Delhi System, which is upper pre-Cambrian, there are contemporaneous as well as intrusive basic igneous rocks intercalated with sediments of both the Alwar and Ajabgarh Series. Amidst the thin quartzites and phyllitic mica-schists, there are layers of amphibolites. Streaks of amphibolites considered to be either thin sills or contemporaneous flows or ash beds are also found in the Alwar quartzites. These rocks consist of green amphibole, clear plagioclase and interstitial quartz with sphene, apatite and biotite as accessory or varietal constituents.

The Gwalior System and the exposures at Ranthambhor are away from the Aravalli mountain belt so that their age is in doubt, but they are perhaps equivalent to the Delhis. In the Ranthambhor syncline there are five horizons of basic igneous rocks of doleritic or basaltic character. The rocks which are more or less uniform-grained do not contain vesicular bands and may represent contemporaneous intrusions as sills. The thickest sill is 200 ft. thick in which the rock is mostly dolerite without olivine or glassy matter. The felspars and augite are partly altered.

According to N. L. Sharma (1953) there are only three important groups of basic intrusives (dykes and sills) in the Aravalli-Delhi succession of Rajasthan. The earlier group, consisting of amphibolites, epidiorites and hornblende-schists is intrusive into the Aravallis but is of pre-gneissic age. The next group consisting of meta-dolerite is post-gneissic but of pre-Erinpura granite

age, the Erinpura granite being slightly younger than the Delhi system. The third group mainly consists of dykes and occasional sills which are all of post-Erinpura granite age and may be pre-Malani.

The Bijawar series occurring in the Bijawar district in Central India (V.P.) consists of a sedimentary sequence of quartzites, conglomerates, sandstones, siliceous limestones and hornstone-breccia with basic lava flows, tuff-beds and dykes. The dyke rocks are doleritic with micro-pegmatitic patches but some are of the nature of olivine dolerites. The Bijawars are generally correlated with the Cuddapahs and Delhis.

V. S. Dubey (1950) has proposed a threefold division of the Bijawars. He regards the Lower Cuddapah and Delhi rocks as Lower Bijawars; the Middle Bijawars as including those with contemporaneous basic igneous activity as in the Bijawars of the type area and the Upper Cuddapah of the Andhra area; the Upper Bijawars as including the Semris (generally referred to the Lower Vindhyans) of the Son valley which contain acid flows and tuffs, and the Lower Vindhyans of Rajasthan with their Malani igneous suite, as well as the ultrabasic pipes in Bundelkhand containing diamonds. He advocates that the base of the Cuddapahs should be fixed at about 900 million years (Holmes' upper Pre-Cambrian to correspond to Cuddapah); that the Erinpura granites and their pegmatites giving an age of 735 million years, and also the Malani igneous rocks should all be included in the Cuddapahs. A point in favour of this idea is that all Cuddapah and post-Cuddapah igneous activity would be relegated to the Cuddapahs and that the Vindhyans of Northern India would be the equivalents of the Kurnools which do not show any contemporaneous volcanism. There is also a well marked unconformity between the Semris and what are now regarded as the Upper Vindhyans.

Semri Series: The rocks immediately underlying the Vindhyans of the Son Valley, consisting of shales and sandstones, contain dykes of doleritic character. These rocks show both monoclinic as well as rhombic pyroxenes together with andesine, some ilmenite and pyrite and a few cloudy interstitial patches. Amygdular cavities are occasionally present. The Semris are separated from the Vindhyans by a distinct unconformity; they have been somewhat metamorphosed, in contrast with the unmetamorphosed Vindhyans which overlie them. It is, therefore, possible to assign a Cuddapah age to them.

Malani Igneous Rocks: The Lower Vindhyans of Rajasthan are intruded by acid igneous rocks (Jalor and Siwana granites) whose extrusive phases are rhyolites, tuffs, felsites, etc., which are found over a fairly large area in Western Rajasthan. The upper Vindhyans, which contain a large thickness of sandstones, shales and limestones do not contain any igneous rocks and there is also a hiatus between the lower and upper Vindhyans.

Lower Vindhyans Semris	Dolerite dykes. Malani rhyolites, felsites, etc. Pre-Malani dolerites.

Erinpura granite

Bijawar & Delhi	Meta-dolerites and epidiorites, mainly contemporaneous.

Granitoid Gneiss and Banded Gneissic complex

Aravalli System	Sills of amphibolite, epidiorite etc.

Bihar and Orissa: In the Singhbhum district of Bihar and in the neighbouring Sundargarh district (Gangpur) of Orissa, the pre-Cambrian succession is fully developed and it has been studied in recent years. The major units of the succession are shown below.

Newer Dolerite	Dolerite dyke systems traversing the Chota Nagpur and Singhbhum granites.
Arkasani granophyre	Acid granophyric rock formed mainly in the shear zone of North and East Singhbhum.

Singhbhum Granite

Dhanjori and Dalma volcanics		Basic lavas (flows and sills) and tuffs, all metamorphosed to amphibolites, epidiorites and schists.
Iron-ore Series	Iron-Ore Stage	Phyllites, quartzites etc., with basic igneous rocks in the lower portion.
	Chaibasa Stage	Schists of different types partly derived from tuffs and lavas.

In the Gangpur series which occupies the greater part of the Sundargarh district there are a few important sills composed of metamorphosed basic igneous rocks—amphibolites, epidiorites and hornblende-schists. They are generally schistose and contain hornblende, actinolite, tremolite, zoisite-clinozoisite, plagioclase, some quartz; also magnetite, ilmenite, calcite and sphene. Occasionally the amphibolites are garnetiferous. These are probably to be correlated with the Dalma volcanics of Singhbhum as they are earlier than the Chota Nagpur granites (Krishnan, 1937).

In the succeeding Iron-ore Series the lower stage (Chaibasa Stage) contains mica schists, quartz schists, hornblende schists, and tuffs, the last two apparently of igneous origin. The tuffs consist of angular grains of felspar, quartz, slate, etc. In the lower beds of the succeeding Iron-ore Stage also there are well-marked sills of ultramafic igneous rocks. Sometimes the tuffs contain large fragments of foreign rocks and are occasionally finely schistose. Most of the chloritic phyllites in North Singhbhum are considered by J. A. Dunn (1929) as representing original tuffs. There are zones of epidiorite and amphibolite as well as of amphibole and chlorite schists which are interbedded with other rocks and may therefore be sills. Along the major thrust zone there are layers of talc rocks and serpentines.

The Dalma and Dhanjori lavas form a unit in the succession and they overlie the Iron-ore stage. In the east, this volcanic belt is marked by coarse agglomerates representing fragmental products which show numerous angular or rounded pieces of foreign rocks several inches across. The common type of lava is an amphibolite or epidiorite with hornblende, andesine, epidote and leucoxene. Other rock types found here are chlorite schists, ferruginous phyllites and epidosites. The lavas have been severely folded and metamorphosed. They are older than the granites, in which they are found as inclusions, which have been partly assimilated to produce hybrid types.

Finally, there are the basic dykes traversing the granites, in Singhbhum, Keonjhar and Ranchi. They are of the nature of quartz dolerites with interstitial granophyric patches. The thicker dykes, which may be occasionally several hundred yards wide, show coarse gabbro in the centre. The rocks show only a slight alteration and resemble similar dyke rocks in the Cuddapah formation in Andhra (Krishnan, 1935).

Assam: Shillong Series: In the Shillong Series, occurring on the Assam plateau, there occur sills and lenses of amphibolite, hornblende schists, chlorite schists, etc., which are metamorphosed basic lavas. The Khasi greenstones intrusive into the Shillong Series have also been metamorphosed and folded in pre-Cambrian times.

PERMO-CARBONIFEROUS TO TRIASSIC

Panjal Volcanics: Thick lava flows occur in the Pir Panjal as well as in the Kashmir valley. The earliest manifestations of volcanic activity were of the explosive type, consisting of agglomerates and ash beds. They are exemplified by the Agglomeratic Slates which are greywacke-like and grade into slate. They contain angular fragments of quartz, felspar, quartz-porphyry, granite, limestone and slate. They overlie the Syringothyris Limestone or Fenestella Shales and are to be found even in the upper part of the same limestones which are of Middle Carboniferous age.

The volcanic flows, which are later, have ages varying from Lower Permian to Upper Trias. The flows overlap beds of different ages in different places. In some localities they underlie the Zewan beds or the Gangamopteris beds. Beyond the Wular lake they are associated with various horizons of the Trias. They consist of bedded flows of green, brown and dark colours and include rhyolites, trachytes, dacites, andesites and basalts. They are to a large extent altered, the felspars and ferro-magnesians being changed to epidote and chlorite. Their equivalents are found in the lavas overlying the infra-Trias of Hazara which attain a thickness of 100 ft. or so, and are of Carboniferous to Permian age. The volcanism which gave rise to these rocks is probably connected with the Hercynian revolution which is known to have affected large areas in Central and North-east Asia and raised up the Karakoram and Hindukush ranges.

It is likely that there are other areas of similar volcanic rocks in parts of the Himalayas and Tibet. It is known that the Everest Pelitic Series in Mt. Everest (of Carboniferous age) is profusely injected with granitic rocks which are probably connected with the Hercynian revolution. Little is known regarding the geology of the Trans-Himalayan range in Tibet which is known to contain granites and metamorphic rocks of Mesozoic and earlier ages, for they have contributed materials to the Mt. Kailas conglomerate, forming part of Mt. Kailas. It is probable that igneous activity

in the Trans-Himalayan region as well as some of the materials now inextricably mixed with the rocks of the Central and Sub-Himalayas are of Hercynian age.

GONDWANA

The Gondwana strata are mainly shallow water estuarine and continental sediments. Their age ranges from Permo-Carboniferous to Jurassic, but the plant beds of Umia in Kutch are also included in the Gondwanas. Two types of igneous rocks are intrusive into these strata, viz., mica lamprophyres and dolerites, as dykes and sills. Contemporaneous volcanism is seen in the Rajmahal hills in North Bengal where the Rajmahal Traps are intercalated with plant-bearing sediments of lower or middle Jurassic age.

The Rajmahal Traps are mainly of doleritic character but there are also a few acid flows in them. The lavas attain a thickness of about 2000 ft. and have been extruded as a series of flows between some of which there are sedimentary beds containing well-preserved plant and animal fossils. The typical Rajmahal Trap resembles the Deccan Trap closely.

Doleritic rock of similar character is found as dykes traversing the coal-bearing lower Gondwana sediments in all the coal-fields of the Damodar Valley. The dykes are up to 150 ft. thick and some of them are several hundred yards long. Occasionally they form sills following the bedding planes. As they cut across the faults without showing any displacement they must have been intruded after the main faults were formed. Occasionally they are found to be affected by a few cross-faults which are evidently later than the major faults. The age of the main faulting is considered to be post-Mahadeva, i.e., Rajmahal.

The usual rock types in the above mentioned dykes are normal dolerites without olivine, but a few dykes contain olivine dolerite. The felspars are somewhat turbid and the pyroxene is monoclinic showing ophitic relationship to the felspars. Glassy interstitial patches are noticed in some rocks.

Mica lamprophyre (usually called mica peridotite) is seen mainly as sills in the coalfields of Damodar valley and the Darjeeling foot-hills. This rock generally follows the coal seams, often intimately penetrating them as thin anastamosing veins, lens-like

masses, and thin flat sheets at the junction of coal seams and sandstones, or in the coal seams themselves. These rocks must have been emplaced in a highly fluid state at a moderate temperature, as they have destroyed the coals at the contacts converting them to a type of coke (called *jhama* by the miners) which is rich in phosphorus. The phosphorus may perhaps have been contributed by the intrusive which is rich in that element. In some places the intrusive rock has filled faults and may therefore be partly contemporaneous with the faults and partly slightly younger. It is definitely older than the dolerite which is sometimes found cutting across the lamprophyre sills or dykes. The lamprophyre is a grey to brown or greenish rock of coarse- or fine-grained texture and is generally highly altered. It consists largely of olivine which is almost always completely altered to serpentinous matter, some altered felspar, calcite or dolomite, quartz, mica, and chlorite, and abundant grains of apatite. As the rock is rich in potash it is likely that the original felspar was largely orthoclase.

CRETACEOUS

Exotic Blocks of Hundes: The Mesozoic rocks of the exotic facies found in Johar and Hundes must originally have been deposited somewhere in Tibet but are now found thrust over the earlier rocks to the south. They are intimately penetrated by intrusive igneous rocks which have metamorphosed them to some extent at the contacts. It would appear that the igneous rocks were injected during thrust movements into all available spaces like bedding planes and joints. The intrusives comprise ultramafic rocks such as peridotites and pyroxenites and basic rocks like dolerite, basalt, spilite and greenstone. The ultramafic rocks which are generally serpentinised and found associated with radiolarian cherts of Middle or Upper Cretaceous age, are thought to be of upper Cretaceous age.

Southern Tibet: In the Kampa-Dzong region there are serpentinised ultramafic and greenstone intrusives in the Cretaceous rocks. These may be of Upper Cretaceous or Lower Eocene age.

Kashmir: In the Borzil-Dras region of Kashmir, serpentinites and other ultramafic rocks are intercalated with or intrusive into the Cretaceous rocks. They are associated with sediments containing *orbitolinae* which are upper Cretaceous or early Eocene. Ac-

cording to D. N. Wadia, the associated granites in the region are presumably of Eocene age.

Baluchistan: There are abundant manifestations of volcanism in the Calcareous zone of Baluchistan. According to H. Crookshank (1952) the earliest manifestations contain agglomerates and tuffs found as contemporaneous materials in the Belemnite shales of Neocomian age. The agglomerates contain rolled pebbles, probably derived from islands formed at this time (presumably an Island arc). These are followed by pillow lavas and spilites with shale intercalations in the Parh Limestone (Barremian to Aptian). Similar basic rocks, apparently products of contemporaneous volcanism, as also ash beds are found in the Olive Shale (Senonian) and the Pab Sandstone (Senonian-Maestrichtian). It may, therefore, be said that the geosyncline in Baluchistan witnessed igneous activity through a greater part of the Cretaceous. Mountain building probably commenced here in Upper Cretceous times though an island arc seems to have come into existence in the early Cretaceous.

Indo-Burma Frontier: Evidences of late Cretaceous volcanism also exist in the hills of the little known Indo-Burma frontier tract, where both ultramafic and basic intrusive rocks are known to occur. The activity seems to have continued in the Arakan region and in the Andamans, in the succeeding Eocene period.

The Deccan Traps: The Deccan Traps, which cover nearly 200,000 sq. miles of western and west central parts of Peninsular India, are composed of layers of volcanic flows erupted from fissures. They reach a maximum thickness of some 7000 ft. near the Bombay coast, and are mainly of the composition of dolerite and basalt. Ash beds are often seen in the upper parts of the traps. Some flows are vesicular, the amygdular cavities being filled with chalcedony and other forms of silica and also various zeolites and calcite. The dolerites are composed of labradorite and monoclinic pyroxene (pigeonite). Interstitial glassy matter rich in iron is present in the basaltic types while granophyric or micrographic patches are occasionally present. Olivine is found in some types which are, however, comparatively rare. In parts of Kathiawar, Kutch and Gujarat there are a few places where the traps are associated with acid, intermediate (rhyolites, trachytes, etc.), and ultrabasic rock types derived through differentiation of the original magma.

There are numerous dykes traversing the lava sheets in certain areas. In Kathiawar the dyke systems are radially arranged around volcanic centres. Dykes are common in eastern Kathiawar, Narmada valley, Gujarat, Satpura region in Madhya Pradesh, and N. W. Hyderabad.

In Sind and in the Godavari valley in Andhra, the Deccan Traps are overlain by beds containing *Cardita* (*Venaricardia*) *beaumonti*. They are, therefore, thought to have been formed in late Cretaceous, extending into Paleocene and Lower Eocene. The intercalated inter-trappean beds contain plant and molluscan fossils of Lower Eocene age. The volcanism which gave rise to the Deccan basalts is of the fissure-eruption type caused by the formation of tension cracks in the crust connected with a later phase of the disruption of Gondwanaland.

TERTIARY AND QUATERNARY

Baluchistan: There is an igneous belt just west of the Calcareous zone of Baluchistan associated with the Khojak zone which is largely of Oligocene age. Further west, near the Afghanistan and Persian frontier, there are also products of geologically-recent volcanism, including the sulphur deposits of the Nushki desert region.

Burma: Abundant evidence of contemporaneous igneous activity are found in the Irrawaddian (mainly pliocene) and in the Pleistocene of the median Tertiary belt of Burma. The lava flows and associated ash beds have been found to have been folded by the tectonic movements at the end of the Tertiary era. The volcanic activity should have continued up to the sub-recent times as there are numerous fairly well-preserved volcanic craters, the volcanic materials from which are occasionally intercalated with Pleistocene and sub-recent fossils.

The localities at which this activity is seen are the Jada Mine area, Kabwet and other places in Shwebo and Mandalay, Mount Papa, the Pegu Yomas in Prome and Tharrawaddy. The products of the activity are agglomerates, rhyolites, trachyandesites, trachytes, dolerites and basalts, found as contemporaneous flows with occasional occurrences of pyroclastics. There are several volcanic craters in these places probably arranged along fault lines. This line connects Barren Island and Narcondam which were active volcanoes during the last century.

Another set of volcanic craters of Plio-Pleistocene age is to be found around Tang Yuch in Yunnan and Loi Han Hun in Northern Shan States, in which basaltic flows, ash and tuff beds are seen.

Summarising, it may be stated that volcanic activity is generally manifested in the later stages of the growth of a geosyncline beginning with the formation of an island arc at the commencement of the period of compression. As the compression proceeds, contemporaneous volcanic products are either injected into the sedimentary basin or ejected at the surface. These products are largely of basic character but acidic materials are also to be found, especially in ash beds and in material ejected at the surface. In the earlier phases of compression ultramafic rocks are formed along the mountain roots and these often rise to considerable heights into the sedimentary basin. They include peridotites, pyroxenites, eclogites and serpentines. Granitic intrusions come in only in the later stages of compression when the depressed parts of the geosyncline have had time to ultimately melt and form granite and granitic gneiss. The accompanying table indicates the various geological periods in India when there was volcanic activity.

Recent	Volcanic rocks of Barren Island, Narcondam & Baluchisthan-Persian frontier.
Pleistocene Pliocene	Volcanic cones of the median belt of Burma, Yunnan & N. Shan States.
Paleocene to Upper Cretaceous	Deccan Traps. Basic rocks of Indo-Burma frontier. Basic rocks and ash beds of Dras, Kashmir. Volcanics in Cretaceous of Kampa-Dzong.
Lower Cretaceous	Basic ultramafic rocks intrusive into Exotic blocks of Hundes. Agglomerates, basic flows and sills in Baluchistan.
Gondwana (Jurassic?)	Mica Lamprophyre and Dolerite-basalt dykes & sills in Gondwanas; Rajmahal Traps.
Mid-Carboniferous to Trias	Basic volcanic in Everest Pelitic Series. Panjal volcanics of Kashmir (Agglomerates, rhyolites, felsites, meta-dolerites etc.).

Lower Vindhyan	Malani rhyolites and felsites; volcanics in Semri Series, Diamond-bearing ultramafics & agglomerates.
Delhi-Cuddapah	Amphibolites, ash beds etc., in Delhis; Basic sills in Cuddapahs of Andhra, in the Bijawars of Central India; Newer Dolerite of Singhbhum.
Upper Dharwarian	Rhyolites, felsites, meta-dolerites & chlorite schists; Basic rocks in Iron Ore Series of Bihar; Dalma & Dhanjori volcanics.
Lower Dharwarian	Amphibolites, hornblende-schists, greenstones, chlorite schists etc., of Mysore; Basic sills in the Aravallis and in the Shillong Series; Khasi greenstones, etc.

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The International Geophysical Year *

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Geophysics is characteristically an observational science in all its phases, where nature performs the experiments and men observe as a preliminary to understanding. In this way it differs from fields of science which can be investigated under laboratory controls.

The International Geophysical Year, the greatest cooperative scientific undertaking of all time, is dedicated simply to the observation of man's environment—the land on which he lives, the oceans which he fishes and sails, and the atmosphere that he breathes and the influences from outside the earth, particularly the sunshine that supports all life. It is designed to help us understand our environment better so that we may be better able to withstand some of its majestic vicissitudes—earthquakes, ocean tidal waves, atmospheric storms and the vagaries imposed by solar flares and sunspots. It is the vastness of the geophysicist's laboratory—the earth—that lends to geophysics its fascination, that causes the study of geophysics to be a combination of the ideas conjured up in the word combinations, science and art, invention and imagination, research and exploration, doubt and the probing of unknown reaches where as yet we cannot go. It is this same vastness of our environment which necessitates an international cooperative effort on its problems. The oceans belong to no one nation; solar activity and the ionosphere affect all on earth and are problems of such a scale that they transcend international boundaries; droughts, cold and great storms recognize no political lines and blow aside curtains even of iron. This is why the excellent partnership between forty-eight participating nations, and ten or twenty more who will assist in the program, has been achieved. Indeed, the International Geophysical Year of 1957-58 is the third

* Lecture delivered at the Madras University in November 1956.

such cooperative attack on geophysical problems. The first, in 1883, was a joint international effort to collect as simultaneously as possible geophysical data on the then little-known north polar region. This first International Polar Year contributed mightily to our knowledge of geomagnetism and aurora. So successful was it that a Second International Polar Year was held almost exactly fifty years later. With the advances of engineering during those fifty years, the poles had become reasonably accessible and the observations led to many advances in understanding. Now, almost exactly twenty-five years since the second international effort, further advances in engineering and in instrumentation permit us to probe hitherto unknown regions of our environment which should leave no part of our planet geophysically unknown. Instruments which extend, add to, and complement man's own senses and powers of observation play one of their most dramatic roles in the field of geophysics. The deep recesses of the ocean, the impenetrable interior of the lithosphere, and the rarefied reaches of the upper atmosphere are challenging regions for modern imaginative instrumentation.

Because many features of our environment on earth, particularly in the lower and upper atmosphere, are caused or influenced by the sun, one of the unifying features of all the programs in the Geophysical Year is the observation of solar activity. Indeed, it is because this period will be one of near maximum sunspots that it was chosen. One might say that nature picked the date.

The enthusiasm for the International Geophysical Year and the magnificent cooperation of all, not only in our country but in the many others, has far exceeded the most optimistic expectations of those who took part in the early planning. Over five thousand scientists and engineers have been, and are still, preparing throughout the world for the eighteen months of observation, and the total research effort is estimated at about 250 million dollars, with the direct United States appropriation by Congress of nearly 40 million dollars. Just as important as the direct participation of these research workers is the interest and support of lay people. Without this interest the necessary governmental support of the plan conceived by private scientists could hardly come about.

The United States program is but one, though a great one, of the total world effort. It contains thirteen groupings of disciplines and areas of activity. These thirteen may, in turn, be grouped

under three main categories: *physics of the upper atmosphere* (rocket studies, the satellite studies, ionospheric physics, cosmic rays, geomagnetism, aurora and airglow, and solar activity); *lower atmosphere and surface and interface geophysics—the earth's heat and water budget* (meteorology, oceanography, and glaciology); and thirdly, *the earth's crust, interior* and figure (seismology, gravity, and longitude and latitude program). In bearing its share of the coverage of the earth, our program will cover not only the continental limits of the United States and Alaska but will cooperate directly in observations with nations of North and South America and will extend into the waters of the Atlantic, Pacific, and equatorial Pacific Oceans. In addition, the United States will make a tremendous contribution through its program in the Antarctic.

It is interesting to indicate the coherence and essential entity of the United States International Geophysical Year Program. All the elements are global in character and all the fields of geophysics in the program interact with one another and most bear ultimately on a relationship to solar energy fluctuations and disturbances to a greater or lesser degree. Furthermore, fields selected are those in which the solution of major problems will be aided by simultaneous and synoptic observation during the eighteen-month IGY period so that the inter-relationships of the fields, one with another, and with solar activity may be studied. Emphasis is also given to little known regions. The interactions between the fields can best be indicated in terms of major categories.

Interaction of hydrosphere, lithosphere and atmosphere at the interfaces between land and air and land and sea

The interaction is shown by the continual cycle of water from the oceans carried to the continents by the atmospheric winds and back to the sea in the rivers, there to be re-evaporated. "All rivers run into the sea, yet the sea is not full." Also in this process it is couneracting mountain-building forces in the restless interior and crust of the earth which battle with erosion processes to keep land's head above water.

Man's primary dependence on weather is for heat and water, both of which are distributed by the winds of the atmosphere driven by the sun's energy. The total water on earth is primarily in the oceans; on the land in lakes and rivers and in glaciers and ice deposits; and in the atmosphere as vapor and as condensate

in clouds. To understand the over-all interaction in the budget of water and heat it is clearly necessary that a simultaneous closely-knit program in meteorology, oceanography and glaciology be carried out.

Meteorology in the International Geophysical Year will emphasize making up deficiencies of data in the Southern Hemisphere where the relatively unknown ocean regions together with Antarctics comprise over ninety percent of the total area of that hemisphere. It is in this meteorological incognita that we will establish about sixty stations during the IGY. These, in conjunction with already established networks on the inhabited land areas, will give us a suitable network of data for the first time. A second unknown region in meteorology upon a global scale is the high atmosphere up to about a hundred thousand feet. Balloon-borne instruments will fill this gap. The third major objective of the meteorological program is to provide new information on the large-scale global wind circulation. Because the global circulation is predominantly zonal, it will not be able to pass these fences unobserved. Three fences of stations along three meridians will be established from pole to pole to obtain data from the surface up to high altitudes. One of these lines of stations is primarily continental and another almost entirely oceanic.

Just as every heat engine has a warm and a cold side to its workingfluid, so does the atmosphere and the Antarctic program will provide us with daily weather maps of that huge continent which constitutes the major cold pole in the atmospheric heat engine. These, together with the supplemental data for the rest of the Southern Hemisphere, should go far toward a better understanding of the general circulation of the atmosphere.

The oceans which cover four-fifths of the surface of the globe form a huge thermo-accumulator or bank where heat is deposited and stored and withdrawn in times of cold, thus exerting a tempering effect on world climate. The reason that the ocean is able to store great quantities of heat is due to the stirring by the winds. It is the winds too that primarily drive the great oceanic surface circulations which distribute this heat in a manner profoundly affecting climate. The oceanographic program centres on two main objectives—first, to study the very deep ocean currents about which practically nothing is known. This may provide clues for long-range weather and climatic trends and for an understanding of how the fertility of the ocean with its sink of nutrients at the bottom

is maintained by the exchange of water between its surface and the depths. While studying the deeps, a part of the effort will be devoted to delineating the shape of the sometimes craggy ocean floor and the constitution of the earth beneath that floor. Another major effort will be to measure sea level at about a hundred new stations located to fill out the existing tide gauge stations. Sea level is a misnomer perpetrated to confuse. Sea surface height varies from place to place, changes with the season, with the amount transported by the great ocean currents and due to other factors. This study will, among other things, provide valuable information on the transport of water near the surface of the sea.

The meteorological and oceanographic programs will join in attempts to establish how fast the ocean can absorb the great quantities of carbon dioxide which industrialization is pouring into the atmosphere at ever-increasing rates. If this adjustment is not quick enough, man, by changing the infra-red radiation balance, may accelerate the warming up of climate with consequent opening up of water-ways, rise in sea level, and he may unwittingly be embarked on a collosal geophysical control experiment.

Whenever at the surface of the earth the interface with the atmosphere is an ice sheet, the ice, by reflecting incident radiation, where other surfaces absorb, exerts a profound influence on climate. Glaciology is an essential part, therefore, of these interface problems. The amount of glaciation in the past has given us indications of major fluctuations of climate in geological time and an understanding of the retreat or advance of glaciers at present should aid in assessment of climatic trends. In IGY the United States program will attempt to observe not only the condition of that greatest of all ice sheets in Antarctica but will also study many glaciers and ice deposits in the Northern Hemisphere.

The upper reaches and influences from beyond

The denser atmosphere in the troposphere and lower stratosphere is the protective shield against radiations from extra-terrestrial sources, some of which would otherwise be lethal and, as we have seen, it too is the vehicle in the regulation of our heat and water. But many phenomena, fundamental in our environment, occur primarily in the high atmosphere over fifty miles above the earth, where the atmosphere is so rare that an

ordinary vacuum tube would burst outwards. It is in this region, of course, that the sun's activity in the form of flares and eruptions has the most profound effect on the ionosphere, on aurora and on geomagnetism, and it is at these heights that certain of the cosmic rays which are absorbed by the denser atmosphere below can be studied best. The basic study, therefore, in the upper reaches is that of the solar activity in the period near maximum of sunspot activity. The surface of the sun will be scanned in the visible frequencies and its noise will be listened to at radio frequencies on a world-wide, twenty-four hour watch basis to count the number and size of the sunspots and of the flares which shoot out hundreds of thousands of miles and to observe the activity of the corona. The systematic recording of solar events will provide the basis of correlations with ionospheric, geomagnetic, auroral and cosmic ray perturbations. A warning service and special world intervals are planned for those periods of unusual solar activity and solar eclipses when other high atmosphere observations of the associated unusual consequences should be accelerated.

When Marconi sent the first radio signal from Europe to our continent, it was not understood how the radio waves bent around the spherical earth. The existence of the radio wave mirror, the ionosphere, which made this first and, indeed, makes all long-distance radio transmission possible, was only demonstrated experimentally in 1925. Since then the importance of the ionosphere and its study has been intensified.

The objective of the ionospheric program will be to develop a network of both vertically-scanning and oblique-scanning stations which can provide ultimately a global picture of the ionospheric layers much like a world weather map. From such a global picture, prediction of ionospheric conditions with its ultimate application to frequency selection and "routing" of long-range communications is possible. The number of vertical incidence ionospheric stations now operating around the world will be doubled in IGY. The United States likewise will bear its share by doubling the stations which it now operates. The additional stations are carefully located to fill out a world-wide network. Particular emphasis will be placed on the 70° to 80°W. longitude fence (in the same position as one of the meteorological fences); a program in the region of the geomagnetic equator; and a program in Antarctica of particular interest in the long Antarctic night when all radiation from the sun and its influence on the ionosphere is absent.

Both northern lights and southern lights will be watched in the two polar regions. Data on the aurora can tell us a great deal about the streams of ionized particles from the sun which give rise to these magnificent visible displays. Spectroscopic and visual photographic observations of the distribution of aurora made simultaneously with the observations of the face of the sun and of the earth's magnetism will provide the three elements needed for further understanding. At the same time the ionization associated with the aurora, and which also affects radio communications, can be studied by radar. Aurora borealis and aurora australis, if observed simultaneously at the opposite poles, can establish time correlations and show whether occurrences are simultaneous. Time correlation studies such as these will aid in understanding the effects of the earth's magnetic field upon aurora.

The airglow, or weak light, which shines palely from the night sky, which is believed to be caused (in contradistinction to aurora) from chemical reaction in the high atmosphere, varies in time, place and intensity. Airglow will be measured in IGY with highly sensitive instruments distributed along somewhat the longitudinal chain as the ionospheric stations; that is, from the Arctic through North, Central, and South America to the Antarctic. A few other stations will also be located at different longitudes. While not as dramatic in its manifestation as some other atmospheric phenomena, the processes causing airglow may be stimulated by the introduction of catalysts foreshadowing another conceivable geophysical control.

The earth's magnetic field

The eighty magnetic observatories, now spread over the world, record continuously the earth's behaviour as a great magnet. Its magnetic field, while fairly stable and unchanging, not only shows slow variations which, like the permanent field, are connected with dynamo-like interactions of the crust and the interior of the earth, but also rapid fluctuations which are believed to be caused by external influences. What causes this geomagnetic storminess and how it relates to other variations in the upper atmosphere—solar, ionospheric, and auroral—is the main objective in IGY in this field of science. Twelve stations outside the United States (in Alaska, Antarctic and the equatorial Pacific), together with eight stations in the United States, arranged in an

east-west north-south cross, will be the networks of the program in geomagnetism.

In cosmic rays, emphasis will be placed on measurements at high altitudes where the masking effect of our layer of protective atmosphere is at a minimum. Here, too, inter-relations of the various high atmospheric phenomena are intimate but not entirely understood so the disturbances and great variations of cosmic ray showers and variations of the magnetic field of the earth each throw light on the understanding of the other. The magnetic field is a way of analyzing energy of cosmic rays and cosmic rays, in turn, give clues to the distribution of the magnetic field and in particular the location of the geomagnetic equator, and the relationship of both of these to solar flares and periods of sunspot maximum. The origin and nature of cosmic rays are still uncertain and reflect its uncertainty both on the geometrically immense field of astrophysics and the geometrically minute field of the structure of the atomic nucleus.

Crust, core and shape

Seismology, gravity measurements, and the determination of latitude and longitude are three inter-related programs to tell more about the interior of the earth, the disturbances which occur in the interior, and the precise shape and size of our planet. Internal stresses and strains give rise dramatically to earthquakes, ranging from those of small intensity to those which rupture the crust and devastate human communities by their shattering forces and by the resultant floods and tidal waves. Seismology joins with meteorology in the field of microseisms where atmospheric disturbances through oceanic waves cause small tremors to jar the earth. Shall man-made earthquakes by the use of explosives enable seismological techniques to aid glaciology in the determination of the thickness of ice in Antaratica and the as yet unknown profile of the underlying continent, and seismology joins in the same with oceanography to provide knowledge of what is beneath the ocean bottom? The size of the earth, more precise distances between continental and island points separated by oceans (now uncertain by anywhere from 200 feet to a mile) and problems of uniform time and irregular earth turning rate will be some of the objectives of the latitude and longitude program which will use both astronomical techniques, a new technique involving the moon, and possibly the artificial satellite as well.

The shape of the earth is quite complex, departing from a true sphere, and measuring the pull of the earth (which varies from place to place) can tell us much about the shape and the density of materials in the earth's crust. IGY will concentrate on the use of pendulums which are the absolute way to determine gravity in order to provide a gravity standard throughout the world. To tie together the various international networks measurements on ships, on the bottom of oceans, on submarines, and in the Antarctic will all contribute to a greatly enlarged and more comparable world picture of the pull of the earth.

Rockets, rocoons, and rockairs

It is only within the last ten years that the engineering tools have been provided to make direct measurements at altitudes above the ceiling of balloons and up to heights of 200 miles and more. The United States has planned an impressive program using some 600 rockets during the period to measure directly a number of the events in the various inter-locking fields of interest in the high atmosphere. Surface launched rockets to reach 200 miles and up, rocoons (which are smaller rockets launched from balloons and which may be launched from shipboard), and rockairs (which are small rockets released from high-flying airplanes) will all be used—the rocoons for those parts of the program which can be satisfied with a sixty mile altitude and the rockairs to be ready for periods when the sun breaks out in flares and increased activity. In all cases provision will be made for special firings when the sun decides to perform some particularly interesting experiment. While we still make good use of balloons for the lower altitudes, the need to go higher and higher has made the use of rockets common enough to be referred to as a conventional rocket program; but the trouble about conventional rockets is that they only spend a very short time at high altitudes at one small part of the upper atmosphere before they fall back to earth. And for many of the proposed investigations a longer time and greater coverage is needed. It is for this reason that the ambitious and difficult satellite program was added. The need for satellites for high altitude geophysical research has been urged by several international bodies. In contradistinction to the conventional rocket, the satellite will circle the earth in its orbit and make geophysical measurements with world-girdling scope and at greater leisure as it hurtles along at 18,000 miles an hour. Even taking into account the expense and effort needed to make these initial

satellites successful, if one were to assess the cost of the measurements per unit time at high altitude, the satellite would emerge as the cheapest way of obtaining this data.

So much has been written about the satellite that it will suffice here to summarize the main features of this ambitious experiment. Vanguard, the vehicle to carry the satellite up, over 70 feet long and weighing about 11 tons, will be launched from Cape Canaveral in Florida approximately at the end of next year. The first stage of the three-stage rocket carrying the twenty-pound, twenty-to-thirty-inch spherical satellite, is a reworked and improved finless Viking. Liquid oxygen and alcohol and gasoline will thrust the assembly off the earth with a force of 27,000 pounds. Starting up vertically a movable thrust chamber in gimbals will control the direction and gradually tilt the system so that two minutes (140 seconds) and 35 miles up from take-off, when this first stage separates from the system, it will be travelling roughly 45° to the vertical at from 3,000-4,000 miles an hour. The second stage will then ignite immediately and burn nitric acid and dimethyl-hydrazine for an additional two minutes in another gimbal mounted thrust motor. At about 140 miles high, when the system has reached 11,000 miles per hour, the second stage will burn out but will continue with the assembly to control the direction by means of a set of jet reactors while it coasts onward and upward to about 300 miles altitude, 700 miles from the launching point. During this coasting time, the jet reactors will place the last stage in its correct position, and the final stage rocket will be spun to provide directional stability in the manner in which the spin of a rifle bullet provides stability.

The second stage has then done its work and separates. The third stage, an unguided solid propellant rocket, will then fire to speed up the satellite into its orbit at 18,000 miles per hour, which is required so that the centrifugal force offsets the earth's gravitational pull. When the third stage burns out, the bottle will be separated from the little satellite sphere at a relative velocity of only a few feet per second compared to the absolute velocity of 26,000 feet per second. The satellite and this separated bottle will really then form two satellites circling the earth. If the satellite could be launched precisely horizontally, then the launching point could be the point nearest to the earth. But because of inevitable errors it must be launched from a height of 300 miles if the nearest point to the earth is to be not less than 200 miles; and if the satel-

lite dips lower than 200 miles at its nearest point, its life time will be seriously shortened. The farthest point of the elliptic orbit away from the earth will be less than 1500 miles. The orbit and speed is such that the satellite will circle the earth in a plane chosen to be inclined 40° to the equatorial plane. Every part of the earth between latitudes 40° N. and 40° S. will pass under the satellite once every 24 hours, or every 15 revolutions of the satellite. The lifetime of the "little moon" depends on how thin the very rare atmosphere at these heights really is, and so wide are the limits of our ignorance in this respect that estimates of the lifetime range from a minimum of a few days to a maximum of a few years. The most important and immediate result from the first satellite will be the closing of these limits and a real knowledge of the density of the atmosphere at these great heights. The end of the satellite's life comes as the drag of the atmosphere slows down its circumferential speed and as gravity accelerates its motion inwards toward the earth. When it spirals inwards and reaches the denser atmosphere, it will burn up like a shooting star and perhaps leave a trail which itself may give us valuable new data on the winds. Even in a short lifetime the satellite can provide information which it will relay by radio, not only on the atmospheric density but on the earth's shape and size, on the sun's shine in the ultra-violet, on cosmic rays, on the distribution of mass in the earth and numerous other matters.

There is one facet of the program which appeals to me as being of particular and unique significance. This is the effort to excite the interest of amateurs, both here and abroad, to watch the satellite either visually or with radio receivers. The involvement of large groups of amateurs actually performing most important work in the International Geophysical Year will give a further impetus to the general understanding and support of geophysics.

Throughout the program I have stressed the importance that is attached to extending observations in all the geophysical fields to the little-known region of Antarctica. Eleven nations will cooperate directly in the Antarctic program with the United States portion being no less than six stations, including the important one at the South Pole which will serve as a central point not only for its own important scientific work but to aid all other participants. To be ready for the period of intensive observation our expeditions have already established aircraft facilities and carried

instruments, supplies, and other materials to Antarctica in anticipation of the Geophysical Year period.

From the scope of these programs and their inter-relationships, it is clear that a great deal of attention must be given to the best ways in which the millions of observations which will result can be recorded and processed to give maximum availability and usefulness. A center to do this will be put in the United States to supplement centers elsewhere in the world.

Uses and outcome

When we are asked what uses will result from this tremendous effort, it is a difficult question to answer, not because the uses are few but because they are so many. The direct and obvious contributions of the increased knowledge toward questions such as climatic trends, melting of glaciers with the consequent rise in sea level, improved long-range radio transmission, improved understanding and possibly prediction of violent events (such as storms, earthquakes, tidal waves and the like) which affect the daily life and activity of everyone including world commerce and industry, and transport by land, sea and air. But in addition to these and many others that you may add, suggestions for the answers of thousands of other questions will emerge and results of the International Geophysical Year will lay a new groundwork of understanding that will be practically exploited in world-scale engineering applications in the future. There are many instances of man exerting, through lack of understanding, inadvertent and sometimes harmful control of his environment. It is not out of the question to suggest that with the far greater understanding that will result from this tremendous international effort he should be able to better predict, and therefore exert some planned beneficial control on some of the environmental factors which determines his living.

I have emphasized that the International Geophysical Year achieved its successful co-operation through the efforts of non-governmental international organizations. This is not surprising because co-operation is essentially voluntary. But it is well also to stress that it gains strength by the tremendous support in our country, as in others, of the unique government laboratories in the various fields, without which support it is quite clear that many of these programs would not be possible. It is interesting, too, to

note in the International Geophysical Year, it is the most modern engineering tools and vehicles and instruments which made them possible—the rockets, the satellites, the calm occupation of the Antarctic using every boon of modern engineering to overcome the rigors of nature in that inhospitable part of the world. The new understanding of fundamental geophysical processes which should result will in turn lead to applications which we cannot foresee.

Finally, I believe that the International Geophysical Year may set a pattern in world science which might well be emulated in other areas. The medical people are beginning to talk of an International Geopathological Year. If this and others can have a measure of the success which we anticipate in the IGY, the sum total cannot fail to make its impact felt toward the improvement of international co-operation in general.

J. Madras Univ., B, 27, No. 1. Centenary Number, 1957.

Zur Vollständigen Reduzibilität Der Linearen Darstellungen Halbeinfacher Gruppen und zum E. E. Levischen Satz.

VON

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Für die vollständige Reduzibilität der linearen Darstellungen halbeinfacher Gruppen kennt man den Weylschen Integrationsbeweis und rein algebraische Beweise — im Infinitesimalring — von H. Casimir und B. L. v. d. Waerden[1] und von J. H. C. Whitehead.[2]

Der hier folgende Beweis, den ich 1947 auf einer Konferenz in Strassburg vorgetragen habe, bringt methodisch kaum etwas Neues. Am meisten ähnelt er dem Whiteheadschen. Trotzdem möchte ich es wagen, ihn zu publizieren.

Zum Schluss beweise ich nach dem Whiteheadschen Vorbild in diesem Zusammenhang auch den Satz von E. E. Levi.

1. Γ ist eine halbeinfache Lie-Algebra der Dimension r, $\operatorname{ad} Z \cdot X = [Z, X]$, $(X, Y) = \operatorname{sp} (\operatorname{ad} X \operatorname{ad} Y)$ ein nicht-entartetes inneres Produkt, das gegenüber den Transformationen der adjungierten Algebra invariant ist:

$$([Z, X], Y) + (X, [Z, Y]) = 0.$$

ad Z und das innere Produkt lassen sich in der k-fachen Tensoralgebra T^k über Γ definieren; man braucht das nur für die $X_1 \times \ldots \times X_k$† zu tun:

$$\operatorname{ad} Z\,(X_1 \times \cdots \times X_k) = \operatorname{ad} Z \cdot X_1 \times \cdots \times X_k + \cdots + X_1 \times \cdots \times \operatorname{ad} Z \cdot X_k,$$
$$(X_1 \times \cdots \times X_k,\ Y_1 \times \cdots \times Y_k) = (X_1, Y_1) \cdots (X_k, Y_k).$$

Dieses innere Produkt ist nicht-entartet und gegenüber den ad Z invariant.

† Für das Tensorprodukt habe ich das Zeichen $\times$ verwendet, da das übliche "Rad" in der Druckerei nicht vorrätig war.

1. *Mathematische Annalen*, 111 (1935); 1-12.

2. *Proc. Camb. phil. Soc.*, 32 (1936): 229-237; *Quart. J. Math.*, 8 (1937): 220-237.

Der symmetrische Teil S^k von T^k besteht aus den linearen Kombinationen der

$$X \times X \times \cdots \times X.$$

Er ist gegenüber den ad Z invariant.

Zu einer auf S^k definierten linearen Funktion f definiert

$(F, P) = fP$ für alle $P \varepsilon S^k$

eindeutig ein Element $F \varepsilon S^k$. Ist f gegenüber den ad Z invariant, also

$$f(\mathrm{ad}\, Z \cdot P) = 0 \qquad \text{für alle } P \varepsilon S^k,$$

so ist auch $(F, \mathrm{ad}\, Z \cdot P) = 0$ „ ,

also $(\mathrm{ad}\, Z \cdot F, P) = 0$ „ ,

also

$$(1{\cdot}1) \qquad \mathrm{ad}\, Z \cdot F = 0.$$

Ein gegenüber den ad Z invariantes f erhält man, wenn eine lineare Darstellung

$$X \to \hat{X}$$

von Γ in einem Raum R gegeben ist und man

$$f(X \times X \times \ldots \times X) = \mathrm{sp}\ (\hat{X}^k)$$

setzt. Speziell wichtig ist der Fall, dass $k = 2$, $R = \Gamma$ und $\hat{X} = \mathrm{ad}\, X$ ist. F ist dann der Casimir-Operator G.

Durch die lineare Darstellung von Γ im Raum R,

$$X \to \hat{X},$$

wird eine lineare Abbildung

$$P \to \hat{P}$$

von T^k in den Raum R^R der linearen Abbildungen von R in sich induziert mit der Eigenschaft

$$X_1 \times \cdots \times X_k \to \hat{X}_1 \cdots \hat{X}_k.$$

Gehört F zu einem invarianten f, so gilt nach (1·1)

$$[\hat{Z}, \hat{F}] = 0;$$

ist die Darstellung irreduzibel, so ist $\hat{F}$ also eine Multiplikation mit einem Skalar, de sich für den Fall des Casimir-Operators mittels des höchsten Gewichtes p und der Summe δ der positiven Wurzeln zu

$$(p, p + 2\delta)$$

berechnet,[3] also (ausser für die Nulldarstellung) positiv ist. Die Spur von $\hat{G}$ ist demnach ausser bei der Nulldarstellung immer positiv.

Man kann das Vorige in verschiedenen Richtungen verallgemeinern.

Gegeben seien k lineare Darstellungen von Γ im *selben* Raume R

$$X \to \varrho_i(X), \quad i = 1, \cdots, k.$$

Durch

$$X_1 \times \cdots \times X_k \to \varrho_1(X_1) \cdots \varrho_k(X_k)$$

wird wieder eine lineare Abbildung

$$P \to \hat{P}$$

von Γ in R^R induziert. Will man nun 1·1 anwenden, so muss man beachten, das

$$\sum_i X \times \cdots \times \operatorname{ad} Z \cdot X_i \times \cdots \times X_k \to \sum_i \varrho_1(X_1) \cdots [\varrho_i(Z), \varrho_i(X_i)] \cdots \varrho_k(X_k)$$

zu setzen ist.

3. H. Freudenthal, *Proc. Kon. Akad. Wet.* A, **57** (1954): 369-376.

Noch allgemeiner: Man braucht nicht einmal zu wissen, dass die $X \to \varrho_i(X)$ Darstellungen von Γ sind. Es genügen lineare Abbildungen von Γ in R^R; nur darf man dann in

$$\sum_i \varrho_1(X_1) \cdots \varrho_i([Z, X_i]) \cdots \varrho_k(X_k)$$

$\varrho_i([Z, X_i])$ nicht durch $[\varrho_i(Z), \varrho_i(X_i)]$ ersetzen.

2. Seien nun zwei lineare Darstellungen von Γ im selben Raum R gegeben,

$$X \to \hat{X} \varepsilon R^R,$$

$$X \to \tilde{X} \varepsilon R^R.$$

Durch

$$X \times Y \to \hat{X}\tilde{Y} - \tilde{X}\hat{Y}$$

wird eine lineare Abbildung von T^2 in R^R bestimmt, bei der

$$G \to \Phi$$

abgebildet wird.

Wir wählen in Γ zwei Basen

$$X_1, \ldots, X_r \text{ und } X^1, \ldots, X^r,$$

so dass

$$X_i = \text{sp } (\text{ad } X_i \text{ ad } X_j) X^j$$

ist (Einstein-Konvention!). Dann wird

$$G = X_i \times X^i = X^i \times X_i,$$

$$\Phi = [\hat{X}_i, \tilde{X}^i] = [\hat{X}^i, \tilde{X}_i],$$

und (1·1) liefert

$$[[\hat{Z}, \hat{X}_i], \tilde{X}^i] + [\hat{X}_i, [\tilde{Z}, \tilde{X}^i]] = 0.$$

Auf den ersten Summanden wenden wir "Jacobi" an und erhalten:

$$[[\hat{X}_i, \tilde{X}^i], \hat{Z}] + [[\tilde{X}_i, \hat{Z}], \hat{X}^i] = [\hat{X}_i, [\tilde{Z}, \tilde{X}^i]],$$

also

$$(2\cdot1) \quad [\hat{Z}, \Phi] = [\hat{X}_i, [\tilde{X}^i, \tilde{Z} - \hat{Z}]].$$

3. Wir betrachten nun eine Darstellung von Γ in R,

$$X \to \tilde{X},$$

mit einem invarianten Teilraum L. Es gibt dann einen linearen Teilraum M und eine Darstellung von Γ in R,

$$X \to \hat{X},$$

so dass $L + M = R$, $L \cap M = (0)$,

$$\hat{X}M \subset M,$$

$$\tilde{X} = \hat{X} \text{ auf } L,$$

$$(\tilde{X} - \hat{X})\, M \subset L$$

ist. Es ist unsere Aufgabe, die Äquivalenz beider Darstellungen zu erweisen. Dabei dürfen wir noch die Annahme machen:

$\tilde{X}$ irreduzibel auf L,

$\tilde{X}$ irreduzibel auf M.

Der Raum der linearen Abbildungen von R in R, bei denen M in L und L in 0 abgebildet wird, heisse P. Es ist $PP = 0$, also $[P, P] = 0$. Die $\tilde{X} - \hat{X}$ sind in P. Wir suchen ein $K \varepsilon P$, so dass

$$(1 - K)\, \hat{X}\, (1 + K) = \tilde{X} \text{ für alle } X \varepsilon \Gamma$$

ist; wegen $(1-K)\ (1+K)=1$ genügt das zum Beweise. Die Bedingung für K kann auch

$$(3\cdot1)\quad [\hat{X},\ K]=\tilde{X}-\hat{X}$$

geschrieben werden.

In **P** definiert man eine Darstellung von Γ,

$$X\to\overset{\vee}{X}$$

durch

$$\overset{\vee}{X}P=[\hat{X},\ P]\ (=[\tilde{X},\ P]).$$

Dann kann man $(2\cdot1)$ auch

$$(3\cdot2)\quad \overset{\vee}{Z}\Phi=\overset{\vee}{X}_i\overset{\vee}{X}^i\ (\tilde{Z}-\hat{Z})=\overset{\vee}{G}(\tilde{Z}-\hat{Z})$$

schreiben, und die Forderung $(3\cdot1)$ wird

$$(3\cdot3)\quad \overset{\vee}{X}K=\tilde{X}-\hat{X}.$$

Mit $3\cdot2$ haben wir das Ziel $3\cdot3$ schon beinahe erreicht; wir müssen $\overset{\vee}{G}$ noch wegschaffen.

Nach $(1\cdot1)$ ist

$$(3\cdot4)\quad \overset{\vee}{G}\overset{\vee}{X}=\overset{\vee}{X}\overset{\vee}{G}.$$

Wir werden versuchen, die Gleichung

$$(3\cdot5)\quad \overset{\vee}{G}K=\Phi$$

zu lösen; setzen wir das in $3\cdot2$ ein, so ergibt sich nämlich

$$(3\cdot6)\quad \overset{\vee}{G}\overset{\vee}{X}K=\overset{\vee}{G}(\tilde{X}-\hat{X}),$$

und das heisst, dass

$$(3\cdot7)\quad \overset{\vee}{X}K-(\tilde{X}-\hat{X})$$

durch $\overset{\vee}{G}$ annulliert wird, also jedenfalls gelegen ist im Teilraum $\mathbf{N}$ von $\mathbf{P}$, der zum Eigenwert 0 von $\hat{G}$ gehört.

Wegen (3·4) ist $\mathbf{N}$ bei den $\overset{\vee}{X}$ invariant. $\overset{\vee}{G}$ hat in $\mathbf{N}$ die Spur 0. Nach 1 ist $X \to \overset{\vee}{X}$ da die Nulldarstellung, also

$$(3\cdot8)\quad \overset{\vee}{X}\mathbf{N} = 0,$$

$$\hat{X}P = P\hat{X} \text{ für } P \varepsilon \mathbf{N},\ X \varepsilon \Gamma.$$

Wegen $P\,M \subset L$ und der Irreduzibilität von $\hat{X}$ in L und M bedeutet das:

$P = 0$, oder P vermittelt eine Äquivalenz zwischen den Darstellungen in L und M. Im ersten Fall ist $\mathbf{N} = (0)$, im zweiten Falle besteht es aus den skalaren Vielfachen eines $P \neq 0$.

Im ersten Fall besitzt $\overset{\vee}{G}$ ein Inverses, 3·5 ist lösbar, und 3·7 verschwindet, womit der Forderung 3·3 genügt ist. Im zweiten Fall ist 3·5 wenigstens mod $\mathbf{N}$ lösbar, aber das führt nach 3·8 immer noch zu 3·6 und dann zu

$$(3\cdot9)\quad \overset{\vee}{X}K = \tilde{X} - \hat{X} + \alpha(X)P.$$

Berechnet man

$$[\overset{\vee}{X}\overset{\vee}{Y}]K = \overset{\vee}{X}\overset{\vee}{Y}K - \overset{\vee}{Y}\overset{\vee}{X}K,$$

so erhält man wegen 3·8 in 3·9 rechts

$$[\hat{X}, \tilde{Y} - \hat{Y}] - [\hat{Y}, \tilde{X} - \hat{X}]$$

$$= [\tilde{X}, \tilde{Y} - \hat{Y}] - [\hat{Y}, \tilde{X} - \hat{X}]$$

$$= [X, \tilde{Y}] - [X, \hat{Y}].$$

Also verschwindet α auf der Kommutatoralgebra von Γ, die aber mit Γ identisch ist.

Hiermit ist der Forderung 3·3 Genüge geleistet und der Beweis der vollständigen Reduzibilität erbracht.

4. Γ sei eine Lie-Algebra linearer Transformationen des linearen Raumes R. Γ lasse einen linearen Unterraum L invariant. Man kann dann einen linearen Unterraum M von R finden und eine homomorphe Abbildung $X \to \hat{X}$ von Γ auf eine geeignete Lie-Algebra $\hat{\Gamma}$ linearer Transformationen von R, so dass

$$(4\cdot1)\quad L + M = R,\ L \cap M = (0),$$

$$(4\cdot2)\quad \hat{\Gamma} M \subset M,$$

$$(4\cdot3)\quad X = \hat{X} \text{ auf } L,$$

$$(4\cdot4)\quad (X - \hat{X})\ R \subset L$$

ist. $\hat{\Gamma}$ ist bis auf Äquivalenz eindeutig bestimmt.

Wir nehmen nun an, dass $\hat{\Gamma}$ halbeinfach und auf L sowie auf M irreduzibel ist.

Wir behaupten dann: Man kann $\hat{\Gamma}$ auch als Unter-Lie-Algebra von Γ wählen.

5. Der Raum der linearen Abbildungen, die R in L und L in 0 abbilden, heisse wieder **P**.

Dann ist also $X - \hat{X} \varepsilon \mathbf{P}$. Es gilt $\mathbf{P}\mathbf{P} = 0$, also $[\mathbf{P}, \mathbf{P}] = 0$.

Der Kern von $X \to \hat{X}$ heisse Δ. Für $X \varepsilon \Delta$ ist $\hat{X} = 0$, also $X \varepsilon \mathbf{P}$. Somit $\Delta \subset \mathbf{P}$, Δ abelsch.

$\tilde{\Gamma}$ sei ein linearer Unterraum von Γ mit

$$\tilde{\Gamma} + \Delta = \Gamma,\ \tilde{\Gamma} \cap \Delta = (0).$$

Zu $X \varepsilon \Gamma$ gehört ein wohlbestimmtes $\tilde{X} \varepsilon \tilde{\Gamma}$ mit $\tilde{X} \equiv X \bmod \Delta$. Es ist dann auch $\tilde{X} - \hat{X} \varepsilon \mathbf{P}$.

$$[X, \tilde{Y}] = [\tilde{X}, \tilde{Y}] + A(X, Y) \text{ mit } A(X, Y) \varepsilon \Delta.$$

Wir werden das gesuchte neue $\hat{\Gamma}$ konstruieren, indem wir (für alle Z) $\hat{Z}$ ersetzen durch

$$\hat{Z} + [\hat{Z}, Q] \text{ mit geeignetem } Q \varepsilon \mathbf{P}.$$

4·3-4 wird dadurch nicht beeinträchtigt. Wegen

$$[\hat{X} + [\hat{X}, Q], \hat{Y} + [\hat{Y}, Q]]$$
$$= [\hat{X}, \hat{Y}] + [\hat{X}, [\hat{Y}, Q]] - [\hat{Y}, [\hat{X}, Q]]$$
$$= [\hat{X}, \hat{Y}] + [[\hat{X}, \hat{Y}], Q]$$

ist auch das neue $\hat{\Gamma}$ eine Lie-Algebra. Man muss nur dafür sorgen, dass

$$\hat{Z} + [\hat{Z}, Q] \varepsilon \Gamma$$

ist. Ein neues M, das 4·2 genügt, findet sich dann von selber wegen der vollständigen Reduzibilität von $\hat{\Gamma}$.

Wir wählen zwei Bases für $\hat{\Gamma}$ wie in 2, und entsprechend für $\tilde{\Gamma}$ und setzen

$$\Phi = [\hat{X}_i, \tilde{X}^i] = [\hat{X}_i, \tilde{X}^i - \hat{X}^i] \varepsilon \mathbf{P}.$$

Dann wird

$$[[\hat{Z}, \hat{X}_i], \tilde{X}^i] + [\hat{X}_i, [Z, \tilde{X}^i]] = 0.$$

$$[[\hat{Z}, \hat{X}_i], \tilde{X}^i] + [\hat{X}_i, [\tilde{Z}, \tilde{X}^i] + A(Z, X^i)] = 0.$$

Anwendung von "Jacobi" im ersten Summand ergibt

$$[\hat{Z}, \Phi] + [\hat{X}_i, [\tilde{X}^i, \hat{Z} - \tilde{Z}]] + [\hat{X}_i, A(Z, X^i)] = 0.$$

Wie in 3 definiert

$$\overset{\vee}{G}\, C = [\hat{X}_i, [\hat{X}^i, C]] = [\hat{X}_i, [\tilde{X}_i, C]] \text{ für } C \varepsilon \mathbf{P}$$

eine lineare Abbildung $\overset{\vee}{G}$ von $\mathbf{P}$ in sich, deren Nullraum aus den skalaren Vielfachen eines P (eventuell $= 0$) besteht. P kommutiert mit den $\hat{Z}$. $\overset{\vee}{G}$ kommutiert mit den $\hat{Z}$ und lässt Δ invariant. Man kann also die Gleichungen

$$\Phi \equiv \overset{\vee}{G}\, Q \bmod P, \quad Q \varepsilon \overset{\vee}{G}\, \mathbf{P},$$

$$A(Z, X^i) \equiv \overset{\vee}{G}\, Q^i \bmod P, \quad Q^i \varepsilon \Delta \cap \overset{\vee}{G}\, \mathbf{P}$$

befriedigen und erhält wegen $[\hat{Z}, P] = [\hat{X}, P] = 0$:

$$[\hat{Z}, \overset{\vee}{G}\, Q] + \overset{\vee}{G}(\hat{Z} - \tilde{Z}) + [\hat{X}_i, \overset{\vee}{G}\, Q^i] = 0.$$

Wegen der Vertauschbarkeit von $\overset{\vee}{G}$ mit den $\hat{Z}$ ist dann auch

$$\overset{\vee}{G}([\hat{Z}, Q] + \hat{Z} - \tilde{Z} + [\hat{X}_i, Q^i]) = 0,$$

$$\tilde{Z} - [\hat{X}_i, Q^i] = \tilde{Z} - [\tilde{X}_i, Q^i] \varepsilon \Gamma,$$

also $\hat{Z} + [\hat{Z}, Q] + \alpha(\hat{Z})\ P \varepsilon \Gamma$

für eine gewisse lineare Funktion α. Bildet man von zwei solchen Ausdrücken den Kommutator, so sieht man, dass $\alpha([\hat{X}, \hat{Y}]) = 0$ ist, also bereits

$$\hat{Z} + [\hat{Z}, Q] \varepsilon \Gamma,$$

was wir erreichen wollten.

6. Der Satz von E. E. Levi besagt:

Ist Γ eine Lie-Algebra und Δ ihr maximales auflösbares Ideal, so gibt es eine (halbeinfache) Unter-Lie-Algebra Ω mit $\Gamma = \Delta + \Omega$, $\Delta \cap \Omega = (0)$.

Γ mod Δ ist jedenfalls halbeinfach.

Annahme 1: Γ mod Δ ist obendrein einfach.

Sei Δ_o $(\neq (0))$ ein minimales auflösbares Ideal von Γ.

Annahme 2: $\Delta = \Delta_o$ (also sicher abelsch).

Wir wollen 4 anwenden, wobei ad Γ die Rolle von Γ aus 4 spielen soll. Δ ist bei ad Γ invariant und entspricht L. Für $X \varepsilon \Delta$ ist (ad X) $\Delta = 0$, also ad $\hat{X} = 0$. Dem $\hat{\Gamma}$ aus 4 entspricht also Γ mod Δ, das halbeinfach ist. Wegen der Minimalität von Δ ist ad Γ auf Δ irreduzibel. $\hat{\Gamma}$ auf M ist im Wesentlichen ad (Γ mod Δ), also auch irreduzibel.

Die Voraussetzungen von 4 sind erfüllt. Das ergibt für ad Γ eine Zerlegung $\Delta' + \Omega'$, wo Δ' maximal auflösbares Ideal, Ω' eine Lie-Algebra und $\Delta' \cap \Omega' = (0)$ ist. Die Originale von Δ', Ω' bei dem Homomorphismus $X \to \text{ad } X$ seien Δ_1, Ω_1. Dann ist $\Gamma = \Delta_1 + \Omega_1$, $\Delta_1 \cap \Omega_1$ ist das Zentrum von Γ, $\Omega' \cong \Omega_1$ mod $(\Delta_1 \cap \Omega_1)$ ist halbeinfach, also gibt es eine Lie-Algebra Ω mit $\Omega_1 = \Omega + (\Delta_1 \cap \Omega_1)$, $\Omega \cap \Delta_1 = (0)$. $\Gamma = \Delta_1 + \Omega$. Damit ist der Satz unter den Annahmen 1-2 bewiesen.

Wir machen nun statt Annahme 2 die

Annahme 2': $\Delta \neq \Delta_o$.

Wir führen einen Induktionsschluss hinsichtlich dim Γ. Insbesondere dürfen wir die Richtigkeit des zu beweisenden Satzes für Γ mod Δ_o voraussetzen. Es gibt also eine Lie-Algebra Ω' mit $\Gamma = \Delta + \Omega'$, $\Delta \cap \Omega' = \Delta_o$. Ω' mod Δ_o ist mit Γ mod Δ isomorph, also halbeinfach, also ist Δ_o maximal auflösbares Ideal von Ω', also existiert bereits (wegen dim $\Omega' <$ dim Γ) eine Lie-Algebra Ω und eine Zerlegung $\Omega' = \Delta_o + \Omega$, $\Delta_o \cap \Omega = (0)$. Nun ist $\Gamma = \Delta + \Omega$ die gewünschte Zerlegung von Γ.

Statt Annahme 1 machen wir nun

Annahme 1′: Γ mod Δ ist nicht einfach.

Wir verfahren wieder induktiv. Es gibt nun echte Ideale Γ_1, Γ_2 von Γ mit $\Gamma = \Gamma_1 + \Gamma_2$, $\Gamma_1 \cap \Gamma_2 = \Delta$. Induktiv ist die Existenz einer halbeinfachen Lie-Algebra Ω_1 mit $\Gamma_1 = \Delta + \Omega_1$, $\Delta \cap \Omega_1 = (0)$ gegeben. Die adjungierte Darstellung von Γ_1 induziert in Γ_2 eine Darstellung von Ω_1. Wegen der vollständigen Reduzibilität und wegen $[\Omega_1, \Gamma_2] \subset [\Gamma_1, \Gamma_2] \subset \Delta$ gibt es einen linearen Unterraum K von Γ_2 und eine Zerlegung $\Gamma_2 = \Delta + K$, $\Delta \cap K = (0)$, mit $[\Omega_1, K] = (0)$. K erzeugt eine Lie-Algebra K′, für die auch noch $[\Omega_1, K'] = 0$ gilt. $K \subset K' \subset \Gamma_2 = \Delta + K$, also gibt es einen linearen Unterraum $\Delta' \subset \Delta$ mit $K' = \Delta' + K$. Nun ist $[\Delta', \Delta'] \subset \Delta$, aber auch $\subset K'$, also $\subset \Delta'$. Daher ist Δ' maximal auflösbares Ideal von K′, und $K' = \Delta' + \Omega_2$ mit $\Delta' \cap \Omega_2 = (0)$, $[\Omega_2, \Omega_2] \subset \Omega_2$. Ferner war $[\Omega_1, K'] = (0)$, also $[\Omega_1, \Omega_2] = 0$. $\Omega = \Omega_1 + \Omega_2$ ist die gesuchte Lie-Algebra.

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The Elementary Theory of Population Growth

BY

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This article demands an apology. I ignore much previously published work, since the biological journals in which it has been published are not available in this Institute, and are scattered through various libraries in India.

I shall only deal with the following elementary problem. "How quickly would a population increase or decrease if mortality and fertility as a function of age remained constant over a period of several generations?" This may seem hopelessly artificial. However, animals introduced into a new area do, in fact, sometimes increase in geometrical progression for five or ten generations. Lack gives examples. And we can reasonably ask how quickly the population of India would increase at the present levels of fertility and mortality, even if we hope that both will fall in the next twenty years. We can make similar calculations for hypothetical fertility and mortality rates.

Annual plants or animals:

Many plants and animals have a maximum life of one year. They reproduce during a short period of a few days or weeks, and then die. There is usually some stage in the life cycle when they are most easily counted. For example, the Lepidoptera which occasionally increase to such an extent that they defoliate European pine woods often spend the winter in the soil as pupae. The pupae can be counted in some hundreds of separate square metres of soil, and the density in different years compared. It would be vastly harder to count all the caterpillars even on a single tree, and impossible to count all the flying insects, though their numbers might be estimated by systematic capture, marking, release and recapture. A biologist would like to be able to estimate the number of eggs laid in a given area, which is difficult, or the number of seeds produced, which is much easier, and then to follow up their

Lack, D. (1954). *The natural regulation of animal numbers*, Oxford.

mortalities from various causes through a year; and also to estimate the number of eggs laid per adult female. This has only been done thoroughly in one case. Varley (1947) studied a gall-forming insect, examining the content of over 50,000 galls. Only about 1% of the eggs laid developed into adults, and these laid about 200 eggs each on an average. If it were possible to follow up the life of several thousand female eggs, tabulating the number of female eggs, r_k, laid by the female arising from the k-th egg, we should find that was usually zero, but occasionally several hundred, but that the average value R was not far from unity, though it might be 5 in a very favourable year, and 0·2 in a very unfavourable one. If we followed up females as they emerged as adults, counting the number r'_k of their female progeny emerging next year, we should probably find that in most years r'_k was only zero in about half of all cases, and very rarely exceeded ten, its average value R′ being again near to unity.

When an insect is studied in an area where its ancestors have lived for some time and where conditions are fairly steady, the values of R or R′ may be large or small in any particular year, but their geometric mean must be very close to unity, or the mean of their logarithms very close to zero. For example, if the geometric mean of R were 1·01 over 1000 years, the density would increase more than 20,000 times. Long before this occurred, hunger, disease, predation, or some other cause would prevent further increase.

But if an insect is put in a new environment, either because it is introduced into a new area, or because a new crop or a new enemy is introduced into the area where it has lived for some time, R may differ from unity, and remain fairly steady for some years. If d_n is the density per unit area in year x (e.g., 23,000 per hectare in 1953); then

$$d_{n+1} = \mathrm{R}d_n,$$

$$\text{and } d_{n+k} = \mathrm{R}^k d_n,$$

$$\text{or } d_n = \mathrm{A}\mathrm{R}^n.$$

The density increases in a geometric progression if R exceeds unity, and falls off in a geometric progression if R is less than unity. Thus, with a new crop R might be about 2. The density would double in each year, and increase 1000 times in ten years, so that the insect became a serious pest.

Birds:

It is not difficult to count the number of nests and, therefore, the number of adult pairs, of many bird species in a given area. This has been done in many parts of Europe and North America. It could and should be done in India both for conspicuous birds such as vultures and cattle egrets, and for small birds with conspicuous nests such as weaver birds. One can also count the numbers of eggs and young birds in a number of nests, and estimate the mortality by putting metal rings on the legs of nestlings, and counting the numbers of these rings collected in subsequent years. If a small reward is offered for each ring sent to a given address this method becomes possible as soon as a rural population is literate. The fraction of rings returned in Britain varies from under 1% for some small birds to 15% for fairly large ones. It is found that the mortality is high in the first year, but after this it remains constant. Birds do not die of old age in nature. A starling (a bird related to your mynah, of about the same size) is no more likely to die in the next year when it is six years old than when it is one year old, though since only about half the adult birds survive till next year, not many live till the age of six years. The annual survival varies from about 35% for some small birds to 80% or more for some sea birds which lay one egg per year.

Let us consider a very simple case. Suppose that a bird can breed at the age of one year, as most of our song-birds can, that a fraction s of adults survives till next year, and that on an average a pair produce f progeny which survive for one year. Then if the sexes occur in equal numbers, and there are P_n adult pairs in year n,

$$\begin{aligned} P_{n+1} &= s\,P_n + \tfrac{1}{2} f\,P_n \\ &= (s + \tfrac{1}{2} f)\,P_n. \end{aligned}$$

For a fraction s survive till year $n+1$, and in addition there are $\frac{1}{2}f$ new pairs for each pair in the preceding year. So $P_n = A\,(s+\frac{1}{2}f)^n$.

Thus, the population increases or decreases in a geometrical progression. In fact, after a year of heavy mortality f increases since more food is available, so the population density soon returns to a nearly constant level. But in the long run $s + \frac{1}{2}f = 1$ nearly. So, for example, if $s = \frac{1}{2}$, $f = 1$. The statistics bear this out. Thus, for the starling s is lower in Switzerland than in England, but the average number of eggs laid is higher.

Large birds do not become mature and begin to nest till they are more than a year old. The age is about 4 for some British sea birds, and 9 for the albatross. The mathematical theory is more complicated, and has been given by Capildeo and Haldane (1954). Let me give an example of its application. The fulmar (*Fulmarus glacialis*) is a sea bird which only lays one egg per year and matures at 4 years. About 20 years ago the British ornithologist James Fisher asked me a ridiculous question. "If fulmars were immortal, and we started with a population of ten nesting pairs, how many nesting pairs would there be after 5, 10, 15 and 20 years?" I gave him the answer. It turned out that the number of pairs nesting on a particular headland of an island off Scotland had increased more rapidly than the rate given by my calculation. Thus, it was certain that birds had immigrated into the community counted. This is of interest, because some animal communities will not admit immigrants.

Continuous reproduction

Many animals, including human beings, can reproduce at any time of the year, and, what is more important, at any age within a certain range. If fertility is merely a function of age, the following analysis holds. Let t be the time, e.g., 1953-1956.

Let x be the age from fertilization, or from some other well defined event, e.g., emergence from a pupa or first menstruation. For the moment we shall consider this event to be birth.

Let $\mathrm{N}(t)$ be the number of females in an area at time t.

Let $\mathrm{B}(t)\,\delta t$ be the number of females born (or reaching the age taken as zero) between times t and $t + \delta t$.

Let l_x be the probability of a female surviving to age x.

Let $\mathrm{F}(x)\,\delta x$ be the probability that a female will bear a female child between the ages x and $x + \delta x$.

Let $\mathrm{R} = \int_0^\infty \mathrm{F}(x)\, l_x dx$.

Let $\varrho = log_e\, \mathrm{R}$.

Let $f(x) = \mathrm{R}^{-1}\mathrm{F}(x)\, l_x$.

Then $\mathrm{N}(t) = \int_0^\infty \mathrm{B}(t - x)\, l_x dx$. .. (1)

For at time t a fraction l_x of the females born x years ago, that is to say at time $t-x$, are still alive.

Consider the females born x years ago. Only a fraction of them are alive now, and the number of females born between ages x and $x+\delta x$ to n females born x years ago is $nl_x\mathrm{F}(x)\delta x$. So in the course of their lives these females bear n R daughters. $\mathrm{F}(x)$ is called the gross fertility at age x, $l_x\mathrm{F}(x)$ is called the net fertility, and R is called the net reproductive index. If R exceeds unity the population will ultimately increase, even though it may fall temporarily if most of the females are immature. Similarly, if R is less than unity the population will ultimately fall.

To find out how quickly it will increase or decrease we note that $\mathrm{B}(t) = \int_o^\infty \mathrm{B}(t-x)\mathrm{F}(x)l_x dx$

$$= \mathrm{R}\int_o^\infty \mathrm{B}(t-x)f(x)\,dx \qquad .. \quad (2)$$

For the females born between times $t-x$ and $t-x+\delta x$ have a net fertility $\mathrm{F}(x)l_x$. We see that $\int_o^\infty f(x)\,dx = 1$, so $f(x)$ is a distribution function. It can be called the maternal age distribution function. It is not, however, the distribution function for the ages of mothers bearing daughters in a given year. It is the distribution function for the ages at which all women born, say, in 1900, bore daughters. The two distributions are only the same if the population is stationary. In an increasing population such as that of India the mean age of mothers who bore children in 1955 is less than the mean of the maternal age distribution because the fraction of young mothers is greater than it would be in a stationary population.

To solve equation (2) we note that it is a linear equation. For it is easily shown that if $\mathrm{B}_1(t)$ and $\mathrm{B}_2(t)$ are solutions, $\mathrm{B}_1(t)+\mathrm{B}_2(t)$ is also a solution. Let $\mathrm{B}(t) = \mathrm{A}e^{-\lambda t}$ be a solution. Then

$$\mathrm{A}e^{-\lambda t} = \mathrm{R}\int_o^\infty \mathrm{A}e^{\lambda(x-t)}f(x)\,dx,$$

$$\text{or} \quad \int_o^\infty e^{\lambda x}f(x)\,dx = \mathrm{R}^{-1} \qquad .. \quad (3)$$

The left hand side of equation (3) is the moment generating function of the distribution $f(x)$. It is a monotone increasing function of real λ, so it has only one real root of λ, though in

general an infinity of complex roots. To solve it we take logarithms. The natural logarithm of the moment generating function is the cumulant generating function. So

$$K_1\lambda + \frac{K_2\lambda^2}{2!} + \frac{K_3\lambda^3}{3!} + \frac{K_4\lambda^4}{4!} + \ldots = -\varrho + 2k\pi i.$$

Here K_1, K_2, K_3, etc., are the cumulants of the distribution of maternal ages. K_1 is the mean, K_2 the variance, K_3 the third moment about the mean, and so on. k is any integer. This equation is due to Lotka.

The general solution is as follows. The real root is

$$\lambda_o = \frac{-\varrho}{K_1}\left[1 + \frac{K_2\varrho}{2!K_1^2} + \frac{(3K_2^2 - K_1K_3)\varrho^2}{3!K_1^4} + \frac{(15K_2^3 - 10K_1K_2K_3 + K_1^2K_4)\varrho^3}{4!K_1^6} + \ldots\right] \quad (4)$$

The k-th pair of complex roots are

$$\lambda_{\pm k} = \lambda_o + \frac{2k^2\pi^2}{K_1^3}\left[K_2 + \frac{(3K_2^2 - K_1K_3)\varrho}{K_1^2} + \right]$$

$$\pm\frac{2k\pi i}{K_1}\left[1 - \frac{K_2\varrho}{K_1^2} - \frac{(3K_2^2 - K_1K_3)(3\varrho^2 \times 4k^2\pi^2)}{6K_1^4} + \ldots\right] \quad (5)$$

In the long run $B(t)$ approximates to $A_o e^{-\lambda_o t}$. When this is so $B(t)$ increases or diminishes exponentially with time. It increases by a factor R in a time T given by $e^{-\lambda_o T} = R$, so

$$T = K_1\left[1 - \frac{K_2\varrho}{2K_1^2} - \frac{(3K_2^2 - 2K_1K_3)\varrho^2}{12K_1^4} - \frac{(6K_2^3 - 6K_1K_2K_3 + K_1^2K_4)\varrho^3}{24K_1^6} - \ldots\right] \quad (6)$$

T may be called the equivalent generation time. It is generally less than the mean maternal age K_1 if ϱ is positive and the popula-

tion increasing, and greater if the population is diminishing. The general solution is

$$B(t) = A_o e^{-\lambda_o t}[1 + \sum_{k=1}^{\infty} B_k e^{-\alpha_k t} \cos(\beta_k t + C_k)] \quad .. \quad (7)$$

where A_o, B_k, C_k are real constants and $\lambda_k = \lambda_o + \alpha_k + \beta_k i$, as in equation (5). The complex root represents damped oscillation about the exponential increase or decrease. The slowest oscillation has a period $K_1(1 - \frac{K_2 \varrho}{K_1^2} - \ldots.)$, or about the mean maternal age, and is damped by a factor of about $e^{\frac{-2\pi^2 K_2}{K_1^2}}$ per period. The quicker oscillations are more strongly damped.

It is easy to give solutions when $f(x)$ approximates to a known distribution. Thus, if the maternal age distribution is normal,

$$\lambda_o = \frac{\sqrt{K_1^2 - 2K_2\varrho} - K_1}{K_2}$$

If the maternal age distribution is a Gamma or Pearson Type III distribution, of the form $f(x) = \frac{a^c}{\Gamma(c)} x^{c-1} e^{-ax}$ which is nearly true for some insects,

$$\lambda_o = a(1 - R^{1/c}),$$

$$\lambda_k = a\left[1 - R^{1/c}\left(\cos\frac{2k\pi}{c} + i \sin\frac{2k\pi}{c}\right)\right]$$

So if c is a positive integer the number of complex roots is finite. It can readily be shown that the expression for $N(t)$ is of the same form as that for $B(t)$. We should like to know the value of R and the form of $f(x)$ for the human population of India as a whole, and for various populations within it. There are, unfortunately, no very satisfactory data. Das Gupta, Som, Majumdar and Mitra (1955) have studied couple fertility on the basis of sample survey data. Their results are inadequate for two reasons. The ages of the parents are very uncertain. The years of birth of children

are less so. Further, a very large fraction which could even exceed half, of all infant deaths was not reported.

The results are remarkable. The data summarized in their Diagram (6·1) and (6·2) make it clear that $F(x)$ per couple rises sharply after marriage or the wife's fifteenth birthday, reaches a steady level of about 0·12 (daughters per year) at about 2 years after this time, and does not fall off greatly till the marriage has lasted for 22 years. Even then the fall is not very great before 30 years. Mr. Ajit Ray is now undertaking a similar investigation among the Juangs, a tribal people of Orissa believed to number about 12,000. His results will be of great value, being the first such study of a relatively large endogamous primitive community. Since they are based on the work of a single anthropologist rather than of a number of enumerators trained primarily for economic work, it is possible that they will be more accurate than the sample survey data.

However, the sample survey data could, I think, be utilized to obtain rough estimates of R, and the first few moments of the distribution of maternal ages. If $F(x)$ is taken as approximately constant from 16 to 45 years, it must be multiplied by $l'x$, which is not the probability of a woman surviving for x years after marriage, but that of a woman and her husband both surviving, since the remarriage of widows may be neglected.

If it is concluded that a large number of dead children are forgotten, as seems to be the case with the sample survey data, it might be better to consider female lives as starting at 15, and ask the question. "Among a group of women who reach the age of 15, what is the average number of daughters who reach the age of 15?". We could also ask "Among a group of women who marry, what is the average number of daughters who live long enough to marry?". Either of these would give an approximate value of R, though the second would give rather too low a value if the daughters, on an average, marry later than the mothers.

I think it might be desirable to try both the above methods, along with the methods which are usual for European and American data.

It would be possible to calculate the increases in R and λ_o resulting from hygenic measures reducing death rates both before and after marriage. It is not, I think, desirable that the population of India should increase at much more than its present rate.

The possible effects of birth control could be calculated if we supposed for example, that, after the birth of the first two (or three) children, the fertility of couples were reduced to half.

Discussion

I am fully aware of the inadequacy of this discussion. Fertility and mortality change with time. One, but only one, of the causes responsible for changes in them is change in population density. In animal populations it may be possible to allow for such changes by averaging over a number of generations, though if R is very variable this average will be too high, and it is better to average ϱ. In human populations the values of R in the past can be roughly calculated from the total population and approximate values of l_x, K_1, and so on. It is much harder to predict changes in the future. In India it seems likely that R will increase as the result of lower mortality, but it may decrease in twenty years or so as a consequence of industrialization.

I would, however, urge Indian workers to obtain approximate figures on population growth and decline for Indian animals both under natural and artificial conditions. It is not difficult to study insect reproduction in the laboratory, but surprisingly little accurate work has been done on it. Varley's (1947) work is a model of what can be done under natural conditions, but it would be better to spread such work over at least five years.

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A Hyper-Isogonal Cremona Involution of Conic-Envelopes

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I. The well-known Cremona quadratic involution of point-pairs PP′ in the plane consists of pairs PP′ conjugate w.r.t. all conics S through four points I_0 I_1 I_2 I_3 forming a proper quadrangle. It is clear that the conics S are all the conics apolar to the four squared points I_0^2, I_1^2, I_2^2, I_3^2; it follows that the system of conic-envelopes apolar to all the conics S is the system of all linear combinations $\lambda_0 I_0^2 + \lambda_1 I_1^2 + \lambda_2 I_2^2 + \lambda_3 I_3^2$. If ABC is the harmonic triangle of $I_0I_1I_2I_3$, it is clear that the system of conics inscribed in ABC belongs to Σ. Hence the system Σ may also be considered as the set of all linear combinations.

$$\mu_1 (BC) + \mu_2 (CA) + \mu_3 (AB) + \mu_0 \Omega,$$

where Ω is any particular point-pair PP′. In particular, if the pair of circular points belongs to the involution, and is denoted by Ω, $I_0I_1I_2I_3$ become the in-and ex-centres of ABC, and the quadratic involution reduces to the system of isogonal point-pairs PP′ of the triangle ABC.

II. It is a remarkable fact, which does not appear to have been noted before, that this involution of point-pairs induces a Cremona quadratic involution among all the ∞^5 conic-envelopes in the plane. This may be gathered from the following considerations. The family Σ of conic-envelopes contains 4 linearly independent members. Thus, the product of any two conics Σ is a 6-dimensional subspace of the 14-dimensional space of all quartic envelopes. Also, this subspace contains $4 + 6 = 10$ linearly independent members, and hence generates a 9-dimensional flat subspace of the space of all quartic envelopes. Hence there is a linear ∞^4-system of quartic curves σ, apolar to all products of pairs of conic envelopes Σ. Hence if S is *any* conic-envelope in the plane, there is in general a unique conic S′, such that that SS′ is inpolar to the 5 linearly independent curves determining the system σ. The involution between the conic-envelopes S, S′ is precisely the involution referred to, as induced by the quadratic involution defined by $I_0I_1I_2I_3$.

To obtain explicitly the transformation from S to S′ we may without loss of generality use trilinear coordinates with ABC as reference-triangle, and assume that the basic Cremona involution is the involution of isogonal pairs. The system Σ has then the linearly independent members:

$$mn,\ nl,\ lm,\ \Omega \equiv l^2 + m^2 + n^2 - 2\,mn\cos A - 2\,nl\cos B - 2lm\cos C,$$

(where l, m, n are tangential coordinates). The ten linearly independent products of two Σ's are then:

$$m^2n^2,\ n^2l^2,\ l^2m^2,\ l^2mn,\ lm^2n,\ lmn^2,\ \Omega mn,\ \Omega nl,\ \Omega ml,\ \Omega^2.$$

The 5 linearly independent quartics apolar to these 10 are immediately verified to be:

$$y^4 - z^4,\ z^4 - x^4,\ y^3z - yz^3,\ z^3x - zx^3,\ x^3y - xy^3.$$

The quartic curves of the system σ, have, therefore, the equation

$$\alpha x^4 + \beta y^4 + \gamma z^4 + p(y^3z - yz^3) + q(z^3x - zx^3) + r(x^3y - xy^3) = 0 \quad (\alpha + \beta + \gamma = 0).$$

If

$$S \equiv AL^2 + Bm^2 + Cn^2 + 2\,Fmn + 2\,Gnl + 2\,Hlm$$
$$S' \equiv A'l^2 + B'm^2 + C'n^2 + 2\,F'mn + 2\,G'nl + 2\,H'lm,$$

the conditions that SS′ is apolar to all curves σ become:

$$AA' = BB' = CC'$$
$$BF' + B'F = CF' + C'F$$
$$CG' + C'G = AG' + A'G$$
$$AH' + A'H = BH' + B'H.$$

Thus, the equations to the transformation $S \to S'$ are

$$A' = BC$$
$$B' = CA$$
$$C' = AB$$
$$F' = AF$$
$$G' = BG$$
$$H' = CH.$$

The geometrical significance of these equations is immediate. For, the tangents from the vertex A of the triangle of reference to the conics S, S′ are given by:

$$Bm^2 + 2Fmn + Cn^2 = 0.$$
$$B'm^2 + 2F'mn + C'n^2 = 0.$$

and the transformation equations say that the latter two tangents are the reflections of the former in the bisectors of the angle A.

The transformation equations also show that

Discriminant $S' =$ ABC $\times$ Discriminant S.

Hence when S or S′ is a pair of points, so is the other, as is indeed obvious from the geometrical interpretation for the transformation $S \rightarrow S'$. Another remarkable property is that *S′ is a circumconic of the triangle of reference whenever S is.* For writing

$$A = a_1^2,\ B = a_2^2,\ C = a_3^2,\ F = -a_2a_3,\ G = -a_3a_1,\quad H = -a_1a_2,$$

we have

$$A' = a_2^2a_3^2,\ B' = a_3^2a_1^2,\ C' = a_1^2a_2^2,\ F' = -a_1^2a_2a_3,$$
$$G' = -a_1a_2^2a_3,\ H' = -a_1a_2a_3^2$$

Thus, *if S is the polar conic of a point P* (a_1, a_2, a_3), *S′ is the polar conic of the isogonal conjugate P′,* (a_2a_3, a_3a_1, a_3a_2).

III. This conic-transformation may be applied to prove a rather intriguing property.

Let (p, q, r), $(p'\ q'\ r')$ be a pair of the quadratic involution defined by the four points I_0ABC. Then $(pl + qm + rn)$ $(p'l + q'm + r'n)$ must be apolar to $x(y - z)$, $y(z - x)$ and $z(x - y)$. This gives

$$qr' + q'r = rp' + r'p = pq' + p'q.$$

Solving these equations we obtain:

$$p' : q' : r' = p(q + r - p) : q(r + p - q) : r(p + q - r).$$

We know that the transformation of the point-pair $(pl + qm + rn)$, $(p'l + q'm + r'n)$ must also be a point-pair. This transform is

$$\Sigma qq'rr'l^2 + \Sigma pp'(qr' + q'r)mn = 0.$$

Substituting the values of p', q', r', this point-pair becomes:

$$\Sigma q^2r^2\ \{p^2 - (q-r)^2\}l^2 + 2pqr\ \Sigma mnp^2(q + r - p) = 0.$$

If we subtract from this $p^2q^2r^2(l + m + n)^2$ (which is the expression for the squared point I_0), we obtain

$$-\Sigma l^2q^2r^2(q-r)^2 + 2pqr\ \Sigma mn\ \{p^2(q + r - p) - pqr\} = 0.$$

We may immediately verify that this is the tangential equation of a conic circumscribed to the triangle of reference. For instance, putting $l = 0$, we verify that

$$p^2q^2r^2\{p^2(q + r - p) - pqr\}^2 = p^4q^2r^2\ (q-p)^2(p-r)^2$$
$$= \text{product of coefficients of } m^2,\ n^2.$$

The meaning of this is evidently the curious fact that the isogonal conjugates of the two points (pqr) $(p'q'r')$, which are themselves mates in the quadratic involution determined by I_0ABC, are two points which are *the points of contact of tangents from I_0 to a conic circumscribed to ABC*. It is clear that we may also assert the converse, namely, *the isogonal conjugates of the two points of contact of tangents from I_0 to a circumconic are mates in the quadratic involution defined by I_0ABC*. It may be noted that the actual relation between the two points of contact would be a Cremona involution of much higher degree than the Second.

This theorem gives a relation between the quadratic involutions defined by $I_0I_1I_2I_3$ and I_0ABC. A deeper elucidation of this relation is a matter for further research.

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The Theory of Turbulence

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I

It is a matter of common knowledge that hydrodynamic flow manifests itself in one of two qualitatively distinct forms: the laminar and the turbulent. In the laminar state, the velocity is a continuous function of position and time and is uniquely determined by the equations of motion and certain prescribed boundary conditions. In contrast to this, in the turbulent state the velocity exhibits variations with position and fluctuations with time which do not seem to be related in any discernible way with the particular conditions under which the motion has been set up. Whether the laminar or the turbulent state prevails depends on circumstances; but, it may be stated that as a general rule, motions in systems with large linear dimensions are more likely to be turbulent than not. Somewhat more precisely, we may state that the condition for the occurrence of one or other of the two states depends on the value of a certain non-dimensional constant, the Reynolds number,

$$R = \frac{\varrho v l}{\mu} \tag{1}$$

where ϱ is the densitiy, μ is the coefficient of viscosity and v and l are measures of the velocities present and the linear dimension of the system, respectively. If R so defined exceeds a certain critical value, then the laminar state which can prevail at lower values of R ceases to be stable and other modes of motion compatible with the equations of the problem are excited; and as the Reynolds number increases still further, more and more new modes are excited. A state of turbulence which thus comes to be established is characterized by two essential facts: first there are many different modes of motion which are simultaneously present; and second, there is a constant exchange of energy between the different modes.

And it is a prime requirement of a theory of turbulence that it describe quantitatively, in some way, the different coexisting modes and the mechanics of the energy exchange between them.

II

Since a state of turbulence follows the breakdown of a laminar state, it might appear natural to begin the study of turbulence by examining the stability of simple patterns of laminar motion—such as that between two parallel plates or between two rotating cylinders—and establishing the critical Reynolds number for the onset of instability. Several successful investigations along these lines are now available. It is perhaps not altogether surprising that these investigations have not helped in any material way our basic understanding of the turbulence phenomenon: for, the fundamental characteristic of turbulence is the continuing exchange of energy between the different modes of motion and no linear theory (as a theory of stability necessarily is) can be expected to throw light on a phenomenon which basically owes its existence to the non-linearity of the governing equations. For these reasons it would appear more profitable to start the discussion by an analysis of a state of turbulence already well established, rather than attempt to trace all at once the development of turbulence from the initial state when the laminar state ceases to be stable to the final state as the Reynolds number tends to infinity.

III

We have seen that a state of turbulence eventually results from the instability of a laminar state. But there are many other ways in which turbulence can be set up. In nature, for example, one of the commonest causes for turbulence is thermal instability; and in the laboratory, one of the easiest methods of producing turbulence is by letting air flow through a grid in a wind tunnel. On account of this very diversity of ways in which turbulence arises and can be created, it is convenient to have a fairly specific model which we may visualize and to which we may refer in our subsequent discussion.

IV

The model we shall have in mind is the following: A volume of air (say) is confined inside a large vessel. The air inside this vessel is kept continually stirred by a number of fans attached

to the inside walls. In this manner we set up violent motions near the walls. It is, however, clear that these motions will pervade the whole vessel and we shall concentrate our attention on the state of motions which prevails in the central regions of the vessel far from the walls and the fans.

V

In a general way we may characterize the motions in the model we have described as follows: We supply energy, principally, to air elements—"eddies"—of the dimensions of the vessel. This energy is in some way transferred to air elements of smaller scales; and in a steady state which we envisage, the rate at which energy is dissipated in the air in the form of heat must equal the rate at which energy is supplied to the system. The process which is taking place is clearly described by the following rhyme due to L. F. Richardson:

> Big whirls have little whirls
> That feed on their velocity,
> Little whirls have lesser whirls
> And so on into viscosity.

In other words: energy is supplied at a constant rate to the big whirls; these whirls transfer their energy to the smaller whirls; the smaller whirls transfer their energy to still smaller whirls; and this process continues until the energy is finally dissipated as heat by viscosity in the smallest whirls in which the sheer is the greatest.

The foregoing remarks bring out certain basic facts which must be kept in mind in discussions of turbulence: That an external agency supplying energy to the systems must be pre-supposed; that the energy is principally supplied to the largest eddies; that there is a cascade of this energy to the smaller eddies; that in a stationary state energy must be dissipated at the same rate at which it is supplied.

The fundamental problem of turbulence is this: What is the distribution of energy among the eddies of various sizes in stationary turbulence; and what is the mechanics of the cascade process which transfers energy from the larger to the smaller eddies?

VI

We shall now formulate more precisely the questions raised in the preceding paragraph. It is clear that for this purpose it is convenient to express the velocity field as a superposition of plane waves i.e., analyze it in a Fourier series in the form:

$$\mathbf{u}(\mathbf{r}, t) = \Sigma\, \mathbf{u}(\mathbf{k}, t)\ \exp\ (i\,\mathbf{k}\,.\,\mathbf{r}) \tag{2}$$

where the summation is over all k.

The mean square velocity is given by

$$< |\mathbf{u}|^2 >_{Av} = \tfrac{1}{2}\, \Sigma < |\mathbf{u}(\mathbf{k})|^2 >_{Av} \tag{3}$$

From equation (3) it is clear that $< |\mathbf{u}(\mathbf{k})|^2 >_{Av}$ is a measure of the intensity with which the mode of motion characterized by the wave numbers (k_x, k_y, k_z) occurs in the system; it defines the spectrum of turbulence. The specification of $< |\mathbf{u}(\mathbf{k})|^2 >_{Av}$ is therefore the first question to which we must address ourselves. It is, however, necessary to emphasize that a description of turbulence in terms of the spectrum only cannot be a complete one. A description in terms of the spectrum only would be complete, if, as in the case of the black-body radiation, there are no phase relationships between the different Fourier components of the field. In turbulence this is not the case. Phase relationships must exist: without them there would not be the exchange of energy between the different Fourier components which is in the essence of the phenomenon.

VII

Before we can proceed further, it is necessary to write down the fundamental equations which govern the motions and to point out the terms in the equations which are responsible for the different processes which are taking place.

In writing down the equations of motion we shall suppose that the fluid is incompressible. This may be considered as an adequate approximation so long as the velocities are small compared to the velocity of sound.

The equations governing an incompressible viscous fluid are the equation of continuity,

$$\text{div}\ \mathbf{u} = 0, \tag{4}$$

and the equation of motion

$$\frac{\partial \mathbf{u}}{\partial t} + (\mathbf{u}.\ \text{grad})\ \mathbf{u} = -\ \text{grad}\ \tilde{\omega} + \nu \nabla^2 \mathbf{u} \qquad (5)$$

where $\mathbf{u}$ is the velocity, $\tilde{\omega} = p/\varrho$, where p is the pressure and ϱ is the density, and $\nu = \mu/\varrho$ is the coefficient of kinematic viscosity. These are the Navier-Stokes equations.

In equation (5) we distinguish three terms: the term $\nu \nabla^2 \mathbf{u}$ which expresses the role of viscosity, the term $-$ grad $\tilde{\omega}$ which expresses the role of pressure and the term $(\mathbf{u}.\ \text{grad})\ \mathbf{u}$ which is non-linear and which is called the *inertial term*. We shall see that these three terms play very different roles for turbulence.

VIII

By expanding $\mathbf{u}$ in a Fourier series as in equation (2), we first find from the equation of continuity that the vector $\mathbf{u}(\mathbf{k})$ is normal to the wave vector $\mathbf{k}$; thus,

$$\mathbf{u}(\mathbf{k}) \,.\, \mathbf{k} = 0. \qquad (6)$$

Next, the equation of motion gives

$$\frac{\partial \mathbf{u}(\mathbf{k})}{\partial t} = -\,i \sum_{\mathbf{k}'} \{\mathbf{u}(\mathbf{k}') \cdot \mathbf{k}\}\mathbf{u}(\mathbf{k}-\mathbf{k}') - i\,\tilde{\omega}(\mathbf{k})\mathbf{k} - \nu|\mathbf{k}|^2\mathbf{u}(\mathbf{k}) \qquad (7)$$

where $\tilde{\omega}(\mathbf{k})$ is the coefficient of exp $(i\,\mathbf{k}.\mathbf{r})$ in the Fourier expansion of $\tilde{\omega}$:

$$\tilde{\omega} = \Sigma\ \tilde{\omega}\ (\mathbf{k})\ \exp\ (i\,\mathbf{k}.\mathbf{r}) \qquad (8)$$

Multiplying equation (7) by $\mathbf{u}^*(\mathbf{k})$ and making use of the orthogonality of $\mathbf{u}^*(\mathbf{k})$ and $\mathbf{k}$, we get

$$\tfrac{1}{2}\frac{d}{dt}|\mathbf{u}(\mathbf{k})|^2 = -\,i \sum_{\mathbf{k}'} \{\mathbf{u}(\mathbf{k}')\cdot\mathbf{k}\}\ \{\mathbf{u}(\mathbf{k}-\mathbf{k}')\ .\ \mathbf{u}(-\mathbf{k})\}$$

$$-\,\nu|\mathbf{k}|^2|\mathbf{u}(\mathbf{k})|^2 \qquad (9)$$

From equations (7) and (9) several important conclusions follow:

(1) Viscosity plays no role in the cascade process. Indeed, the only effect of viscosity is to reduce the amplitude of each Fourier component independently of the others. Thus,

$$\mathbf{u}(\mathbf{k};\ t) = \mathbf{u}(\mathbf{k};\ 0)\ \exp\ (-\nu|\mathbf{k}|^2 t) \qquad (10)$$

Accordingly, the amplitude of the Fourier component with a wave number $|\mathbf{k}|$ decreases with a mean life $\tau_k = (\nu|\mathbf{k}|^2)^{-1}$. The smallest eddies are, therefore, the shortest lived.

(2) The pressure term does not also play any direct role in transferring energy between the different Fourier components. This is evident from the non-appearance of $\tilde{\omega}$ in equation (9). But the pressure term does play a role in exchanging energy between different directions belonging to the same $|\mathbf{k}|$. It thus operates in the direction of bringing about a *state of isotropy*.

(3) It is the inertial term which couples the different Fourier components belonging to different $|\mathbf{k}|$ and makes possible an exchange of energy between them. It is, therefore, the term which is responsible for the cascade process. The manner in which the inertial term operates to bring about a statistical distribution of energy among the different Fourier components is the central problem in the theory of turbulence.

IX

From our discussion up to the present, it would appear that the character of turbulence will depend on circumstances and in particular on the manner in which it is created and maintained. Certainly, no arguments have been advanced to indicate that this will not be the case. Nevertheless, we may ask: Are there aspects of the turbulence phenomenon which are universal? And by "universal" in this connection, we mean that the aspects in question are the same for all manner of ways of producing and maintaining turbulence. This question, whether there might not be universal aspects of the turbulence phenomenon is not unreasonable in view of the following circumstances:

We have stated that turbulence is mostly maintained by supplying energy to the largest eddies. The distribution of energy among these largest eddies will, of course, depend on the particular way

energy is supplied and reflect among other things the anisotropy implicit in the whole process. Nevertheless, in virtue of the inertial term this energy cascades to smaller eddies; and moreover, the pressure term operates to efface the anisotropy of the larger eddies. Accordingly, we may expect that as we go down the hierarchy of eddies we will gradually lose correlation with the structure and distribution of the larger eddies. In other words, in the part of the volume away from the confining walls (in the model described in § IV, for example) we may expect to find a high degree of homogeneity and isotropy. The prevalance of homogeneity and isotropy among these smaller eddies is clearly necessary before we can contemplate on whether or not universal characteristics exist.

X

Now it is one of the surprising facts about turbulence that isotropy is the most common and most readily realized. For example, in a wind tunnel conditions of isotropy have been shown to prevail already near the grid through which the air is flowing. Similarly, isotropy has been established in atmospheric turbulence. In view of the importance of this prevalence of homogeneity and isotropy for our whole discussion, we shall define precisely what we mean by these terms and how their occurrence has been established.

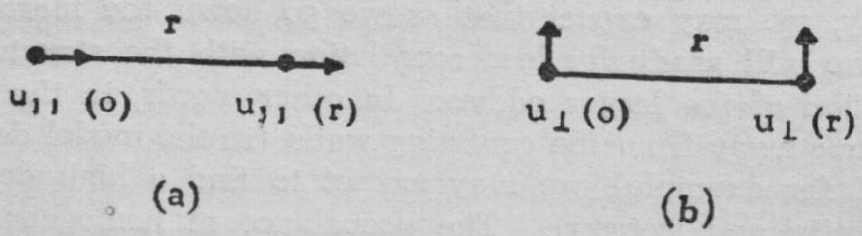
XI

In isotropic turbulence it is clear that the mean square energy associated with a component of the velocity in some direction must be independent of the direction. However, for our purposes we shall need a more exacting criterion. We shall state the requirements as follows :

Consider the velocities $\mathbf{u}_a\ (P)$ and $\mathbf{u}_b\ (P')$ at a given instant of time at two points P and P' and in the directions of two unit vectors $\mathbf{a}$ and $\mathbf{b}$ (see Fig. 1); and consider the average value* of $< u_a\ (P)\ u_b\ (P') >$. Let $\mathbf{e}$ denote the vector joining P and P'. We shall define isotropy and homogeneity to imply that in this case the quantity $< u_a\ (P)\ u_b\ (P') >$ *is invariant to all rotations and translations of the vector configuration* $(\mathbf{a}, \mathbf{e}, \mathbf{b})$ *as a rigid*

* Averages here and elsewhere are to be understood as ensemble averages in the sense of Gibbs. Also, angular brackets will be used to denote such averages.

body. The extension of this condition for the correlations between several components of the velocity at P and at P' is evident and need not be formally stated.



(a) (b)

FIG. 1. The longitudinal (a) and the transverse (b) velocity correlations.

XII

We shall now state the kind of deductions that can be made from the equations of motion and continuity and the requirements of isotropy and homogeneity. For this purpose, we shall consider first the fundamental covariance

$$Q_{ij} = < u_i(P, t)\ u_j(P', t) >. \tag{11}$$

It can be shown (cf. von Karman and Howarth, 1938; Robertson, 1940; for an account of these and related developments, see Chandrasekhar, 1953) that the equation of continuity together with the conditions of isotropy and homogeneity require that this tensor be of the form

$$Q_{ij} = \left(\frac{1}{r}\frac{\partial Q}{\partial r}\right)\ e_i e_j - \left(r\frac{\partial Q}{\partial r} + 2Q\right)\ \delta_{ij} \tag{12}$$

where Q is a scalar function which depends (apart from t) only on the distance r $(= |\mathbf{e}|)$ between the two points P and P'. The implication of this particular form Q_{ij} can be most readily seen by considering the longitudinal and the transverse correlations (see Fig. 2)

$$\begin{aligned} &< u_{||}(0)\ u_{||}\ (r) > = f(r) \\ \text{and } &< u_{\perp}\ (0)\ u_{\perp}\ (r) > = g\ (r). \end{aligned} \tag{13}$$

From equation (12) it follows that

$$f(r) = -2Q \text{ and } g\ (r) = -\left(r\frac{\partial Q}{\partial r} + 2Q\right). \tag{14}$$

Accordingly, these two correlations are related by

$$g = f + \tfrac{1}{2} r \frac{\partial f}{\partial r}. \tag{15}$$

This relation between the two correlations is a definite prediction and it has been directly verified by experiments in a wind tunnel. This verification is a strong confirmation that under the conditions of the experiments isotropy in the precise sense did prevail.

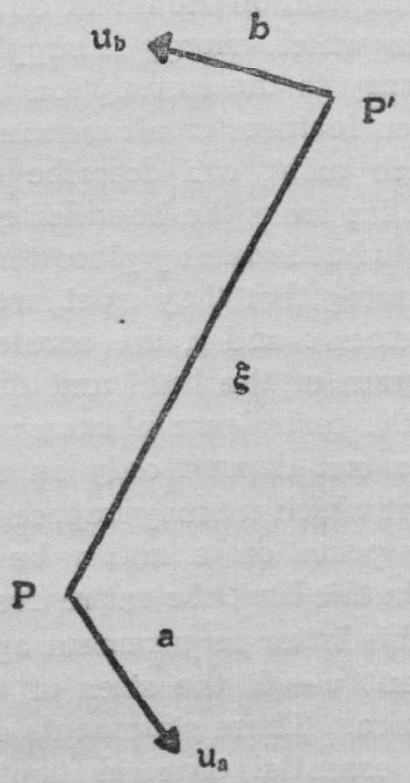


FIG. 2. For illustrating the requirements of homogeneity and isotropy.

A further confirmation of the prevalance of isotropy is provided by measurements on the rate of dissipation of the kinetic energy per unit mass ε. By combining the equations of motion with the requirements of isotropy it can be shown (cf. Taylor, 1935) that

$$\varepsilon = -15\,\nu \left(\frac{\partial^2 f}{\partial r^2}\right)_{r=0} \tag{16}$$

i.e., the rate of dissipation is simply related to the curvature of the longitudinal correlation function at the origin. Again, this relation has been verified by G. I. Taylor who was the first to introduce the concept of isotropy in turbulence which we have described.

XIII

We now return to the question of § IX whether there might not be some aspects of the turbulence phenomenon which are truly universal. There is no assurance that such aspects exist. The prevalence of homogeneity and isotropy in turbulence as it occurs in nature and in the laboratory, suggests that there might be such aspects. And if such aspects exist, then what can one say about them from entirely general considerations; in particular, on what parameters can such universal aspects depend? Clearly, it is necessary that they do not depend on the linear dimensions of the system and on the manner in which energy is supplied; a dependence on these would be contrary to the assumption that the aspects under discussion are common features of all manner of turbulence. Universal aspects, if such exist, can describe only processes taking place in the medium far from the boundaries and which are, in a limit, independent of what is taking place there. To describe such processes, again assuming that they exist, we have at our disposal only the two parameters ε and ν, the constant rate of dissipation of energy by each gram of the fluid and the kinematic viscosity. *Thus, any aspect of the turbulence phenomenon which might have a universal character can depend only on these two parameters.* A further requirement which we may impose is that *the dependence of such universal aspects on* ε *and* ν *be such that the limit* $\nu \to 0$ *exists; and that in this limit the entire description of the aspects depends only on* ε. This latter requirement appears entirely reasonable since in the limit $\nu \to 0$, the sizes of the energy dissipating eddies also tend to zero. These statements concerning the universal aspects are the essential physical content of the principles which Kolmogoroff has enunciated. However, we have left open the question whether or not universal aspects exist; in particular, we have left open the possibility that Kolmogoroff's principles may not be applicable to *every* aspect of the turbulence phenomenon. Indeed, one of the questions which we shall have to settle concerns exactly the selection of the function to which we may apply the principles.

XIV

We shall now illustrate the ideas outlined in § XIII by considering the quantity

$$\psi(r, t) = \tfrac{1}{2} < (u_1' - u_1'')^2 >, \qquad (17)$$

where u_1' and u_1'' are the x-components of the velocity at two points on the x-axis separated by a distance r and at two times, an interval t apart. (In this and in the subsequent sections, t will always signify the difference of the times at which the respective field quantities are measured at two points). If (cf. Fig. 2)

$$< u_1(0,\ 0,\ 0;\ 0)\ u_1(r,\ 0,\ 0;\ t) > = f(r,\ t), \tag{18}$$

then the quantity we are considering is related to this longitudinal correlation function (cf. equation (13)) by

$$\psi(r,\ t) = < u_1^2 > - f(r,\ t). \tag{19}$$

We now ask: Can we suppose that the specification of $\psi(r,\ t)$ lies within the framework of a universal theory? A decision on this can be reached only by applying Kolmogoroff's principles to this quantity and seeing whether we are led to any contradictions.

Consider first $\psi\ (r,\ 0)$ which is directly related to the longitudinal correlation function considered in § XII. Now this is a quantity of the dimensions of a square of a velocity. Since $(\nu^3/\varepsilon)^{\frac{1}{4}}$ and $(\varepsilon\nu)^{\frac{1}{4}}$ are the only combinations of ε and ν which are of the dimensions of a length and a velocity, respectively, Kolmogoroff's principles, if applicable to $\psi\ (r,\ 0)$ would require that

$$\psi(r,\ 0) \equiv (\varepsilon\ \nu)^{\frac{1}{2}}\ \Psi_1\ (r\ \varepsilon^{1/4}/\nu^{3/4}) \tag{20}$$

where $\Psi_1(x)$ is a universal function of the argument. If in the limit $\nu \to 0$, $\psi(r,\ 0)$ should cease to depend on ν, then the condition on $\Psi_1\ (x)$ is:

$$\Psi_1(x) \to C_1\ x^{2/3} \text{ as } x \to \infty, \tag{21}$$

where C_1 is a constant: for we can then conclude from equations (20) and (21) that

$$\psi\ (r,\ 0) \to C_1\ \ (\varepsilon\ r\ )^{2/3} \text{ as } \nu \to 0; \tag{22}$$

and this is independent of ν as required. But this last result throws doubt on whether Kolmogoroff's principles can be applied to $\psi(r,\ 0)$. For it is a strict consequence of the definition of ψ that

$$\psi(r,\ 0) \leqslant < u_1^2 >, \tag{23}$$

which follows from Schwarz' inequality :

$$< u_1' \; u_1'' > \leqslant < u_1^2 >. \tag{24}$$

This doubt on the applicability of Kolmogoroff's principles to $\psi(r, 0)$ is further strengthened when we consider

$$\psi(0, t) = \tfrac{1}{2} < \{ u_1 \; (0, 0, 0; 0) - u_1 \; (0, 0, 0; t) \}^2 >, \tag{25}$$

which is related to the correlation of the velocities at the same point at two different times. Since $(\nu/\varepsilon)^{\frac{1}{2}}$ is the only combination of ε and ν which is of the dimensions of a time, Kolmogoroff's principles applied to $\psi \; (0, t)$ would require that

$$\psi(0, t) \equiv (\varepsilon \nu)^{\frac{1}{4}} \; \Psi_2(t \; \varepsilon^{1/2}/\nu^{1/2}), \tag{26}$$

where $\Psi_2(x)$ is a universal function of the argument. If $\psi(0, t)$ should be independent of ν as $\nu \to 0$, then

$$\Psi_2(x) \to C_2 x \text{ as } x \to \infty, \tag{27}$$

where C_2 is a constant; for, we can then conclude from equations (26) and (27) that

$$\psi(0, t) \to C_2 \; \varepsilon \; t \text{ as } t \to \infty; \tag{28}$$

and this is independent of ν as required.

The unboundedness of $\psi(r, 0)$ as $r \to \infty$ and $\psi(0, t)$ as $t \to \infty$ implies that the function $\psi(r, t)$ cannot be specified without reference to the size, r_0, of the largest eddies present i.e., of the linear dimension of the system. Thus, one might say (as is customary in the literature) that the applicability of (22) is limited by $r \ll r_0$ and that this limitation will guarantee that

$$C_1 \; (\varepsilon \; r)^{\frac{2}{3}} \ll < u_1^2 >; \tag{29}$$

and similarly that (28) cannot be applied for $t > r_0/\{ < u_1^2 > \}^{1/2}$. But it is difficult to see how in the framework of a universal theory (assuming that such a theory exists) we can restrict the consequences of that theory in terms of parameters such as r_0 and $< u_1^2 >$ which are not related to ε and ν in any way.

XV

Arguments similar to those advanced in § XIV apply also to the spectrum of turbulence, $F(k)$. The definition of $F(k)$ is contained in

$$< | \mathbf{u} |^2 > = \int_0^\infty F(k)\, dk; \tag{30}$$

$F(k)$ is accordingly of dimensions $[L^3/T^2]$ and Kolmogoroff's principles would require that

$$F(k) \equiv (\nu^5 \varepsilon)^{1/4}\, \Phi\ (k\nu^{3/4}/\varepsilon^{1/4}) \tag{31}$$

where $\Phi\ (x)$ is a universal function of the argument. From the requirement that $F\ (k)$ becomes independent of ν as $\nu \to 0$, it readily follows that

$$F(k) \to \text{constant}\ \varepsilon^{2/3}\ k^{-5/3}. \tag{32}$$

This is the Kolmogoroff spectrum. But it makes the integral defining $F(k)$ diverge. It cannot, therefore, be valid for $k \to 0$; the description of the largest energy containing eddies cannot be within the scope of the universal theory.

That neither $f(r)$ nor $F(k)$ can be uniquely defined in the framework of a universal theory is further supported by the fact that the relation (cf. Heisenberg, 1948; equation (50) in this paper)

$$f(r,\ 0) = (\tfrac{1}{2}\pi)^{1/2} \int_0^\infty dk\, F(k)\, J_{3/2}(kr) / (kr)^{3/2} \tag{33}$$

between them, where $J_{3/2}$ denotes the Bessel function of order 3/2 is not meaningful when $F(k)$ is given by Kolmogoroff's spectrum: the Hankel-transfer on the right-hand side of equation (33) does not exist.

XVI

If the spectrum $F(k)$ and the fundamental correlation function $f(r,\ t)$ are both outside the framework of a universal theory, then the question arises as to what the function is for which we may seek a universal characterization.

It appears that it is meaningful to consider the space derivative

$$\chi(r, t) = -\frac{\partial f(r,t)}{\partial r}, \tag{34}$$

of the longitudinal correlation function and that in the framework of stationary homogeneous and isotropic turbulence it can be specified uniquely without reference to $< u_1^2 >$ or r_0. Several arguments in support of this can be advanced.

(1) Kolmogoroff's principles applied to χ $(r, 0)$ would require that

$$\chi(r, 0) \equiv \left(\frac{\varepsilon^3}{\nu} \right)^{1/4} \quad X_1 \, (r\varepsilon^{1/4}/\nu^{3/4}) \tag{35}$$

where X_1 is a universal function of the argument and that

$$\chi(r, 0) \to \text{Constant } \varepsilon^{2/3} \, r^{-1/3} \text{ as } \nu \to 0; \tag{36}$$

and this does not diverge as $r \to \infty$. The divergence of the function at $r = 0$ need not concern us: it is due to the neglect of viscosity. [Also this divergence at $r = 0$, precludes us from considering χ $(0, t)$ at this point.]

(2) In contrast to equation (33) the relation (cf. Chamberlain and Roberts, 1955).

$$r^2 \frac{\partial^3 f}{\partial r^3} + 7r \frac{\partial^2 f}{\partial r^2} + 8 \frac{\partial f}{\partial r} = -\int_0^\infty k F(k) \sin kr \, dk, \tag{37}$$

represents a mathematically meaningful relation: The sine transform of k F (k) for the Kolmogoroff spectrum does exist; and it is clear that multiplication by k in the Fourier space is equivalent to differentiation with respect to r in ordinary space.

(3) The restriction to χ in the framework of a universal theory is in agreement with certain general ideas which have been advanced by Heisenberg and Schlüter that it is physically more significant to think of turbulence in terms of the vorticity field rather than in terms of the velocity field. If ω $(=$ curl $\mathbf{u})$ denotes the vorticity then it can be shown that the longitudinal vorticity

correlation $<\omega_1'\ \omega_1''>$ defined analogously to $<u_1'\ u_1''>$) is related to χ in the manner

$$<\omega_1'\,\omega_1''> = \frac{\partial \chi}{\partial r} + \frac{4}{r}\,\chi; \tag{38}$$

and this correlation as we observe depends only on χ.

XVII

Suppose then that $\chi\ (r,\ t)$ is a function which can be determined in the framework of a universal theory. Then what should be required of a physical theory of stationary homogeneous and isotropic turbulence? We should hope that the theory starting from the equations of motion and continuity and some clearly defined physical postulates leads to a differential equation for $\chi\ (r,\ t)$ which satisfies the following conditions:

(1) The equation for χ allows a Kolmogoroff transformation i.e., if we measure

$$\begin{aligned} &\chi \text{ in the unit } (\varepsilon^3/\nu)^{1/4}, \\ &r \text{ in the unit } (\nu^3/\varepsilon)^{1/4}, \\ &t \text{ in the unit } (\nu/\varepsilon)^{1/2}, \end{aligned} \tag{39}$$

the equation for χ becomes dimensionless. This would prove that, consistent with Kolmogoroff's principles, χ is of the form

$$\chi \equiv (\varepsilon^3/\nu)^{1/4}\ X\,(r\,\varepsilon^{1/4}/\nu^{3/4},\ t\varepsilon^{1/2}/\nu^{1/2}), \tag{40}$$

where $X\ (x,\ y)$ is a universal function of the arguments.

Now in the limit of zero viscosity the function $X\ (x,\ y)$ should have the behaviour

$$X\,(x,\ y) \to x^{-1/3}\ o\,(y/x^{2/3}). \tag{41}$$

This behaviour of $X\,(x,\ y)$ for $\nu \to 0$ can be deduced as follows: Noting that $y/x^{2/3}$ is the only combination of the arguments $x = r\ (\varepsilon/\nu^3)^{\frac{1}{4}}$ and $y = t\ (\varepsilon/\nu)^{\frac{1}{4}}$ which is independent of ν, and expressing $X\ (x,\ y)$ in the equivalent form

$$\chi\,(r,\ t) \equiv (\varepsilon^3/\nu)^{1/4}\ Y\,(r\varepsilon^{1/4}/\nu^{3/4},\ \varepsilon^{1/3}t/r^{2/3}), \tag{42}$$

we observe that the condition that the right-hand side become independent of ν as $\nu \to 0$ is

$$Y(x, y/x^{2/3}) \to x^{-1/3}\, \sigma(y/x^{2/3}) \text{ as } x \to \infty, \qquad (43)$$

$y/x^{2/3}$ remaining fixed.

These consequences of Kolmogoroff's principles will require of a differential equation for χ the following additional properties:

(2) The equation for χ when ν has been set equal to zero admits a solution of the form

$$\chi = r^{-1/3}\, \sigma(t/r^{2/3}), \qquad (44)$$

where σ satisfies the boundary conditions enumerated below.

(2a) σ is an even function of the argument. This follows from the basic requirement that in stationary turbulence the tensor $< u_i' u_j'' >$ depends on the times at which the velocity components are measured at the two points, only through the absolute value of their difference. We must, therefore, require that

$$\frac{d\sigma}{dx} = 0 \text{ at } x = 0 \quad (x = t/r^{2/3}). \qquad (45)$$

(2b) A second condition on σ is that it be bounded at the origin. Actually, it can be shown that σ should tend to a *positive constant* σ_0 as $x \to 0$:

$$\sigma \to \sigma_0 > 0 \text{ as } x \to 0. \qquad (46)$$

The reasons for requiring $\sigma_0 > 0$ are somewhat involved and we shall not go into them. (The interested reader might refer to Chandrasekhar, 1956).

(2c) A final condition is that

$$\sigma \to 0 \text{ as } x \to \infty. \qquad (47)$$

This condition is necessary: otherwise χ would not be meaningful in a universal theory.

We now return to the boundary conditions on $\chi\ (r, t)$. Clearly the evenness of $\chi\ (r, t)$ in t requires that

$$\frac{\partial \chi}{\partial t} = 0 \text{ for } t = 0 \text{ for all } r. \tag{48}$$

Further, we must also require that

$$\chi(r, t) \to C r^{-1/3}\, \sigma(t/r^{2/3}) \text{ as } r \to \infty \text{ for any finite } t, \tag{49}$$

where C is a numerical constant to be determined.

It remains to consider how we are to introduce in the theory the fact we have hitherto not considered namely, that we are dealing with a system in which there is a continuing dissipation of energy at the rate ε. At first sight it may seem difficult to introduce ε in a universal theory since this is also the rate at which energy is supplied to the system and we wish specifically to avoid any reference, direct or indirect, to the manner in which it is supplied. However, we can introduce ε without violating the framework of a universal theory by making use of Taylor's relation (cf. equation (16))

$$\left(\frac{\partial^2 f}{\partial r^2}\right)_{r=0,\ t=0} = -\frac{\varepsilon}{15\nu}. \tag{50}$$

as a further boundary condition. For χ expressed in Kolmogoroff's units (39) this condition implies that

$$\left(\frac{\partial \chi}{\partial r}\right)_{r=0,\ t=0} = \frac{1}{15}. \tag{51}$$

This last condition together with the earlier conditions (48) and (49) should determine C and make the solution for χ unique.

Let it be stated once again that if χ should be the quantity which it is meaningful to consider in a universal theory and if a deductive physical theory should lead to a differential equation for χ then it should be compatible with all the conditions we have enumerated. There is certainly no assurance that it will be possible to satisfy all these many conditions. Conversely, if a theory should lead to a differential equation for χ and it satisfies

all these conditions then we may regard these facts as providing an *a posteriori* justification for the ideas we have set forth.

XVIII

We shall now describe a theory (Chandrasekhar 1955, 56) which leads to a differential equation for χ and which satisfies all the requirements of the preceding paragraph.

Various correlations between the velocity components at two different points ($\mathbf{r}'$ and $\mathbf{r}''$ say) and at two different times (t' and t'' say) are first defined under conditions of stationary homogeneous and isotropic turbulence. These correlations depend only on the vector joining P and P' and on the difference in the times t' and t''. Thus,

$$Q_{ij} = < u_i\,(\mathbf{r}',\, t')\;\; u_j\,(\mathbf{r}'',\, t'') >,$$

$$T_{ij;\,k} = < u_i\,(\mathbf{r}',\, t')\;\; u_j\,(\mathbf{r}',\, t')\;\; u_k\,(\mathbf{r}'',\, t'') >,$$

$$\text{and } Q_{ij,\,kl} = < u_i\;(\mathbf{r}',\, t')\;\; u_j\,(\mathbf{r}',\, t')\;\; u_k\,(\mathbf{r}'',\, t'')\;\; u_l\,(\mathbf{r}'',\, t'') >, \qquad (52)$$

are isotropic tensors depending only on $\mathbf{e} = \mathbf{r}' - \mathbf{r}''$ and $t = t' - t''$.

We have already seen that the equation of continuity and the conditions of homogeneity and isotropy restrict Q_{ij} to the form (12) and define it in terms of a single scalar function $Q\;(r, t)$. Similarly, it can be shown that the tensor $T_{ij;\,k}$ (symmetrical in i and j and solenoidal in k) is also uniquely expressible in terms of a single scalar function $T\;(r, t)$ in the form

$$T_{ij;\,k}\;(\mathbf{e}, t) = \frac{2}{r}\frac{\partial T}{\partial r}\,\mathrm{e}_i\,\mathrm{e}_j\,\mathrm{e}_k - \left(r\,\frac{\partial T}{\partial r} + 3\,T\right)(\mathrm{e}_i\,\delta_{jk} + \mathrm{e}_j\,\delta_{ik})$$

$$+ 2\,T\,\mathrm{e}_k\,\delta_{ij}. \qquad (53)$$

With regard to quadruple moment $Q_{ij;\,kl}$ the assumption is made that it is related to the second-order moments Q_{ij} as in a normal distribution i.e.,

$$Q_{ij;\,kl} = Q_{ik}\,Q_{jl} + Q_{il}\,Q_{jk} + Q_{ij}\;(0,\,0)\;\;Q_{kl}\;(0,\,0). \qquad (54)$$

This is the basic statistical hypothesis of the theory; and it is the only one that is made.

XIX

A basic assumption such as (54) requires some justification; at least the motivation for making it should be explained.

First, let it be stated that the use of quasi-Gaussian hypotheses, (similar to (54)) in the solution of physical problems is by no means novel. It underlies, for example, a systematic solution of the Boltzmann equation in the theory of gases; and it underlies theories of the liquid state as that developed by Kirkwood (to whom we owe the descriptive phrase "breaking the bars"). But what are the physical grounds for making such an assumption? We might answer this as follows :

Whenever a physical system exhibits a property which is the result of many different, apparently unrelated causes acting simultaneously, then the assumption that a Gaussian distribution underlies the property is a normal and a natural one. Thus, in the dynamical theory of gases and in the theory of Brownian motion such an assumption (even, if not in so undefended a form!) is made; and the resulting theories provide a description adequate for most purposes. In these instances when one wishes a description of higher accuracy (as in the theory of non-uniform gases) one finds that the quasi-Gaussian hypothesis generally leads to the desired next approximation.

In the phenomenon of turbulence the statistical element is provided by the inertial coupling between the different Fourier components of the velocity field. However, in a theory of turbulence we cannot assume, even as a first approximation, that the relevant distributions are Gaussian. For, in a Gaussian distribution the triple correlation $T_{ij;\,k}$ (like all odd moments) will vanish. But to make $T_{ij;\,k} = 0$ is to abolish turbulence! For, turbulence owes its origin to the phase relationships between the different Fourier components; and the vanishing of the odd moments is equivalent to asserting that no such relationships exist. Thus, by multiplying the equation of motion

$$\frac{\partial u_i'}{\partial t'} + \frac{\partial}{\partial x_k'}\,(u_i'\,u_k') = -\frac{\partial \tilde{\omega}'}{\partial x_i'} + \nu\,\nabla^2 u_i' \qquad (55^*)$$

* The primes distinguish that the values of the variables at $\mathbf{r}'$ and t' are meant.

by u_j'' at $(\mathbf{r}'', t'')$ and averaging we obtain an equation relating the double (Q_{ij}) and the triple correlations $(T_{ik};\ _j)$. The term in the triple correlation arises from the inertial term in the equation of motion. If $T_{ij};\ _k = 0$, then in the equation for Q_{ij} the role of the inertial term would have been eliminated. But as we have already emphasized in § VIII, it is the inertial term which holds the key to the turbulence phenomenon. Consequently, in a theory of turbulence we cannot make the customary Gaussian assumption.

If the Gaussian assumption cannot be made, then the quasi-Gaussian assumption appears as the next simplest one to make. But can we justify the assumption on any other grounds besides those of simplicity? And, when indeed can such an assumption be justified?

Consider the latter question first: For a one-dimensional Gaussian distribution,

$$f(x) = \frac{1}{\sqrt{\pi}} e^{-x^2} \tag{56}$$

we have

$$<x^2> = 0\cdot5,\ <x^3> = 0 \text{ and } <x^4> = 0\cdot75. \tag{57}$$

Suppose that in some instance these three moments are

$$<x^2> = 0\cdot45,\ <x^3> = -\,0\cdot05 \text{ and } <x^4> = 0\cdot70. \tag{58}$$

It is possible that these moments are realized in a problem in which the non-vanishing of the third moment is of primary interest. Nevertheless, if the third moment is small (relative to the second and the fourth moments, as in the example) the skewness of the distribution required by the non-vanishing of the third moment may not be so great that the 2 : 3 Gaussian ratio between the second and the fourth moments is seriously in error. If this should be the case, then the quasi-Gaussian hypothesis may be adequate to determine the fourth moments (in a first approximation) in terms of the second. Clearly it should not be used for any direct evaluation of the third moments.

In using the quasi-Gaussian approximation in a theory of turbulence we are implicitly supposing that the relevant distributions,

even though skew, are not so skew that the fourth moments cannot be estimated (in a first approximation) in terms of the second moments by the Gaussian relation between them. We naturally ask if there is any empirical evidence for such a supposition. Without going into any detail it may be stated that the available evidence is not unfavourable. Thus, the triple correlations are indeed, small compared with the double correlations: the maximum value of $h = \langle u_1'^2 u_1'' \rangle / \langle u_1^2 \rangle^{3/2}$ is $-0{\cdot}05$ while $\langle u_1' u_1'' \rangle / \langle u_1^2 \rangle$ varies from 1 to 0 and is about $0{\cdot}4$ when h attains its maximum. Moreover, there have been attempts to verify the relation (54) directly and these attempts have in general confirmed it.

Quite apart from the empirical evidence which supports the use of the quasi-Gaussian approximation, we can perhaps justify its use in the theoretical development of a physical theory of turbulence as follows: We have indicated how from the equations of motion we can derive an equation relating the second and the third moments. We can similarly derive an equation relating the third and the fourth moments; an equation relating the fourth and the fifth moments, and quite generally an equation relating the nth and $(n+1)$th moments. The solution of the turbulence problem is formally equivalent to the solution of the infinite set of moment equations which can thus be derived from the Navier-Stokes equation. When any physical problem leads to such an infinite set of equations, a standard procedure for solving them is to consider the finite system formed by the first n equations and let n increase indefinitely (in principle, that is!). (This method of treating the moment equations is analogous to expanding a desired function in a complete set of orthogonal functions and retaining an increasing number of terms in the expansion.) In considering the infinite system of moment equations to which the Navier-Stokes equation leads, we have seen that breaking the system at the first equation leads to a trivial solution. Consequently, it is reasonable to ask: Do we obtain a physically significant solution by considering the first two equations? The quasi-Gaussian approximation is precisely the assumption which is needed to answer this question. Depending on the results which follow from the solution of the first pair of moment equations coupled through the quasi-Gaussian approximation, we may judge the usefulness of the approach. Similar methods have proved successful in many problems of mathematical physics; there is no *a priori* reason why it should not also prove useful in the solution of the turbulence problem.

XX

As we have explained in § XIX, we can derive from the Navier-Stokes equation an infinite set of moment equations. The quasi-Gaussian approximation (54) couples the first two of these equations and enables us to consider them independently of the others. We shall conclude our discussion of the turbulence problem by a brief summary of the analysis based on these two equations. It is, however, beyond the scope of the present essay to describe any aspect of the mathematical content of the theory. The interested reader might refer to the original papers (Chandrasekhar, 1955, 56 and Backus, 1957).

After suitable transformations and eliminations the theory developed on the premises outlined in §§ XVIII and XIX, leads to the following differential equation for χ:

$$\frac{\partial}{\partial r}\left[\left(\frac{\partial^2}{\partial t^2}-\nu^2\frac{\partial}{\partial r}\mathrm{D}_5\,\mathbf{D}\right)\chi\;\Big/\frac{\partial}{\partial r}\mathbf{D}\chi\right]=-\chi, \tag{59}$$

where we have used the abbreviations

$$\mathbf{D}=\frac{\partial}{\partial r}+\frac{4}{r}\quad\text{and}\quad \mathrm{D}_5=\mathbf{D}\,\frac{\partial}{\partial r}=\frac{\partial^2}{\partial r^2}+\frac{4}{r}\,\frac{\partial}{\partial r}. \tag{60}$$

It can be readily verified that equation (59) is compatible with Kolmogoroff's principles; for, if in accordance with these principles we measure χ, r and t in the units (39), we find that the equation becomes non-dimensional and that

$$\frac{\partial}{\partial r}\left[\left(\frac{\partial^2}{\partial t^2}-\frac{\partial}{\partial r}\mathrm{D}_5\,\mathbf{D}\right)\chi\;\Big/\frac{\partial}{\partial r}\mathbf{D}\chi\right]=-\chi. \tag{61}$$

Next consider the equation

$$\frac{\partial}{\partial r}\left[\frac{\partial^2\chi}{\partial t^2}\Big/\frac{\partial}{\partial r}\mathbf{D}\chi\right]=-\chi. \tag{62}$$

obtained from equation (59) by setting $\nu=0$. As we have seen in § XVII the similarity principles require that in this limit the solution for χ be of the form

$$\chi=r^{-\frac{1}{3}}\,\sigma(t/r^{\frac{2}{3}}), \tag{63}$$

where σ (x) satisfies the boundary conditions (45), (46) and (47). By mathematical processes which are too technical to outline here, it can be shown that equation (62) does allow a solution of the form (63) and further, that σ (x) is uniquely determined (apart from adjustable scale factors) by the boundary conditions. Figure 3 illustrates the resulting solution for σ; and in Figure 4 the evolution of χ with t is illustrated.

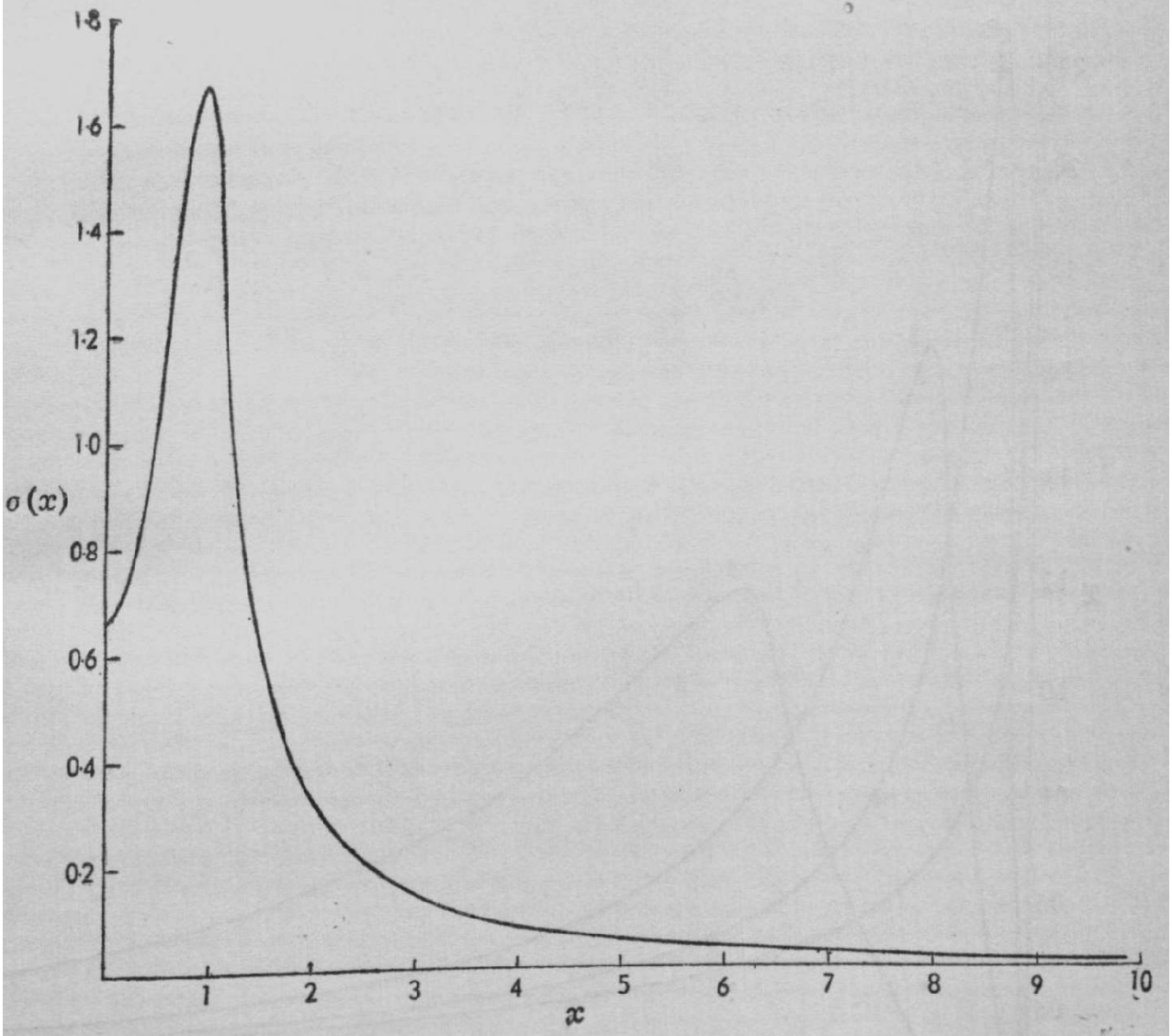


FIG. 3. Illustrating the dependence of the function σ on x.

Finally, it has to be shown that the conditions (48), (49) and (51) uniquely determine a solution of equation (61). While a rigorous proof has not been given to establish the existence of such a solution, it has been possible to demonstrate by numerical methods that a solution satisfying all the conditions of the problem can, in fact, be found.

XXI

The present essay has been devoted to a general discussion of what the central problem of turbulence really is and what the guiding principles are on which one may seek a solution of the problem. It has seemed to the writer that the formulation of the

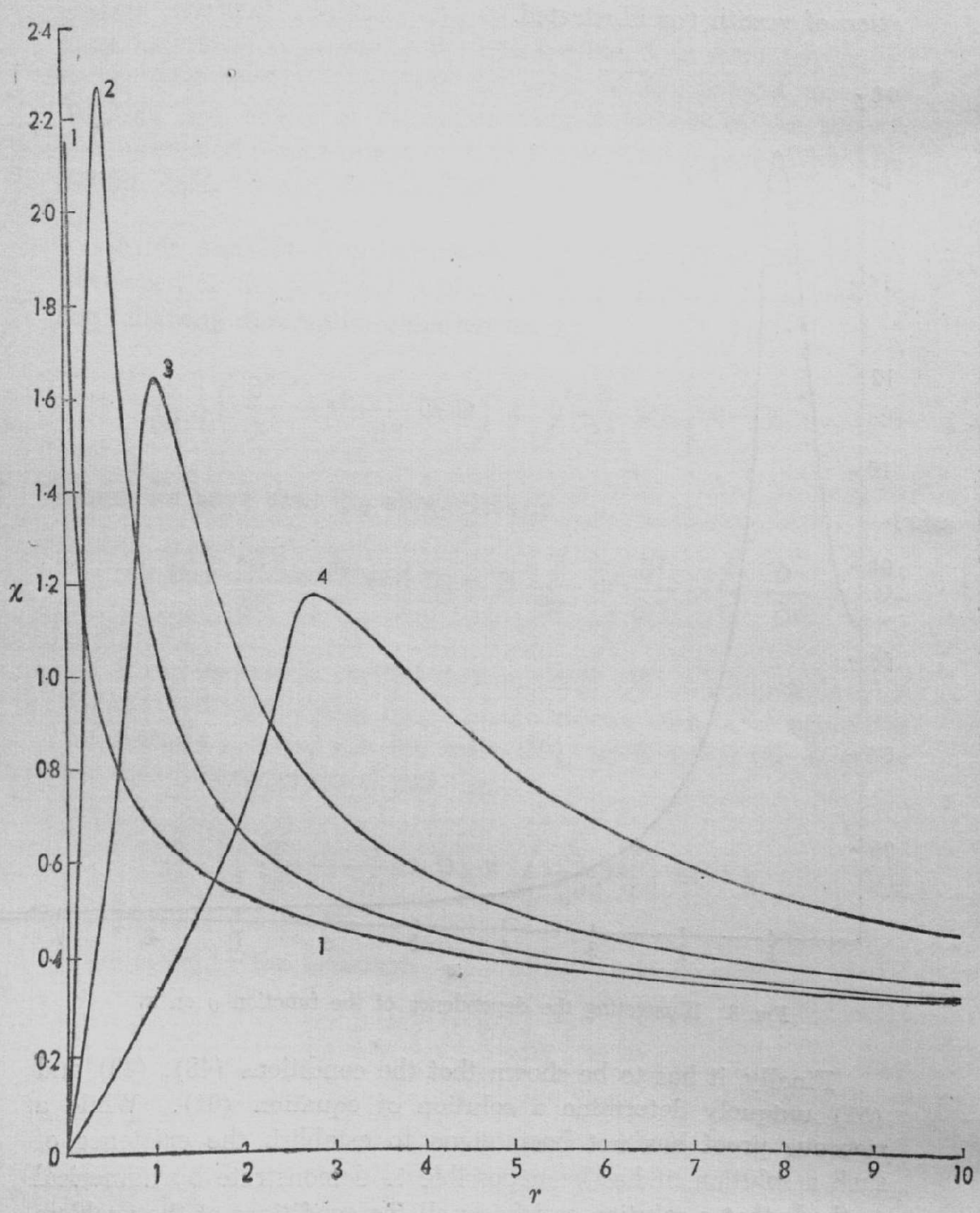


FIG. 4. The evolution of $\chi\ (r, t) = r^{-\frac{1}{3}}\sigma(t/r^{\frac{2}{3}})$ for various values of t. The curves labelled 1, 2, 3 and 4 are for $t = 0$, $0{\cdot}5$, $1{\cdot}0$ and $2{\cdot}0$, respectively.

problem and the clarification of the principles are of some value. While the ideas expressed are not all new, the point of view presented is a personal one: it is the point of view which led the writer to the theory of turbulence he has recently developed. The mathematical theory itself (except for the brief summary in § XX) is not described in this essay.

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Le Theoreme de Convergence en Theorie du Potentiel

PAR

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I. Introduction

1 — Un des théorèmes fondamentaux de la théorie classique du potentiel établit, selon H. Cartan, que l'enveloppe inférieure d'une famille de fonctions surharmoniques $\geqslant 0$ (dans un ouvert) est égale à une fonction surharmonique sur le complémentaire d'un ensemble exceptionnel qui est localement "polaire" ou "de capacité extérieure nulle". Une forme équivalente est que l'enveloppe inférieure d'une famille de potentiels de mesures $\geqslant 0$ pris avec la fonction de Green comme noyau dans un domaine ω, vaut un potentiel de ce type sauf aux points d'un ensemble du type exceptionnel précédent (il est même équivalent de traiter lecas particulier où ω est sphérique et où les masses ne chargent qu'un compact de ω).

Voici un bref historique de ce théorème: RADO [8] considérant des suites décroissantes montra d'abord grâce aux fonctions presque sousharmoniques de SZPILRAJN que l'ensemble exceptionnel est de mesure de Lebesgue nulle; BRELOT [1]prouva ensuite qu'il est de capacité intérieure nulle. Enfin H. CARTAN [4], utilisant essentiellement la notion d'énergie, montra que la capacité extérieure est nulle et que le théorème, considéré à l'origine pour des suites, s'étend à un ordonné filtrant décroissant donc à l'enveloppe inférieure d'une famille quelconque de fonctions surharmoniques positives. Récemment BRELOT [2] y parvint sans l'énergie en reprenant sa démonstration initiale, remarquant que pour les suites les résultats récents sur la capacitabilité de CHOQUET [5] permettent de passer de la capacité intérieure nulle à la capacité extérieure nulle, et ramenant à une suite le cas d'une famille quelconque.

Nous nous proposons ici non seulement de simplifier encore, sans usage de l'énergie, mais, par un effort axiomatique, de donner dans le cas d'espaces et de noyaux beaucoup plus généraux un énoncé qui fournira aussitôt les résultats maintenant classiques. On énoncera d'ailleurs certains résultats préliminaires avec une généralité plus grande qu'il ne serait nécessaire ici.

II. Preliminaires

2—*Définition* 1. Soit E un espace localement compact. On appelle *noyau* toute application G de $E \times E$ dans $\bar{R}_+$ (c.à.d. $[0, +\infty]$) telle que pour tout $x \in E$, $G(x, y)$ soit une fonction mesurable de y pour toute mesure $\mu \geqslant 0$ de Radon sur E.

Pour une telle mesure $\mu \geqslant 0$ sur E (qui peut être définie comme fonctionnelle $\mu(f)$ de f finie continue et nulle hors d'un compact), l'application $G\mu$ de E dans $\bar{R}_+$ définie par $G\mu(x) = \int G(x, y)\, d\mu(y)$ s'appelle le *potentiel* de la mesure μ pour le noyau G.

On notera M_+ l'espace des mesures de Radon $\geqslant 0$ sur E (brièvement *mesures*), muni de la topologie faible bien connue (dite vague selon Cartan et Bourbaki).

On dira qu'une mesure $\mu \geqslant 0$ est portée par un fermé F de E si le complémentaire de E est de μ-mesure nulle; le support $S\mu$ de μ est le plus petit des fermés portant μ.

Lemme 1 : Pour que l'application $\Gamma : (x, \mu) \to G\mu(x)$ de $E \times M_+$ dans $\bar{R}_+$ soit semi-continue inférieurement (brièvement s.c.i.), il faut et il suffit que G soit s.c.i. dans $E \times E$.

Démonstration : Si E′ est la partie de M_+ formée par les mesures ponctuelles ε_x de masse-unité en x, l'application $x \to \varepsilon_x$ de E sur E′ est une homéomorphie qui identifie $E \times E$ avec la partie $E \times E'$ de $E \times M_+$. Donc si Γ est s.c.i, sa restriction à $E \times E'$ l'est aussi. C'est dire que $G\,\varepsilon_y(x)$ ou $G(x, y)$ est s.c.i, dans $E \times E$.

Inversement supposons que G soit s.c.i dans $E \times E$. C'est l'enveloppe supérieure d'une famille ordonnée filtrante croissante (G_i) d'applications de $E \times E$ dans R_+, finies continues et nulles hors d'un compact. Pour tout i, l'application $(x, \mu) \to G_i\mu(x)$ est continue dans $E \times M_+$; en effet elle s'écrit aussi $(x, \mu) \to \mu(G_{i,x})$ où $G_{i,x}$ est l'application $y \to G_i(x, y)$. On voit que c'est la composée de l'application continue $x \to G_{i,x}$ de E dans $C(E)$[1] et de l'application continue $(\mu, f) \to \mu(f)$ de $M_+ \times C(E)$ dans R.

1. C(E) est l'espace vectoriel des fonctions finies continues nulles en dehors d'un compact de E, un voisinage de l'origine étant défini comme devant contenir toutes les fonctions de module $< \varepsilon$ nulles hors d'un même compact.

Mais alors, comme $G\mu(x) = \text{Sup } G_i\mu(x)$, l'application $(x, \mu) \to G\mu(x)$ est s.c.i.

Corollaire: Si G est s.c.i. dans $E \times E$,

(a) Pour toute mesure μ, $G\mu$ est s.c.i dans E,

(b) Soit μ_n une suite de mesures convergeant faiblement vers μ; si pour tout $x \in E$ la suite $G\mu_n(x)$ est convergente, on a $\lim G\mu_n \geqslant G\mu$.

Définition 2. On dit qu'un *compact* K de E est de *G-capacité nulle* (G étant un noyau sur E) si, pour chaque mesure $\mu \neq 0$ portée par K, la restriction de $G\mu$ à K est non bornée.

On dit qu'un ensemble A de E est de *G-capacité intérieure nulle* si tout compact contenu dans E est de G-capacité nulle.

Ces notions offrent peu d'intérêt pour un noyau quelconque car un sous-compact d'un compact de capacité nulle peut alors ne pas être de capacité nulle. Pour éviter de telles singularités, on se restreint aux noyaux "réguliers":

Définition 3: On dit qu'un noyau G est *régulier* si pour toute mesure μ de support compact, la propriété que la restriction de $G\mu$ au support de μ soit finie continue entraîne que $G\mu$ soit finie continue dans E.

On sait d'après EVANS [6], p. 238-241, ou VASILESCO [9], que c'est le cas du noyau classique newtonien ou logarithmique en espace euclidien, donc de la fonction de Green. On étudiera ailleurs des critères généraux de régularité.

Lemme 2: Soit G un noyau s.c.i. sur E et soit μ une mesure non-nulle portée par un compact K de E et telle que la restriction de $G\mu$ à K soit bornée. Alors pour tout $\varepsilon > 0$ il existe un compact $K' \subset K$ tel que

(1) si μ' est la restriction de μ à K', la restriction de $G\mu'$ à K' soit continue

(2) on ait $\int d\mu - \int d\mu' < \varepsilon$

Démonstration: En effet d'après un théorème de Lusin généralisé, comme $G\mu$ est mesurable (puisque s.c.i) il existe $K' \subset K$ tel que $\int d\mu - \int d\mu' < \varepsilon$ et tel que la restriction de $G\mu$ à K' soit continue. Posons $\mu'' = \mu - \mu'$, mesure $\geqslant 0$. Le potentiel $G\mu$ est

somme des fonctions s.c.i. $G\mu'$ et $G\mu''$; sa restriction à K' ètant continue, il en est de même de celles des autres fonctions, en particulier de $G\mu'$.

Corollaire : Si G est s.c.i. régulier et si K est un compact de capacité non nulle, il existe μ (non nulle) sur K, de potentiel fini continu dans E.

C'est dans les cas classiques un résultat de DE LA VALLEE POUSSIN [7], p. 84.

Notons encore que la régularité de G entraine que si K est de G-capacité nulle, tout sous-compact de K l'est aussi.

3 —Enfin arrivons à la partie la plus nouvelle de cette étude, au lemme-clef:

Lemme 3 : Soit E un espace topologique à base dénombrable d'ouverts. Soit $(f_i)_{i\varepsilon I}$ une famille d'applications s.c.i. de E dans $\overline{R}$ (c.à.d. $[-\infty, +\infty]$). Pour tout $J \subset I$, posons $f_J = \inf_{i\varepsilon J}(f_i)$. Alors il existe une partie dénombrable I_o de I telle que si g est une application s.c.i. quelconque de E dans R, la relation $g \leqslant f_{I_0}$ entraîne $g \leqslant f_I$.

Démonstration : L'application $\mu \to \dfrac{\mu}{1+|\mu|}$ est une homéomorphie croissante de $\overline{R}$ sur $[-1, +1]$ et ramène le théorème au cas où les fonctions considérées prennent leurs valeurs dans un intervalle compact fini.

Soit alors $(\omega_n)_{n=1, 2,..}$ une base d'ouverts de E. Pour tout n, il existe dans I un élément i_n tel que :

$$(1)\quad \inf_{y\varepsilon\omega_n} f_{i_n}(y) - \inf_{y\varepsilon\omega_n} f_I(y) < \frac{1}{n}$$

Soit I_o l'ensemble des ces i_n. Considérons $x \varepsilon E$. On va établir le lemme en montrant que

$$(2)\ g \leqslant f_{i_n} \text{ quel que soit n entraîne } (3)\ g(x) \leqslant f_I(x)$$

En effet pour tout $\varepsilon > 0$, il existe dans la suite (ω_n) un ω_p contenant x, tel que $\dfrac{1}{p} \leqslant \varepsilon$ et que

(4) $g(y) \geqslant g(x) - \varepsilon$ pour tout $y \varepsilon \omega_p$.

Alors (1), (2) et (4) donnent

$$(1') \quad \inf_{y \varepsilon \omega_p} f_{i_p}(y) - \inf_{y \varepsilon \omega_p} f_I(y) < \varepsilon$$

$$(2') \quad \inf_{y \varepsilon \omega_p} g(y) - \inf_{y \varepsilon \omega_p} f_{i_p}(y) \leqslant 0$$

$$(4') \quad g(x) - \inf_{y \varepsilon \omega_p} g(y) \leqslant \varepsilon$$

d'où par addition membre à membre

$$g(x) - \inf_{y \varepsilon \omega_p} f_I(y) \leqslant 2\varepsilon \text{ et a-fortiori}$$

$$g(x) \leqslant f_I(x) + 2\varepsilon$$

Il s'ensuit (3) puisque ε est arbitraire.

Définition 4 : On appelle *adjoint* d'un noyau G sur E, le noyau G^* sur E défini par $G^*(x, y) = G(y, x)$. Il est s.c.i. si G l'est.

III. Le Theoreme de Convergence

4 —*Théorème* : Soit E un espace localement compact à base dénombrable d'ouverts[2] et soit G un noyau s.c.i. sur E dont l'adjoint soit régulier.

Soit $(\mu_i)_{i \varepsilon I}$ une famille de mesures sur E, $i_o \varepsilon I$, et J l'ensemble des i tels que $G\mu_i \leqslant G\mu_{i_o}$. On suppose que :

(a) Les potentiels $G\mu_i$ forment une famille ordonnée filtrante décroissante.

2. c.à.d. que E est homéomorphe à un ouvert d'un espace métrique compact.

(b) Pour tout compact K, les masses $\mu_i(K)$ ($i\varepsilon J$) sont majorées indépendamment de $i\varepsilon J$.

(c) Pour toute mesure fixe ν de support compact et de potentiel $G^*\nu$ fini continu, $\int G^*\nu d\mu_i^{CK} \to 0$ selon l'ordonné filtrant croissant des compacts K[3], indépendamment des $i\varepsilon J$ (μ_i^{CK} désignant la restriction de μ_i au complémentaire de K) (la condition (c) est vérifiée si E est compact).

Alors (A) Il existe une mesure μ telle que $G\mu$ vaille Inf $G\mu_i$ sauf sur un ensemble de G^*-capacité intérieure nulle.

(B) De facon plus précise, il existe une suite α_n d'éléments de I tels que

(1) μ_{α_n} converge faiblement vers une mesure μ.

(2) la suite $G\mu_{\alpha_n}$ est décroissante.

(3) $G\mu \leqslant \inf G\mu_i \leqslant \lim G\mu_{\alpha_n}$

ces trois fonctions étant égales sauf sur un ensemble de G^*-capacité intérieure nulle.

Démonstration : Comme $(\inf G\mu_i) = (\inf G\mu_i)_{i\varepsilon J}$, on se ramène au cas $J = I$. Alors $\int G^*\nu d\mu_i$ ou $\int G\mu_i d\nu$ est borné indépendamment de i; car cela est majoré par $\int G\mu_{i_o} d\nu$ ou $\int G^*\nu d\mu_{i_o}$ qui d'après (c) est fini.

La famille $G\mu_i$ répond aux conditions du lemme 3. On considérera une suite (i_n) formant la partie I_o de I qu'introduit ce lemme. On en déduit grâce à (a) une suite $\mu_{i'_n}$ telle que la suite $G\mu_{i'_n}$ soit décroissante, avec $G\mu_{i'_n} \leqslant G\mu_{i_n}$ Grâce à (b) on en tire une suite μ_{α_n} convergeant faiblement vers une mesure μ.[4] Posons $\lim G\mu_{\alpha_n} = \Phi$. D'après le corollaire du lemme 1, $G\mu \leqslant \Phi$ donc grâce

3. c.à.d. que pour ν et ε donnés on a $\int G^*\nu d\mu_i^{CK} < \varepsilon$ pour un compact K convenable, indépendamment de $i\varepsilon J$.

4. Cela résulte de ce que C(E) (considéré note 1) contient ici un ensemble dénombrable partout dense de fonctions φ pour chacune desquelles on pourra, grâce au procédé diagonal, faire converger $\mu_{\alpha_n}(\varphi)$ pour une suite α_n convenable extraite de la suite (i'_n).

au lemme 3, $G\mu \leqslant \inf G\mu_i$ pour tout i. Ainsi (μ_{α_n}) satisfait à (1), (2), et aux inégalités de (3).

Montrons enfin que l'ensemble A des x où $G\mu(x) \neq \Phi$ est de G*-capacité intérieure nulle: Sinon A contiendrait un compact K de G*-capacité non nulle. D'après le corollaire du lemme 2 il existerait une mesure $\nu \neq 0$ portée par K, de potentiel $G^*\nu$ fini continu dans E. Or

$\int G\mu_{\alpha_n} d\nu = \int G^*\nu d\mu_{\alpha_n}$, quantité bornée indépendamment de n, par une constante k finie. Le premier membre tend vers $\int \Phi d\nu$ fini. On aura la contradiction cherchée en montrant que le second membre tend vers $\int G^*\nu d\mu$ ou $\int G\mu d\nu$. Cette limite est évidente si E est compact. Sinon, soit K un compact de E et φ_k une fonction continue $\geqslant 0$ majorée par $G^*\nu$, égale à $G^*\nu$ dans K et nulle hors d'un compact. D'abord

$\int \varphi_k d\mu_{\alpha_n} \leqslant k$ donc $\int \varphi_k d\mu \leqslant k$ et enfin $\int G^*\nu d\mu \leqslant k$. Puis $\left| \int G^*\nu d\mu_{\alpha_n} - \int G^*\nu d\mu \right| \leqslant \int (G^*\nu - \varphi_k) d\mu_{\alpha_n} + \left| \int \varphi_k d\mu_{\alpha_n} - \int \varphi_k d\mu \right| + \int (G^*\nu - \varphi_k) d\mu$. Le premier et le troisième terme sont majorés par $\int_{CK} G^*\nu d\mu_{\alpha_n}$ et $\int_{CK} G^*\nu d\mu$; donc en vertu de la condition (c), ε étant donné, ils sont majorés par ε pour un choix convenable de K indépendamment n. Si K est fixé le second terme est $< \varepsilon$ pour n assez grand. Il s'ensuit que le premier membre tend vers 0 pour $n \to \infty$.

5 — *Applications.* D'autres énoncés voisins résultent immédiatement des remarques suivantes:

(1) Si G est strictement positif, (b) est conséquence de (c);

(2) Si les $\mu_i(E)$ sont bornés, (b) est satisfait, et (c) l'est sous l'une des hypothèses suivantes :

(α) les μ_i ne chargent qu'un même compact,

(β) pour toute ν de support compact dans E non compact, $G^*\nu$ tend vers 0 au point d'Alexandroff de E.

(3) Supposons G tel que, quelles que soient μ et μ', Inf $(G\mu, G\mu')$ soit un potentiel et que $G\mu \leqslant G\mu'$ entraîne $\mu(E) \leqslant \mu'(E)$.

Alors, si pour un indice $i_o \varepsilon I$, $\mu_{i_o}(E)$ est fini, on a la propriété (A) sans les hypothèses (a), (b), (c), mais avec (β).

En effet considérons les enveloppes inférieures d'un nombre fini quelconque des $G\mu_i$, puis leurs enveloppes inférieures ave $G\mu_{i_o}$. On obtient des potentiels de certaines mesures μ_i'. L'enveloppe inférieure des $G\mu_i'$ vaut celle des $G\mu_i$ et les μ_i' satisfont, sous les conditions ci-dessus, à (a), (b), (c).

(4) Appelons G-capacité d'un compact K la borne supérieure des charges $\mu(K)$ des mesures de support dans K telles que $G\mu \leqslant 1$ sur K. La G-capacité intérieure d'un ensemble A est la borne supérieure des G-capacités des compacts contenus dans A; la G-capacité extérieure de A est la borne inférieure des capacités intérieures des ouverts contenant A. Un ensemble A est dit G-capacitable si ses deux capacités précédentes coïncident (Voir [5]).

Supposons que les boréliens soient G-capacitables, comme c'est le cas d'après [5] pour le noyau newtonien dans R^n ($n \geqslant 3$) ou de la fonction de Green dans un domaine euclidien ou même un *espace de Green* (Voir [3]). Alors l'ensemble exceptionnel du théorème, où $G\mu = \lim G\mu_{\alpha_n}$ étant borélien, est de G-capacité *extérieure* nulle.

6 — *Cas des potentiels de Green dans un espace de Green* Ω.[5]

On prendra pour G la fonction de Green de Ω; on a ici $G^* = G$.

Proposition : Soit μ_i une famille de mesures dans Ω, admettant chacune un potentiel de Green $G\mu_i$ non partout infini. Il existe une mesure μ telle que $G\mu = \inf G\mu_i$ sauf sur un ensemble de capacité extérieure nulle.

Démonstration : On se ramène d'abord immédiatement au cas où il existe un $G\mu_{i_o}$ qui majore tous les $G\mu_i$, et où la famille est filtrante décroissante. D'aprés (5, 1), il suffit alors de montrer que les (μ_i) satisfont à (c).

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5. Dans le cas particulier d'un domaine euclidien borné par exemple, on peut, au lieu du lemme topologique 3, utiliser la propriété suivante: soit (v_i) une famille filtrante décroissante de fonctions surharmoniques $\geqslant 0$, sommables dans un domaine euclidien. Si une suite extraite décroissante v_{i_n} satisfait à $\int v_{i_n} d\omega \to \inf \int v_i d\omega$, toute fonction semi-continue inférieurement minorant inf v_{i_n} minore inf v_i; autrement dit la régularisée de lim v_{i_n} minore inf v_i.

Or $G\mu_i^{CK}$ est, dans $\overset{o}{K}$, (intérieur du compact variable K), une fonction harmonique majorée par la plus grande minorante harmonique de $G\mu_{i_o}$ dans $\overset{o}{K}$. Comme celle-ci tend vers O[6], d'ailleurs uniformément sur tout compact, selon le filtre des K, il en est de même de $G\mu_i^{CK}$ indépendamment de i. De sorte que $\int G\mu_i^{CK} dv$ ou $\int Gv d\mu_i^{CK}$ tend vers O pour toute v du théorème, indépendamment de i.

Remarque : La démonstration est particulièrement simple dans le cas où Ω est l'intérieur d'une boule (ou l'extérieur de cette boule complété par le point à l'infini) et où les supports des μ_i sont dans un même compact de Ω. De ce cas particulier on pourrait passer aussi au cas général.

Corollaire : On peut déduire de là, comme on le sait (Voir par example Cartan [4]) qu'une famille de fonctions surharmoniques $\geqslant$ O dans un espace de Green admet comme enveloppe inférieure une fonction égale à une fonction surharmonique, sauf aux points d'un énsemble de capacité extérieure nulle, ou ce qui est équivalent, polaire.[7]

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6. Cette propriété d'un potentiel quelconque se démontre en se ramenant au cas où μ_{i_o} est une masse ponctuelle unité; et dans ce cas ce n'est autre qu'une des propriétés fondamentales de la fonction de Green.

7. Un ensemble est polaire s'il est localement une partie de l'ensemble des infinis d'une fonction surharmoniqu.

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Semiothisa Pervolgata wlk., A Semilooper Pest on Sesbania Aculeata (Dhaincha)*

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Introduction :

Sesbania aculeata is one of the popular and important green manure crops grown largely in Madras State. It is subjected to the attack of number of pests. Ramakrishna Iyer (1940) noted the following pests :—

Azygophleps scalaris F., *Prodenia litura* F., *Alcidodes bubo* F., *Laphygma exigua* H., *Pericyma glaucinans* Hb., *Terias hecabe* L., *Catopsilia pyranthe* L., *Metripes* G., *Coptosoma cribraria* F., *Brachyplatis vahlii* F., and *Semiothisa pervolgata* Wlk. as pests of *Sesbania aculeata.* Among them *Semiothisa pervolgata* Wlk. is a serious pest belonging to the family Geometridae of the order Lepidoptera.

Distribution :

Semiothisa pervolgata Wlk. is distributed in Bengal, Poona and in Southern India (Hampson, 1895). In the Madras State it was noted in Coimbatore in 1936 doing severe damage to Dhaincha for the first time. Soon, an attempt was made to study the pest by Cherian and Rangiah Pillai (1938) and they have given broad outlines on its life history and methods of control. The present contribution gives a more detailed knowledge on its biology, bionomics and latest control measures.

Nature, symptoms of damage and alternate host plants :

It has no alternate host plants and is deemed to be a specific pest of Dhaincha, appearing in a pest form during the months

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of February and March, June to November when the crop is on the field. It assumes a characteristic position on the leaflets in that it clings to the edge of the blade in a looping fashion (Plate I, Fig. 1). When the attack is severe, it completely defoliates the entire plant leaving only the mid-rib and stem resulting in loss of all green matter.

Life history, habits and description of stages.

Copulation and oviposition :

Two days after the emergence of adults, copulation between female and male moths take place. The female moth after copulation rests on the branches of the side shoots and lays eggs. The eggs are laid singly on the leaves and tender shoots. A single female moth has been found to lay as many as 290 eggs at a time.

Egg : (Plate I, Fig. 2). The eggs are blue green in colour and oval in outline with a depression in the middle and measuring 0·577 mm. The egg period varies from 3 to 4 days.

Larva :

1st Stage : The newly hatched caterpillar is studded with tiny hairs all over the body and measures 1·5 m.m. in length. The head capsule is light brown or orange yellow and the thorax and abdomen dark green in colour. On the second abdominal segment, two black dots are very prominently noted. The prolegs are borne on the sixth and last abdominal segments. There are four moults during the larval stage before it undergoes pupation at its fifth moult.

2nd Stage: The larva after first moult measures about 5 to 6 m.m. and is light green in colour. The head region is globular with black dots. The thoracic legs are black in colour. Prolegs are slightly pinkish in tinge.

3rd Stage : After the second moult, the larva measures 1 cm. and turns ashy green with violet coloured intersegmental areas. There are a number of white parallel lines running on the dorsal aspect of the body. Head shield at this stage has a number of black lines and dots and the spiracle is circular with black dots.

PLATE I

Stages of *Semiothisa pervolgata Wlk.*

FIG. 1. Larva on *Sesbania aculeata* plant.

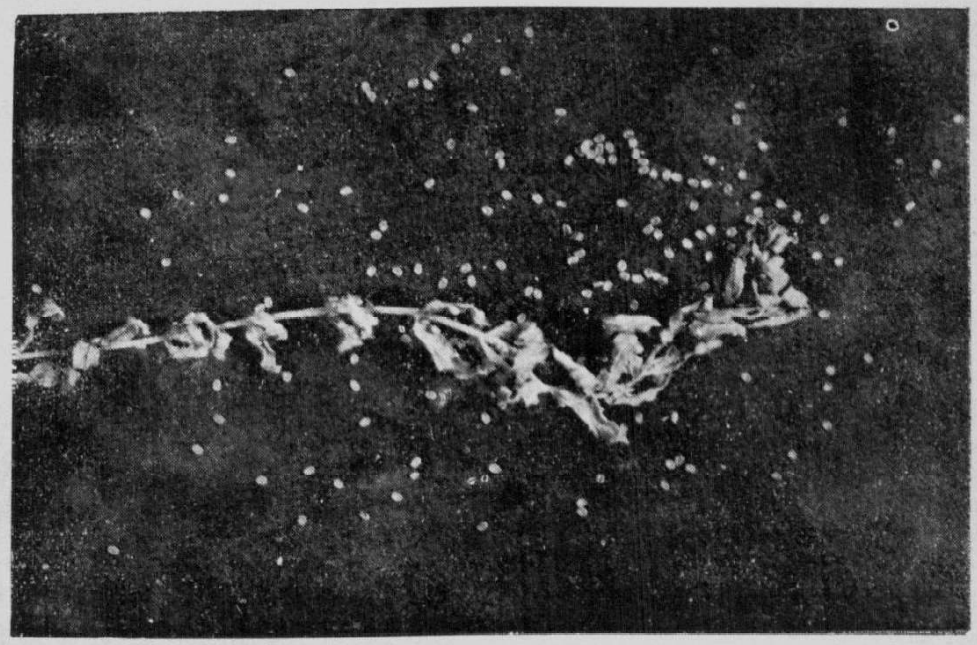


FIG. 2. Egg.

4th Stage: The larvae in this stage attains a length of 1·8 cm. and has a dark green body with parallel white lines more patent on the dorsal aspect of the body. The black dots and lines become darker in colour on the head region. There is distinct yellow line flanking the sides of the larvae. Minute short hairs are seen arising on either side of the mid-dorsal line. Small blackish hairs arise on the lateral sides of the head region.

5th Stage: The larva is pink and measures 2·7 cm. in length. There are horizontal bands of violet patches confined to the dorsal aspect of each segment and the head is bigger and mottled. Body is smooth and wrinkled in appearance with isolated and obscure short hairs. Spiracles are large, oval with pale centre and black rims. Prolegs are stout, fleshy and strong and occur as two pairs.

Pupa (Plate II, Fig. 3). The full grown larva enters the soil for pupation and has larval period of 10 to 12 days. The pupa is cylindrical, reddish brown and measuring 1·3 cm. In fields where the soil is hard, naked pupae have been collected from crackes and crevices. The pupal period is 6 to 8 days.

Adult (Plate II, Fig. 4). A detailed description of the moth of *Semiothisa pervolgata* Wlk. belonging to the sub-family Boarmiinae is described by Hampson (1895), as follows: "Whitish, slightly suffused with brown and striated with fuscous. Fore-wing with obliquely waved ante-medial, medial and post medial dark lines angled below the costa; a black spot at the end of cell; some fuscous suffusion beyond the post medial line; a dark patch on costa and spot below vein 5; a series of marginal black specks. Hind-wing with black spot at the end of cell; ante and post medial waved lines, the latter with fuscous suffusion beyond it and a black spot or spots at vein 2; a marginal crenulate black line. Underside the marginal area largely suffused with rufous and with white patches on it."

Total life cycle. Text Fig. 1. The total life cycle is about 20 to 23 days, the egg, the larval and pupal periods being 3 to 4, 10 to 12 and 6 to 8 days (Table I).

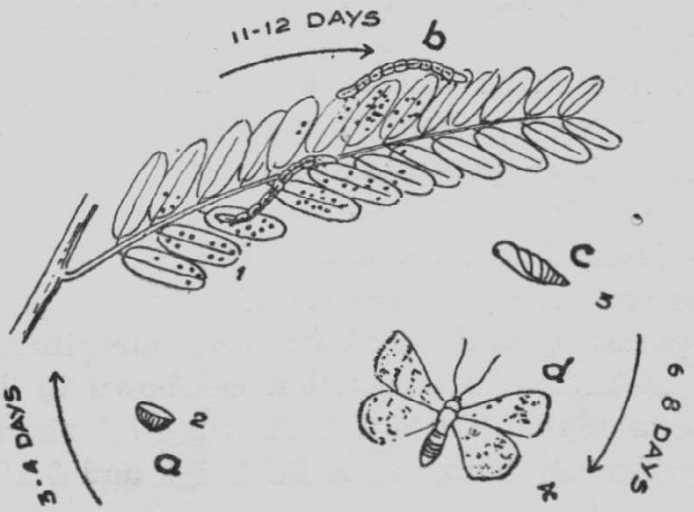
Natural enemies: In nature, the pest is controlled to a larger extent by a braconid, *Apanteles hypsipylae,* Wlk. In June they appear in large numbers and attack the caterpillar in all stages of growth. Medium sized caterpillars are highly susceptible. The caterpillars attain a greenish yellow colour and for all purposes

TABLE 1

Summary of life history of *Semiothisa pervolgata* Wlk. during 1953 to 1955.

S. No.	Date of egg laying.	Date of hatching.	Incubation period.	Date of pupation.	Larval period.	Adult emerged on.	Pupal period.	Total in days.
1.	2— 2—53	5— 2—53	3	15— 2—53	11	22— 2—53	8	22
2.	9— 7—53	12— 7—53	3	24— 7—53	12	31— 7—53	8	23
3.	20— 7—53	24— 7—53	4	3— 8—53	11	9— 8—53	7	22
4.	18— 8—53	21— 8—53	3	31— 8 53	10	6— 9—53	7	20
5.	12— 9—53	16— 9—53	4	25— 9—53	10	30— 9—53	6	20
6.	18—12—53	22—12—53	4	2— 1—54	11	8— 1—54	7	22
7.	7— 2—54	10— 2—54	4	21— 2—54	10	25— 2—54	6	20
8.	16— 7—54	19— 7—54	3	30— 7—54	11	6— 8—54	8	22
9.	14— 8—54	17— 8—54	3	28— 8—54	11	2— 9—54	6	20
10.	10— 9—54	13— 9—54	3	24— 9—54	11	30— 9—54	7	21
11.	18—12—54	22—12—54	4	11—12—54	10	6— 1—55	6	20
12.	4— 8—55	8— 8—55	4	19— 8—55	11	25— 8—55	7	22
	Maximum		4		12		8	23
	Minimum		3		10		6	20
	Average		3·6		10·8		6·9	21·2

appear very active till the full grown grubs come out. Besides these, there are 2 kinds of hyper parasites noted. Two pentatomids



TEXT FIG. 1. Life history stages of *Semiothisa pervolgata* Wlk. (a) Egg; (b) Caterpillar; (c) Pupa; (d) Adult

and a reduvid were noted by Cherian and Brahmachari (1941) to be predaceous on the caterpillar. Both nymphs and adults of *Contheconidia furcellata* Wolff, were observed to attack the larvae. *Andrallus spinidons* F. attacks the larvae but it occurs very late in April, when the pest has almost completed one generation.

Seasonal history :

The seasonal history of this pest was studied in Coimbatore. It is noted during the months of February to April and from June to November. About four broods were observed during the duration of the crop in the field.

Control measures :

Light trap was found to be effective, as it gave encouraging results. Calcium arsenate mixed with lime (1 : 6) applied to the plants on a small scale, both as dust and spray is not very effective and economical to use in the long run. In the present studies the trial on *Semiothisa pervolgata* Wlk. was conducted in the Central Farm Wetlands. Of the recently introduced insecticides the following chemicals DDT, BHC, Aldrin and Dieldrin in varying proportions were tried for each experiment. For each treatment 5 cent plots were taken and initial count of the population on 30 plants at random were taken prior to treatments. The plots for

different treatments were selected at different corners so as to avoid the effect of each chemical on one and the same plot. The following treatments were tried. (1) BHC 5%. (2) D.D.T. 5%. (3) BHC. 0·1%. (4) DDT 0·1%. (5) Aldrin 0·1%. (6) Dieldrin 0·1% and (7) Control. One 5 cent plot was set apart as control and this was located as distant from treated plots as possible. Mortality counts were taken 24 and 48 hours after treatment. The results have been presented in Table 2. It will be observed from the table that BHC 0·1% and Dieldrin 0·1% gave cent percent kill. Next in descending order come BHC 5%, DDT 5%, and DDT 0·1%. In the control there was slight reduction probably due to the effect of parasites on the medium sized caterpillars. The results were put to statistical interpretation as shown in Table 3. It is seen from the table that BHC 0·1%, Aldrin 0·1% and Dieldrin 0·1% were not significantly better than DDT 5% and 0·1%.

TABLE 2

Control trials on *Semiothisa pervolgata* Wlk.

S. No.	Treatments.	Initial Counts.	Counts after 24 hours.	Counts after 48 hours.	% of reduction in population after 48 hours.
1.	BHC 5%	42	5	3	93
2.	DDT 5%	35	12	7	80
3.	BHC ·1%	28	2	Nil.	100
4.	DDT ·1%	20	14	3	80
5.	Aldrin ·1%	30	6	Nil.	100
6.	Dieldrin ·1%	40	8	Nil.	100
7.	Control	45	30	25	

TABLE 3

Statistical analysis of the control trials on *Semiothisa pervolgata Wlk.*

S. No.	Treatments.	Difference in proportion.	Standard error.	Significance.
1.	DDT 5% / BHC ·1%	0·20	·079	No.
2.	DDT 5% / Aldrin ·1%	0·20	·078	No.
3.	DDT 5% / Dieldrin ·1%	0·20	·069	No.

PLATE II

Stages of *Semiothisa pervolgata* Wlk.

FIG. 3. Pupa.

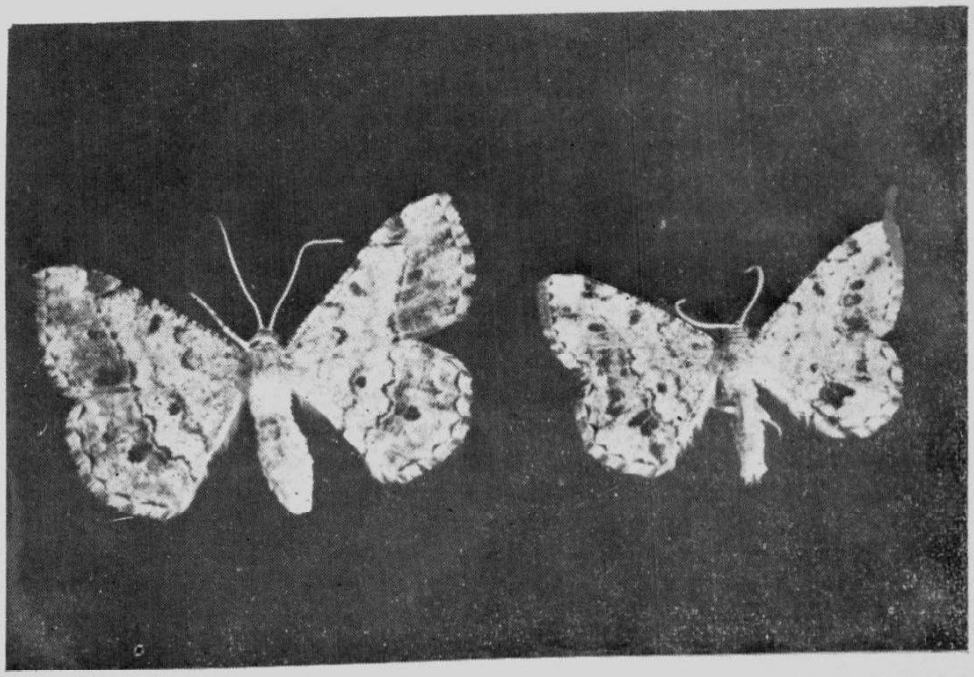


FIG. 4. Adult.

Summary :

Among the cultivated green manure crops in Southern India, *Sesbania aculeata* is frequently affected by a Geometrid pest, *Semiothisa pervolgata* Wlk. It was noted for the first time in Coimbatore during the year 1936 doing extensive damage to Daincha. It attacks the leaflets of the plant and in severe attack only naked stems are left behind, devoid of any green matter.

The moth lays eggs on the tips of leaflets and on buds. A single female is capable of laying upto 290 eggs. The egg period lasts from 3 to 4 days. The caterpillar undergoes 4 moults and at its fifth moult it pupates in the soil after a larval period of 10 to 12 days. The adult moth emerges after a pupal period of 6 to 8 days. The entire life cycle of the pest from egg to adult is completed in about 20 to 23 days. In nature, the pest is controlled to a large extent by a braconid *Apanteles hypsipylae* Wlk., but two kinds of hyper parasites are noted. It can be controlled to some extent by light traps by attracting gravid females. Calcium arsenate dust mixed with lime in proportion of 1 : 6 is also somewhat effective. Application of 5% BHC dust or spray of 0·1% Wettable BHC, Aldrin and Dieldrin completely controls the pest.

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On Strictly Convex Spaces

BY

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In this paper a new definition of strict convexity (s.c.) of a normed linear space is given, and in theorem I it is proved to be equivalent to the definition used by Clarkson (1). In theorem II I have obtained a necessary and sufficient condition that a seminorm is s.c. which is different from the one given by Smiley (2) in that the "betweenness" has been replaced by a single point namely the "mid-point". In theorem III the necessary and sufficient condition that F-space may be normed by a s.c. norm is obtained.

Section 1. Let B be a linear space; if it is normed (seminormed) the norm (seminorm) is denoted by $||\ ||$ ($|\ |$). Let a, b, c, ... x, y, z, ... be the elements of B and θ be the null element.

Def. 1. We define the $||\ ||$ to be s.c. if $||x||=||y||=1$, $||x+y||=||x||+||y||$ together imply $y=x$.

Def. 2. Clarkson (1) defines the $||\ ||$ to be s.c. if $||x+y||=||x||+||y||$, implies $y=\lambda x$ $(\lambda>0)$.

These definitions can be extended to seminorms also.

Theorem 1. Definitions 1 and 2 are equivalent.

Let $||\ ||$ satisfy conditions of def. 1 and let a, b, ε B such that $||a+b||=||a||+||b||$ ..... (1)

If possible let $b=\lambda a$; and $c=a+b$. Then

$$||\frac{c}{||c||}+\frac{a}{||a||}||\leqslant||\frac{c}{||c||}||+||\frac{a}{||a||}||$$

$$\leqslant 2 \quad \text{..... (2)}$$

The equality sign in (2) may be dropped for in that case by def. 1 the result follows:

From inequality (2) $||\, c\,||\, a\,|| + a\,||\, c\,||\,|| < 2\,||\, a\,||\,||\, c\,||$

Similarly $||\, c\,||\, b\,|| + b\,||\, c\,||\,|| < 2\,||\, b\,||\,||\, c\,||$

Hence $||\, c\,(||\, a\,|| + ||\, b\,||) + ||\, c\,||\, (a+b)\,||$

$$< 2\,||\, c\,||\,(||\, a\,|| + ||\, b\,||)$$

i.e. $2\,||\, c\,||^2 < 2\,||\, c\,||^2$ which is a contradiction.

Hence $b = \lambda a$. Substituting λa for b in equation (1) λ is seen to be positive.

Next we shall prove that def. 2 implies def. 1.

Let $||\ ||$ satisfy def. 2, and let

$||\, a+b\,|| = ||\, a\,|| + ||\, b\,||$, so that $b = \lambda a \quad (\lambda > 0)$,

If further $||\, a\,|| = ||\, b\,|| = 1$, then it follows $\lambda = 1$ or $a = b$.

Section 2. If $a, b, c, \ldots\ldots$ be elements of a linear space. then b is said to be the algebraic mid-point of a, c if $b = \frac{1}{2}\,(a+c)$, and is a metric mid-point of a, c, if $d(a, b) = d(b, c) = \frac{1}{2}\,d(a, c)$.

Lemma: In an F-space the algebraic mid-point is also a metric mid-point.

For $d(a, b) = d(a-b, \theta)$ where d is the F-metric.

Theorem 2. The necessary and sufficient condition that a seminorm $(|\ |)$ is s.c. is that any pair of elements has a unique metric mid-point.

Proof: Suppose that the seminorm, $|\ \ |$ is s.c. and let 'b' be the unique metric mid-point of a, c.

i.e. $|\, a-b\,| = |\, b-c\,| = \frac{1}{2}\,|\, a-c\,| = t$.

Hence $|\, a/t - b/t\,| = |\, b/t - c/t\,| = \frac{1}{2}\,|\, a/t - c/t\,| = 1$

and $|\, a/t - b/t\,| + |\, b/t - c/t\,| = |\, a/t - c/t\,|$

Since $|\ \ |$ is s.c. by def. 1 it immediately follows that b is the algebraic mid-point of a, c, which is unique. Hence the condition is necessary. It is also sufficient. For the metric mid point being

unique coincides with the algebraic mid-point. $|x| = |y| = 1$, and $|x+y| = |x| + |y|$. Hence,

$$d(x, \theta) = d(\theta, -y) = \tfrac{1}{2}\, d(x, -y) = 1$$

i.e. $\theta = \frac{1}{2}\ (x - y)$ or $x = y$.

Section 3. If $a, b, c, \ldots\ldots$ are elements of B, b is said to be algebraically between a, c if

$$b = ta + (1-t)c,\ 0 \leqslant t \leqslant 1,$$

and metrically between a, c if

$$d(a, b) + d(b, c) = d(a, c).$$

It is obvious that any point algebraically between x, x is x. In what follows $d(x, y)$ is a metric of the F-type over a linear space.

Theorem 3. The necessary and sufficient condition that $d(x, \theta)$ becomes a s.c. norm is that metric betweenness is identical with algebraic betweenness.

Proof: Let $||x|| = d(x, \theta) = d(\theta, x)$. We shall show $||x||$ thus defined is a norm.

If $||x|| = 0$ then $d(x, \theta) + d(\theta, x) = d(x, x) = 0$

Hence θ is algebraically between x, x. i.e. $x = \theta$.

Since metric betweenness coincides with algebraic betweenness and d is of the F-type for λ, $+ve$ integral

$$||\lambda x|| = d(\lambda x, \theta) = \sum_{p=0}^{\lambda-1} d(\overline{\lambda - px} - \overline{\lambda - p - 1\, x}, \theta)$$

$$= \lambda d(x, \theta)$$

If λ is $a - ve$ integer, let $\lambda = -k$,

$$||\lambda x|| = d(-kx, \theta) = kd(-x, \theta) = |\lambda|\, d(x, \theta).$$

If λ is rational equal to p/q, q positive,

$$q\, ||\frac{p}{q} x|| = ||px|| = |p|\, ||x||.$$

Hence $||\frac{p}{q}x|| = |\frac{p}{q}|\,||x||$.

Next λ be irrational, and let $\frac{p_n}{q_n}$ be a sequence of rationals such that $\underset{n\to\infty}{\text{Lim}} \frac{p_n}{q_n} = \lambda$.

Since $d(\frac{p_n}{q_n}x,\ \theta) = |\frac{p_n}{q_n}|\,d(x,\ \theta)$ by the continuity of multiplication of the F-metric it follows that $d(\lambda x,\ \theta) = |\lambda|\,d(x, \theta)$ when λ is irrational. Thus for real λ,

$$||\lambda x|| = |\lambda|\,||x||.$$

It is easily verified that $||x||$ satisfies the other postulates of the norm. Hence the condition is sufficient. The condition is also necessary for if $||x|| = d(x, \theta)$ is a s.c. norm the metric betweenness coincides with the algebraic betweenness.

In conclusion I wish to thank Dr. M. V. Subba Rao for his help in the preparation of the paper.

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A Note on Some Properties of the Power Function of Chisquare*

BY

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SUMMARY

Some properties of the power function of the chisquare distribution are considered in this paper. It is noted that when the non-central distribution is used as an alternate test criterion for hypothesis H_1 against chisquare itself as H_0, the power is seen to increase with increase in degrees of freedom and decrease with increase in the the value of the parameter λ. When both vary the power decreases if the ratio $\frac{\lambda}{n}$ is greater than 0·2.

Let ξ_i, $i = 1, 2, \ldots n$ be 'n' normal variates with zero mean and unit variance. Then the sum $\sum_1^n \xi_i^2$ follows the well known χ^2-distribution with 'n' degrees of freedom. If, however, the ξ_i's are normal variates with unit variance, but with different means a_i $(i = 1, 2, \ldots n)$, Fisher (1920) derived the distribution of $\sum_1^n \xi_i^2$ as a particular case of multiple correlation coefficient. Later Tang (1938) gave an analytical proof for the same distribution which is known as non-central χ^2 and is usually denoted by χ'^2. This distribution is given by

$$p(\chi'^2) = e^{-\lambda/2} \sum_{j=0}^{\infty} \left(\frac{\lambda}{2}\right)^j \frac{1}{j!} \frac{e^{-\chi'^2/2} \left(\frac{\chi'^2}{2}\right)^{\frac{n}{2}+j-1}}{2^{\frac{n}{2}+j} \Gamma\left(\frac{n}{2}+j\right)} \quad \ldots \quad (1)$$

with 'n' degrees of freedom, where 'λ' is the non-central parameter. If the variance is σ^2, then $\sum_1^n \xi_i^2$ is distributed as $\sigma^2 x'^2$.

* Some of these results are included in a thesis accepted for the M.Sc. Degree by the University of Madras.

Let the test criterion for a hypothesis 'H' depend on the statistic $\chi^2 = \sum_1^n \xi_i^2$ where ξ's are normal variates with zeromean and unit standard deviation and let χ'^2 be the criterion for the hypothesis H_1, with n-degrees of freedom, and a non-central parameter λ. Then the power of the test w.r.t. an alternative hypothesis, H_1 is $\beta(n, \lambda, \alpha) = \int_{\chi_\alpha^2}^{\infty} pn(\chi'^2/\lambda)\, d\chi'^2$ where χ_α^2 is the 100 α% point of the χ^2-distribution with 'n' degrees of freedom, and this is given by

$$\beta(n, \lambda, \alpha) = \sum_{j=0}^{\infty} \frac{e^{-(\lambda/2)}(\lambda/2)^j}{j!\,\Gamma((n/2 + j)} \int_{\chi_\alpha^2}^{\infty} e^{-(\chi'^2/2)} (\chi'^2/2)^{(n/2)+j-1} d(\chi'^2/2) \quad .. \quad (2)$$

since term by term integration is justified on account of uniform convergence of the series.

From (2) it can be seen that the power function depends on $\lambda = \sum_{i=1}^{\infty} a_i^2$, the overall non-centrality so that the power, remains unaltered for any set of a_i's that will give the same λ.

Power as a decreasing function of the degrees of freedom.

Let the significance level α be kept constant. Then (2) can be written as

$$\beta(m, \lambda) = \sum_{j=0}^{\infty} \frac{e^{-(\lambda/2)}(\lambda/2)^j}{j!\,\Gamma(m+j)} \int_{x_\alpha}^{\infty} e^{-x} x^{m+j-1}\, dx \quad .. \quad (3)$$

where $\frac{\chi'^2}{2} = x$, $\frac{\chi_\alpha^2}{2} = x_\alpha$ and $\frac{n}{2} = m$.

Also from the definition of x_α

$$\alpha = \frac{1}{\Gamma(m)} \int_{x_\alpha}^{\infty} e^{-x} x^{m-1}\, dx \quad .. \quad (4)$$

If λ is fixed, and m increases to $(m+1)$ let $\chi_a'^2$ be the $100\,\alpha\%$ point of the χ^2-distribution with $2(m+1)$ degrees of freedom. Thus we have corresponding to (3) and (4), the relations

$$\beta(m+1,\ \lambda) = \sum_{j=0}^{\infty} \frac{e^{-(\lambda/2)}(\lambda/2)^j}{j!\,\Gamma(m+j+1)} \int_{x'_a}^{\infty} e^{-x} x^{m+j} dx \quad .. \quad (5)$$

$$\text{and } \alpha = \frac{1}{\Gamma(m+1)} \int_{x'_a}^{\infty} e^{-x} x^m dx \quad .. \quad (6)$$

Integrating by parts the above integral, (6) becomes

$$\alpha = \frac{1}{\Gamma(m+1)} \left[m \int_{x'_a}^{\infty} e^{-x}\, x^{m-1}\, dx + e^{-x'_a}\, x_a'^m \right]$$

substituting the value of α given by (4), it follows

$$\int_{x_a}^{x'_a} e^{-x}\, x^{m-1}\, dx = (e^{-x'_a}\, x_a'^m)/m \quad .. \quad (7)$$

Since the expression on the right of (7) is positive, m being greater than zero, the integral on the left is positive. Hence it follows that

$$x'_a > x_a.$$

The inequality $\beta\,(m+1,\ \lambda) < \beta\,(m, \lambda)$ follows if

$$\frac{1}{\Gamma(m+j+1)} \int_{x'_a}^{\infty} e^{-x}\, x^{m+j}\, dx < \frac{1}{\Gamma(m+j)} \int_{x_a}^{\infty} e^{-x}\, x^{m+j-1}\, dx \quad .. \quad (8)$$

is true. Integrating by parts the left side of (8). and simplifying (8) becomes

$$\frac{e^{-x_a'}\, x'^{m+j}_a}{(m+j)} < \int_{x_a}^{x'_a} e^{-x}\, x^{m+j-1}\, dx \quad .. \quad (9)$$

By using equation (7) the inequality (9) reduces to

$$\frac{m x'^j_a}{(m+j)} \int_{x_a}^{x'_a} e^{-x}\, x^{m-1}\, dx < \int_{x_a}^{x'_a} e^{-x}\, x^{m+j-1}\, dx.$$

Since $e^{-x} > 0$, the above inequality holds if

$$x'_a{}^j \, (x'_a{}^m - x_a{}^m) < (x''_a{}^{m+j} - x_a{}^{m+j})$$

i.e., if $x'_a{}^j > x_a{}^j$ which is true as already noted. Hence the inequality (8) is true for every j. So $\beta(m+1, \lambda) < \beta(m, \lambda)$. From this result it can be noted that when n increases by 2 or multiples of 2 the power decreases. Since both χ^2-curve and corresponding χ'^2-curve move steadily to the right we infer that the power function must steadily decrease when 'n' increases.

Power as an increasing function of the non-central parameter λ.

Write (2) as a series of integrals so that

$$\beta(m, \lambda) = \sum_{j=0}^{\infty} \frac{e^{-(\lambda/2)} (\lambda/2)^j}{j!} \int_{x_a}^{\infty} \frac{e^{-x} x^{m+j-1}}{\Gamma(m+j)} dx. \quad .. \quad (10)$$

Differentiating (10) w.r.t. λ and arranging the result by combining the part obtained by differentiating $(\lambda/2)^j$ in the $(j+1)^{\text{th}}$ term with the part obtained by differentiating $e^{-(\lambda/2)}$ in the j^{th} term, gives

$$\frac{\partial \beta}{\partial \lambda} = \sum_{j=0}^{\infty} I_j.$$

$$\text{where } I_j = \frac{e^{-(\lambda/2)} (\lambda/2)^j}{2 \cdot j!} \left[\int_{x_a}^{\infty} \frac{e^{-x} x^{m+j}}{\Gamma(m+j+1)} - \int_{x_a}^{\infty} \frac{e^{-x} x^{m+j-1}}{\Gamma(m+j)} dx \right] \quad .. \quad (11)$$

The term by term differentiation is justified since the series (10) is uniformly convergent for $\lambda \geqslant 0$. Integrating the first integral in equation (11) by parts and simplifying we get

$$I_j = \frac{e^{-(\lambda/2)} (\lambda/2)^j}{2 \cdot j!} \left[\frac{e^{-x_a} x_a{}^{m+j}}{\Gamma(m+j+1)} \right]$$

which is always positive for all j. (x_a and λ being always positive).

Hence, $\frac{\partial \beta}{\partial \lambda}$ is positive so that $\beta(m, \lambda)$ is an increasing function of λ.

Changes in Power when n and λ both vary.

Next we shall consider the power when n and λ both vary. Consider when $\frac{\lambda}{n} = \text{const.}$, say c^2. Let the increase in n be $\triangle n$ and the corresponding increase in λ and β be $\triangle\lambda$ and $\triangle\beta$ respectively.

Omitting differences in derivatives of higher order than one we have

$$\frac{\triangle\beta}{\triangle n} \simeq \frac{\partial\beta}{\partial\lambda}\frac{\triangle\lambda}{\triangle n} + \frac{\beta(n+\triangle n, \lambda) - \beta(n,\lambda)}{\triangle n}$$

$$\simeq \frac{\partial\beta}{\partial\lambda}c^2 + \left[\frac{\beta(n+\triangle n,\lambda)\,\beta(n,\lambda)}{\triangle n}\right] \quad .. \quad (12)$$

It has been shown in the preceding section that $\frac{\partial\beta}{\partial\lambda}$ is positive, and the expression in the square bracket is negative. Hence the sign of $\frac{\triangle\beta}{\triangle n}$ will depend on the relative magnitudes of $\frac{\partial\beta}{\partial\lambda}c^2$ and the expression in square brackets. Hence it is clear that $\frac{\triangle\beta}{\triangle n}$ will be positive if c^2 exceeds some particular value. Let this value of c^2 be equal to c_o^2. It is seen from the tables of power function of χ^2-test that the value of $c_o^2 = 1/5$. Hence as 'n' and 'λ' both vary the power increases as long as $\frac{\lambda}{n} > 1/5$. Since $\frac{\lambda}{n} = \frac{(\Sigma a i^2)}{n}$ the above result implies that the power of χ^2-test of H_o against H_1 increases when more observations having the same non-centrality c are included. This is true even when the expectation of all the x_i's under H_1 are not same, but are such that $\frac{\lambda}{n} = c^2$. This is in conformity with the general observation that where the χ^2-test is relevent, the sensivity of the test increases with the size of the sample.

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Theory of Demi-group Structures—I Axioms and Closure Operations for D-structures

BY

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Introduction

In this and subsequent papers we shall be considering commutative demi-groups with unit element over which further structures are imposed. Such structures based on a demi-group we shall call d-structures. The conditions that these satisfy for the main forms of extra structure on the demi-group give the suggestion for the axioms stated for d-structures in this I Part. Then, in a manner very similar to, and in some sense dual to, the development of a theory of operations for c-structures presented in an earlier paper entitled 'Closure operations on c-structures' (Proc. Dutch Acad. Sci., Ser A, 56, No. 4, pp. 317-329), the operational structure for the d-structures is discussed in this part. As a concrete example of such d-structure one may consider demi-groups with neutral elements with a partial order monotone under the operation of the demi-group, or with a uniform structure determined by nuclii around the neutral element with the demi-group operation uniformly continuous from the squared space to the original. In these cases the a-homomorphisms are order-preserving or index-lowering as the case may be; where the index of a point of a demi-group in a nucleus is defined as by S. Kaplan in his paper on 'Extension of Pontrajagin duality' I, (Duke math. jour., 15, ('48) 649-58).

These examples and questions of duality will be treated in greater detail in the II Part of this paper.

1. *Axioms for d-structures.*

The primitive concepts are: the d-structure, the a-structure, the d-homomorphism, and the a-homomorphism. Before listing the

axioms that these satisfy, let us define some preliminary notions that are also necessary.

Def. A set D closed for a binary associative operation (written here as $+$) is called a demigroup. An element O of D is a neutral element for the demigroup if $a + O = O + a = a$, for all a from D; such an element is surely unique when it exists. The demigroup is a semigroup, if the two cancellation laws hold: $a + x = b + x$ implies $a = b$, and $x + a = x + b$ implies $a = b$. If the demigroup has a neutral element and an inverse $-a$ for each element a (such that $a + (-a) = (-a) + a = O$) then we get a group.

Def. Given a class of demigroups D_i with neutral elements O_i, their direct sum $D (= S(D_i))$ is a demigroup whose elements are collections (a_i) of one element from each D_i, with $a_i = O_i$ for all but an atmost finite number of the i; the sum of two such elements (a_i) and (b_i) being computed componentwise, i.e., it is (c_i), where $c_i = a_i + b_i$. The (1-1) mapping of D_i in D defined by: $\delta_i(a_i) =$ the element (a_i) of D with $a_j = O_j$ except for $j = i$ when a_j is the element a_i taken from D_i, is called the ith injection in D. Clearly then any element α of D can be written as $\Sigma\, \delta_i\ (\alpha_i)$.

We may observe that D has $O = (O_i)$ as neutral element, and that when all the demigroups D_i are (i) commutative or (ii) are semi-groups, or (ii) are groups, then so is D.

We start with the axioms now.

B1. To each d-structure **A** is associated a unique a-structure A, called the a-base of **A**; each a-structure is the a-base of some d-structure. To each a-structure A is associated a unique commutative demigroup α, called the D-base of A. The elements of α are also called the elements of A, while if A is the a-base of **A**, and α the D-base of A, then α is also called the D-base of **A**, and its elements are called elements of **A** also.

B2. If **A**, **A**′ are d-structures and f is a d-homomorphism of **A** in **A**′, then f is also a a-homomorphism of the a-base of **A** in the a-base of A′. If A, A′ are a-structures and f is a a-homomorphism of A in A′ then it is a demigroup homomorphism of the D-base of A in the D-base of A′.

It is a consequence of B2 that the d-homomorphisms and a-homomorphisms are special instances of mappings of the elements of the one in the other.

Def. A (1-1) mapping f of the D-base of **A** in that of **A'** (of the D-base of A in that of A') is called a d-isomorphism (a a-isomorphism) if both f and f^{-1} are d-(a-) homomorphisms.

B3. Two d-(a-)structures are identified (considered indistinguishable) if they have the same D-base and the identity mapping of this base in itself is a d-(a-)isomorphism. The composition of two successive d-(a-)homomorphisms is also such a homomorphism.

B4. If f is a a-homomorphism of a a-structure A *on* each of two a-structures B', B'' having the same D-base, then B' and B'' are identical. If α, β are demigroups and f is an isomorphism of α on β, then if A is a a-structure having α as D-base, a a-structure B with β as D-base can be found such that f is a a-isomorphism of A on B. If f is a a-homomorphism of A in B and $f(\alpha)=\gamma=$ a subset of β, where α, β are the D-bases of A, B, then a a-structure C with γ as D-base can be found such that f is a a-homomorphism of A on C and the identity mapping of γ in β is a a-homomorphism of C in B.

Def. If E is a congruence relation on the commutative demigroup α (that is an equivalence relation such that aEb implies $(c+a)\mathrm{E}(c+b)$, for all a, b, c, from α), and β is the quotient demigroup α/E (whose elements are the equivalence classes $\mathrm{E}(a)$, a in α, with $\mathrm{E}(a)+\mathrm{E}(b)=\mathrm{E}(a+b)$) a a-structure B over β (i.e., with β as D-base) is called a 'quotient' of a a-structure A over α by E if the canonical mapping (a in α mapped on $\mathrm{E}(a)$ in β) of α in β is a a-homomorphism of A in B; by the first part of B4 it follows that if a quotient of A by E exists it must be unique. We denote it by A/E, when it exists.

Def. Two d-structures with the same a-base are said to be comparable; when **A**, **A'** are comparable d-structures (over the a-base A) we say **A** is finer than A' (in symbols $\mathbf{A}<\mathbf{A}'$) if the identity mapping of the D-base of A in itself is a d-homomorphism of **A** in **A'**.

B5. The d-structures comparable with a given d-structure **A** (over the a-base A) form a class **C** (A); given a d-homomorphism f of a d-structure **A** in a d-structure **B** from the class **C** (B), there exists a finest among the d-structures **B'** of **C** (B) for which f is a d-homomorphism of **A** in **B'**; we denote this d-structure by $f(\mathbf{A})$.

B6. If f, g are a-homomorphisms of A in B and of B in C, where A, B, C are a-structures, and if **A**, **C** are d-structures on

A, C such that gf is a d-homomorphism of **A** in **C**, then a d-structure **B** on B can be found such that f, g are d-homomorphisms of **A** in **B** and of **B** in **C** respectively.

Def. When α is a commutative demigroup, and $\beta = \alpha/E$ is a quotient demigroup, then a d-structure **B** over β is called a quotient d-structure of **A** over α by **E** if the canonical mapping f of α on β is a d-homomorphism of **A** on **B**, and further $\mathbf{B} = f(\mathbf{A})$.

Again the quotient structure if it exists is unique and is denoted by $\mathbf{A}/\mathbf{E}$; the a-base of $\mathbf{A}/\mathbf{E}$ is the quotient A/E where A is the a-base of **A**.

Def. For a class of d-structures $\mathbf{A}_i$ we define the d-structure **A** to be a direct sum of the class, if the D-base of **A** is the direct sum of the D-bases of the $\mathbf{A}_i$, if the injections δ_i are all d-homomorphisms of the $\mathbf{A}_i$ in **A**, and finally whenever there is a family of d-homomorphisms φ_i of the $\mathbf{A}_i$ in some d-structure **B**, then the mapping φ of **A** in **B**, defined by $\varphi(a) = \varphi((a_i)) = \Sigma\varphi_i(a_i)$, for any $a = (a_i)$ in A, is also a d-homomorphism.

From the definition follows the uniqueness of the direct sum of a class of d-structures, when it exists. We denote the sum of the $\mathbf{A}_i$ by $S(\mathbf{A}_i)$.

B7. For any class of d-structures there exists a direct sum.

Def. If the D-base α of a d-structure **A** is contained in the D-base β of a d-structure **B**, and the identity mapping g of α in β is not only a d-homomorphism of **A** in **B**, but even $\mathbf{B} = g(\mathbf{A})$, then **B** is called an extension of **A**. If there is a d-homomorphism g of a d-structure **A** in a d-structure **B**, and $\mathbf{B} = g(\mathbf{A})$, then **B** is called an envelope of **A**, and g is referred to as a map enveloping **A** in **B**.

B8. If, for a class of d-structures $\mathbf{A}_i$, $\mathbf{B}_i$ is an extension of $\mathbf{A}_i$ for each i, then $S(\mathbf{B}_i)$ is an extension of $S(\mathbf{A}_i)$.

2. *First consequences from the axioms.*

Lemma 1. (i) If the D-base β of a d-structure **B** is isomorphic with a demigroup α, then there is a d-structure **A** with D-base α such that the isomorphism between the D-bases is also d-isomorphism between **B** and **A**.

(ii) If f is a d-homomorphism of a d-structure A in a d-structure **B**, and **E** is the equivalence determined by f for the D-base

of α, then f can be extended to a d-homomorphism of the quotient $\mathbf{A}/\mathrm{E}$ in $\mathbf{B}$.

Proof. The axioms B4 and B6 give (i), while (ii) follows from B4, B5 and B6.

Lemma 2. (i) If f envelopes $\mathbf{A}$ in $\mathbf{B}$ and g envelopes $\mathbf{B}$ in $\mathbf{C}$, then gf envelopes $\mathbf{A}$ in $\mathbf{C}$; (ii) $\mathbf{B}$ is an envelope of $\mathbf{A}$, if, and only if, $\mathbf{B}$ is an extension of an isomorph of a quotient d-structure of $\mathbf{A}$.

Proof. (i) That gf is a d-homomorphism is clear; if $\mathbf{C}'$ is comparable with $\mathbf{C}$, and gf is d-homomorphism of $\mathbf{A}$ in $\mathbf{C}'$, then as f, g are a-homomorphisms of the a-base of $\mathbf{A}$ in $\mathbf{B}$ and of $\mathbf{B}$ in $\mathbf{C}'$, by B6 there is a d-structure $\mathbf{B}'$ comparable with $\mathbf{B}$ such that f, g are d-homomorphisms of $\mathbf{A}$ in $\mathbf{B}'$ and of $\mathbf{B}'$ in $\mathbf{C}'$. Hence as f, g are enveloping mappings, $\mathbf{B}$ is finer than $\mathbf{B}'$, g is a d-homomorphism of $\mathbf{B}$ in $\mathbf{C}'$ and $\mathbf{C}$ is finer than $\mathbf{C}'$. So gf envelopes $\mathbf{A}$ in $\mathbf{C}$.

(ii) An isomorphic mapping, an extension mapping and the canonical mapping on a quotient structure are all enveloping mappings, and so, by (i), an extension of an isomorph of a quotient d-structure is an envelope. While if f envelopes $\mathbf{A}$ in $\mathbf{B}$, and E is the equivalence relation on the D-base of $\mathbf{A}$ defined by f, then it follows from lemma 1, (ii), that f can be considered as a d-homomorphism of the quotient d-structure $\mathbf{A}/\mathrm{E}$ in $\mathbf{B}$. By B4 it can be seen that f is then a a-isomorphism between the a-base of $\mathbf{A}/\mathrm{E}$ and a a-structure C with D-base $\gamma = f(\alpha)$, α being the D-base of $\mathbf{A}$, and the identity mapping of γ in β (the D-base of $\mathbf{B}$) is an a-homomorphism of C in B. As in the proof of (i) above, using B6 we can then see that there is a d-structure $\mathbf{C}$ with C as a-base such that $\mathbf{A}/\mathrm{E}$ is isomorphic with it, and $\mathbf{B}$ is an extension of it.

Lemma 3. If $\mathbf{A}$ is the direct sum of a class of d-structures $\mathbf{A}_i$, and for each i, $\mathbf{A}_i{}^*$ is a quotient d-structure of $\mathbf{A}_i$ by E_i, then the direct sum $\mathbf{A}^*$of the $\mathbf{A}_i{}^*$ is d-isomorphic with the quotient d-structure of $\mathbf{A}$ by an equivalence $\mathrm{E} = \mathrm{SE}_i$ defined on it by: $(a_i)\ \mathrm{E}\ (a_i') <=>$ for all i, $a_i \mathrm{E}_i a_i'$:.

Proof. Let δ_i, $\delta_i{}^*$ denote the injective mappings into the sums $\mathbf{A}$, $\mathbf{A}^*$ respectively. Then if $a_i{}^*$ is the class under E_i containing a_i of $\mathbf{A}_i$, from the fact that the mappings of $a_i \to a_i{}^*$, and $a_i{}^* \to \delta_i{}^*(a_i{}^*)$ are d-homomorphisms it follows, from the definition of $\mathbf{A} = \mathrm{S}\mathbf{A}_i$, that the mapping $(a_i) \to (a_i{}^*)$ from $\mathbf{A}$ to $\mathbf{A}^*$ is a d-homomorphism. As this mapping determines on $\mathbf{A}$ just

the equivalence **E** by axiom B5 and the definition of the quotient structure it follows that this homomorphism splits up into the canonical one into **A**/**E** followed by a 1-1 d-homomorphism on **A***. That the inverse of this 1-1 mapping is also a d-homomorphism is deduced from the fact that the mappings $a_i^* \rightarrow$ the class under E containing $\delta_i(a_i)$, where a_i is any element of a_i^*, are d-homomorphisms.

Corollaries. The direct sum of $\mathbf{B}_i$ is d-isomorphic with (an envelope of) the direct sum of the $\mathbf{A}_i$, if, for each i, $\mathbf{B}_i$ is d-isomorphic with (an envelope of) $\mathbf{A}_i$.

The isomorphism part is evident; the other then follows from B8, lemma 2 (ii) and the above lemma.

3. *The Lattice of comparable d-structures.*

The relation $<$ of 'being finer than' is a partial ordering relation in the class **C** (A) of all comparable d-structures with the common a-base A. We examine this partially ordered set further now.

Theorem 1. The *p.o* set **C**(A) is a conditionally complete lattice; for a family $\mathbf{A}_i$ of its elements that are bounded above the lattice sum is d-isomorphic with the quotient of the direct sum of the $\mathbf{A}_i$ by an equivalence relation E.

Proof. Define on the D-base of the direct sum S($\mathbf{A}_i$) the equivalence relation E by: $(a_i)\mathrm{E}(a_i')$ if $\Sigma a_i = \Sigma a_i'$, these being finite sums in the commutative demigroup A. Then the correspondence f: taking the class under E containing (a_i) to Σa_i in A is a demigroup isomorphism between the quotient demigroup of the D-base of S($\mathbf{A}_i$) by E and the D-base α of A. If the quotient d-structure of S($\mathbf{A}_i$) by E exists then f gives a d-isomorph over A which is $>$ all the $\mathbf{A}_i$; while, conversely, if an upper bound $\mathbf{A}'$ exists for the α_i, f^{-1} gives over the base of S($\mathbf{A}_i$)/E a d-isomorph, which implies then that the quotient structure S($\mathbf{A}_i$)/E exists, and is $>$ than this isomorph. Thus it follows that the existence of an upper bound to the $\mathbf{A}_i$ is equivalent to the existence of the quotient structure, and when this exists, from the quotient structure f gives rise to a d-isomorphic structure on A which is the lattice sum, of the $\mathbf{A}_i$. Since every sub-family of **C**(A) bounded above has a lattice sum, every subfamily bounded below will also have a lattice product (namely the lattice sum of

the family of lower bounds). So the ordered set $\mathbf{C}(\mathbf{A})$ is a conditionally complete lattice.

Lemma 4. (i) If $\mathbf{A} = \mathrm{S}(\mathbf{A}_i)$, then $\mathbf{A} = \mathrm{V}(\delta_i(\mathbf{A}_i))$, V denoting the lattice sum. (ii) If f envelopes $\mathrm{V}(\mathbf{A}_i)$ in $\mathbf{B}$, then $\mathbf{B} = \mathrm{V}(f(\mathbf{A}_i))$.

Proof. (i) As δ_i is a d-homomorphism of $\mathbf{A}_i$ in $\mathbf{A}$, so $\delta_i(\mathbf{A}_i)$ is $< \mathbf{A}$, for each i. If $\delta_i(\mathbf{A}_i)$ is $< \mathbf{A}'$, for a d-structure comparable with $\mathbf{A}$, for each i, then from the definition of the direct sum it follows that the identity mapping of the D-Base of $\mathbf{A}$ on itself is a d-homomorphism of $\mathbf{A}$ in $\mathbf{A}'$. Thus $\mathbf{A} = (\delta_i\ (\mathbf{A}_i))$ is proved.

(ii) if α, β denote the D-bases of $\mathbf{A}$ and $\mathbf{B}$, and the identity mappings of α, β in themselves are denoted by α', β' then, $f = f.\alpha' = \beta'.f$. Hence as α' from $\mathbf{A}_i$ to $\mathbf{A}$, and f from $\mathbf{A}$ to $\mathbf{B}$ are d-homomorphisms, f is a d-homomorphism from $\mathbf{A}_i$ to $\mathbf{B}$, thus $f(\mathbf{A}_i) = \mathbf{B}_i$ exists and it is $< \mathbf{B}$ for each i. If each $\mathbf{B}_i$ is also $<$ some $\mathbf{B}^*$ comparable with $\mathbf{B}$, then f from $\mathbf{A}_i$ to $\mathbf{B}_i$ and β' from $\mathbf{B}_i$ to $\mathbf{B}^*$ being d-homomorphism f is a d-homomorphism of $\mathbf{A}_i$ in $\mathbf{B}'$ for each i; it follows that the mapping $(a_i) \to \Sigma\, f(a_i)$ is a d-homomorphism of $\mathrm{S}(\mathbf{A}_i)$ in $\mathbf{B}^*$. Since $\Sigma\, a_i = \Sigma\, a_i'$ implies that $\Sigma f(a_i) = \Sigma f(a_i')$, from theorem 1 we see that f is a d-homomorphism of $\mathrm{V}(\mathbf{B}_i)$ in $\mathbf{B}^*$; it follows that $\mathbf{B}$ is $< \mathbf{B}^*$, and the lemma is proved completely.

4. *Direct classes of d-structures and their limits.*

Def. A class $\mathbf{B}_d$ of d-structures indexed by a directed set $(\mathrm{D}, >)$ is called a directed class of d-structures if for each pair of indices d, d' with $d < d'$ there is a d-homomorphism $\lambda(d, d')$ of $\mathbf{B}_d$ in $\mathbf{B}_{d'}$, such that when $d < d' < d''$ then $\lambda(d', d'').\ \lambda(d, d') = \lambda(d, d'')$.

Def. If $\mathbf{B}_d$ is a directed class of d-structures, a d-structure $\mathbf{B}$ is called the limit of the directed class if it is the quotient d-structure of the direct sum $\mathrm{S}(\mathbf{B}_d)$ by the equivalence E defined by: $(b_d)\mathrm{E}(b_{d'})$ for (b_d), $(b_{d'})$ from the D-base of $\mathrm{S}(\mathbf{B}_d)$, if there is a $d'' >$ all the d's for which b_d or $b_{d'}$ is not equal to O_d, such that $\Sigma\lambda(d, d'')(b_d)$ and $\Sigma\lambda(d, d'')(b_{d'})$ (summed over the $d < d''$) are equal in the D-base of $\mathbf{B}_{d''}$.

Evidently E is an equivalence relation as D is directed by $>$. So the limit is unique if it exists.

Theorem 2. (i) If $\mathbf{A} = \mathrm{S}(\mathbf{A}_i)$, i in a set I, then $\mathbf{A}$ is isomorphic to the limit of the directed class $\mathbf{B}_d$, where d ranges over the family D of finite subsets of I directed by $\supset$, where $\mathbf{B}_d$ is taken as the direct sum $\mathrm{S}(\mathbf{A}_i)$, i in d, and $\lambda(d, d')$ for $d' \supset d$ is the mapping taking an element (a_i) of $\mathbf{B}_d$ to (a_j') of $\mathbf{B}_{d'}$, with $a_j' = a_i$ for $j' = i$ in d, and $a_j' = \mathrm{O}_j'$ otherwise.

(i) If $\mathbf{A} = \mathrm{V}(\mathbf{A}_i)$, i in I, then $\mathbf{A}$ is isomorphic to the limit of the directed class of d-structures $\mathbf{B}_d$, where d is taken from the family D of finite subsets of I directed by $\supset$, $\mathbf{B}_d$, is $\mathrm{V}(\mathbf{A}_i)$ for i in d, and $\lambda(d, d')$ for $d' \supset d$ is the identity mapping of the D-base of $\mathbf{A}$ in itself.

Proof. (i) It is easy to see that the $\mathbf{B}_d$ do form a directed class of d-structures. The mapping f_d taking (a_i) of $\mathbf{B}_d$ to the element of $\mathbf{A}$ having the same nonzero terms at the same places, is a d-homomorphism. Hence the mapping f taking (b_d) of $\mathrm{S}(\mathbf{B}_d)$ to $\Sigma f_d(b_d)$ in $\mathbf{A}$ is a d-homomorphism and this defines on $\mathrm{S}(\mathbf{B}_d)$ the equivalence E under which $(b_d)\mathrm{E}(b_{d'})$ if and only if $\Sigma \lambda(d, d'')(b_d) = \Sigma \lambda(d, d'')(b_{d'})$, for d'' equal to or containing the union of the d's for which the b_d, $b_{d'}$ are nonzero. Thus, from the definition of the limit and lemma 1 (ii), it follows that the limit of the directed class $\mathbf{B}_d$ exists and f is a 1-1 d-homomorphism of the limit on $\mathbf{A}$. That the reverse of this 1-1 mapping is a d-homomorphism follows from the fact that each of the mappings $\varphi_i: a_i \rightarrow$ the class under $\mathbf{B}$ containing (b_d), where b_d is O_d except for $d =$ the one element set (i) for which it is a_i itself, is a d-homomorphism of $\mathbf{A}_i$ in the quotient structure of $\mathrm{S}(\mathbf{B}_d)$ by E. Then from the definition of $\mathrm{S}(\mathbf{A}_i)$ it follows that the inverse of f from $\mathbf{A}$ to the limit structure is a d-homomorphism. Thus $\mathbf{A}$ is isomorphic with the limit, which exists, of the directed class of the $\mathbf{B}_d$.

(ii) That the $\mathbf{B}_d$ form a directed class is easily verified. For each d $\mathbf{B}_d$ is comparable with $\mathbf{A}$ and $< \mathbf{A}$; hence the identity mapping I of the D-base α of A in itself is a d-homomorphism of $\mathbf{B}_d$ in$\mathbf{A}$. It follows that the mapping f taking $(\mathbf{B}_d)$ of $\mathrm{S}(\mathbf{B}_d)$ to the element Σb_d of $\mathbf{A}$ is a d-homomorphism, defining the equivalence E under which $(b_d)\mathrm{E}(b_{d'})$ means that $\Sigma b_d = \Sigma b_{d'}$ in α, that is to say that $\Sigma \lambda(d, d'')(b_d) = \Sigma \lambda(d, d'')(b_{d'})$ when d'' contains the union of the d's for which b_d, $b_{d'}$ in the two elements given are nonzero. Thus as before the limit of the $\mathbf{B}_d$ exists and there is a 1-1 d-homomorphism of the limit on $\mathbf{A}$. To prove that the inverse of this map is also a d-homomorphism it is enough to

note that the mappings $\varphi_i: a_i \to$ the class under E containing the (b_d) with zero terms except for $d =$ (i) for which $b_d = a_i$, are d-homomorphisms from the $\mathbf{A}_i$ to the limit.

5. *Operations on types of d-structures.*

By adding extra restrictions we get types of d-structures. Of two types t, t' we say t is stronger, and write $t' => t$, if every d-structure of type t' is also of type t. If $t => t'$ and $t' => t$ (or as we also write it $t <= t'$) then the two types t, t' are called equivalent, and we write $t <=> t'$.

From a binary relation R defined for ordered pairs of d-structures, or an operation θ defined for certain classes of d-structures, we can derive operations, also denoted by R or θ, on types of d-structures: $\mathrm{R}(t)$ is the type of d-structure $\mathbf{A}$ for which a d-structure $\mathbf{B}$ of type t can be found such that $\mathbf{ARB}$ is true; $\theta(t)$ is the type of d-structure $\mathbf{A}$ for which a class of d-structures $\mathbf{A}_i$ of type t can be found such that $\mathbf{A} = \theta(\mathbf{A}_i)$. Evidently if $t => t'$, then $\mathrm{R}(t) => \mathrm{R}(t')$ and $\theta(t) => \theta(t')$, for these operations on types. When φ is an operation on types (i.e., associates to each type t another $t' = \varphi(t)$) which satisfies such a monotonicity condition w.r.t. $=>$, then a type t is said to be closed for φ if $\varphi(t) => t$. If t' is a type closed for φ stronger than a given type t and feebler than any type t'' closed for φ stronger than t, then t' is called a φ-closure of t; it is unique upto equivalence of types if it exists. The operation on types φ is a 'closure operation on types' if, for any type of d-structure t, $\varphi(t)$ is a φ-closure of t; if φ is monotone w.r.t. $=>$, a sufficient condition that φ be a closure operation on types is that, for any type t, one has $t => \varphi(t)$ and $\varphi\varphi(t) => \varphi(t)$. Of two operations on types θ, φ we say θ is stronger if, for any type t, $\theta(t)$ is stronger than $\varphi(t)$.

We consider now the interrelations between the various operations on types defined by the following relations and operations on d-structures:

1. $\mathbf{AiB}$ is the relation $\mathbf{A}$ is d-isomorphic with $\mathbf{B}$.
2. $\mathbf{AqB}$ is the relation that $\mathbf{A}$ is the quotient d-structure of $\mathbf{B}$ by an equivalence E.
3. $\mathbf{AxB}$ is the relation that $\mathbf{A}$ is an extension of $\mathbf{B}$ by a mapping f.

4. $\mathbf{A}e\mathbf{B}$ is the relation that $\mathbf{A}$ is the envelope of $\mathbf{B}$ by a mapping f.

5. $\mathbf{A} = \mathrm{S}(\mathbf{A}_i)$, $[\mathbf{A} = s(\mathbf{A}_i)]$ stand for the operations of taking the direct sum of an arbitrary [of a finite] class of d-structures.

6. $\mathbf{A} = \mathrm{V}(\mathbf{A}_i)$, $[\mathbf{A} = v(\mathbf{A}_i)]$ stand for the operations of taking the lattice sum of an arbitrary [of a finite] class of comparable d-structures.

7. $\mathbf{A} = \mathrm{L}(\mathbf{A}_d)$, d in $(\mathrm{D},>)$, stand for taking the limit of a directed class of d-structures $\mathbf{A}_d$.

The results of the earlier section then lead to the following results :

Lemma 5. For any given type of d-structures t, we have

(i) $t => \varphi(t)$, if φ is any one of the operations, i, iq, x, e, S, s, V, v, iL; $i.\ i(t) <=> i(t)$, $i.x(t) <=> x.i(t)$, $i.q(t) <=> q.i(t)$, $e.e(t) <=> e(t)$; $e(t) <=> x.i.q(t)$, $\mathrm{S\,S}(t) => i.\mathrm{S}(\mathrm{T})$, $s.s.(t) => i.s(t)$, $\mathrm{VV}(t) <=> \mathrm{V}(t)$, $v.v.(t) <=> v(t)$, $\mathrm{S}(t) => \mathrm{V}.x(t)$, $s(t) => v.x(t)$. $\mathrm{V}(t) => i.q.\mathrm{S}(\mathrm{T})$, $v(t) => i.q.s(t)$;

(ii) if φ denotes any one of the operations, i, iq, x or e, then $\mathrm{S}.\varphi(t) => \varphi.\mathrm{S}(t)$, $s.\varphi(t) => \varphi.s(t)$, $\varphi.\mathrm{V}(t) => \mathrm{V}\varphi(t)$, $\varphi.v(t) => v.\varphi(t)$;

(iii) $\mathrm{S}(t) => i.\mathrm{L}.p(t)$, $V(t) => i.\mathrm{L}v(t)$, $\mathrm{L}(t) => q.\mathrm{S}(t)$, $L(t) => \mathrm{V}.e(t)$.

Proof. (i) The first two sets of assertions are easily deducible from the definitions involved. Lemma 2 (ii) gives $e(t) <=> x.i.q(t)$; Theorem 1 and lemma 4 (i) take care of the last four assertions, while the others are merely forms of associativity of the direct sum or lattice sum, allowance being given to distinguish isomorphic structures.

(ii) Lemma 3 and its corollary prove the first results, while lemma 4 (ii) gives the results $e.\mathrm{V}(t) => \mathrm{V}.e(t)$ and $e.v(t) => v.e(t)$, while similar proofs can be given to prove the corresponding results $\varphi.\mathrm{V}(t) => \mathrm{V}.\varphi(t)$, $\varphi.v(t) => v.\varphi(t)$, for $\varphi = i, i.q.$ or x.

(iii) Theorem 2 (i) and (ii) give the first two results; the definition of the limit implies that $\mathrm{L}(t) => q.\mathrm{S}(t)$, while this with $\mathrm{S}(t) => \mathrm{V}.x(t)$ leads to $\mathrm{L}(t) => q.\mathrm{V}.x(t) => i.\mathrm{V}.q.x(t) => \mathrm{V}.i.q.x(t) => \mathrm{V}.e(t)$, as each of $i(t)$, $q(t)$ and $x(t)$ implies $e(t)$.

From the above results we deduce the main structure theorem:

Theorem 3. (i) The following are closure operations on types of d-structures: φ, φ.S, $\varphi.s$, V.φ and $v.\varphi$, where φ stands for any one of the operations i, $i.q$, x or e.

(ii) There is a closure operation on types stronger than all combinations of the basic operations i, q, x, e, S, s, V, v and L, and this is equivalent to each of the following: e.S, V.e, e.Ls, i.L.$v.e$ and $i.q$.S.e.

Proof. (i) The lemma 5 proves that the operations mentioned are all closure operations on types.

(ii) $e.\mathrm{S} \Rightarrow e.\mathrm{V}.x \Rightarrow \mathrm{V}.e.x \Rightarrow \mathrm{V}.e \Rightarrow i.q.\mathrm{S}.e \Rightarrow q.e.\mathrm{S} \Rightarrow e.\mathrm{S}$

$e.\mathrm{S} \Rightarrow e.i.\mathrm{L}.s \Rightarrow e.\mathrm{L}.s \Rightarrow e.q.\mathrm{S}.s \Rightarrow e.\mathrm{S}$

$\mathrm{V}.e \Rightarrow i.\mathrm{L}.v.e \Rightarrow i.\mathrm{V}.e.v.e \Rightarrow i.\mathrm{V}.v.e \Rightarrow i.\mathrm{V}.e \Rightarrow \mathrm{V}.i.e \Rightarrow \mathrm{V}.e$

$\mathrm{V}.e \Rightarrow i.q.\mathrm{S}.e \Rightarrow i.q.e.\mathrm{S} \Rightarrow e.\mathrm{S}$

So the five operations given are all equivalent, and as one of them e.S is a closure operation so are all the others. And any one of the single basic operations and hence also all combinations of them imply one of these five and so all of them.

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Solvent Extraction of South Arcot Lignite

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ABSTRACT

Solvent Extraction formed part of the series of investigations taken up on the study of the Utilisation of the South Arcot Lignite. A number of solvents were used in order to establish the best solvent that gives an appreciable quantity of good quality extractive. With one of the solvents, viz., Xylol, a study was made of the effect of temperature and pressure on the yield of extractive, and correlations have been obtained between the yield on the one hand and the time and temperature of extraction on the other. The properties of the various extractives obtained were determined and a sample of refined montan wax from the Xylol extract, was prepared.

Introduction :

Solvent Extraction of coals, lignites and peats has been an attractive line of investigation pursued with the object of either studying the coal constitution or isolating technically important extractives. In this laboratory a sample of the South Arcot Lignite was subjected to solvent extraction with a number of solvents with a view to preparing some commercially important resins and waxes and the results are discussed in this paper.

The extraction of coals with organic solvents involves not only a physical solution of the material but also a chemical decomposition or depolymerisation. Reference has also been made to the possibility of chemical combination of the solvent and the coal substance (Lowry, Vol. I). Hence at each temperature the amount of the extractive obtained is governed by the solubilising power of the solvent at that temperature and the degree of decomposition and the consequent losses due to gasification of the extracted product.

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The extractives obtained are composed of the original coal substance extracted by true solution as well as those resulting from the thermal depolymerisation of the coal substance. Besides, in the case of certain solvents like pyridine a process of peptisation or colloidal dispersion of the coal substance is brought about. Quite a variety of organic substances ranging from more or less familiar low molecular weight compounds to highly complex molecules and colloids are therefore met with in the extractives. The extractive obtained from peats, lignites and semi-bituminous coals are commonly termed Montan Wax, although strictly speaking the term must refer to the refined wax obtained from the Crude extractive. The montan wax as obtained from lignite is essentially a mixture of waxy constituents consisting of monohydric alcoholic esters and high molecular weight acids, along with some resinous and asphaltic materials (Warth). The composition of the crude montan wax is given as: ester of waxy acids 58-59%, free wax acids 17-19%, free monohydric alcohols 3-4% (primary), 1% (secondary), resins 10-12%, and ketones less than 10%.

Montan wax has been of great industrial importance in its applications for electrical insulation, for wax coatings on leather, for polishes, for brewer's pitch, for phonograph records, etc. Crude montan wax may be advantageously used to lower the pour points of oils.

The suitability of the South Arcot Lignite for the production of montan wax and resins and the properties of the waxes obtained under various conditions were thoroughly investigated. Soxhlet extraction of the lignite was conducted with a number of solvents and the best solvent which gives a good quantity of the wax having the desired properties was established. The effect of various conditions like moisture content and particle size of the lignite on the yield on extraction for a particular solvent was also studied. This was followed by a series extractions conducted under pressure. Attempts were made to refine the crude wax.

Experimental: Extraction of South Arcot Lignite at atmospheric pressure.

An air-dry sample of −30, +60 mesh size was used in the extractions. 10 gms. of the sample were taken in a Whatman extraction thimble and extractions were conducted in soxhlet apparatus at atmospheric pressure. A variety of solvents like Benzol, Xylol,

Tetralin, etc., were tried, in addition to a mixed solvent, ethanol-benzene. In each case the extraction was allowed to proceed until the siphoned solution was colourless, thereby ensuring complete extraction of all the soluble components. The yield of extractives was determined by distilling off the solvent under reduced pressure. The last traces of the solvent were removed by placing it in an air oven at a temperature of 105-110°C for several hours until constant weight was recorded. The melting point of each extractive was determined by the open capillary tube method. Table I gives the results.

TABLE I

YIELDS ON EXTRACTION WITH SEVERAL SOLVENTS

No.	Solvent.	No. of siphonings.	% Yield, (on air-dry lignite).	m.pt. of extractive.	Appearance and Colour.
1.	Benzol	16	9·1	78°C	Dark brown, brittle waxy.
2.	Xylol	18	11·8	80	Dark brown, resinlike structure.
3.	Tetralin	16	11·6	95	-do-
4.	Tetrachloroethylene	26	7·7	84	Light brown, waxy.
5.	Chlorobenzene	30	8·1	96	-do-
6.	Pyridine	25	14·8	140	Black, hard and brittle.
7.	Ethanol-benzene 1:3	20	13·7	117	Dark brown resinlike.

Results:

The largest quantity of extractive is given by Pyridine and it is due to its peptising power. This is closely followed by the benzene-alcohol mixture. But these extractives have high melting points indicating a high C/H ratio in the waxes present and as such are somewhat limited in their uses. Tetralin and Xylol give the next largest amount of extractive and the latter gives a product having melting point quite close to that of standard commercial samples. Hence it was concluded that among the solvents tried, Xylol was the best in respect of yield consistent with quality.

The effects of particle size and moisture content of the lignite were studied for some of the solvents. The results show that, as a rule, the yield increases with the fineness of the particle size. But in practice, there was some difficulty due to choking in the case of very fine lignite. This was remedied to some extent by admixing the lignite with clean washed sand.

TABLE II

EFFECT OF THE CONDITION OF THE LIGNITE ON THE YIELD OF EXTRACTIVES

No.	Solvent.	Condition of lignite.		No. of Extractions.	% Yield Extractive on air-dry lignite).
		Size.	Moisture content.		
1.	Rectified Spirit and Benzol, 1:3	-30, 60	Air-dry	20	13·7
2.	Absolute alcohol and Benzol, 1:3	-30, 60	Oven-dry	23	12·5
3.	-do-	-30, 60	Oven-dry 10·4% moisture added	25	13·6
4.	-do-	-30, 60	Oven-dry 15% moisture added	28	13·5
5.	Rectified Spirit and Benzol, 1:3	-4, 6	Air-dry	25	11·5
6.	-do-	-6, 10	-do-	26	12·9
7.	-do-	-80, 120	-do-	30	15·5
8.	Pyridine	-4, 6	-do-	20	12·5
9.	-do-	-6, 10	-do-	20	13·8
10.	-do-	-30, 60	-do-	24	14·5

The moisture content seems to play an important role in controlling the process of extraction. Oven-dry lignite extracted with a mixture of absolute alcohol and benzol yielded a definitely lesser amount of the crude extractive than the air-dry lignite (moisture 10·4 %). Remarkably enough, the yield improved and reached the original value when the oven-dry lignite was moistened with the requisite amount of water to bring the moisture content to that of the air-dry sample. However, no further improvement

was obtained when more water was added to the lignite. These experiments indicate that a small amount of moisture facilitates extraction, although there is no water soluble constituent in the extractives obtained. The exact mechanism of the phenomenon is not definitely known. The results are given in Table II.

Extraction of the lignite under pressure :

Experiments were conducted to determine the yield on equilibrium extraction of South Arcot lignite under various pressures at high temperatures and to study the effect of the variables on the yield. Commercial xylol was used as the solvent. The correlations obtained are sure to be of interest in designing large scale extraction plants.

The pressure extraction was conducted in a one-litre Chas. W. Cook's Tilting type stainless steel autoclave with a rated pressure of 350 atm. and a rated temperature of 350°C. Heating was done by a wire-wound electric furnace surrounding the autoclave. The temperature was controlled to ± 2°C for any length of time by means of a Wheelco Rheotrol Input Controller.

The autoclave was charged with a weighed amount of the air-dry lignite (50 gms) and 500 cc of Xylol. The mixture was kept agitated by rocking the autoclave and then it was heated. After attaining the desired temperature the mixture was maintained at the same temperature for a measured interval of time. Then the mixture was cooled, filtered and washed once with fresh solvent. The extractives were separated by distilling off the solvent under vacuum. The weights of the solvent-free residue and the extractive were determined. Extractions were done at various temperatures for a period of 2 hours. Also a series of runs were taken all at 200°C for periods ranging from $\frac{1}{2}$ hour to 5 hours. The results are tabulated in Table III.

Results:

The weight of the dried extractive was reported as % on the moisture-free lignite. The yields were also calculated following the method of Parr and Hadley by subtracting the weight of the dried residue and the weight of the moisture from the original weight of the lignite. The difference between the computed and experimental values for the yields, though small, increases with temperature because of the possible decomposition at higher

temperatures. This was evident in the runs conducted above 270°C, in which on opening the cooled autoclave, crude gas was found to escape from the autoclave indicating decomposition at higher temperatures.

TABLE III

PRESSURE EXTRACTION OF SOUTH ARCOT LIGNITE*

No.	Extraction Temp. °C.	Time of Extraction hrs.	Wt. of Residue gms.	Moisture-free basis			
				Extract calculated		Extract obtained	
				gms.	%	gms.	%
1.	30	2	41·42	3·68	8·25	3·54	7·8
2.	140	2	39·98	5·12	11·4	5·00	11·1
3.	170	2	38·72	6·38	14·2	6·32	14·0
4.	200	2	38·25	6·85	15·2	6·72	14·9
5.	240	2	36·24	8·86	19·8	8·51	18·9
6.	270	2	34·94	10·16	22·5	9·86	21·8
7.	300	2	34·97	10·33	22·9	9·91	22·0
8.	325	2	33·96	11·14	24·7	10·53	23·4
9.	200	½	40·94	4·16	9·2	4·05	8·9
10.	200	1	39·19	5·91	13·1	5·73	12·6
11.	200	3	38·10	7·00	15·5	6·82	15·1
12.	200	4	37·80	7·30	16·2	7·09	15·5

*Note: 50 gms. of air-dry (9·8% moisture) lignite was taken for each run.

The influence of temperature on yield is represented graphically (Fig. 2). The yield increases with temperature, but not proportionally. Since greater and greater decomposition was observed at higher temperatures, there is reason to believe that the yield may reach a maximum around 400 or 450° C. But owing to the limitation on the maximum permissible temperature of the autoclave used, a direct determination at a higher temperature was not possible.

The yield increases with time, but the graph (Fig. I) indicates that most of the extractives go into solution in the first hour after which there is very little increase.

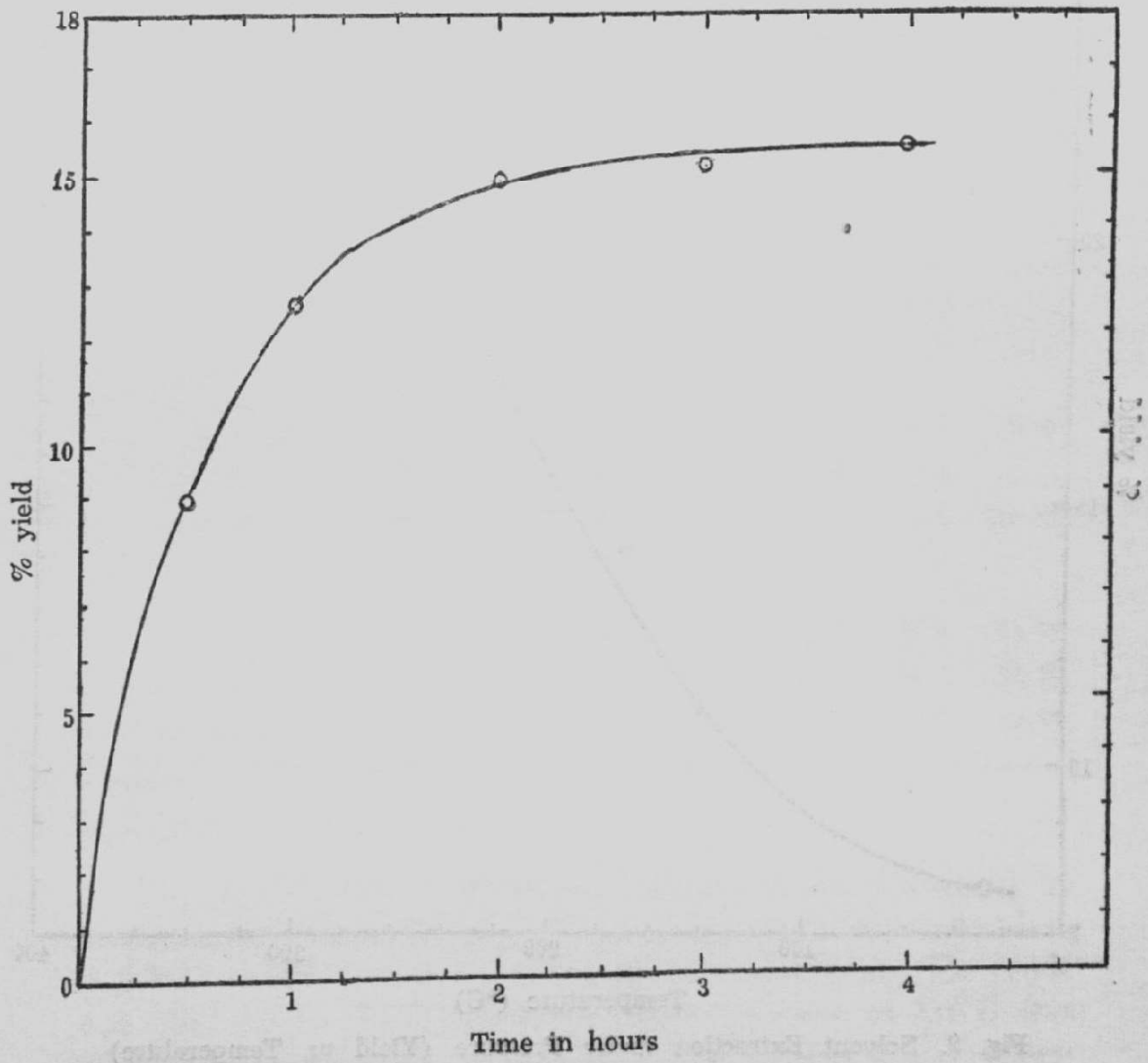

Fig. 1 Solvent Extration under Pressure (Yield *vs.* Time)

Properties of the Extractives:

The extractives as obtained or the crude montan wax as it is often called, is a dark brown solid with a varying degree of hardness, having a conchoidal fracture. It is a mixture of montan ester-waxes, resins and bitumens. The quantity of wax in the extractives was determined by dissolving away the resins in ether at room temperature in which the wax is insoluble. In order to compare the quality of the waxes obtained, their properties like the Acid Value, Saponification Value, Ester Value, etc. were determined. The procedure followed was according to Warth, p. 322. In the estimation of the saponification value the titration

of the dark coloured solution was done by conductometry. The values are given in Table IV.

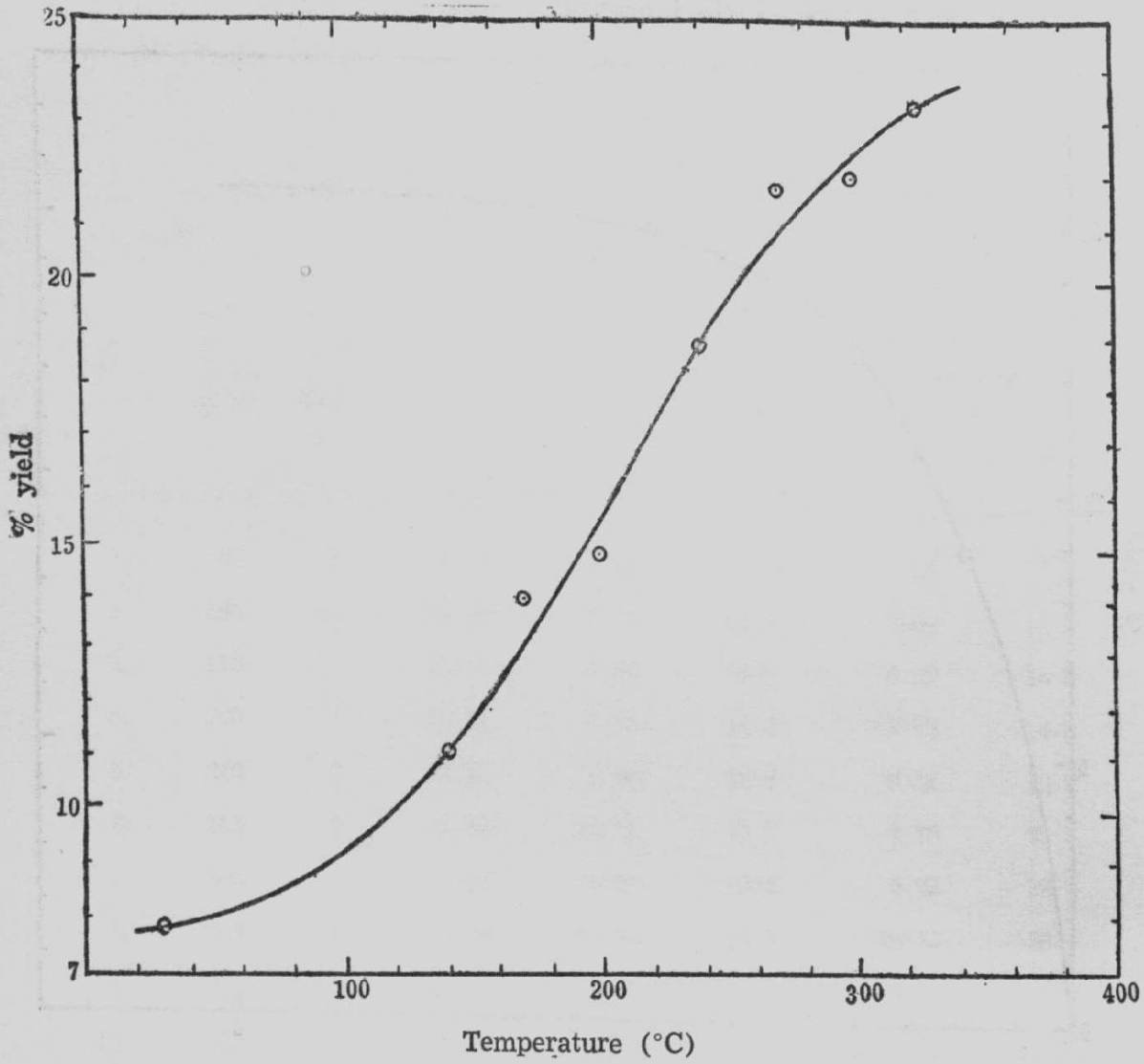

Fig. 2. Solvent Extraction under Pressure (Yield *vs.* Temperature)

A study of the results gives rise to the following conclusions. Larger yields are obtainable on extraction with Pyridine and with mixed solvents like benzene-alcohol. But in these cases the crude waxes have an appreciable resin content with a high melting point. The extractives with Tetrachloro-ethylene and chlorobenzene are comparatively rich in their wax content although the total yield is poor. The Xylol extraction conducted at various temperatures has indicated that although the total quantity of extractives obtained increases with temperature, much of the increase is due to the formation of larger quantities of resins resulting from the thermal depolymerisation rather than to the production of any wax.

TABLE IV

PROPERTIES OF THE EXTRACTIVES FROM SOUTH ARCOT LIGNITE

No.	Solvent Used.	Sap. Value.	Acid Value.	Ester Value.	Iodine Value.	Resin % Ether sol.	Wax % Petr. Ether sol.	M. P. of Wax °C
1.	Benzol	93·6	27·6	66·1	12·9	54·3	74·1	78–80
2.	Xylol	89·4	28·1	61·1	11·4	56·8	66·2	85–86
3.	Tetralin	98·5	32·1	66·4	14·8	45·8	49·5	93–95
4.	Tetrachloro-ethylene	99·0	31·4	67·6	10·7	36·1	69·3	85–89
5.	Chlorobenzene	102·0	24·3	67·7	12·5	31·8	68·4	79·82
6.	Pyridine	184·0	44·8	139·2	19·2	15·8	64·5	120–130
7.	Ethanol-Benzene, 1:3	126·0	36·1	89·9	12·1	42·8	59·3	74·80
8.	Xylol (200°C)	82·3	25·4	56·9	14·7	60·7	56·4	68·70
9.	Xylol (325°C)	78·8	24·5	54·3	15·1	63·4	47·5	66·68

Purification of Montan Wax:

The montan wax as separated from the crude extractives by a subsequent ether extraction of the resins is light brown to black in colour and for certain uses it needs further refining. The crude wax obtained by Xylol extraction under pressure at 325°C was purified by the following procedure. About 10 gms of the extractive was extracted repeatedly with petroleum ether at room temperature until the solution was practically colourless. The residue (resin) left behind was estimated. The Petroleum ether solutions were added together and treated with concentrated sulphuric acid in a separating funnel, and the acid layer run off. The treatment was continued until the wax solution was colourless, after which it was collected in a weighed china dish, the solvent evaporated, and the purified wax weighed. About 30 cc of the acid is used for the purification.

It was possible to prepare a refined montan wax, pale white in colour, amounting to 47·5% of the deresinified wax or 21·1% of the crude extractive. This works out to nearly 4·5% of the

weight of lignite. The properties of the wax are comparable to those varieties available commercially (Warth) although there is still scope for improving the colour. (Table V).

TABLE V

PROPERTIES OF MONTAN WAX OBTAINED FROM SEVERAL SOURCES

No.	Source	Sap. value.	Acid value.	Ester value.	Ether soluble.	M.P. °C.	Remarks.	Ref.
1.	South Arcot Lignite	84	36	48	20	73-75	Brownish white.	Authors
2.	German Lignite	76	33	43	14	74-82	Dark brown.	Cawley & King.
3.	Devon Lignite	75	30	45	40	73-83	-do-	-do-
4.	Standard Montan Wax	105	100	5	15	75	——	Berle & Lunge.
5.	Double Bleached Montan Wax	105	100	5	15	76	——	-do-
6.	Eisenreich	95	83	2	—	77	Yellowish brown.	Warth.

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[*Note*: The authors are grateful to the University of Madras, for giving permission for publishing this paper the subject-matter of which formed part of a thesis on the basis of which the M.Sc. degree in Technology was awarded to one of us (S.S.)].

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On a Function Connected with $\varphi(n)$

BY

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1. Let $\varphi(n)$ denote Euler's function representing the number of numbers less than and prime to n. Let $\varphi_1(n) = \varphi(n)$; $\varphi_r(n) = \varphi(\varphi_{r-1}(n))$, $r = 2, 3, \ldots$. For a given n, $\varphi_r(n)$ decreases as r increases, and hence there is a least value of r, say r_1, such that

$$\varphi_r(n) = \varphi_{r+1}(n) = \ldots = 1.$$

S. S. Pillai proved that

$$\left[\frac{\log\ (n/2)}{\log\ 3}\right] + 1 \leqslant \mathrm{R}(n) \leqslant \frac{\log\ n}{\log\ 2} + 1.$$

where $r_1 = \mathrm{R}(n)$. In this paper I consider an analogous function $\mathrm{S}(n)$ given by

$$\mathrm{S}(n) = \varphi_1(n) + \varphi_2(n) + \ldots + \varphi_{r1}(n).$$

It is shown here that

$$\mathrm{R}(n) \leqslant \frac{\log\ n}{\log\ 2} \text{ if } n \text{ is even}$$

and

$$\leqslant \frac{\log\ (n-1)}{\log\ 2} + 1 \text{ if } n \text{ is odd,}$$

which is an improvement over Pillai's result;

Also

$$\mathrm{S}(n) \leqslant n - 1 \text{ if } n \text{ is even}$$

$$\leqslant 2n - 3 \quad \text{,,} \quad \text{odd,}$$

while $\mathrm{S}(n) \geqslant 2^{[\log\ (n/2)/\log 3]+1} - 1$.

I also consider the solutions of the equation $\mathrm{S}(n) = n$, and show that each one of the following values of n provides a solution:

$$n = n_1,\ 3n_2,\ 3n_3,\ \ldots.$$

where $n_1 = 3^k,\ k = 0, 1, 2, \ldots.$

$n_2 = 1 + 4n_1$

$n_3 = 1 + 12n_2$

$n_4 = 1 + 12n_3, \ldots.$

provided these are primes. The question whether these are the only primes remains open.

2. *Theorem* 1. If n is even,

$$R(n) \leqslant \frac{\log n}{\log 2};\ S(n) \leqslant n - 1.$$

If n is odd, $R(n) \leqslant \dfrac{\log (n-1)}{\log 2} + 1$

$$S(n) \leqslant 2n - 3.$$

Proof: When n is even, $\varphi_1(n) \leqslant (n/2)$;

$$\varphi_2(n) \leqslant \frac{1}{2}\varphi_1(n) \leqslant \frac{1}{4}n; \ldots.$$

$$\varphi_r(n) \leqslant \frac{1}{2^r}n.$$

If $r = r_1 = R(n)$ we get

$$1 \leqslant \frac{1}{2^{r_1}}n,$$

$$r_1 \leqslant \frac{\log n}{\log 2} \qquad \ldots \quad (1)$$

If n is a power of 2, r_1 is actually $\left(\dfrac{\log n}{\log 2}\right)$

Again $S(n) = \varphi_1(n) + \varphi_2(n) + \ldots + \varphi_{r1}(n)$

$$\leqslant \frac{1}{2}n + \frac{1}{4}n + \ldots + \frac{1}{2^{r_1}}n;$$

$$= n\left(1 - \frac{1}{2^{r_1}}\right)$$

$$\leqslant n\left(1 - \frac{1}{n}\right), \text{ by (1)}$$

Hence $S(n) \leqslant n - 1.$

If n is a power of 2, it is easily seen that $S(n) = n - 1$.

Let n be odd. Then $\varphi_2(n) \leqslant \frac{1}{2}\varphi_1(n)$;

$$\varphi_3(n) \leqslant \frac{1}{2}\varphi_2(n) \leqslant \frac{1}{2^2}\varphi_1(n); \ldots.$$

$$\varphi_r(n) \leqslant \frac{1}{2^{r-1}}\varphi_1(n)$$

Hence $$1 \leqslant \frac{1}{2^{r_1-1}}\varphi(n),$$

$$r_1 \leqslant \frac{\log \varphi(n)}{\log 2} + 1 \leqslant \frac{\log (n-1)}{\log 2} + 1 \quad .. \quad (2)$$

If n is a prime of the form $2^k + 1$, we get the equality sign. Again

$$\begin{aligned} S(n) &\leqslant \varphi(n)\,(1 + \tfrac{1}{2} + \ldots + (1/2^{r_1-1})) \\ &= 2\varphi(n)\,(1 - (1/2^{r_1})) \\ &\leqslant 2(n-1)\,(1 - (1/2^{r_1})) \\ &\leqslant 2(n-1)\,(1 - (1/2n-1)), \text{ by } (2) \\ &= 2n - 3. \end{aligned}$$

Actually $S(n)$ attains this value if n is a prime of the form $2^k + 1$.

3. Before considering the lower bounds of $R(n)$ and $S(n)$ we prove the following:

Theorem 2.* For a given value of $R(n)$, the maximum value of n is $2 \cdot 3^{R(n)-1}$.

The proof of this depends on the

Lemma. $R(pn) > R(n) + (\log n/\log p)$, if p is a prime > 3. If $p = 3$ it becomes an equality.

Proof by induction: Suppose the result is true for all primes $q < p$ so that

$$R(qn) \geqslant R(n) + \frac{\log q}{\log 3} \quad .. \quad (3)$$

* This is equivalent to Pillai's theorem II; but the proof by induction given here is different from Pillai's.

for all $n > 0$ and all $q < p$. We will prove the same for p. Let $p - 1 = 2^a q_1^{b_1} q_2^{b_2} \ldots .$

Then $R(n) = 1 + R[\varphi(n)]$ .. (4)

assuming as we may that $n > 2$; for if $n = 2$, the result of the lemma is obvious.

Also $R(pn) = 1 + R(\varphi(pn))$.

Assume that p is prime to n as a first case. Then

$$\begin{aligned} R(pn) &= 1 + R[(p-1)\varphi(n)] \\ &= 1 + R[2^a q_1^{b_1} q_2^{b_2} \ldots . \varphi(n)] \\ &= 1 + R[q_1^{b_1} q_2^{b_2} \ldots . \varphi(n)] + a, \end{aligned}$$

since $\varphi(n) \geqslant 2$ and $R(2m) = R(m) + 1$ if m is even;

Hence $R(pn) \geqslant 1 + R[\varphi(n)] + a + \Sigma \left(\frac{\log q_i}{\log 3}\right) b_i,$

by using (3) repeatedly,

$$= R(n) + a + \left\{\frac{\log (p-1)/2^a}{\log 3}\right\}$$

$$= R(n) + \frac{\log (3/2)^a (p-1)}{\log 3}$$

$$\geqslant R(n) + \left(\frac{\log p}{\log 3}\right)$$

if $(3/2)^a (p-1) \geqslant p$ which is true since $a \geqslant 1$, and $p \geqslant 3$. Since the result is true for $p = 3$, the lemma follows in the case when p is prime to n. Next let $(p, n) \neq 1$. Then in each of the pairs $(\varphi_1(pn); \varphi_1(n)); (\varphi_2(pn); \varphi_2(n)); \ldots .$ p occurs upto a stage, say in the first k pairs, and in the pair $(\varphi_{k+1}(pn); \varphi_{k+1}(n))$ the first member contains p while the second does not.

Let us call $\varphi_{k+1}(np) = pu$; $\varphi_{k+1}(n) = u$; then $(p, u) = 1$. Hence by the above result,

$$R(pu) > R(u) + \left(\frac{\log p}{\log 3}\right) \quad .. \quad (4)$$

But

$$R(pn) = k+1+R(pu)$$
$$R(n) = k+1+R(u)$$

Hence adding $k+1$ to both sides of (A) we get

$$R(pn) > R(n) + \left(\frac{\log p}{\log 3}\right), \quad p > 3,$$

so that the lemma follows in this case also. To prove Theorem 2, let $R(n)=t$ (fixed), and consider all possible values of n satisfying this. The greatest of such n's, say n_1, must not contain a prime factor p greater than 3, for then

$$R(n_1) = R\left(\frac{n_1}{p}\cdot p\right) \geqslant R\left(\frac{n_1}{p}\right) + \left(\frac{\log p}{\log 3}\right)$$
$$\geqslant R\left(\frac{n_1}{p}\right) + \left[\frac{\log p}{\log 3}\right] + 1$$
$$= R\left(\frac{n_1}{p}\cdot 3^k\right),$$

where $K = \left[\frac{\log p}{\log 3}\right] + 1$, since $R(3u) = R(u) + 1$ for all u.

But $\frac{n_1}{p}\cdot 3^k > n_1$. Hence n_1 is not the largest solution.

Let now $n_1 = 3u$; then $t = R(3u) = R(u) + 1$ so that $R(u) = t-1$. Assuming that the theorem is true for all values of $R(n) < t$, it follows that the maximum value of u is 2.3^{t-2}. Hence $3u = 2.3^{t-1}$ and the theorem follows by induction. From the theorem it follows, as proved by Pillai, that

$$R(n) \geqslant \left[\frac{\log n - \log 2}{\log 3}\right] + 1. \qquad .. \quad (5)$$

Theorem 3. $S(n) \geqslant [\log (n/2)/\log 3] + 1$.

For $S(n) = \varphi_1(n) + \varphi_2(n) + \ldots + \varphi_t(n)$; $t = R(n)$

Now $\varphi_t(n) = 1$; $\varphi_{t-1}(n) = 2$; $\varphi_{t-2}(n) \geqslant 2^2$; $\varphi_{t-3}(n) \geqslant 2^3$; ...

Hence $S(n) \geqslant 2^{t-1} + 2^{t-2} + \ldots + 2 + 1$

$= 2^t - 1.$

The theorem now follows by using (5).

(4) *$S(n)$ as a function of $R(n)$.* We will show that

$$2^{R(n)} - 1 \leqslant S(n) \leqslant 3^{R(n)}.$$

For a given $R(n) = t$, proceeding as above we get $S(n) \geqslant 2^t - 1 = 2^{R(n)} - 1$. When n is a power of 2, this becomes an equality. Again, by Theorem 2,

$$\varphi_t(n) = 1;\ \varphi_{t-1}(n) = 2;\ \varphi_{t-2}(n) \leqslant 2\cdot 3;\ \varphi_{t-3}(n) \leqslant 2\cdot 3^2 \ldots.$$

$$\varphi_1(n) \leqslant 2\cdot 3^{t-1}$$

Hence for a given t,

$$S(n) \leqslant 2\cdot 3^{t-1} + 2\cdot 3^{t-2} + \ldots + 2\cdot 3 + 2 + 1$$
$$= 3^t.$$

$S(n)$ actually reaches this value when $n = 2\cdot 3^{t-1}$.

4. Let us finally consider the solutions of the equation $S(n) = n$. By Theorem 1, n can only be odd. It is easily verified that two sets of solutions are $n = 3^k$ $(k = 0, 1, \ldots)$ and $n = 3p$ where p is a prime $= 4.\ 3^k + 1$, $k = 1, 2, \ldots$; we will prove

Teorem 4. If $S(3p) = 3p$ then $S((p-1)/4) = ((p-1)/4)$, where p is a prime > 3.

Proof. $3p = S(3p) = 2(p-1) + S(2(p-1))$

$$= 2(p-1) + 2S(p-1) + 1,$$

using $\cdot S(2u) = 2S(u) + 1$ if u is even

$$= S(u) \qquad \text{if } u \text{ is odd.}$$

Hence $S(p-1) = \dfrac{p+1}{2}$

Since $p > 3$, $\dfrac{p-1}{2}$ is even,

Hence $S\left(\dfrac{p-1}{2}\right) = \left(\dfrac{p-1}{4}\right)$.

Since $S(n)$ is always odd, $\frac{p-1}{4}$ must be odd, $p=4k+1$ (k odd);

$\frac{p-1}{2}=2k$, k odd $S\left(\frac{p-1}{2}\right) = S\left(2\cdot\frac{p-1}{4}\right) = S.\left(\frac{p-1}{4}\right)$

$= S$ since $\frac{p-1}{4}$ is odd.

Hence the theorem follows.

With the help of this result, we can derive an infinite sequence of classes of solutions of $S(n) = n$.

One is $n=3p$, where $\frac{p-1}{4}=3^a$ or $p=4\cdot 3^a+1=p_1$ say.

Another is $n=3\ p_2$ where p_2 is a prime given by $\frac{p_2-1}{4}=3\,p_1$ or $p_2=12p_1+1$.

.... ..

This sequence of solutions can be continued indefinitely. It is conjectured that this sequence exhausts all solutions.

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Structure and Physiology of the Timber Borer *Bankia Indica* with Special Reference to Boring Habit [1]

BY

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ABSTRACT

The role played by the shell, the adductor muscle and the foot in the boring process of *Bankia indica* is described after giving an account of their minute structure. A detailed account of the structure of the mantle and gills together with a description of the ciliary currents is given. The alimentary canal is described in detail with notes on the histology of the different regions so as to throw light on the adaptations to the peculiar diet to which the borer is committed. The course of the food particles through the alimentary canal is described. The results are discussed in relation to the findings of other workers.

1. *Introduction*

The Teredinidae as a group of molluscs has not only specialised towards a timber-boring habit but also has adapted their physiology to the exploitation of wood into which they bore as an important source of nourishment. These features have been attended with modifications in the structure of the organism to suit the new conditions of habitat and food. The present study is an attempt to elucidate several points in their anatomy which have a direct bearing on the boring habits of this bivalve. Such a study is an essential prerequisite for the formulation of the means for the control of these pests.

2. *Previous Work*

Historical: The most notable work on teredine anatomy is that of Sigerfoos (1908) who studied *Bankia gouldi, Teredo dilatata* and *Teredo navalis;* though studies on the anatomy have been

1. Part of a thesis which formed the basis of award of the degree of Doctor of Philosophy of the Madras University.
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made by earlier authors like Deshayes (1848), Quatrefages (1849), Grobben (1888), Menegaux (1890), Pelseneer (1891), and Beauk (1899). Important among the recent contributions on the anatomy of ship-worms are those of Potts (1923), on the structure and functions of the liver, Lazier (1924), on the morphology of the digestive tract, Yonge (1926 d), on the digestive diverticula, Miller (1922, 1924), on the shell and the mechanism of boring, Coe (1933, 1933a, 1934, 1935, 1936, 1943) on the sexual phases, and Purchon (1941) on the ciliary mechanism.

3. *Material and Methods*

The material for the following study was obtained from the wooden floats used for fishing off the Madras coast and also from test planks. The course of the ciliary currents in live specimens was studied with the help of current indicators like mercuric sulphide, powdered carmine and carborundum dust of various grades. Bouin-Dubosq was the fixative chiefly used, but Zenker's fluid, corrosive sublimate-acetic, osmic fluids and formaldehyde were also employed. Paraffin sections were cut at a thickness of from 4 to 10 microns and were stained with either Delafield's or Iron-Haematoxylin and counter stained with eosin or orange G. Mucicarmine was handy for the study of mucous glands.

4. *General Organisation of the Body*

Like other teredines, *Bankia indica* has a long, slender, cylindrical and soft body (Plate I). Covering only a small portion of the anterior end of the visceral mass, the irregularly shaped shell functions as the main organ for the excavation of the burrow and accommodates within it only the discoidal foot, the anterior part of the alimentary canal and the adductor muscles. The shell valves when in contact gape widely in front and behind for the projection of the foot and extension of the body respectively. The edges of the shell are not brought together because of the modified contact on the dorsal and ventral articulations (*vide* shell).

Fig. 1 shows the general arrangement of the organs in *B. indica* as seen from the right side when the shell and mantle are removed. Due to the lengthening of the body, the visceral mass occupies approximately a third of the body while the ribbon-like posterior part of the ctenidium (G) extends from the visceral mass to the base of the siphons (ES, IS). Most of the organs lie behind the posterior adductor muscle (PA) instead of anterior to

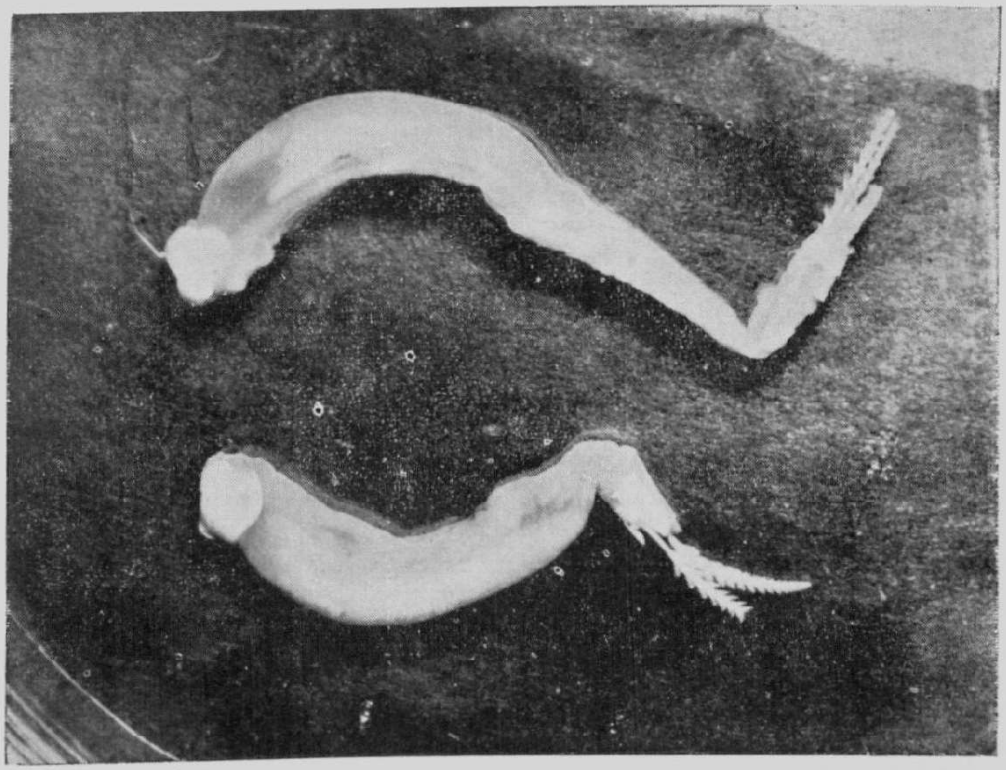
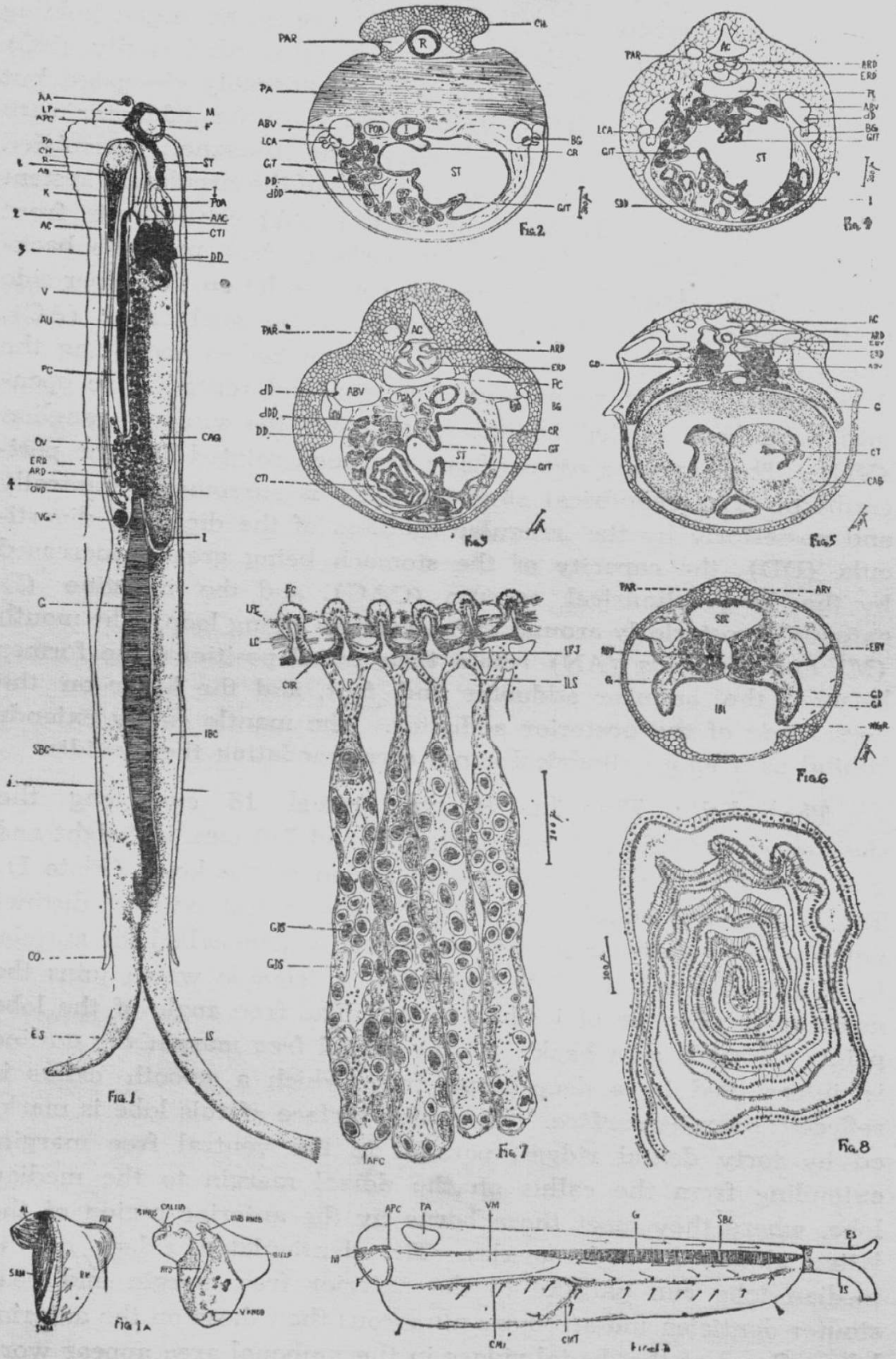


PLATE 1

Two entire Shipworms (*Bankia indica*)

it as in most lamellibranchs. The discoidal foot (F) occupies the anterior most portion in relation to its use as an organ holding the shell in position while boring. The pericardial cavity (PC) with its contained heart has not only considerably elongated but also has undergone rotation around the posterior adductor with the result it occupies a position above the intestine. Associated with these the relation of the various parts of the circulatory system has been changed. The posterior aorta (POA) continues in front of the ventricle while the anterior aorta (AAC) proceeds backwards. The kidney (K) and ureters come to lie on the upper side of the pericardial cavity and ventral to the anal canal (AC). The gonads are creamy white conspicuous bodies occupying the dorso-median and lateral portions. The renal and reproductive openings have been shifted behind as also has the visceral ganglion (VG). The digestive organs have not been rotated like the pericardium. The cylindrical stomach (ST) is surrounded ventrally and posteriorly by the irregular tubuless of the digestive diverticula (DD), the capacity of the stomach being greatly increased by the long cylindrical caecum (CAG), and the intestine (I) extendnig posteriorly around the caecum as a long loop. The mouth (M) and the anus (AN) retain their usual positions, the former between the anterior adductor and foot, and the latter on the dorsal side of the posterior adductor. The mantle cavity extends behind as a long cylindrical canal accommodating the ctenidia.

The shell: Fig. 1A. In an animal 18 cm. long the shell measures only 6·6 mm. in length and 7·6 mm. in height and therefore covers only a small front region of the body (Plate I). The valves are equal, each being inequilateral with a distinct anterior lobe, a median lobe, and a posterior lobe called the auricle. In outline the anterior lobe resembles a triangle which joins the median lobe by one of its long sides. The free angle of the lobe points forwards as a beak. At the dorsal free margin the outline is interrupted by a deep sinus from which a smooth callus is reflected over its surface. The outer surface of this lobe is marked by forty dental ridges, parallel to the ventral free margin, extending from the callus on the dorsal margin to the median lobe, where they meet those borne by the anterior region of the latter almost at right angles. The denticulated ridges of the median lobe run parallel to the anterior free margin and bear stouter denticles and are more numerous than those on the anterior lobe. Some of the dental ridges in the umbonal area appear worn out by friction. The dental ridges are turned upwards and con-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Figures 1—8

tinued into the middle region where the ridges become smooth and low and extend into the posterior region where they finally cease. The middle region of the median lobe is further marked by a shallow groove running down from the umbonal region widening as it reaches the ventral margin of the shell. The posterior region of the middle lobe is broadest, thick and convex. Posteriodorsally it extends into the auricle (Posterior lobe). The surface of the latter is marked by curved lines of growth which run parallel to its margin. This region of the shell is thin at the edges and slightly reflected outwards at the margin.

When viewed from within, the middle lobe appears to be pushed to the outside forming a trough in which is lodged a part of the anterior region of the animal. The front edge of the posterior lobe forms a shelf inside the shell projecting inwards into the hollow of the median lobe. Towards the extreme ventral edge of this lobe which is slightly bent inwards appears a sperical knob-like thickening of such a form as to permit a movable articulation with a knob borne by the opposite valve. This articulation is a feature peculiar to teredines and some pholads and marks the specialisation of the shell. Dorsally the margin of the middle lobe curves inwards to form a pronounced umbonal region, the tip of this region, the umbonal knob is of such a shape as to permit a freely movable articulation with the similar process of the opposite valve. The elastic ligament is reduced to a small squarish structure, since the two valves do not open and close. The movement of the shell is around the axis formed by the umbonal knob in the dorsal side and the ventral knob.

Hanging from the umbonal knob on the inner side of the shell is a broad sickle-shaped blade-like process, the apophysis accommodated within the concavity of the middle lobe. This plate serves for the origin of the pedal retractor group of muscles.

The three layers, periostracum, prismatic layer and the inner nacreous layer though present, differ from those of other bivalves in being relatively thickened and in the penetration of the prismatic layers by laminae of nacre. Over the dental ridges the periostracum is extremely thin, evidently worn out by friction in burrowing. It is evident that the shell of *Bankia indica* is highly specialised for performing boring movements.

The pallets: These peculiar calcareous structures which occur at the base of the siphons on either side beneath the muscular

collar of all the Teredinidae have been shown (Bartsch, 1922) to have a toxonomic value. The two pallets are pressed against each other so that the straight inner sides become apposed and the semicircular outer surfaces form a complete conical plug effectively closing the circular opening when the siphons are retracted within the burrow. If the salinity of the sea water should greatly change, or if poisons are introduced or the timber is taken out of water the pallets are thurst into the opening, closing the burrow as a shield protecting the soft animal inside. The retraction and protrusion of the pallets are brought about by a set of retractor and protractor muscles respectively. The adductors of the pallets facilitate the separation of the right and left blades of the pallet before the retraction of the pallets to enable the siphons to extend. The pallet consists of a series of upto 19 funnel-shaped, cone-in-cone structures which constitute its blade when the overall pallet length is 16·8 mm., which have been secreted and cemented in succession to a calcareous cylindrical handle. The cone-in-cone units of the blade are not of the same size, the largest one being found towards the base and the smallest at the distal tip, hence the blade tapers from base to tip. They are clearly separated, their distal margins covered with yellowish brown periostracum which is entire and separated from those of its neighbours, and drawn out into pointed processes laterally. The pallets are formed as a result of the secretory activity of the inner epithelium of the mantle at the posterior region and are formed in special pockets of the mantle. The handle is the first to be secreted and this is followed by the formation of the cone-in-cone units which are added on at the base as the animal grows. Hence the distal one is the oldest, since it is the earliest to be formed while the younger and larger ones are added in a series at the proximal part.

5. *The Internal Anatomy*

1. *The adductor muscles.* In sections the posterior adductor is about eleven times larger in area than the anterior adductor. Below the posterior adductor runs the terminal part of the intestines which bends over it in front, to terminate at the anus above it. Applied to the posterior face of the adductor lies the kidney. The adductor, unlike those of *Ostrea* and *Pecten* is homogeneous and consists of coarse muscle strands reddish in colour, stretching transversely across the anterior part of the body, from valve to valve. The muscle bundles are attached to nearly the entire inner surface of the outwardly turned auricles of the

shell and near its attachment to the shell the muscle becomes spread out so that it is thicker there than elsewhere. The anterior adductor is comparatively small and is attached to the outwardly turned anterior edge of the shell-valves just in front of the dorsal knob. This is also a compact bundle of similar striated fibres.

From their position of attachment on the valves and their size it is evident that these two muscles instead of contracting simultaneously as in typical dimyarians do so alternately drawing together the front and hind parts of the shell-valves causing the two valves to swing upon each other back and forth on the pivot formed by the dorsal and ventral knobs. This type of movement effects in boring. The large size and mode of attachment of the posterior adductor suggest that the valves can be brought together more vigorously in the posterior region than in the anterior and that this vigorous adduction is responsible for the powerful outward thrust of valves at the anterior part. The divarication of the front ends of the shell is of significance in the boring operations, the relatively feeble adduction through the contraction of the anterior adductor being sufficient to bring back the shell to the normal position (Miller, 1924).

1. *The foot*: The short, cylindrical foot of *Bankia indica* is situated at the extreme anterior end of the visceral mass ventral to the mouth (Fig. IF). This organ serves to grip at the burrowing end while the shell is rotated round this point of attachment during the drilling process. The front surface of the foot which is thus applied to the wood is cordate in outline, with a smooth cetral disc-like area or 'sole' (SO) bounded by elevated, wrinkled, peripheral, glandular margins (PR) which are ciliated and thrown into folds. This shape and structure are quite in keeping with its habit of using this organ for attachment. The median notch, marking the divisions between the muscle insertions of the right and left halves of the foot of *Teredo navalis* is absent in *Bankia indica*.

The distal end of the crystalline style-sac with the contained style lies close behind the epithelium of the central disc (Fig. 19) and extends backwards through the length of the foot. The anterior extensions of the digestive diverticula surround the sac of the crystalline style. The rest of the interior of the foot is packed with loosely arranged connective tissue containing muscle strands and blood spaces.

Histology: The folded rim of the foot is covered externally by an epithelium composed of very closely set columnar ciliated cells containing large nuclei. Large numbers of club-shaped vesicular glands sometimes grouped in bunches of varying sizes which are sub-epithelial open on the surface of the foot by slender ducts (Fig. 13) through small crypts. Basic stains act quickly on the granular cytoplasm of the gland. A deeply staining oval nucleus containing dark chromatin masses is present in each cell. Vacuolated cells also occur among the epithelial cells with large vacuoles surrounded by a thin layer of cytoplasm. The central disc-like 'sole' is formed of an outer layer of regularly arranged columnar non-ciliated epithelial cells with large round nuclei (Fig. 22). Beneath the outer epithelium of the foot, forming the main bulk of the organ lies loose connective tissue through which strands of muscles run in various directions. The histology of the connective tissue is not unlike that of turgescent tissues of other protractile organs. The cells are irregular with large intercellular lacunar spaces which can be filled with blood.

The musculature of the foot: The typical lamellibranch plan consisting of anterior and posterior retractors, as well as the protractors is modified in *Bankia indica*. The two anterior retractors run on either side of the foot from their origin on the anterior umbonal region of the right and left valves to their insertion on the peripheral ridge just beneath the ciliated epithelium (Fig. 19). The posterior pedal retractors are semicircular bands springing from the whole length of the apophyses and ending along the margins of the foot. The fibres are arranged all round the foot in such a way that their contraction must result in lessening the diameter of the foot. The protractors are comparatively smaller than the retractors and are attached to the anterior umbonal region close to the origin of the anterior retractor.

The Movements of the foot: are mainly two-fold, extension and retraction. The retraction may result in the withdrawal of the foot as a whole or of the 'sole' of the foot alone, producing a cupping action and facilitating attachments to the end of the burrow. The extension of the foot is mainly brought about by the flow of blood into the pedal sinuses and consequent turgescence. The blood flow is brought about by the contraction of the protractor muscle fibres which compresses the anterior part of the body driving the blood into the lacular spaces of the connective tissue of the foot. Due to the relaxation of the retractor muscles the

opening into the pedal sinus gets closed. Now the small fibres of the intrinsic muscles through their contraction diminish the diameter of the foot creating an area of increased pressure which tends to distend the foot sometimes 3 to 4 mm. beyond the anterior margins of the shell. The retraction of the foot is brought about by the contraction of the posterior retractor which increases the diameter of the foot, which thereby opens up the passage of blood into the foot and permits the blood to flow out. This is also aided by the rhythmic contractions of both the retractors which drive away the blood from the lacuna of the pedal mass due to general contraction of the foot. The constant changes of the contour and the thickness of each of the folds of the epithelium forming the rim of the foot are caused by the intrinsic muscle fibres in those regions.

3. *The mantle.* This 'fold of the integument of the dorsal part of the molluscan body' (Cooke, 1895) consisting of two membraneous flaps lining the two valves of a typical bivalve, has in *Bankia indica* and other ship-worms fused along its edges resulting in the formation of a long and delicate tube open in front for the protrusion of the foot and produced into long siphons behind for the ingress and egress of the water. The mantle is thin, soft, translucent, and not of uniform thickness. Anteriorly, the mantle is in close relation with the shell, it extends underneath the shell and ends in a slightly thickened edge below the front edges of the shell. On the hind edges of the shell the mantle is raised up and extends over the umbonal region as a thickened muscular ridge. This duplication of the mantle, called the 'cephalic hood' by Quatrefages (1849) was considered by him as the organ directly concerned with boring into wood. Sigerfoos (1908) and Kofoid (1921) were of opinion that it fitted into the burrow like a washer around the 'head' preventing fine particles of wood from entering the burrow. In all the animals examined the 'cephalic hood' (Figs. 1 & 2 CH.) appeared very calloused suggesting frequent rubbing against the walls of the burrow. The rest of the mantle except the siphons is thin and the organs of the pallial cavity are seen through it in the fresh animal and it is responsible for the secretion of the nacreous layers of the shell, the calcareous protective tubing, the apophyses and the dorsal and ventral knobs. Since, the entire surface of the mantle is thus capable of deposition of calcium, the opening area of the burrow, the oldest region, occupied longest by the animal has received the maximum deposition of calcium.

In cross sections of the animal, the mantle, on either side of the median line ventrally shows two prominent longitudinal thickenings (Figs. 3, 4, 5, & 6). The exact significance of these two ridges which are absent in *Teredo navalis* and *Teredo megotara* (Purchon, 1941) is not clear. It is probable that these folds from incomplete longitudinal partitions of the mantle chamber helping to some extent in regulating the movement of sediments accumulating in the mantle cavity.

The anterior edge of the mantle is thickened and consists of three lobes typical of lamellibranchs (Fig. 12). The outer lobe (OM), formed by the turning upwards of the mantle is nearest the shell while the inner lobe (IM) formed by the extension of the inner epithelium of the mantle projects forwards away from the shell. Between these two the edge forms the middle lobe (MM). Between the middle lobe and the outer lobe runs the periostracal groove (PEG) which is shallow in *Bankia indica*. This groove is lined by small flat secretory cells with deeply staining contents, which become continuous with those of the middle lobe and the tall and columnar cells of the outer lobe lying adjacent to the shell. The epithelium of the inner lobe consists of cells about half as high as those in the middle lobe. The middle lobe which is primarily sensory in function and bear tentacles and even eyes in forms like *Pecten* and *Spondylus* is devoid of any special sensory structures except at the tips of the siphons where sensory papillae are found. That part of the middle lobe forming the periostracal groove and its duplication over the shell are secretory. The fold of the outer lobe over the shell is responsible for the ridges and denticles resulting in the growth of the shell by marginal increment. Strands of periostracum (PE) are formed from the secreting cells at the bottom of the periostracal groove, where the epithelium is thicker than on the sides. Several such fibres coalesce to form thick strands which pass over the edge of the shell to the outer surface. The outer epithelium of the outer lobe lying below the shell secretes the prismatic part of the shell.

The Siphons: The Siphons are formed by the fusion of the mantle edges at the posterior end. The inner lobe of the mantle takes part in the formation of the siphons. The inhalant and exhalant siphons are joined together for about half their lengths, their openings fringed with sensory tentacles and papillae which perceive even suspended particles. Since the animal is buried inside the wood the siphons are the only visible soft parts of the

body in the natural habitat. The surface of the inhalant and exhalant siphons are mottled with small scarlet or brown spots. The interior of the former is uniformly of a light brownish hue except for six slightly raised white longitudinal patches running lengthwise. The inhalant siphon is, in normal life extended to about 5 cms. beyond the opening in an animal whose burrow measured 12 cms., the exhalant siphon is shorter, thinner and bears sensory papillae which are absent in *B. gouldi*. The withdrawal of the siphons into the burrow is brought about by the contraction of the siphonal retractors while their extension is through the relaxation of these muscles and by the flow of blood into the sinuses. Due to the arrangement of their intrinsic muscles the siphons can also shorten their length and diameter and thus increase or decrease the force of the current passing through them. As the inhalant siphon is wider than the exhalant, the inflowing current is less powerful in this form than in deposit-feeding forms like *Abra* and *Scorbicularia* (Yonge, 1949). In this respect the present form resembles forms like *Gari* and *Donax* (Yonge, 1949). The exhalant siphon is capable of bending and twisting movements. When faecal filaments are forced out, the exhalant extends sometimes even beyond the inhalant siphon and the excurrent water is ejected so forcibly and in such a swirling manner as to prevent the contamination of the neighbourhood. Both inhalant and exhalant are capable of constricting their distal end, and are sensitive to touch, light and salinity of the medium.

Structure of the Siphonal walls: (Fig. 11). In cross sections the siphons show a musculature far more specialised than the rest of the mantle. There are four layers of circular fibres separated by three layers of longitudinal fibres. Of the three longitudinal fibres the middle one is very thick and of the circular layers the outer is thinnest being two fibre thick. In the disposition of these muscle layers the present form resembles *Donax* (Yonge, 1949) except in the relative thickness of the various layers.

The histology of the mantle: The mantle consists of three layers. The outer layer is epithelial in character. It becomes thicker anteriorly and gets folded in the region of the cephalic hood where the cells are distinctly tall and columnar. Iron haematoxyln reveals large nuclei with prominent nucleoli arranged towards the base of the granular cells (Fig. 20). The rest of the epithelium is of secretory cells responsible for the formation of the calcareous lining of the burrow (Fig. 14). The inner epithelium is composed

of very flat cells the boundaries of which are not discernible. The protoplasm shows a faint reticular structure. On the side facing the mantle cavity the cells bear cilia which are especially long along the lateral margins of the mantle opposite the marginal and branchial grooves. The cells along these tracts of the mantle are tall, columnar and resemble those of the outer epithelium. The middle layer is of loose connective tissue. It is filled with numerous blood spaces and traversed by muscle bundles and nerve fibres. Longitudinal and circular muscle bundles can be made out beneath the outer epithelium while oblique fibres run between the two epithelia. Among the cells of the connective tissue of a piece of mantle freshly removed are spherical nodular bodies. Deshayes (1948) thought them to be non-nucleated mucous cells. Hancock (1845), on the other hand, believed them to be silicious particles responsible for the formation of the burrow. Sigerfoos determined them as deposits of calcium used for the formation and thickening of the calcareous tube which lines the burrow. These being opaque in transmitted light and white in reflected light and they being soluble in water together with their becoming russet when treated with iodine suggest their being of the nature of glycogen. They are probably reserve food stored in the mantle as glycogen. Alcohol in which specimens have been preserved turns turbid due to the re-precipitation of this glycogen.

Glands of the mantle : Numerous mucous glands are found in the anterior region of the inner layer of the mantle. These mucous cells appear distended with mucous and bulge slightly beyond the general epithelial surface. They bear short cilia at their free ends and have their nuclei on the proximal part of the cell. Large numbers of them occur crowded together along two lateral tracts which unite into a median tract behind the visceral mass and extend to the base of the inhalant siphon. Mucous glands are also concentrated along the dorso-median line of the roof of the supra-branchial cavity (Fig. 6). The composite gland composed of numerous vesicular acini lined by flattened, non-ciliated slightly granular cells described by Sigerfoos in *Bankia gouldi*, is absent in *Bankia indica*.

The ciliation of the mantle and course of water-currents in the mantle cavity. Tests with carmine particles and carborundum powder showed the course of water currents set up by the different groups of cilia. The cilia borne by scattered groups of cells on the inner mantle surface beat in such a manner that

particles falling on them are swept towards two distinct ciliated tracts which lie close to the branchial grooves in the anterior region and thence running backwards as a single tract behind the visceral mass. The cilia along these tracts beat vigorously backwards pushing the particles entangled in the mucus to the base of the inhalant siphon. Of the particles brought into the body by the inhalant siphon the coarse ones sink down by gravity over the gill-surface to the floor of the mantle cavity. The small particles impinging on the mantle surface are swept and fed on to the current caused by the ciliated tracts and sent to the base of the inhalant siphon and collected in a small space close to the base of the latter. When a sufficient mass of these is thus accumulated the particles are forcibly ejected by the contraction of the mantle as 'Pseudo-faces'. The finer particles form a slender rope with the mucous and are carried along the marginal and branchial groove towards the mouth. In the region of the labial palps the scattered groups of cilia function in a different manner. They beat forwards instead of downwards and backwards as those of the inner face of the mantle and drive forwards particles. In this way the particles of wood scattered during boring are prevented from passing backwards and clogging the burrow. The epithelium of the visceral mass of *Bankia indica* is non-ciliated unlike that of *Teredo norvegica* (Purchon, 1941) and hence there are no water currents over the visceral mass.

6. *The Alimentary Canal.*

The morphology of the digestive tract of teredines have been studied by Deshayes (1848), Beauk (1899), Sigerfoos (1908) and Lazier (1924) while Potts (1923), and Yonge (1926) have described and commented on the digestive diverticula of *Teredo Sp.* and *T. norvegica* respectively. In furnishing a complete account of the anatomy of this form *B. indica* hitherto not studied, it is hoped that comparisons with other forms will be facilitated.

The alimentary canal of *Bankia indica* measures 10 cm. in a specimen having a length of 11 cms. and is similar to those of *Bankia gouldi* and *Teredo navalis* in its general course. (Fig. I). The *mouth opening* is a medium transverse crescentic slit (M) whose concavity is dorsal, located below the anterior adductor muscle and above the foot and bounded laterally by the labial palps (LP). The labial palps in *B. indica* are small but distinct structures, the outer ones being larger than the inner ones. The

epithelium lining the mouth opening is a continuation of the epithelium of the inner surface of the labial palp and is composed of ciliated columnar cells among which occur large numbers of mucous glands and wandering phagocytes. Some of the epithelial cells have vacuolated cytoplasm containing prominent nuclei which vary from a round to an elliptical shape and are scattered through the centre and base of the epithelium. The cells covering the inner surface of the palp are tall while those of the outer surface are flat and non-ciliated with no mucous glands. The epithelium is underlined by a well defined basement membrane and the surrounding connecting tissue consists of large irregular cells with a few strands of circular and longitudinal muscle fibres lying immediately below the basement membrane. The mouth leads into the *oesophagus* which is a dorso-ventrally flattened ciliated tube measuring 1·1 mm. This narrows posteriorly and has its walls thrown into longitudinal folds of which a dorso-median ridge divides the tube imperfectly into a right and left portions. *The oesophagus* has a wall composed of three layers (Fig. 16) continuous through the whole length of the gut, an outer layer of loose connective tissue, a basement membrane and a columnar ciliated epithelium. The connective tissue is composed of loose irregular cells between which pass strands of muscle fibres and blood sinuses. The epithelium has a border cuticle and interspersed among the cells are many mucous glands as well as phagocytes some of which appear to migrate through the basement membrane. The oesophagus leads into a spacious, long, almost cylindrical *stomach* (ST) whose posterior part is slightly dilated. The stomach is divisible into a proximal part which lies high in the visceral mass receiving the oesophagus anteriorly and a distal part which lies lower in the visceral mass behind the posterior adductor communicating with the caecum and intestine.

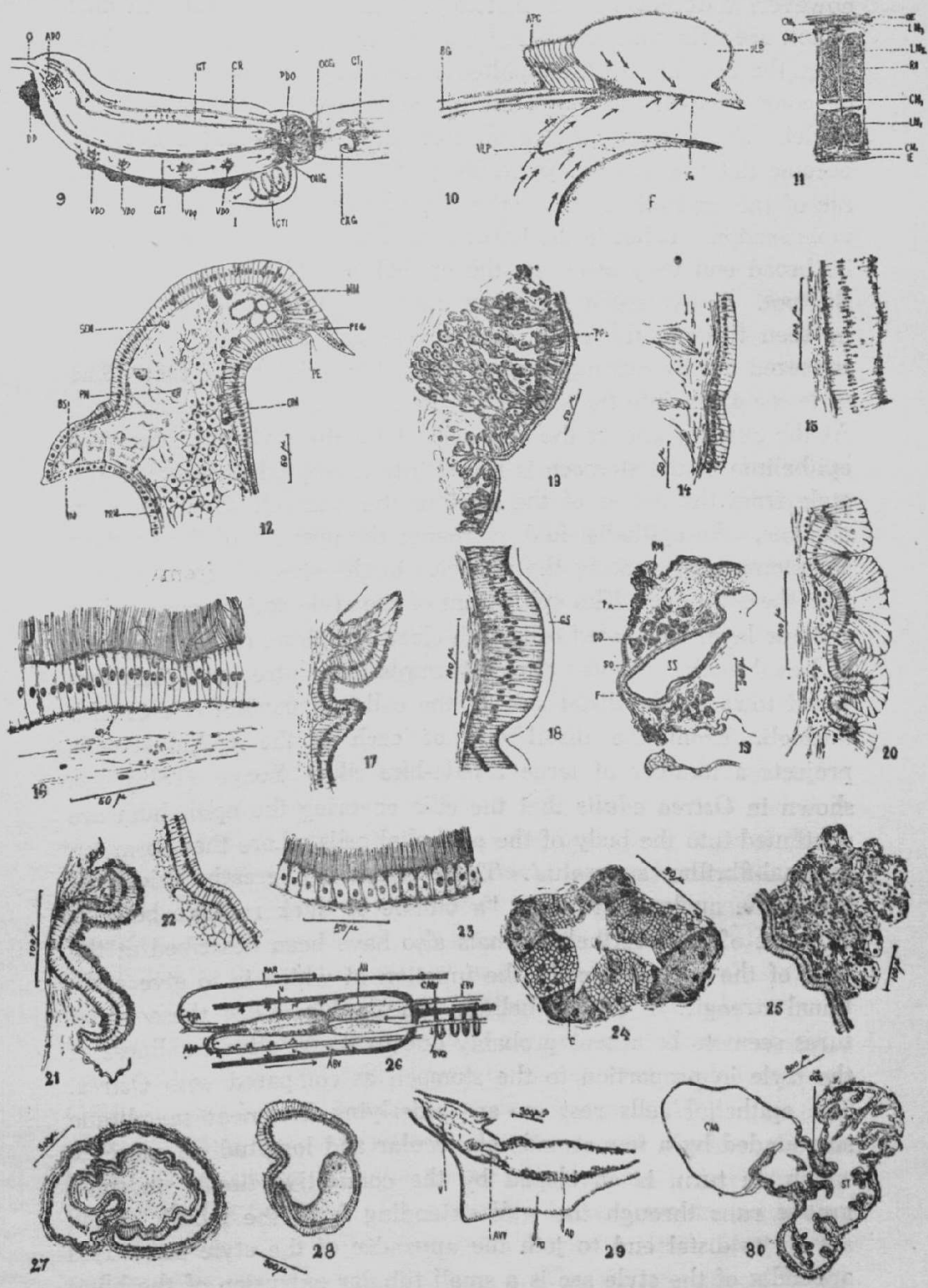
When casts of the stomach were made, the presence of three longitudinal ridges on the ventral wall extending almost the whole length of the stomach was revealed. Lazier (1924) has named these ridges in *Teredo navalis* as (1) *gastro-intestinal typhlosole* (2) *gastric typhlosole* and (3) *ciliated ridge*. 1. The *gastro-intestinal typhlosole* (GIT) starts in the lateral pouch (*vide infra*) and passes across the stomach ventral to the oesophageal opening and runs posteriorly along the ventral wall towards the right side of the stomach and continues into the posterior opening of the digestive diverticula and encircles the lumen of the posterior lobe of the digestive diverticula. This typhlosole emerges at the dorsal

edge of the hepatic opening and then passes posteriorly and ventrally into the intestine where it enlarges and coils forming a large coiled typhlosole of the intestine. (Fig. 3 CTI and Fig. 8). It is continued further as a low ridge and eventually disappears at the place where the intestine takes the bend at the posterior end of the caecum. From here to the anus there is no typhlosole for the midgut and hindgut. It is interesting to note that in *Bankia gouldi*, a small typhlosole extends throughout the whole intestine. Deshayes (1848) and Quatrefages (1849) on the other hand, describe the intestine of the forms they studied as being of uniform diameter indicating the absence of a coiled typhlosole. 2. The *gastric typhlosole* (Figs. 2, 9, GT) originates in the dorsal caecum and passing dorsal to the oesophageal opening runs side by side with the gastro-intestinal typhlosole and extends up to the posterior hepatic opening. 3. The *ciliated ridge* (CR) runs parallel to the free edge of the gastric typhlosole and ends on the right side of the caecal orifice. The cells of the gastro-intestinal typhlosole, gastric typhlosole and ciliated ridge are columnar, and vacuolated bearing long cilia. Their nuclei are prominent and sub-spherical.

Besides these three major folds is a row of transverse folds between the ciliated ridge and the gastro-intestinal typhlosole extending from the posterior end of the gastric typhlosole to the right fold in the caecal orifice. Another row of folds and furrows starts from the left fold in the caecal orifice and takes a spiral course to the right fold. These are visible through the surface of the stomach. Lazier (1924) believes that these irregular corrugations serve to increase the surface or the capacity of the stomach. But observations on living animals, however, suggest that this elaborate mechanism found especially towards the opening of the caecum at the junction of the latter with the stomach is only for sorting and guiding the food particles into the three channels namely, the caecum, the intestine and the openings of the digestive diverticula. The stomach is lined by an epithelium whose cells are not of uniform height, those on the roof being shorter when compared to the tall and columnar cells lining the floor. The whole epithelium is ciliated except for that region covered by the gastric shield. The ciliation is especially dense on the typhlosoles and the ciliated ridge. The nuclei are oval and prominent and are orientated along the central part of the epithelium (Fig. 15). The cells beneath the gastric shield do not bear any cilia and (Fig. 18) the nuclei which they bear vary from a round to an elliptical shape and are orientated at different levels, but the majority occur

towards the central part of the epithelium. Mucous glands do not occur but wandering phagocytes have been found both among the epithelial cells as well as among the connective tissue.

Diverticula of stomach: The stomach bears five distinct kinds of diverticula, namely 1. *the caecum* (CAG), 2. *the sac of the crystalline* style (Fig. 19 SS), 3. the *dorsal caecum*, 4. the *lateral* pouch, and 5. the *digestive diverticula*. (1) The *caecum* is a long cylindrical blind tube extending from the stomach to the posterior end of the visceral mass and forms the largest part of the alimentary canal. The intestine makes a long loop around it (See Fig. 1.). Dorsally it is covered by the gonads and dorsolaterally by the ctenidia and branchial groove. This organ is usually filled with wood particles. The ventral wall of the caecum is unfolded to form a two coiled typhlosole extending through the entire length of the caecum (Fig. 5, CT). This fold is of independent origin and not homologous with the fold of the intestine and increases the surface area of the caecum. The surface of the typhlosole is further thrown into many smaller folds resulting in the formation of additional longitudinal folds. The free edge of the typhlosole is further rolled up, considerably increasing the internal surface area of the caecum. The wood particles held between the folds of this typhlosole are found in a state of dissolution suggesting that a certain amount of digestion of wood takes place in the caecum. In *B. gouldi* the interior of the caecum is lined by a ciliated mucous membrane (Sigerfoos, 1908). However, in *B. indica* such a ciliated membrane appears absent as in *T. navalis* and there is no sign of ciliary movements in the living animal. The gastro-caecal orifice (Figs. 2, 9, 30, OCG) is guarded by two lateral folds of the wall, that on the right side being continuous with the caecal typhlosole. In transverse sections through this orifice a series of vertical incomplete partitions are seen and there is considerable development of musculature, both longitudinal and circular around the orifice. A transverse valvular fold also is present opening towards the caecum orientated towards the left side. The lateral folds divide the orifice incompletely so that two separate openings are provided probably one for the entrance and the other for the exit of wood particles in and out of the caecum. The cavity of the caecum is lined by non-ciliated flat epithelial cells the cell walls of which are very difficult to distinguish giving the appearance of a syncitium. The nuclei of the cells are spindle shaped and lie flat in the cells. The cells are so flat that they are hardly higher than the diameter of their nuclei. The cell boundaries are not perceivable



Figures 9—30

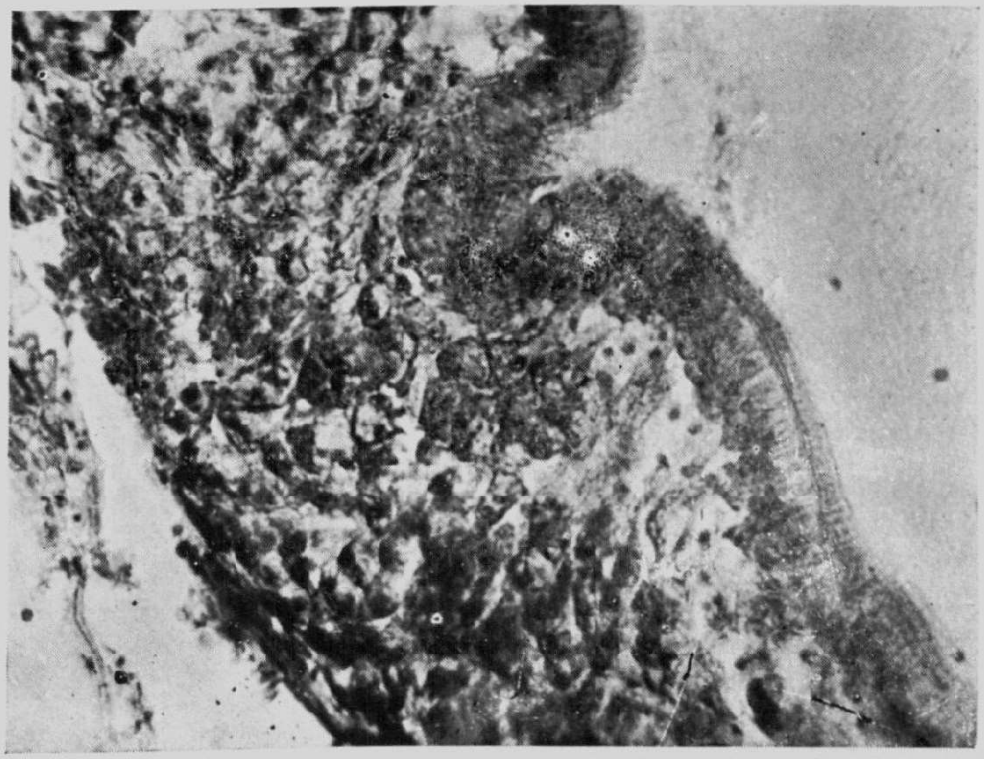
but probably exist. The distal tip of the folded typhlosole is, however, different in the histological details. The cells in this region are columnar and broad with vacuolated cytoplasm. In between the two folds of the typhlosole there is a considerable amount of connective tissue composed of cells bearing deeply staining nuclei. These cells become shorter and shorter and eventually become flat towards the proximal part of the typhlosole. 2. The *sac of the crystalline style* (Fig. 19, SS) is pyriform and oval in cross section. It lies in the lower side of the cavity of the foot with its broad end very close to the epithelium of the central disc of the foot. In live specimens when the foot is turgid, the style could be seen through the epithelium of the foot and could easily be squeezed out by making a puncture in the 'sole' of the foot. The style-sac opens into the stomach at its left side on the anterior end. At the anterior end of the stomach where the style sac arises the epithelium of the stomach is raised into a ring which protects the style from the action of the fluid in the stomach which is more alkaline. An epithelial fold overhangs the opening of the sac into the stomach preventing the particles in the stomach from getting into the style sac. The epithelium of the style sac is composed of a single layer of ciliated cells with clear cytoplasm containing large and oval nuclei, situated almost towards the centre or slightly disposed towards the distal side of the cell and contain one or two nucleoli. From the distal end of each of the epithelial cells projects a number of large bristle-like cilia. Yonge (1926) has shown in *Ostrea edulis* that the cilia covering the epithelium are continued into the body of the epithelial cells where they form 'an internal-fibrillar apparatus'. These fibres are greatly thickened below the nucleus and form "a bundle of thick rod-like bodies". A series of intra-epithelial canals also have been described at the base of the epithelial cells, the function of which is to give additional strength to the epithelium. In *Bankia indica* these structures seem to be absent probably due to the relative smallness of the style in proportion to the stomach as compared with *Ostrea*. The epithelial cells rest on an underlying basement membrane surrounded by a few strands of circular and longitudinal muscles, which, in turn, is enveloped by the connective tissue. A small groove runs through the wall extending from the mouth of the sac to its distal end to join the *appendix* of the style sac. This appendix of the style sac is a small tubular extension of the blind end of the style sac. The cells composing the groove are different from the rest in histological details but identical to those of the

non-ciliated granular cells lining the appendix of the style sac. The cytoplasm is granular with their nuclei disposed towards the base. The wall of the appendix has a lining epithelium closely resembling that of the groove. The cells forming it are not of uniform thickness those arranged on its roof being very thin and flat. The granular, nonciliated nature of the cells suggest that they are secretory and contribute to the formation of the style. The *crystalline style* is a gelatinous flexible body of a glassy transparency shaped like a club. Its widest part is near the central disc of the foot while the tapering part projects across the cavity of the stomach. Styles carefully dissected out showed spiral markings suggesting the rotation of the style by the cilia lining the sac as observed by Nelson (1918). The style in *B. indica* is smaller than those of *Martesia* and *Mya*. Since the style matter is continuously added on, the style was found retaining its bulk in all the specimens assisting both in digestion by the liberation of enzymes as well as in the mechanical process of pushing the particles of wood in the stomach. Transverse section of the *crystalline style* shows that it is composed of a homogeneous substance. The inner core of food materials reported for forms like *Anodonta and Lampsilis* (Nelson, 1918) is absent in *Bankia*. The *gastric shield* (Fig. 18) which in other bivalves functions as a protective shield against the abrasive action of the tip of the rotating style covers only a very small area of the stomach wall in *B. indica*. In freshly opened specimens the distal end of the style was found to be drawn out as a long flabby thread and it is difficult to conceive that the head of the style in a state of dissolution is pressed against the gastric shield as in other forms. This shield is closely applied to the gastric epithelium and can be easily dissected out of the animal. In sections the gastric shield is seen as a cuticular laminated structure in close relation with the underlying epithelium. According to Gutheil (1912) this structure is secreted by the epithelial cells that lie below. Pelseneer (1906), described it as the cuticular lining of the stomach and Johnstone (1899), called it as a gelatinous layer. Nelson (1918), who considered it to be chondrin named it as the gastric shield. Yonge (1926), is of opinion that the gastric shield is not a secretion but 'formed by the fusion of cilia originally in response to the irritation caused by the head of the style'. Berkely (1935), found the gastric shield composed of chitin. 3. The *dorsal caecum* is an outpocketing of the anterior, dorsal wall of the left side of the stomach. The epithelium of this region is folded and ciliated

internally. Since this sac also is filled with particles of wood it is probable that it serves as a receptacle of large particles of wood too many of which would block the proximal region of the stomach and impede normal functioning. 4. The *lateral pouch* is a distinct outpocketing of the stomach wall in between the dorsal caecum and the sac of the crystalline style. Since the wall of this pouch shows several muscle fibres it is probable that this sac by its pulsations aid in the movement of food in the stomach. A double fold of epithelium overhangs this pouch. The pouch is lined by tall non-ciliated columnar cells, which are not of uniform thickness. The epithelial lining is folded and rest on a thin basement membrane. Longitudinal and circular muscles and irregular connective tissue cells surround the epithelium.

5. The *digestive diverticula*: Yonge (1926 a) to whom we owe our knowledge of the assimilative function of these organs designated the term digestive diverticula. Since either 'liver' or hepatopancreas implies an assumption regarding its homology and since these diverticula of the stomach are accepted as digestive in function they are referred to as digestive diverticula.

These digestive diverticula in *Bankia indica* are composed of numerous small blind tubules, and occupy the ventral region and posterior side of the stomach and also extend into the foot and communicate with the stomach by ducts (See Figs. 1, 2, 3, 4, 9, 24, 25 and Ph.M. 1). The diverticula are distinguishable into three groups according to their distribution around the stomach. Most of the anterior diverticula occupy the interior of the foot forming a system of blind tubules, opening into the anterior part of the stomach by a number of opening (Fig. 9 ADO) on the right wall and by a single large opening in the anterior wall of the lateral pouch. Below the stomach closely applied to its wall and opening into it by four orifices form the second group (VDO). The posterior mass which is by far the largest is distinguishable into a large more dorsal right half and a small left half. The right part is brownish green and lies on the right side of the posterior part of the stomach while the left part is fawn coloured composed of few tubules and containing particles of wood. Although these two regions of the digestive diverticula differ in structural details they open into the stomach by a single large circular opening (PDO) on its ventral wall posteriorly. Neither Sigerfoos (1908) nor Lazier (1924) have described the histological structure of the *digestive diverticula* in detail. Our present knowledge is due to



PHOTOMICROGRAPH 1

Section through the foot showing the glandular development

the contributions of Potts (1923) and Yonge (1926 d). As in other teredines the unspecialised diverticula are composed of thick-walled columnar cells forming the tubules considered by Potts (1923) as excretory, whereas the specialised diverticula consist of thin walled wide tubules which are specially designed for the reception and intra-cellular digestion of wood. The lumen of the former is not of uniform diameter and in cross sections they appear as cruciform or triradiate (Figs. 4, 25, and Ph. M. 2). The cells forming these diverticula are irregular in shape and of varying heights with indistinct cell outlines. Two types of cells are distinguished in the epithelium of the tubule, those occupying the extremities of the arms (Fig. 24 c) of the lumen stain deeply and contain roundish nuclei. The others are lightly-staining larger cells with large vacuoles, dirty brown or yellowish pigment bodies, refringent granules, rod shaped as well as irregularly shaped bodies. Since fragments of these larger cells are found free in the lumen of the organ it is possible that they are cast out of the epithelium when loaded with the refractile and pigmented bodies. The darkly staining small cells found at the bases of the larger cells and at the extremities of the arms are considered (Yonge, 1926 a) as young cells which by dividing are able to make good the loss resulting from the casting off of larger cells. This suggestion is supported by the fact that division stages can be observed in the small cells and that this fact has been recorded by Gutheil (1912) in *Anodonta* and Yonge (1923) in *Mya arenaria*. The *specialised digestive diverticula* are different in histological structure. The epithelium is formed of circular, oval or irregular cells whose boundaries are indistinct or ill-defined. They contain clear vacuoles and circular nuclei containing darker nucleoli. Many cells with wood fibres within them are found loose or in groups within the cavity of the diverticula. These cells are amoeboid in character and movements. Ciliated columnar epithelium consisting of granular cells with basal nuclei line the duct and ductules of the digestive diverticula. They are placed over a thin basement membrane supported by a layer of connective tissue (Fig. 25). Enveloping this is a layer of circular muscle strands which pass round the duct.

The digestive diverticula have a sheath of muscle fibres which may aid in the circulation of food through the lumen. The ducts being lined by ciliated epithelium help the movement of particles. These facts and the presence of wood particles in the fawn

coloured region lead to the inference that this region of the digestive diverticula serve as a place where wood can be received and acted upon by enzymes (*vide* physiology of digestion). The occurrence of two functionally different kinds of diverticula has been reported in other ship-worms as well and Sigerfoos (1908), Potts (1923) and Yonge (1926, a) considered that these diverticula are specialised for the digestion of wood.

The *intestine* is dilated into a bulb (Fig. 1. CTI) at the place of its origin at the postero-lateral part of the stomach, runs a short course anteriorly, and then makes a sharp backward turn upon itself and approaches the median ventral part of the body and proceeds backward beneath the caecum, bends over it at its posterior end as a loop and continues forwards through the gonadal tissue and passes below the posterior adductor muscle and then bends over it in front and turns back to open into the anal canal at the anus.

Sections through the intestine show the presence of a *typhlosole.* The anterior bulbous part contains a typhlosole which is very much coiled (Fig. 8) which becomes a simple fold along the ventro lateral side of the intestine (Fig. 27) where the intestine takes a sharp backward turn and continues as a low ridge (Fig. 28) along the inner wall of the intestine till it makes the loop around the posterior end of the caecum. From that point to the anus there is no trace of a typhlosole and the midgut leads indistinguishably into the hindgut. It is noteworthy that in *Bankia gouldi* Sigerfoos has described a typhlosole extending throughout the intestine. On the other hand in *T-navalis* (Lazier, 1924), typhlosole is present only in the anterior portion of the intestine while Deshayes (1848) and Quartrefages (1849) describe an intestine in the species they studied as having uniform diameter thereby indicating the absence of a typhlosole. The intestine is lined by an epithelium of columnar ciliated cells whose oval nuclei are basal in location. The epithelial cells lining the wall of the hindgut are columnar and ciliated. There appears to be no muscle fibres on its wall.

It has been shown that there is a definite correlation between the length and also the function of the midgut and position where the faeces are extruded. Graham (1932) has observed that the very long coiled gut in *Patella vulgaris* is probably concerned exclusively with the formation of firm faeces. It is necessary that an animal which discharges its faeces into the mantle

cavity should avoid the fouling which would ensue, were these extruded in a loose or particulate condition. As absorption does not take place in the midgut, probably this region is mainly concerned with the preliminary moulding of the faeces.

The course of food particles through the alimentary canal

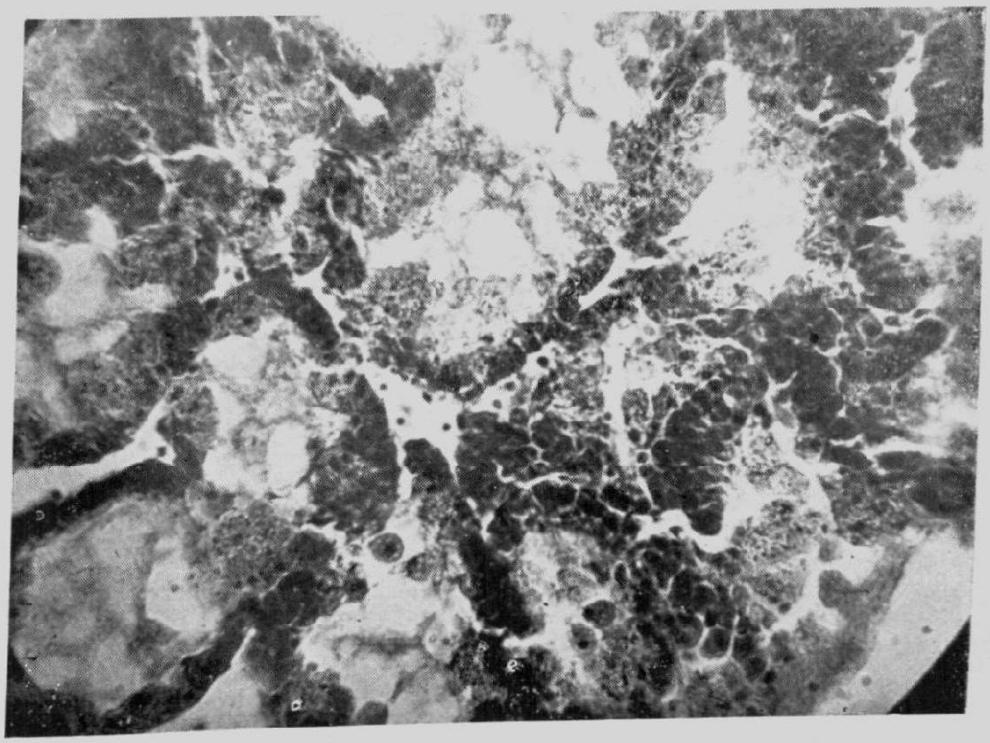
A study of the alimentary tract of living animals removed from the wood and examined under a binocular microscope with current indicators like carmine and carborundum revealed that besides ciliary action muscular movements comparable to peristalsis also aid in the progress of food materials. The lateral pouch on the proximal part of the stomach has muscular walls and their rhythmic bellowing action and elasticity help in driving the food particles with force towards the distal part of the stomach. The stomach is not ciliated throughout, while the typhlosole and the ciliated ridge are covered with long cilia. The typhlosole contains besides the ciliated cells numerous mucus-secreting glands which exude large quantities of mucus. The particles entering the stomach come under the influence of the rows of cilia (Fig. 9). The major part of it is carried away right into the caecum by the strong currents set up by the cilia on the typhlosole and the ciliated ridge. The smaller particles are found freely entering the ducts of the digestive diverticula where whirl-pools of ciliary currents are noticed. The lighter and finer particles are directed towards their openings. At the distal end of the stomach the coarser and heavier particles come in contact with the transverse folds between the ciliated ridge and the gastro-intestinal typhlosole and a major part of the food-mass is separated and sent towards the caecum. Another strong current is developed towards the large posterior hepatic orifice where particles are seen streaming in. At the region of the folds and furrows a section of the mass especially those comprising the larger particles are separated from the main food stream and is carried into the intestinal opening and finally into the midgut.

Besides these movements facilitated by ciliary action the ingested materials are subjected to the mechanical and chemical action of the rotating style and mixed with the enzymes liberated during the dissolution of the style. The ducts of the digestive diverticula bear abundant cilia and an internal investment of long cilia on the thin walls of the diverticula (Potts, 1923). Besides these the digestive diverticula are provided with muscle fibres so

that alternating contractions and expansions not always simultaneous in the different regions maintain a rapid circulation of food during and after the meal. In addition to all these it is to be presumed that the twisting and turning movements of the body while boring considerably augment the movement of particles in the stomach and to some extent in the caecum and prevent chocking of the regions with wood particles. The caecum is not ciliated in *B. indica* and feeble waves of muscular movements are noticeable both on the walls as well as on the typhlosoles. These muscular movements resembling peristalsis is evidently more suited for this storage organ which is usually found so full of wood chips that the cilia cannot efficiently function. The movements of materials in the intestine are obviously by cilia and the currents move towards the anus at an average rate of 1 mm. per two minutes. They are steadily propelled out of the hind-gut into the anal canal where they come under the influence of the exhalant current and are flushed out through the exhalant siphon by the contraction of the mantle which presumably exerts a squeezing action on the suprabranchial cavity, thereby bringing about the expulsion of the faecal matter in the form of long filaments some distance clear of the opening of the burrow. Observations on living animals inside the opened burrow sealed with a glass slide revealed that after burrowing for sometime the animal contracted the anterior ⅓ of the body. This contraction of the anterior end would naturally bring the head of the crystalline style far into the distal end of the stomach and even right into the caecal opening. This probably helps the animal not only to push in the particles of wood into the caecum preventing blockage of the gastrocaecal orifice, but also in conveying the style ferments into the caecum.

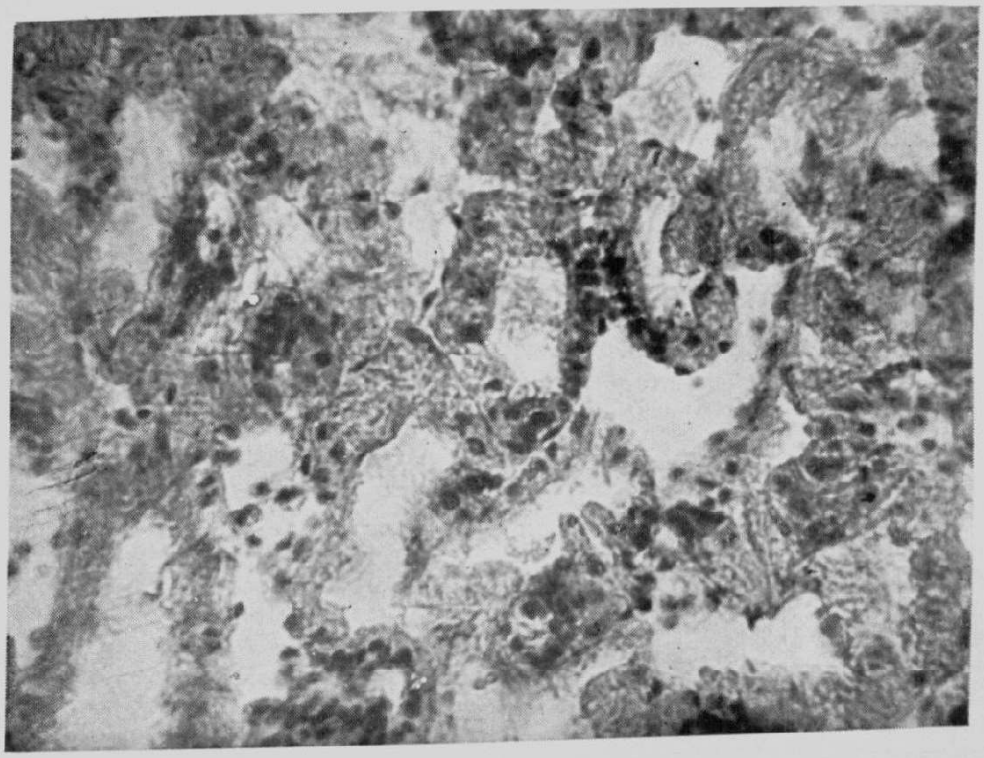
6. *The Gills*

Ridewood (1903), Sigerfoos (1908), and Purchon (1939, 1941) have described the ctenidia (Gills) of *Teredo navalis, Bankia gouldi* and *Teredo norvegica,* while the mechanism and function of the ciliary currents of the lamellibranch ctenidium have been elucidated by Wallengren, (1905), Kellog (1915), Yonge (1923, 1926, 1928) and Atkins (1936, 1937, 1937 a, 1937 b, 1938, 1938 a), further, the significance of the muscular activity in the ctenidia has been described by Setna (1930) and more recently by Atkins (1943).



PHOTOMICROGRAPH 2

Section through the digestive diverticula



PHOTOMICROGRAPH 3

Section through the specialised region of the digestive diverticula showing the particles of wood inside the cells

The ctenidia are outer demibranchs (Fig. 1. G), creamy white in colour, ribbon-like in form and conspicuous structures extending from the labial palps to the base of the siphons, measuring 9·5 cms., in a specimen 10 cms. long. Of this length nearly 2·9 cms. is occupied by only a groove, the branchial groove where no branchial filaments occur with the result that the anterior nine filaments (APC) are separated from the major part of the ctenidium situated towards the posterior part of the body. All the posterior filaments are of the usual 'V' shape found in demibranchs with a small groove at the apex of the 'V'. The grooves of the adjoining filaments form a long marginal groove for the demibranch, which anteriorly becomes continuous with the branchial groove thereby establishing connection with the anterior nine filaments on each side of the body. Sigerfoos (1908) considers this groove as formed by the great broadening of the 10th or 11th filament in *B. gouldi* during the post-larval development. The anterior nine filaments are disposed in such an oblique manner as to cover the base of the sickle-shaped apophysis situated behind the base of the labial palps at the sides of the foot dorsally, towards which the minute food particles are propelled along the marginal and branchial grooves. The major part of the ctenidium measuring 5·1 cms. lies behind the visceral mass embracing the posterior tip of the cylindrical caecum to some extent. The ctenidial axes of the right and left demibranchs converge behind the caecum into a singal common dorso-median axis bearing the filaments of the demibranch of the two sides.

Vertical sections taken through the entire gill across the axis bring out differences in the shape of the demibranchs between this species and *Bankia gouldi*. In this form the lamellae of the demibranchs pass laterally first and then vertically downwards unlike those in *Bankia gouldi* and *T.megotara* where they pass vertically downwards. The thickness of the visceral mass as well as the location of the gill-axis are probably responsible for this difference. However, this feature is likely to be of taxonomic importance since the section passing through *Teredo navalis* commented on by Ridewood (1903) also is strikingly different.

Each ctenidium consists of homorhabdic branchial filaments inserted uniformly on a vascular axis of loose connective tissue traversed by muscle fibres. They descend into the mantle cavity forming the lamella and the distal portions of the filaments folded upon themselves outwards forming the reflected limb of the

lamella. The interfilamentar junctions are organic indicating a synaptorhabdic condition and the intrafilamentar junctions are vascular representing a typical eulamellibranch condition.

The suprabranchial chamber is a long cylindrical canal disposed dorsally in the median line behind the visceral mass. Anteriorly, where the gills diverge and cease they become narrowed into slender canals lying above the branchial groove (Fig. 2. LCA) on the outer surface of the branchial vein in the substance of the mantle.

The *microscopic structure* of each branchial filament shows that its epithelium consists of three kinds of cells.

1. *The ciliated cells.* The frontal edge of the filament is occupied by five large cells bearing the shortest and weakest cilia (Fig. 7. FC). Two cells in the latero-frontal regions bear long and fine cilia in such a way that they interdigitate with their fellows of the adjacent filaments (LFC). These cells are tall and thin containing darkly staining nuclei placed towards the proximal part of the cell. The third type of cells occupy the lateral regions of the filaments. They are, however, separated from the latero-frontal cells by a non-ciliated region. The cilia are stiff and dense (LC) though not so long as those of the latero-frontal cells, the cells themselves are short and cubical in form containing large oval nuclei with scattered deeply staining reticular chromatin. Near the inter-lamellar regions on the hind end, the cells are short and bear thin short cilia (AFC) as in *Teredo* (Ridewood, 1903). In the filaments of the anterior portion of the ctenidium the cilia are not distinguishable into the four types so characteristic of the posterior part of the ctenidium. Further, the blood space of the filament is much more spacious in these than those in the posterior ones. In this respect it differs from *Bankia gouldi.* There are five cells in the frontal regions of the anterior filaments just as in the posterior filaments in the present form. It may be noted that in *B. gouldi* the frontal cells of the anterior filaments are more numerous, the central ones of which do not bear any cilia.

2. The non-ciliated cells form a single row one cell thick between the latero-frontals and the lateral cells in the filaments of the posterior part of the ctenidium. These cells differ from the rest in that they are short and contain lightly staining small nuclei.

3. The *secretory cells* are typical mucous cells found here and there in the free external borders of the epithelium. They are so distended with mucus that their nuclei are pushed towards the side. These glands are more abundant in the epithelium of the filaments of the anterior part of the ctenidium.

Internally the filament is filled with phagocytes, connective tissue where blood freely flows, and also by a glandular tissue known as the glands of Deshayes (Fig. 7 GDS).

A chitinous skeleton (CHR) is seen best developed in the posterior part of the gill and consists of thin chitinous rods which are thickened at the filamentar junctions.

The *glands of Deshayes*: (Fig. 7 GD) These glandular tissue occurs in the gill laminae as two distinct types of structures composed of two types of cells differing in histological details. A highly irregular ramification of filamentous structures lie in the epithelial walls of the gill laminae. They contain numerous deeply staining nuclei strewn indiscriminately in a mesh of granular cytoplasm which does not show any cell boundaries. The second type differs markedly from these in that they are composed of spherical bodies which take a deep stain with Delafields haematoxylin. The spherical bodies occur in groups of two, three or four and are distributed irregularly in the internal space of the gill lamina. The details of their distribution are shown in the figure. In *Bankia indica* these glandular structures are developed only in the gills. The umbonal portion of the gland found in *Teredo dilatata* and *Bankia gouldi* is not met with. From these glands located in the gills on either side a narrow cylindrical duct starts and passes forwards by the side of the branchial groove; (Fig. 4. dD); its lumen containing granular cells which take stain deeply and open on either side of the mouth. The arrangement of the different cellular elements of the glands and their disposition in the gill lamina differ considerably in *Bankia indica* when compared with that in *Bankia gouldi* where it is described fully by Sigerfoos (1908).

8. *The Ciliary Action of the Gills*

The part played by the vertical rows of four different types of cilia borne by the gill filaments in the collection and disposal of particles suspended in water accumulating in the mantle chamber was studied by using carmine particles and carborundum powder,

Large sized particles too heavy for ciliary action were dropped down to the ventral regions of the mantle cavity. Medium sized and fine particles are, however, swept over the surface of the ctenidium by the action of the cilia. The ciliary action of the gills is two-fold, since the four rows on each filament are responsible for the circulation of sea water through the body and also responsible for the collection of particles of food from the sea water and passing them on to the mouth. Since the former is primary it may be described first. As Wallengren (1905), and Orton (1912) have shown, the lateral cilia are mainly responsible for the in-going water current. The lateral cilia play upwards in a rhythmic manner making the water flow from infrabranchial cavity to suprabranchial cavity from where it finds exit through the exhalant siphon. Thus the water which enters the body through the inhalant siphon is lifted up into the suprabranchial cavity before it reaches the anterior end of the animal. Here, in the suprabranchial cavity, the water flows backwards aided by the backwardly beating cilia on the roof of the cavity and the pulsations of the heart, in a direction opposite to that of the incurrent water, till it reaches the exhalant siphon at the hind end, from where it is expelled out of the mantle chamber by muscular contractions of the mantle. During the course of this circulation inside the mantle cavity, the food particles suspended in it impinge on the surface of the gill filaments. On each filament long fine cilia belonging to the latero-frontal band beat slowly with effective strokes in a direction towards the frontal face of each filament. The movements of these cilia are not so rapid but they are vigorous. These cilia strain the particles of food and even push them by contact. To a slight extent water will be displaced out of the infra-branchial chamber in the process of straining movements of the latero-frontal cilia. The particles which are prevented from passing between the filaments are pushed from either side of the filament towards the frontal surface, get entangled in the mucous secreted in that region and are swept down along the filaments by the action of the frontal cilia beating downwards. At the ventral margin of the gills, the strong cilia borne by the marginal groove carry these particles forwards, along the edge of the gill. As a large number of mucous secreting cells occur in the groove a thick rope of mucus traps the medium and fine particles which alone are carried oralwards propelled by the forwardly beating rows of cilia. At the front edge of the posterior part of the ctenidium i.e., at the junction of the marginal groove with the branchial

groove large sections of the mucous thread weighed by the medium sized particles drop down to the floor of the mantle chamber due to the narrowness of the lumen of the branchial groove. Meanwhile the very fine particles which were not so stuck to the mucus are freely propelled oralwards by the cilia of the branchial grooves. These fine particles reach the mouth along the ciliated groove of the anterior gill filaments and between the outer and inner labial palps. It is noteworthy that only the finest particles of suspended matter are thus conveyed to the mouth. The medium sized particles though collected have been forced to be rejected due to the unusual lengthening of the body and development of the caecum both of which contributed to the narrowing of the branchial groove. The medium sized particles along with the heavy ones dropped down from the gills through sheer gravity are expelled from the mantle cavity, through the combined action of the cilia borne on the ciliated tracts and also by muscular contraction of the mantle as 'pseudo-faeces' through the inhalant siphon (*vide supra*). This peculiar type of selection of particles by the gill and subsequently by the branchial groove observed in *Bankia indica* is probably accounted for by the absence of the fine grooves which aid in selection of food particles in the palps of forms like *Ostrea*.

Thus, even though the essential food collecting mechanism apparently is represented and functions as in typical lamellibranchs the enormous development of the caecum and elongation of the body have resulted in the attenuation of the branchial groove into a passage of such a length and narrow diameter that the food-particles that can travel along this to the mouth become limited in size and quantity. This involves such a rigorous elimination that the major part of the plankton and suspended particles conveyed by the water currents have to be rejected as 'pseudo faeces'. Thus, the function the gill discharges as a region of reception of food in other lamellibranchs is minimised owing to the high specialisation towards the exploitation of the wood which they bore into as food. This accounts for the absence of plankton and diatoms in the alimentary canal. As a direct consequence of the specialisation for a wood-boring habit and a diet of wood particles, this mollusc has become so incapable of making full use of the water-borne, planktonic food that when the wood supply is exhausted, it becomes helpless even though it is equipped with enzymes which can digest planktonic food (*vide*, physiology of digestion).

Rate of travel of particles: A few observations were made on the rate of travel of particles along the filaments towards the

marginal groove and also along the groove towards the mouth. These observations were of the speed of small particles conveyed along these paths. Rate of travel when examined with the help of thin bits of lead foil was found to be very regular as there are no conflicting currents. Most of the speeds recorded fell close to 0·3 mm./second along the marginal groove and 0·5 mm./second along the filaments at 30° C.

Branchial movement: The gills are capable of considerable muscular movements. When particles are deposited on the surface of the gills, writhing and swaying were observed. Such movements have been noticed by Kellog (1910) in *Pecten*. Touching a part of the gill with a needle causes contorsions for a distance along the lamellae which are mainly brought about by elongation followed by contraction. Transverse movements of the filaments were also noticeable. Probably this may be important for permitting large particles to fall off thereby removing more effectively the sediment falling on the gill-surface.

9. *Discussion*

The foregoing study of the anatomy of *Bankia indica* notably illustrates the remarkable co-ordination between structure and function. The foot of this lamellibranch illustrates how plastic the organic body proves when an animal adopts an entirely new habit. It is so specialised that it is just a sub-circular disc with pedal muscles of other lamellibranchs modified for assisting the boring into wood. Notwithstanding the correct suggestion of Home (1806), that the foot adheres to the wood acting as a centre-bit while the animal is boring with the shell, there has been from time to time repeated claims that the foot is directly concerned with the boring activities (Dall, 1895; Hedley, 1901; Kuhlman, 1914). These authors considered that the shell effects a grip on the sides of the burrow supporting the foot against the wood. The reason for this idea perhaps might be due to the fact that in other pelecypods such as *Mya, Solen,* and *Cardium,* the excavation is accomplished by the foot, the shell serving only for protection. A study of the histological structure of the organ in the present form, as well as the observation of this organ in living specimens lead the present author to conclude that its primary function is to effect a cupping action at the end of the burrow to hold the shell in position while boring. The peripheral wrinkled region with its ciliary covering helps to conduct the grated particles of wood to the mouth, since the labial

palps are reduced and the mucous secretion from this region assist in cementing the stray particles into a coherent mass. The unique shape of the shell, its elaborate denticulations, the development of additional faces such as the posterior lobe for the insertion of the posterior adductor muscle, the acquisition of secondary articulations to achieve a special type of movement, the loss of the hinge teeth, the reduction of the ligament and the histological peculiarities suggest that the shell in *Bankia* has a specific function to perform. When one considers that the shell in these forms is not only a boring organ but a feeding organ as well, Hedley's suggestion (1901), that the shell in Teredinidae is degenerate and that the next step would be its complete disappearance sounds strange. It is true that the shell has lost the protective character, but in this specialised lamellibranch in which protection to the naked body is afforded by nacre lined burrow, the shell is assigned the role of scraping off such fine particles of wood as will be easy to feed upon and the form, both of the foot and shell suggests a long specialised evolution. The pronounced development of the posterior adductor, its fibres being suited for powerful and frequent contractions, its insertions being adapted for facilitating powerful outward thrust of the denticulated regions of the valves indicate that this muscle is part of a mechanism which has been perfected as a whole. It is obvious, no one organ can be elaborated or simplified without the other organs associated with it undergoing appropriate changes. One is inclined to wonder whether such a complex mechanical pattern, can be produced by just an opportune gene pattern or by the gradual addition of details of structure.

Although the teredines have especially a basic palp structure consisting of an outer and inner one, they show various degrees of reduction, variation in size, shape and functional efficiency. While the outer and inner palps may be recognised in all the species studied, the relative proportions and the development of each of these vary considerably and the variations can be attributed to habit specialisations. The palps of *Teredo norvegica* are the least reduced and are reported as capable of exercising quantitative selection (Purchon, 1941), while those of *Teredo megotara* (Purchon, 1941), *T. navalis* (Lazier, 1924), *Bankia gouldi* (Sigerfoos, 1908) and *Bankia indica,* are greatly reduced and incapable of quantitative selection on the suspended food particles drawn towards the mouth. The presence of a long palp in *T. norvegica* is explained by its habitat in waters where turbidity is high whereas in others such a selection is unnecessary since they are found in

comparatively clear waters where turbidity is practically nil. Further the manipulation of the wood-chips during burrowing is mainly done by the peripheral folds of the foot. That the boring life is not a causative factor for the reduction of the palp is evident since other boring forms of the family *Pholadidae* do not, show reduction of the palps except *Xylophaga* (Purchon, 1941) which is also found in the offshore, less turbid waters.

The caecum of the *Teredinidae* is the enlarged 'dorsal pouch' of other eulamellibranchs evolved as a result of the response to the boring habit and utilisation of wood as food. The large quantities of saw-dust that accumulate while excavating has had to be stored so that no strain is placed upon the remainder of the alimentary canal. The structure of the gastro-caecal orifice indicates that there is provision for the simultaneous ingress and egress of wood-particles into and out of the caecum. The caecum helps to control the arrival of food particles like a safety valve. The development of a specialised region of the digestive diverticula for the intracellular digestion marks another adaptational feature evolved in the morphology of the teredines. Side by side with these unique specialisations there have been physiological adaptations as well, such as the relative dominance of the carbohydrase system correlated with a woody diet, and the elaboration of additional enzymes like cellulase and enzymes to deal with pentosans for the digestion of cellulose and other constituents of wood. (Nair, 1955.)

The gills of *Bankia* show considerable reduction in relation to its boring life, being represented only by a single demibranch. The orientation of the gill is noteworthy since nine filaments are placed far in front on either side of the labial palps and the main ribbon-shaped posterior part behind the visceral mass, the two being connected by a narrow branchial groove. Even though the food collecting mechanism is retained, the function of the gills as a region of reception of food particles has been subordinated since the amount of nourishment needed form the plankton is very little being a small amount of nannoplankton to suppliment the scarce nitrogenous constituents present in the wood.

However, this perfection towards the utilisation of the wood necessitated the growth and elongation of the body throughout life in various directions inside the wood. Even the major part of the planktonic food collected by the gills finds the small passages of the branchial groove too narrow in *Bankia indica* with the result a

good portion of the mucous rope with the cemented food has had to be rejected as pseudo-faeces. Thus, the dependence of these borers on wood has become so much that the borer community colonising a timber block will have to perish when the wood supply is exhausted. However, to some extent the dependence of these marine forms on wood which is a terrestrial product both for food and shelter, and the vicissitudes attending the free-swimming larvae in finding out the adult habitat have been compensated by a prolific fecundity, the development of sperm-conserving devices, and a protracted free-swimming period during which time dispersal is effectively accomplished.

10. *Acknowledgements*

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EXPLANATION OF FIGURES

Figures 1—8

Fig. 1. The general disposition of the organs in the mantle cavity as seen from the right side.

Fig. 1. A. Outer and inner views of the shell.

Fig. 1. B. The course of ciliary currents in the mantle cavity.

Figs. 2-6. Series of transverse sections of an adult animal along the lines indicated in fig. 1.

Fig. 2. Section through the posterior adductor and cephalic hood passing through the line marked 1 in fig. 1.

Fig. 3. Section through the coiled typhlosole of the intestine passing through the line marked 2 in fig. 1.

Fig. 4. Section through the posterior digestive diverticula passing through the line marked 3 in fig. 1.

Fig. 5. Section through the caecum passing through the line marked 4 in fig. 1.

Fig. 6. Section through the branchial region passing through the line marked 5 in fig. 1.

Fig. 7. Section through four gill laminae showing the ciliation of the filaments, position and structure of the glands of Deshayes.

Fig. 8. Transverse section through the coiled typhlosole.

Figures 9—30

Fig. 9. The ciliation and the course of food currents inside the stomach.

Fig. 10. Anterior part of the ctenidium and the labial palps showing the course of particles.

Fig. 11. Transverse-section through a part of the inhalant siphon showing the musculature.
Fig. 12. Transverse-section through the mantle edge.
Fig. 13. Transverse-section through the peripheral glandular rim of the foot showing the glands.
Fig. 14. Transverse-section through the outer mantle epithelium.
Fig. 15. Transverse-section through the stomach epithelium.
Fig. 16. Transverse-section through the oesophageal wall.
Fig. 17. Transverse-section through the gastro-intestinal typhlosole.
Fig. 18. Transverse-section through the stomach epithelium below the gastric-shield area.
Fig. 19. Sagittal-section through the foot showing the position of the style-sac.
Fig. 20. Longitudinal-section through the cephalic hood.
Fig. 21. Transverse-section through the branchial groove.
Fig. 22. Transverse-section through the central non-ciliated 'sole' of the foot.
Fig. 23. Transverse-section through the epithelium of the style-sac.
Fig. 24. Transverse-section through a tubule of the digestive diverticula.
Fig. 25. Transverse-section through a part of the digestive diverticula.
Fig. 26. Diagram showing the heart, the main arteries and veins.
Fig. 27. Transverse-section through the intestine showing the typhlosole.
Fig. 28. Transverse-section through the intestine showing the reduced typhlosole.
Fig. 29. Sagittal-section through the pericardium showing the ventricle and the auricle (one side only).
Fig. 30. Transverse-section through the gastro-caecal orifice.

KEY TO LETTERING OF FIGURES

AA—Anterior adductor
AAC—Anterior aorta
ABV—Afferent branchial vein
AC—Anal canal
ADO—Anterior openings of the digestive diverticula
AFC—Abfrontal ciliated cells
AL—Anterior lobe of the shell
AM—Adductor Muscle
AN—Anus
APC—Anterior part of the ctenidium
APS—Apophysis
ARV—Afferent renal vein
ARD—Afferent renal duct
AU—Auricle of the heart
AUR—Auricle of the shell
AVV—Auriculo-ventricular valve
BC—Border cuticle
BG—Branchial groove
BM—Basement membrane
BS—Blood space
BV—Blood vessels
BVG—Blood vessels of the gills
BY—Byssus thread
C—Crypts of the darkly staining young cells
CA—Caecal artery
CAG—Caecum of the stomach
CABV—Common afferent branchial vein
CH—Cephalic hood
CHR—Chitinous supporting rod
CMT—Ciliated mantle tract
CM—Circular muscle
CO—Collar of the mantle
COT—Connective tissue
CR—Ciliated ridge
CT—Caecal typhlosole
CTI—Coiled typhlosole of the intestine
DD—Digestive diverticula

dD—Duct of the gland of Deshayes
dDD—Duct of the digestive diverticula
DLP—Outer labial palp
EBV—Efferent branchial vein
EN—Endodermal cells
ERD—Efferent renal duct
ES—Exhalant siphon
F—Foot
FC—Frontal cilia
FV—Food vacuoles
G—Gills
GA—Ascending limb of the gill
GD—Descending limb of the gill
GDS—Glands of Deshayes
GG—Groove at the ventral margin of the gill
GIT—Gastro-intestinal typhlosole
GS—Gastric shield
GT—Gastric Typhlosole
I—Intestine
IBV—Infra branchial cavity
IE—Inner epithelium
IFJ—Inter-filamentar junction
ILS—Inter-lamellar space
IM—Inner lobe of the mantle edge
IS—Inhalant siphon
K—Kidney
LC—Lateral cilia
LCA—Lateral canals
LFC—Latero-frontal cilia
LG—Lateral oral groove
LM—Longitudinal muscle
M—Mouth
MA—Mantle
MC—Mantle cavity
MES—Mesodermal cells
MF—Muscle fibres
MG—Mid gut
MGR—Margnial groove
MM—Middle lobe of the mantle
MUG—Mucous glands
N—Nuclei of the epithelial cells
O—Oesophagus
OCG—Gastro-caecal orifice
OD—Opening of the digestive diverticula
OE—Outer epithelium
OM—Outer lobe of the mantle edge
OMG—Opening of the stomach into the midgut
OV—Ovary
OVD—Oviduct
P—Phagocytes
PA—Posterior adductor
PAR—Pallial artery
PB—Pre-oral band of cilia
PC—Pericardium
PDO—Posterior opening of the digestive diverticula (posterior hepatic opening)
PEG—Periostracal groove
PGL—Gland cells of the foot
PN—Pallial nerve
POA—Posterior aorta
POM—Posterior median part of the shell
PR—Peripheral region of the foot
PRM—Pallial retractor muscle
R—Rectum
RA—Radial muscles
RM—Retractor muscles
SAM—Anterior median part of the shell
SBC—Supra branchial cavity
SDD—Specialised region of the digestive diverticula
SEM—Sub-epithelial mucous glands
SH—Shell
SHG—Shell-gland
SMM—Middle-median part of the shell
SOL—Non-ciliated central 'sole' of the foot
SS—Style sac
ST—Stomach
T—Typhlosole
Umb-knob—Umbonal knob
V—Ventricle
VDO—Ventral orifices of the digestive diverticula
VG—Visceral ganglion
VLP—Inner labial palp
VM—Visceral mass.

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A Review of Plant Viroses in India

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Although much work has been done on fungal diseases of plants in India, similar work on plant virus diseases has been little attended to. The discovery of sandal spike, cotton stenosis and later on various growth abnormalities of potato, tobacco and other economically important crops, created a new field for plant pathologists and attention began to be focussed on plant virus diseases.

During the Forty-first Indian Science Congress, under the chairmanship of Dr. B. C. Kundu, many scientists expressed the need for a general survey of plant virus diseases in India, and urged for a greater interest in the field of plant virus research. The present work is aimed at such a need. In this review, since the identity of many viruses have not been established, the diseases have been divided based on the symptoms they produce, and the hosts described under each. Wherever identified, the synonyms of the virus based on Smith's and Holmes's systems of classification are given. At the end of this review is given a table of the physical properties of sap-transmissible viruses and a host index.

Most of the information presented here is gathered from the "Review of Applied Mycology", the rest from various Indian and a few foreign journals. In this review 70 diseases affecting 59 species belonging to 19 families of Angiospermae are reported.

MOSAIC DISEASES

'Mosaic' is a foliar symptom characterised by mottling of the leaves. Very often this main symptom is followed by other subsidiary symptoms like enation, distortion, vein-banding, sterility, and leaf-curl and accordingly several types of mosaic are recognised.

1. *Mosaic of Sugarcane* (*Saccharum officinarum* L.).

The virus : *Saccharum* Virus 1, Smith.

Marmor sacchari Holmes

The first record of the mosaic disease of sugarcane was made by Dastur (1923) during the year 1921 from Pusa. In 1925, typical infection was observed on the cane Co 213 at Pusa (Mc Rae, 1926a) and a search was made in other cane growing provinces of India. It was shown that the disease was prevalent in Assam, Bihar, Bombay, Central Provinces, Madras, Punjab and United Provinces (McRae, 1926a). During 1925, the disease was a serious epidemic throughout North India on the local variety Hemja, while in South, during 1925-29, Godavari district showed an average of 10% infection and Carnatic district 2%. Occasionally however, in these two districts the incidence was as high as 50% (McRae, 1926a; Sundararaman, 1928a). While it is difficult to establish the source of the disease in India, there is evidence to show that the disease was in existence for a long time in India among the thin varieties of cane (McRae, 1926a). The fact that mosaic of sugarcane was first observed in Java in 1892, and that Java varieties POJ 213 and Java 213 are highly susceptible to the disease, lends support to the view that the disease is imported from Java (Dastur, 1926; Barber, 1926). It is suspected that the disease at Pusa, was introduced from Lyallpur (Punjab) on Co 223 which in Punjab is reported to be highly susceptible (McRae and Subramaniam, 1928).

Symptoms :—

The characteristic primary symptom of the disease is the mottled appearance of the young leaves of the crown, with patches of light and dark green colour, alternating with each other in a definite manner. Mottling is most prominent on the first partly unfolded leaf of the crown, becoming faint in the older leaves below. The patches are of various sizes, but mostly linear and arranged parallel to the midrib of the leaf. The intensity of mottling varies with the cane variety and season, being prominent during the spring and summer (Chona, 1947). Some varieties do not manifest any signs of infection until late, while in others the symptoms disappear early (McRae, 1930). Eventhough the secondary symptoms are rare on the Indian varieties of sugar-

cane, they have been observed in a few cases. These include:— whitish or yellowish linear blotches on the internodes of the dark-rinded varieties, which gradually develop into narrow linear cankers resulting from the death of the rind tissue, shrinkage of the internode and in advanced cases stunting (Sundararaman, 1928a).

Histopathology :—

The pale areas of the infected leaf are thinner and the cells in such areas smaller than those in healthy parts. In the infected tissue, the chloroplasts are smaller, less regular in shape and fewer in number (Sundararaman, 1928a). Cytological studies revealed the presence of protozoan-like bodies in the infected canes in all parts of the plant body including the xylem (Dastur, 1926; 1933).

Transmission :—

Under experimental conditions, the disease has been found transmissible by sap inoculation, the sap from infected leaves when rubbed on to healthy leaf causing the disease. Injection of the infective juice into the stem using hypodermic needles and inserting plugs of infected tissue into the setts before planting were also successful (Sundararaman, 1928b).

Transmission of the disease through setts is very common, and this is perhaps one of the ways by which infection takes place in the field. The disease being systemic, the young shoots coming from the infected setts show mosaic symptoms. Seed transmission does not take place (Sundararaman, 1936; Chona, 1944) and the exact identity of the insect vector has not been established, although there are reports of the presence of *Aphis maidis* on the infected canes (Sundararaman, 1928a). Experiments at Pusa and Delhi reveal that infection is maximum in the months of May and June, hot weather favouring transmission of the disease and that the incubation period of the virus ranges from 7 to 10 days (Chona, 1944). The maximum spread of infection in the field is shown to be between the third and fifth months of growth (Sundararaman, 1928b). Secondary infection, if any, begins when the crop is three months old, and spreads rapidly during the next three months followed by a marked decrease (Sundararaman, 1928a).

Susceptible and Resistant varieties :—

While it is difficult to draw a distinction between the susceptible and the resistant varieties of sugarcane in India, it can be generally said that only a few varieties have been found to be resistant to the disease both under experimental and natural conditions. Forty-seven seedling canes, 9 crosses at the Coimbatore breeding station, 2 thin cane varieties and 38 tropical varieties are reported to be susceptible (McRae, 1928). The highly susceptible varieties include: A 2, B 147, B 208, B 254, B 688, B 6308, Chunnee, Co 203, Co 205, Co 210, Co 213, Co 223, Co 232, Co 281, Co 290, Chittal, D 625, D 1135, Fiji B, HM 313, HM 320, HM 332, HM 554, Java Hebbal, J 213, J 247, Khari, Local Reed, Local Striped, Mauritius 16, M 55, Purple Mauritius, Poovan, Q 813, Red Mauritius and Vellai (McRae, 1926b; 1928; Sundararaman, 1928a; 1929). *Saccharum spontaneum* varieties SES 320, Taiwan II, Thick gigas, and Tobange were found to contain sugarcane mosaic virus, though the symptoms were not very clear externally.

Among the resistant varieties, Co 214 is reported to be the only variety which has escaped infection in nature and also artificially (Chona, 1944). The POJ varieties are fairly resistant to the disease, and among the thin canes, Co 205, Co 214, Co 244 and Kassoer are resistant (Sundararaman, 1928b). POJ varieties 2714 and 2727 and a few Fiji varieties are also fairly resistant to the disease. Co 205, a highly susceptible variety at Pusa, is found to be resistant at Coimbatore (Sundararaman, 1931).

Many commercial varieties of sugarcane exhibit varied degree of susceptibility to the disease. Certain varieties which escape infection under one set of conditions succumb to the disease under a different set of conditions (Krishnaswami, 1936). The nature of epidermal structures on the leaf of sugarcane has been shown to determine the degree of resistance of the variety. Thus, the resistant Kassoer variety is shown to possess sharp, stiff unicellular bristles on the surface of the leaf which protect the stomata and the surrounding regions from insect invasion. On the other hand, a number of susceptible varieties like Poovan, Co 1, Red Mauritius, Purple Mauritius and Vellai are extremely soft leaved and lack the above said arrangement of the Kassoer variety, (Venkataraman and Thomas, 1936).

The varieties having the "*Saccharum spontaneum*" blood have been shown to be generally resistant to the disease, and some varieties are atleast tolerant. A few immune varieties exist, and they could be used as parents for the production of resistant hybrid canes (Dutt *et al.*, 1936; Krishnaswami, 1936).

Thomas (1931) applied a practical method of estimating the disease resistance of sugarcane to mosaic disease. This consisted of growing large number of varieties and strains under identical conditions of environment and exposure to identical conditions of infection to mosaic. This is followed by a close study of an equal number of individuals in each variety, a careful estimation of the incidence of infection and the evaluation of the extent of the disease and the loss due to it among the different varieties under trial.

Effect of the disease on host :—

Though sugarcane mosaic is said to be very destructive in other cane growing countries, in India, the disease is of little economic importance (Chona, 1947). Careful experiments at Pusa reveal, that at 100% infection, there is only 10 to 12% loss compared to 5 to 50% in other countries. The quality of the juice from infected canes is not much affected by the disease. Co 213, suffering from mosaic showed 11% less germination, 14·8% less in stripped cane, and 8·9% less in sucrose than healthy canes (McRae, 1932). The disease reduced the weight of the cane, but the quality of the juice, as measured by brix, glucose and sucrose remained unimpaired (McRae, 1932; McRae and Subramaniam 1933, 1934; Sarkar and Dutt, 1934; Sundararaman, 1933).

Physical properties and the strains of the Virus :—

The Indian strain of Sugarcane Mosaic Virus resembles strain I B of Summers, and is known to exist in three different strains designated X, Y, and Z (Chona and Rafay, 1950). The thermal death point of strain X in expressed juice is 60°C, that of Y 55°C and that of Z 45°C. X and Y are able to withstand a dilution of 1:100 and can remain viable for 6 hours at 30 to 32°C, whereas Z cannot remain infective at a dilution of 1: 50 and hardly survives for 2 hours at 30 to 32°C. These strains are not filterable through Chamberland L or Seitz filters and even the filterate through an ordinary filter paper is non-infectious. The virus

can be centrifuged for half an hour at 3000 r.p.m. without sustaining appreciable loss of virulence, though further centrifuging reduces its infectivity. *In vitro,* the virus loses viability in a few hours at room temperature, but at 5 to 6°C, it remains infective for about 15 days. Storage temperature is the most important factor in the *in vitro* longevity of the virus (Chona and Rafay, 1950).

Copper sulphate 1:1,500, hydrochloric acid 1:1,000, nitric acid 1:800, hydrogen sulphide 1:25 and formalin 1:25 inactivated the virus. Zinc powder has no effect on the virulence (Rafay, 1935).

The different strains readily cross-infected other hosts like *Sorghum,* maize and *Euchlaena mexicana.* The same strain produced widely different symptoms on different cane varieties.

Acquired Immunity :—

Immunity from the disease, after infection is reported for a few varieties. These varieties recover from the disease and the possibility of the plant acquiring a permanent immunity is postulated (Krishnaswamy, 1936; Sundararaman, 1936).

Control Measures :—

Selection of good, disease-free setts from tolerant varieties, regular and intensive roguing of clumps that may show mosaic and avoiding growing of jowar or maize near sugarcane plots, are some of the control measures recommended (Chona, 1947). Regular monthly inspections of the crop should be made especially between the second and sixth months of growth, and all suspected plants eradicated. When the incidence is very high, making roguing impossible, the crop should be allowed to mature, and then milled for the juice, care being taken not to use the setts of the diseased cane. Roguing has given excellent results at Coimbatore especially in reducing secondary infection (Sundararaman, 1928a).

Other Plants affected by Sugarcane Mosaic Virus :—

Maize, *Sorghum, Euchlaena mexicana* and *Eleusine coracana* have been shown to be susceptible to sugarcane mosaic virus (McRae and Subramaniam, 1928; Mitra, 1936; Venkatarayan, 1946). Often, these plants when grown near sugarcane plots, show mottl-

ing of the leaves, and this is also experimentally shown by inoculation of the infective sap from mosaic infected sugarcane into the leaves of these plants. The reverse infection from these plants back into healthy sugarcane is, however, seldom achieved (Mitra 1936).

2. *Mosaic of Potato* (*Solanum tuberosum* L.)

The Viruses:

1. *Solanum* Virus 1, Smith (Potato Virus X) *Marmor dubium* var. *annulus* Holmes.

2. *Solanum* Virus 2, Smith (Potato Virus Y) *Marmor cucumeris* var. *upsilon* Holmes.

3. *Solanum* Virus 3, Smith (Potato Virus A) *Marmor Solani* Holmes.

Potato is commonly affected with several types of mosaic, caused by one or more strains of the Potato Viruses. Mosaic of potato was first noted in North India in 1936 (Galloway, 1936), along with other types of virus diseases like leaf-roll. McRae (1923) reported its presence in Bombay Presidency. Annual inspection for seed certification showed, that in Kumaon hills, Chakratta hills and Simla hills, the percentage of infection with negligible and severe mosaics were respectively, 55 to 70 and 24·5 to 46, 24 to 37 and 2·5 to 23, 8 to 20 and upto 40 (Anonymous, 1948). Most of the Indian varieties and many foreign varieties suffer from various types of mottling which have been classified into, 'negligible mottle', 'mild mosaic', 'borderline mild mosaic', 'severe mosaic', and 'borderline severe mosaic' (Vasudeva, 1946). The incidence varied in different months, and masking of symptoms has been found to be very common. In North India the masking of symptoms is negligible in the first crop, whereas in the second crop, masking occurs during the second and third weeks of March, when the mosaic incidence is lowest (Vasudeva, 1946).

Potato Virus X:—

A relatively virulent strain of Potato Virus X (*Solanum* Virus 1) has been studied on Majestic potatoes in United Provinces (Vasudeva and Lal, 1945a). The infectivity of the virus is lost at a dilution of 1:100,000 and 10 minutes' exposure to 75°C. The virus is still active after passage through Chamberland filters

of grades L3 to L5 and after 202 days' storage, though its virulence is slightly affected after 234 days' storage (Vasudeva and Lal, 1945a).

A severe strain of ringspot virus has been recorded on potato plants of Gola variety (Sharma and Raychaudhuri, 1956). The virus is inactivated by extracts from chilli, spinach, and strawberry, the chilli extract being more inhibitory than the others. Growth products of *Aspergillus niger* grown in three different media and thiouracil at a concentration of 500 mg./litre inhibited the virus multiplication, and this aspect has been made use of in curing the mosaic of Aran Victory potato, by treating the plants with thiouracil at a concentration of 125 mg./litre for 7 days (25 c.c. per day per plant). Ultra violet light inhibited the virus, the rate of inhibition being dependent on the time of exposure. It is also shown that the UV light affects only the virus particles and not the host metabolism (Sharma and Raychaudhuri, 1956).

Potato Virus Y :—

Phulwa variety of potato, is known to suffer from mosaic by potato virus Y in North India (Pal, 1941, 1943) and the causative virus was first identified by Dr. R. N. Salaman. Observations at Delhi show that the disease can be controlled by roguing (Pal, 1941).

The disease caused by this strain is very serious and is characterized by reduction in size of leaves, stunting and dwarfing of the shoots, puckering of the leaves and production of necrotic spots on both sides of leaves. In the advanced cases, the lower leaves become brittle and yellow, and ultimately they fall off. The apical leaves remain highly mottled and curved ventrally. If the infection takes place when the plant is very young, no tubers are formed. In Craig's Defiance variety of potato the loss in yield is estimated as 77·5 to 87·5% (Khanna and Ganguly, 1953).

Vasudeva and Lal (1945b) isolated the pure strain of potato virus Y *Solanum* Virus 2) from Phulwa plants showing negligible mosaic and veinal necroses or severe mosaic only. The pure strain is transmissible to tobacco varieties Harrison's Special and White Burley, *Nicotiana glutinosa, N. sylvestris, N. rustica, Datura stramonium, Solanum nodiflorum, Petunia hybrida* and President potato variety. Potato Virus Y can be isolated from

a mixture containing viruses X, Y and A, by passage through *Petunia hybrida*.

The virus loses its activity after exposure to 50°C and at 54°C, it becomes innocuous. It is inactivated by storage of the standard extract for 24 hours at room temperature, and at a dilution of 1:1,000. The virus is held back when passed through Chamberland filters L to L5 (Vasudeva and Lal, 1945b).

Potato Virus A :—

Phulwa and Dharjeeling Red Round potato varieties were found to be infected with Potato Virus A (*Solanum* Virus 3), when the former exhibited yellowing of leaf margins or a brilliant deep yellow mottle, and the latter extreme leaf crinkling (Vasudeva and Ramamoorthy, 1946). The mild mosaic caused by this virus, does not cause as high yield reductions as Potato virus Y (52% due to Potato virus A, compared to 72 to 92% due to Virus Y). Nevertheless, it is one of the most widely prevalent virus diseases of potatoes in India (Vasudeva and Azad, 1952).

Strain of Potato Virus X :—

Solanum virus 6, a strain of Potato virus X, producing necrotic spots on the leaves of Phulwa potato plants has been identified for the first time in India (Vasudeva and Gupta, 1946). The spots are of different sizes, and are found scattered both on young and old leaves, these being more on the lower leaflets. In extreme cases, the leaf-margins or even the entire leaflets may dry up. Transmission is achieved by sap inoculation using carborundum and also by grafting. The strain can infect tobacco variety White Burley, *Capsicum annuum, Datura stramonium,* potato variety President and Epicure.

Control Measures :—

Selection of good seed free from infection, timely sowing, and periodical roguing, often prove very useful in reducing the intensity of infection (Khanna and Ganguly, 1953; Pal, 1941).

3. *Mosaic of Sann hemp (Crotalaria juncea L.)*

Raychaudhuri (1947a) reported a mosaic disease of Sann hemp, *Crotalaria juncea* L. from Delhi.

Symptoms :—

Foliar mottling, stunting and in severe cases malformation, are the main symptoms of the disease. The dark green patches on the leaves alternate with light green ones and the upper side is often raised with corresponding depressions on the lower side. Anatomically, the mesophyll tissues are thinner and the intercellular spaces fewer than the healthy leaves. In advanced stages, the differentation into mesophyll and spongy tissues is lacking in the infected leaves. The cells are roughly isodiametric in transection and their chloroplast somewhat indistinct. The vascular tissues of the mottled leaf show hypertrophy of a few phloem cells.

Physical properties of the Virus :—

The virus has been purified using the method described by Bawden (1950) and a solution of very fine acicular crystals obtained.

The thermal death point of the virus lies between 68° to 70°C, has a longevity *in vitro* of 71 to 76 days and withstands a dilution of 1:1,000 to 1:5,000 (Raychaudhuri, 1947a, 1948). Das Gupta *et al.* (1951) showed that crystalline preparations under electron microscope revealed spherical nature of the particles composing the virus, with a diameter ranging from 26 to 40 mμ, while De (1951) reported a mean diameter of 37·4±6mμ. Ultra violet absorption spectra of Sann hemp mosaic virus and Cowpea mosaic virus reveal, that the two viruses are distinct strains with differences in the tyrosine-tryptophane components of the viruses and hence the inability of the two viruses to cross infect the hosts (Raychaudhuri and Ramamoorthy, 1949). The virus operates over a pH range of 1·55 to 9·87, and at a dilution of 1:10, it is inactivated by treatment with 0·5% copper sulphate (Anonymous, 1952a).

Transmission and Host Range :—

Sann hemp mosaic virus can easily be transmitted to healthy plants by sap-inoculation and the symptoms appear within 6 to 8 days. When the young leaves are punctured with insect needles dipped in inoculum, the symptoms appear within 3 to 4 days (Raychaudhuri, 1947a). No evidence of seed transmission is available. The virus does not infect cowpea (Raychaudhuri and

Ramamoorthy, 1949) but can infect *Crotalaria striata* (Anonymous, 1947).

Southern Sann hemp mosaic :—

Capoor (1950) described a mosaic disease affecting *Crotalaria juncea* from Bombay, though the presence of the disease was noted as early as 1940 at Poona. The disease resembles the Delhi type described by Raychaudhuri (1947a) in general symptoms but differ greatly in host range and physical properties of the virus. It is, therefore, given a new name to distinguish it from the Delhi type, "Southern Sann hemp mosaic" (Capoor, 1950).

Symptoms :—

Like the Delhi type, the Southern sann hemp mosaic virus causes foliar stunting, slight malformation and puckering in the form of raised, blister-like green areas alternating with pale green ones. Occasionally however, thin elongated enations are seen on the ventral side of the leaves. The affected plants produce little seeds and are dwarfed (Capoor, 1950).

Transmission and Host Range :—

The virus can readily be transmitted by sap-inoculation and transmission by seed is not known. Besides *Crotalaria juncea*, the virus infects six others species of *Crotalaria, Lathyrus sativus, Phaseolus vulgaris, P. lunatus, Vigna unguiculata* and *V. sesquipedalis*. Local lesions develop on *Nicotiana glutinosa, N. sylvestris, Datura stramonium* and *Capsicum frutescens*. In tobacco, the virus is carried symptomlessly at low temperature, producing faint mottling at higher ones; tomato, *Solanum nigrum* and *S. nodiflorum* carry the virus symptomlessly (Capoor, 1950; Anonymous, 1954a).

Physical properties of the virus :—

The Southern sann hemp mosaic virus withstands a dilution of $1 : 10^{-5}$, survives after 10 minutes' exposure to 90°C, and the crude sap withstands 557 days' storage. The infectivity is not lost by treatment with 90% ethanol for 24 hours.

Mosaic of Crotalaria mucronata :—

Raychaudhuri and Pathanian (1950) reported a mosaic disease of *Crotalaria mucronata* from Delhi.

Symptoms :—

The symptoms first appear on young leaves in the form of pale areas. An occasional downward curling of the leaf-tips appear later, and typical mosaic mottling develops on the lamina, with a few pale interveinal areas on the older leaves. This disease, unlike the one reported on *Crotalaria juncea* (Raychaudhuri, 1947; Capoor, 1950), is rarely followed by foliar stunting and blistering.

Transmission and Host Range :—

The disease, like that of *Crotalaria juncea,* is sap-transmissible, and the virus infects besides *Crotalaria mucronata,* C. *juncea,* cowpea, *Phaseolus mungo* and *P. aureus.* Symptoms appear within 7 to 12 days after inoculation.

Physical properties of the Virus :—

The thermal death point of the virus lies between 80° and 85°C, and has a longevity *in virto* of 134 to 142 days at 12° to 15°C. It can withstand a dilution of 1:20,000 to 1:30,000 in crude sap.

4. *Mosaic of Cowpea* :—

(a) Mosaic of *Vigna catjang* Walp.

In 1942, cowpea plants (*Vigna catjang* Walp.) growing in the Agricultural College Farm, Poona, showed malformed leaves with dark green areas interspersed with light green ones. The chlorosis became severe at a later stage. The affected plants bore only a few small pods, with broad, light and dark green stripes on the surface, and contained only a few small seeds (Capoor, *et al.,* 1947).

Vasudeva (1942) reported a similar disease affecting Punjab 1 cowpeas, when the plants were inter-cropped with Desi cotton at Lyallpur. The affected plants showed three types of symptoms : — (i) Stunting especially of upper parts with thick foliage, showing vein-clearing, mottling and alternating areas of dark and light green colour, and sometimes raised blister-like undulations, (ii) Uniform foliar chlorosis without distortion, and (iii) Very conspecuous mottling with pale to vivid yellow areas, later turning brown to reddish brown spots.

Anatomically, the diseased leaf showed fusion of the epidermis with the cuticle at various places, paucity of palisade, containing few chloroplasts, discontinuous and irregular tissues, abnormal spongy parenchyma, thickening and enlarging of xylem vessels and sclerenchyma, and disarrangement of vascular bundles. In some cases, the spongy parenchyma was replaced by a large number of elongated cells (Vasudeva, 1942).

Transmission and Host Range :—

The virus is sap-transmissible and better infection is obtained by using 600 mesh carborundum. *Aphis medicaginis* caused 36% infection under glass-house conditions, and about 4% of the commercial cowpea seeds produced infected seedlings. The virus is found to infect besides cowpea, *Phaseolus lunatus* and *Canavelia ensiformis* (Capoor *et al.*, 1947).

Physical properties of the Virus :—

The virus withstands 10 minutes' exposure to 85°C, but at 90°C, it is inactivated. The dilution end point in the crude sap is between 1:50,000 to 1:10,000. Infectivity is lost after 15 days' storage at 24°C, but not after 9 days.

Affinities :—

The virus is considered to be a distinct one from cowpea and Asparagus bean mosaic viruses on account of its distinct physical properties.

(b) Mosaic of *Vigna cylindrica* Skeels :—

The virus: *Marmor vignae* var. *catjang*—Capoor and Varma, 1956.

Chavali cowpea (*Vigna cylindrica* Skeels) is reported to suffer from a mosaic disease. Symptoms of mosaic appear 7 to 19 days after carborundum-sap inoculation. Seeds from infected chavali cowpea carried 17·3 to 22·7% infection. *Aphis medicaginis* caused 25% infection while *Myzus persicae* and *Aphis gossypii* produced 12 and 4 percent infection respectively. The host range included *Crotalaria juncea* and soyabean. *Cyamopsis tetragonoloba* developed numerous necrotic lesions (Capoor and Varma, 1956).

Physical properties of the virus :—

The infectivity of the virus is retained for 19 days at 24°C in sap, but is lost from dried leaves in 6 days. The dilution end point of the virus is found to be between 1:50,000 and 1:1,00,000.

Affinities of the virus :—

The virus is distinguished from that described by Dale, and from Cucumber mosaic virus, by its inability to infect *Dolichos lablab* and cucurbits. It is not possible to decide its relation with the mosaic of cowpea described by Vasudeva (1942), and so a new name is proposed, *"Marmor vignae* var. *catjang"*—a variety of McLean's Cowpea mosaic virus (Capoor and Varma, 1956).

5. *Enation Mosaic of Dolichos lablab* L. :—

A disease of *Dolichos lablab* characterised by mosaic and chlorotic streaks on the leaves is reported from Poona. The size of the leaf blade is markedly reduced by the suppression of growth in the interveinal areas and enations are seen on the lower side (Capoor and Varma, 1948a).

Transmission and Host Range :—

Transmission of the disease is achieved by sap inoculation, and the virus is found to infect a large number of other leguminous plants including beans, *Phaseolus vulgaris* and White Burley tobacco. *Phaseolus vulgaris* reacts with a lethal systemic necrosis and White Burley tobacco shows primary lesions in the form of thin, white rings of necrotic tissue enclosing green areas, and a secondary systemic mottling, which develops 65 days after inoculation (Capoor and Varma, 1948a).

Physical properties of the Virus :—

The dilution end point of the crude sap lies between 5×10^{-6} and 3×10^{-6}. It can resist 10 minutes' heating at 90°C, but is inactivated at 95°C (Capoor and Varma, 1948a).

Affinities of the Virus :—

Though resembling the tobacco mosaic virus in its physical properties, the Enation mosaic virus of *Dolichos lablab* differs from it appreciably in its host range and by the type of symptom it

produces on tobacco. Cross-immunity tests with the two viruses reveal, that neither confers reciprocal protection, showing that the two are not related to each other (Capoor, 1954; Capoor and Varma, 1948a). The virus resembles Bean viruses 4 and 4A in thermal inactivation, and Pea Streak virus in dilution end point. Hence the virus is named "Dolichos enation Mosaic virus" (Capoor and Varma, 1948a).

6. *Yellow Mosaic of Dolichos lablab* L:—

During 1950 at Poona, and since then in many places like Deccan, Gujerat and Khandesh, a disease was observed on *Dolichos lablab,* characterised by broad yellow patches on the leaves (Capoor and Varma, 1950a).

Transmission :—

The disease is transmitted by the white-fly *Bemisia tabaci* Gen., and the virus has an incubation period of 14 to 20 days inside the vector. Juice inoculation and seed transmission gave negative results.

Affinities of the Virus :—

Phaseolus lunatus suffers from a similar disease, transmitted by the same vector, but several tests revealed the inability of the flies to transmit the *Dolichos lablab* virus to *Phaseolus lunatus,* though the yellow mosaic of *Phaseolus lunatus* was innocuous to *Dolichos lablab.* In its transmission, solely by means of the insect, it differs from the Enation mosaic of *Dolichos lablab* and hence a new name is proposed for the virus, "Yellow mosaic of *Dolichos lablab*" (Capoor and Varma, 1950a).

7. *Yellow Mosaic of Double-bean* (*Phaseolus lunatus* L.)

In 1939, a mosaic disease was observed affecting the *Phaseolus lunatus,* in Poona. The disease is reported to be prevalent also in Delhi.

Symptoms :—

The disease is characterized by a mosaic pattern, which starts as faintly discoloured patches gradually becoming yellow, 20 days after infection (Capoor and Varma, 1948c).

Transmission and Host Range :—

Sap inoculation and seed transmission giving negative results, the disease is successfully transmitted by the white-fly *Bemisia tabaci* and by bud-grafting.

In addition to *Phaseolus lunatus,* the virus also infects *Phaseolus limensis, P. vulgaris, P. aureus, Dolichos biflorus* and *Canavelia ensiformis.*

Affinities of the Virus :—

The disease is shown to be similar to that described by Uppal as "Mosaic of *Dolichos biflorus*" and is named, "Double bean Yellow mosaic" (Capoor and Varma, 1948c).

8. *Mosaic of Phaseolus mungo* L. :—

A disease causing mosaic of *Phaseolus mungo* has been reported from New Delhi, (Anonymous, 1952a). The disease is sap transmissible and the virus loses its infectivity after 10 minutes' exposure to 60°C, and is inactivated by treatment with 50% alcohol, 0·5% copper sulphate, and 5% carbolic acid. It remains infective for 26 hours at 15° to 37°C (Anonymous, 1952a).

9. *Sterility Mosaic of Pigeon Pea* (*Cajanus cajan* L).

During 1946, 7% of pigeon pea (*Cajanus cajan* L.) in Bombay, and in 1947, 20% in Gujerat, were found to be affected by a disease named, "Pigeon pea sterility mosaic" (Capoor, 1952).

Symptoms :—

The affected plants are characterized by a general pallor, profuse upright vegetative growth, and an absence of flowering branches. The leaves of the affected plant are small, pale and often with a mosaic pattern.

Transmission :—

Out of 75 grafts with healthy plants, the disease appeared in 34 plants, the time taken for the appearance of the disease being 35 to 40 days. Out of 22 plants inoculated with the infective sap, only one showed symptoms after 81 days (Capoor, 1952).

10. *Yellow Vein Mosaic of Bhendi* (*Hibiscus esculentus* L; *Abelomoschus esculentus* Moench).

The Virus: *Hibiscus* Virus 1.

Ochrovena hibiscae n. gen. n. sp. (Capoor and Varma, 1950b).

Kulkarni (1924) reported the presence of a mosaic disease of bhendi (*Hibiscus esculentus* L) in Bombay, causing serious havoc to the cultivators. Subsequently it was found that the disease was widely prevalent in Bombay and Poona (Varma *et al.*, 1950; Uppal *et al.*, 1940).

Symptoms:—

The first visible symptom of the disease is the clearing of the small veins and then of the larger ones, the ill-defined yellowish green to pale green areas latter extending into the mesophyll. In severe infection, the young leaves develop a general chlorosis, rather than actual mosaic patterns. Growth after infection is stunted, the leaves are undersized, petioles short, the flowering sparse and the fruits few (Uppal *et al.*, 1940). It is shown that the appearance of the symptom could be delayed if the soil is depleted of nitrogen or phosphorous (Anonymous, 1942).

Transmission and Host Range:—

"Yellow vein mosaic of bhendi" is shown to be transmissible only by grafting and by the white-fly *Bemisia tabaci*, and is one of the diseases in which the vector-virus relationship has been studied in some great detail (Varma, 1952). It is shown, that to produce 100% infection at least 10 flies are necessary. A preliminary 4-hour fasting period, followed by a feed of one hour on diseased plants, makes the insect acquire the virus and improves their efficiency as vectors. A 30 minutes' feed on healthy plants is adequate for transmission. Post-acquisition fasting up to 4 hours is also effective, and the flies can transmit the disease after 15 minutes' feed on healthy plants. Flies fed on diseased plants for 12 to 24 hours remain infective for life. They can acquire the virus from the youngest leaf, a few days before the disease symptoms appear. The minimum incubation period of the virus inside the fly is 7 hours, but very few insects become infective at the end of this period.

The female flies are shown to be slightly more efficient as vectors than the males. The virus is shown not to pass to the eggs from viruliferous females. It is shown that the flies can carry the "Yellow vein mosaic of bhendi" and that of Pumpkin simultaneously. (Capoor and Varma, 1950b; Varma, 1955a, b).

"Yellow vein mosaic of Bhendi" is transmitted to a number of Malvaceae, like *Hibiscus tetraphyllus, H. manihot, H. cannabinus,* and *Abelomoschus moschatus* by grafting and by the white-fly, while *Hibiscus sabderifa* showed infection only by grafting (Anonymous, 1954d). Partial recovery from the disease is observed in *Hibiscus cannabinus* and *H. sabderifa* (Capoor and Varma, 1950b).

Changes in the physiology of the infected host :—

Govindjee *et al.* (1956) showed that the infected plants contained more leucine, phenylalanine, valine and methionine, asparaginic acid and asparagin than the healthy ones, and in addition, a nin-hydrin-positive substance.

Control Measures :—

Destroying all plants during April and May, growing none in June-July, and roguing the newly raised plants from August is a schedule for cultivation, reported to have given excellent results. Roguing should be done periodically at weekly intervals. Eradication of *Hibiscus tetraphyllus,* a common weed, often prove very useful in reducing the disease incidence. Spraying the crop with fish-oil-rosin soap to control the white-flies once every three weeks, and keeping the field clear of weeds are some more recommendations (Capoor and Varma, 1950b; Varma *et al.,* 1950).

11. *Mosaic of Soya bean* (*Soja max* or *Glycine soja* Sieb and Zucc).

Chattopadhyay and Das (1954) in their review of common virus diseases of crop plants in West Bengal, includes Soya bean as suffering from a mosaic disease.

12. *Mosaic of Cucurbits* :—

(a) *Mosaic of Bottlegourd* (*Lagenaria vulgaris* Ser).

The virus : *Cucumis* virus 2B, Smith

Marmor astrictum var. *lagenari* Holmes (Capoor and Varma, 1948e).

In 1939, a mosaic disease affecting *Lagenaria vulgaris* Ser. was first observed in the Agricultural College Farm, Poona and subsequently in Bombay Province. The disease is sap-transmissible and after 10 to 14 days of inoculation, pale and dark green patches appear on the leaves, which gradually become diffused. The affected leaves are much stunted, veins are thinner, dwarfed and weak. The fruits are small, very light, with some patchy discolouration (Capoor and Varma, 1948e).

Physical properties of the Virus :—

The virus withstands 10 minutes' heating at 95°C, but not at 98°C, and it retains its infectivity at a dilution of 1:1,000,000 and 308 days' storage in the laboratory. It resists 24 hours' treatment with 95% ethanol, and remains viable after dessication (Capoor and Varma, 1948e).

Host Range :—

The virus affects, besides *Lagenaria vulgaris,* cucumber, melon, watermelon, *Citrullus vulgaris* var. *fistullosus, Luffa acutangula* and *Trichsanthes anguina;* Vegetable marrow and *Momardica charantia* are symptomless carriers (Capoor and Varma, 1948e).

The virus is newly described from India, and is considered to be a strain of *Cucumis* Virus 2 (Cucumber Green Mottle mosaic and is named as "*Cucumis* Virus 2B" (Smiths' classification) and *Marmor astrictum* var. *lagenari* (Holmes's classification) by Capoor and Varma (1948e).

(b) *Mosaic of* Lagenaria leucantha *Rus.*

The virus :

Cucumis virus 2C, Smith.

Marmor astrictum var. *subobscurum Holmes.*

n. var. (Vasudeva *et al.*, 1950)

During 1947-'48, a new strain of *Cucumis* virus 2 (Cucumber Green Mottle mosaic virus) was reported from Delhi, on *Lagenaria leucantha,* by Vasudeva *et al.* (1950).

Symptoms : —

The first visible symptom of the disease is irregular light and dark green mottling, developing into mosaic. Vein-banding fol-

lows soon, and the young leaves show dark green blisters, on a crinkled pale green surface. The plants become stunted, and the flower and fruit production is reduced.

Transmission and Host Range : —

The virus is sap-transmissible and the symptoms appear 12 to 18 days (in winter) and 6 to 8 days (in summer) after inoculation. It produces visible symptoms on *Cucrbita moschata* and cucumber. On *C. moschata,* minute green spots appear 15 to 18 days after inoculation, followed by small, yellow circular spots along the veins. On cucumber, small yellow circular spots on young leaves are seen 7 to 10 days after inoculation, these spots later enlarging until the whole leaf appears pale. The virus is carried symptomlessly in *Datura stramonium, Luffa acutangula,* watermelon and *Momardica charantia* (Vasudeva *et al.,* 1950). Tobacco variety White Burley, *Solanum nodiflorum, S. nigrum* also suffer no visible infection, but yet with the virus in them, while the strain does not affect tomato or chilli (Vasudeva and Nariani, 1952).

Physical properties of the Virus :—

The virus has a thermal death point of 86° to 88°C, and dilution end point of 1: 1,000 in standard extract and 1: 10,000 in pure extract. It has a longevity *in vitro* for more than 90 days (Vasudeva *et al.,* 1950).

The inoculum is infective over a pH range of 1·88 to 9·96, though pH stability is influenced by temperature and storage. In standard extract the virus is resistant to 1-2 % nicotine sulphate, 40 % hydrogen sulphide, 0·5 % potasium permanganate, 0·5 % carbolic acid, 0·5 % silver nitrate, 0·1 0·25 % mercuric chloride, 50% alcohol, 50% glycerine and 50% acetone. It is inactivated by 24 hours' treatment with 50% hydrogen sulphide, 1% potassium permanganate, 1% carbolic acid, 1% silver nitrate, 0·5 % formalin and exposure to ultraviolet light for 2 hours, (Raychaudhuri *et al.,* 1951).

Leaf extracts of *Datura stramonium* and chilli, at a dilution of 1: 100, and those of *Solanum nigrum* and tomato at a dilution of 1: 1,000 inhibits the virus. The toxic principle of *Datura stramonium* is destroyed by heating the extract to boiling point and dilution with water, but not by dessication (Vasudeva and Nariani, 1952).

Affinities of the virus : —

The virus differs from "Cucumber Green Mottle mosaic virus" in its ability to cause visible symtoms on watermelon, and its ability to infect *Datura stramonium*. It also differs from Vasudeva and Lal's (1943) "Bottlegourd mosaic virus" in its greater longevity *in vitro*, higher thermal death point and tolerance to dilution. It also differs from "*Cucumis* virus 2B" of Capoor and Varma (1948e) in its higher thermal death point, dilution end point and host range. It is therefore given a new name, "*Cucumis* virus 2C" (Smith) or *Marmor astrictum* var. *subobscurum* (Holmes) n. var. (Vasudeva *et al.*, 1950).

(c) A similar disease, apparently caused by the same virus has been reported on *Lagenaria siceraria* Standl. (Vasudeva *et al.*, 1954).

(d) *Mosaic of bottlegourd* (*Lagenaria vulgaris* Ser.)

The virus :

(*Cucumis* virus 3, Smith. (Vasudeva *et al.*, 1943).

Symptoms : —

The plants affected with the disease, show chlorotic streaks, dark green blisters and wrinkling of the leaves with wavy margins. The older leaves shrivel and drop within 7 weeks; flowering and fruiting are sparse.

Transmission and Host Range : —

The virus is sap-transmissible, and infects besides *Lagenaria vulgaris*, cucumber, *Momardica charantia*, Melon, Watermelon and Vegetable marrow.

Physical properties of the virus : —

The virus is destroyed by 6 hours' storage at room temperature, and loses its infectivity at a dilution of 1: 500, and succumbs to 10 minutes' at 60°C. It does not traverse Chamberland filters L to L5, and its virulence is lost by passage through filter-papers. The virus is named, "*Cucumis* virus 3", (Vasudeva and Lal, 1943).

(e) *Mosaic of Melon* (*Cucumis melo* L.)

A suspected strain of "*Cucumis* virus 1", is reported to be capable of transmission through seed. When seeds of various

members of Cucurbitaceae were raised in sterilised soil, under insect proof conditions, one plant of *Cucumis melo* (melon) showed symptoms of virus infection, with puckering and interveinal mottling of the first and subsequent leaves. Further trials indicated that there was 56% seed transmission of the disease.

The disease has been found transmissible to *Cucumis sativus, Momardica charantia, Datura stramonium, Vigna sinensis,* and tobacco variety German Samson, by mechanical means. This is the first report of seed transmission of virus disease from India, (Vasudeva and Pavgi, 1943).

(f) *Yellow Vein Mosaic of Pumpkin* (*Cucurbita pepo* L.)

In 1951, a disease affecting pumpkin was observed in Poona (Anonymous, 1954c) and is named, "Yellow vein mosaic" of pumpkin.

Symptoms : —

The disease is characterized by pronounced vein-clearing, and irregular chlorotic patches on the lamina. Internodes turn yellow, and the under-sized fruits show yellow patches on the exterior (Varma, 1955b).

Transmission and Host Range : —

The virus is transmitted by the white-fly *Bemisia tabaci,* (Varma, 1955b) the vector needing 15 minutes to acquire the virus. The incubation period of the virus within the vector is 9 hours (Anonymous, 1954d). The insect is reported to be able to carry the viruses causing the Yellow vein mosaic of bhendi and pumpkin, simultaneously (Varma, 1955b).

The virus is transmissible to cucumber and Vegetable marrow, but does not affect any member of Malvaceae.

(g) *Mosaic of Cucumber* (*Cucumis sativus* L.)

The virus extract of Cucumber mosaic (strain not mentioned) is reported to lose its virulence when passed through sand and pulp filter or fuller's earth, (Uppal, 1934). Adsorption of the virus readily occur when Kaolin or fuller's earth, even in minute quantities is added to the virus extract. The virus is active between pH ranges 5 to 9, and inactive below 5. Adsorbed to fuller's earth,

the virus is freed *in vitro,* in an active state, by altering the pH from 6 to 6·7. It is concluded that the adsorption and inactivation of Cucumber mosaic virus are reversible phenomena (Uppal, 1934).

13. *Mosaic of Cardamom* (*Elettaria cardamomum* Maton)

Synonyms of the disease: "Marble" disease,
"Katte" disease.

Cardamom (*Elettaria cardamomum* Maton) is reported to suffer from a mosaic disease (otherwise called "Marble" or "Katte" disease) in Bombay (McRae, 1923), Delhi (Anonymous, 1955b), Kanara (Kulkarni, 1924), Madras (Anstead, 1924), Mysore and Travancore (Uppal *et al.,* 1945).

Symptoms : —

The first visible symptom appears as stripes of deep green tissue over the surface of the leaf, these running from the mid rib to the leaf margin. Diseased plants appear pale from a distance. When the disease is fully developed, the stripes become evenly distributed. Under natural conditions, the plants are susceptible at all stages of growth, though the disease is seldom seen in nursery beds and appears only after transplanting. The affected plants quickly deteriorate; newly formed clumps from infected shoots are small, and show gradual reduction in longevity, ultimately followed by death. Many clumps wither before beginning to yield (Uppal *et al.,* 1945; Varma and Capoor, 1953).

Transmission : —

That the disease is not soil-borne has been proved (Uppal *et al.,* 1945). Sap inoculation gave negative results (Kulkarni, 1924; Uppal *et al.,* 1945; Varma, 1956). The disease is transmitted by the banana aphid, *Pentalonia nigronervosa* Coq. (Varma and Capoor, 1953). The symptoms of the disease appear on healthy plants 21 to 46 days after feeding by the vector (Uppal *et al.,* 1945).

Although banana is the most important host for the aphid, it has been found to breed on cardamom during the winter months, and sometimes on members of Aroideae (Renga Ayyar, 1954; Varma and Capoor, 1953). In areas, where cardamom is propagated vegetatively, infected rhizomes transmit the disease. The virus is found not to affect banana (Varma and Capoor, 1953).

It has been shown that a single feed by the vector makes it capable of transmitting the disease, and that the percentage of infection depends upon the number of insects transmitting the disease (Varma, 1956). Aphids pick up the virus within a feeding time of 15 minutes, and viruliferous aphids transmit the disease within 5 minutes. The maximum time for the insect to acquire the virus and transmit it, has been found to be one hour. The minimum incubation period for the virus in the aphid is 30 minutes, though the ability of the vector to transmit the disease is increased after one hour. The vector is found to loose all the virus it has acquired immediately after feeding either on a healthy banana or cardamom plant.

No congenital transmission has been reported. But, it is shown that nymphs could also act like the adult, as vectors and transmit the disease (Varma, 1956).

Control Measures : —

Total eradication of all old and diseased plants, constant and ruthless roguing at regular intervals, and transplanting healthy seedlings raised under disease-free conditions, are some of the control measures recommended (Galloway, 1935; Varma and Capoor, 1953).

Other plants affected by similar disease : —

(a) *Amomum subulatum* (Greater cardamom), is reported to suffer from, "Foorkey" disease, stated to be very destructive in Darjeeling area (Anonymous, 1955e). Diseased plants produce many stunted shoots, which fail to flower, and the leaves are slightly chlorotic. The disease is transmitted by the same vector as that of cardamom mosaic, namely *Pentalonia nigronervosa.* The disease is said to be quite distinct from the mosaic of cardamom, as evidenced by the distinct symptoms the two viruses produce on wild *Amomum* sp. (Anonymous, 1955b).

(b) A wild species of *Amomum* is reported to suffer from a disease identical with the mosaic disease of cardamom, transmitted by *Pentalonia nigronervosa,* and controllable by roguing (Anonymous, 1955b).

14. *Mosaic of Papaya* (*Carica papaya* L.).

A mosaic disease affecting papaya, (*Carica papaya* L.) is stated to have been prevalent in Bombay during 1940-'50, and in Poona during 1947 (Capoor and Varma, 1948d; Kamat, 1955).

Symptoms : —

Dot-like, solid necrotic spots all over the blades of the newly opened leaves are the first symptoms visible 20 days after inoculation. The foliage of a fully developed plant presents marked signs of mosaic, with blister-like patches of green tissue scattered at random over the surface of the pale green lamina, which is occasionally dwarfed and malformed. The petiole becomes very short. Elongated water soaked areas are seen on the stem and petiole. The fruits are few, abnormally small and are elongated, Numerous circular water soaked lesions, with a solid central spot covering their surface are seen on these fruits (Capoor and Varma, 1948d).

Transmission and Host Range : —

The disease is sap-transmissible. *Aphis gossypii*, collected from cotton and brinjal, *A. malvae*, from *Lagenaria vulgaris*, *Aphis* sp. from *Euphorbia prolifera* and *Myzus persicae* from *Brassica* sp., are also capable of transmitting the disease and it is supposed that one or more of these insects is responsible for the disease in the field. *Myzus persicae* is shown to be a more efficient vector than *Aphis gossypii* (Anonymous, 1954c). Seed transmission has not been reported.

Papaya mosaic virus is shown to infect a large number of papaya varieties like, Washington Special, Honey Dew, Hawaiian, Ranchi, Ceylon, Bombay, Poona Long and Poona Round (Capoor and Varma, 1948d). Among other plants susceptible to the virus are Cucumber, Vegetable marrow, Musk melon, Watermelon, *Luffa acutangula*, *Lagenaria vulgaris*, *Trichosanthes anguina*, *Impatiens balsamina*, sunflower and *Tropaeolum majus* (Anonymous 1954c; d).

Physical properties of the Virus : —

The virus is quite unstable and apparently of the non-persistent type. The infectivity of the virus is lost after 28 hours' storage *in vitro* but it withstands 10 minutes' of 53°C, not 55°C. It is found to be infectious at a dilution of 1: 1,000 (Capoor and Varma, 1948d).

Affinities of the Virus : —

Though capable of infecting a large number of Cucurbitaceous hosts, the virus is shown to be distinct from the mosaic viruses on

Lagenaria and *Cucurbita moschata* by cross-inoculation experiments (Anonymous, 1954c.).

15. *Mosaic of Egg plant (Brinjal) (Solanum melongena* L.)

Egg plant (*Solanum melongena* L.) is reported to suffer from a mosaic disease in Bombay (Kulkarni, 1924) and Delhi (Raychaudhuri, 1947b).

Symptoms : —

Diseased plants are characterised by bright green mottling of the leaves, followed by puckering and crinkling of the lamina. Occasionally, fine, pale, straw-coloured rings, which are either concentric or irregular, develop on the leaf. In some cases, the midrib or its branches extend beyond the lamina, while in others the entire leaf is changed into a filamentous structure. Flowers and fruits are greatly reduced (Raychaudhuri, 1947b).

Transmission and Host Range : —

The disease is transmissible by sap inoculation, by grafting, and by *Myzus persicae* (Anonymous, 1954c; Raychaudhuri, 1947b). *Empoasca devastans* is suspected to act like vector (Raychaudhuri, 1947b). The disease is transmissible to tobacco plants.

Physical properties of the Virus :—

The thermal death point of the virus is found to be between 60° and 70°C. It is found to be infective even after 15 days' storage at 35°F, but is inactivated at a dilution of 1: 1,000 (Anonymous, 1954c).

Affinities of the Virus :—

The virus is shown to be the same as that causing 'Little leaf' of *Vinca* and *Cucurbita moschata* mosaic (Anonymous, 1954b).

16. *Mosaic of Tobacco* (*Nicotiana tabacum* L.)

In spite of the fact that, tobacco mosaic is widely prevalent in all tobacco growing centres in India, information regarding the disease is meagre. McRae (1926b) reported the disease from Bombay, and Kulkarni (1924) found that it is one of the most destructive diseases of tobacco in Deccan, Karnatak and Gujerat.

Experiments to differentiate *Dolichos* enation mosaic virus and Tobacco mosaic virus, which agree in physical properties, by means of cross-immunization tests, reveal that the two are distinct viruses, the presence of one not inhibiting the entry of the other (Capoor, 1954). Chromatographic studies of the infected plant show the absence of a ninhydrin reacting substance, and the presence of 2 new bands corresponding to asparagine and histidine-lysine (Laloraya *et al.*, 1955).

17. *Mosaic of Tomato* (*Lycorpersicon esculentum* Mill.)

Incidence of tomato mosaic has been recorded from Bombay (McRae, 1923; Kulkarni, 1924), Pusa (Desai, 1933a) and Delhi (Vasudeva and Lal, 1947). According to Vasudeva and Lal (1947), no variety of tomato in India is resistant to the disease.

Symptoms : —

The diseased plants appear stunted, the young leaves are small, crinkled, deformed, pale yellow and brittle. Sometimes small necrotic areas are seen on the leaves, while in a few cases the leaf margin becomes brown. The diseased plants produce very few, small fruits (Desai, 1933a; Kulkarni, 1924). In Delhi, during 1941-'42, the disease incidence was 24·5 % (Vasudeva and Lal, 1947).

Desai (1933a) believed that the virus represented a filterable cyclostage in the life-history of a bacterium, which he found to be closely associated with the disease.

Control Measures : —

Spraying the plants with insecticides does not cause any reduction in the incidence, although it reduced the insects. Pruning and mixed cropping similarly does not reduce the infection. Roguing is recommended only when the incidence is low. Timely sowing is thought to favour less infection (Vasudeva and Lal, 1947).

A suspected strain of Tomato Aucuba mosaic virus : —

During 1949-'50, a mosaic disease affected tomato plants at the Indian Agricultural Research Institute, New Delhi, causing 2·7 to 4·5 % infection in experimental plots. It showed to be a strain of tobacco mosaic virus, differing from the typical virus, in that it withstood heating for 10 minutes' at 98°C, and produced local necro-

tic lesions, instead of stripes on leaves of eggplant, mottling and vein-banding on tobacco and interveinal mottling on *Solanum nigrum* (Das and Raychaudhuri, 1954).

18. *Mosaic of Chilli* (*Capsicum* sp.)

The occurrence of mosaic disease of chilli, is reported from Bombay (McRae, 1923; Kulkarni, 1924) and Delhi (Raychaudhuri and Jha, 1954).

Symptoms : —

Mosaic mottling and occasional malformation of the leaves are the main symptoms described for *Capsicum fruitescens* L. In some cases, the leaves get reduced to small tendrilar structures. Diseased plants often have diminutive leaves, and few flowers and fruits. The yield of affected plants is very low (Kulkarni, 1924; Raychaudhuri and Jha, 1954).

Transmission and Host Range : —

The disease is transmissible by sap inoculation, by *Aphis gossypii* and thrips. *Aphis gossypii* is thought to be responsible for the natural spread of infection. Transmission through seed does not take place.

The virus has been shown to produce mosaic symptoms on White Burley tobacco, *Nicotiana glutinosa, Nicotiana* selections 78 and 78 A, *Solanum nigrum, Petunia hybrida, Carthamus tinctorius, Cucumis melo* var. *utilissimus* and *C. sativus*. President and Craig's Defiance potato varieties, and *Datura stramonium* carry the virus symptomlessly.

Physical properties of the Virus :—

The virus is inactivated when exposed to 60°C for 10 minutes, and diluted to 1: 30,000; it loses its infectivity when stored for 22 days at 15° to 33°C and 23 days at 7° to 10°C. 50 % nicotine sulphate, 1% silver nitrate, 50% alcohol for 4 hours, 0·25% carbolic acid for 24 hours, and ultra violet radiation for 6 hours also inactivate the virus.

Effect of the virus on Meiosis and Seed fertility : —

Mosaic infection in chilli, causes meiotic disturbances, delayed embryo-sac development and few seeds per fruit. Only the

gametes free from virus-infected tissue are functional. The virus is shown not to penetrate the seed embryo, and the presence of some virus inhibiting substances is suspected to be present in the seed. These two reasons probably explain the non-seed transmission of the chilli mosaic virus (Swaminathan, 1956).

Affinities of the Virus :—

The virus causing mosaic in chilli, is considered to be a distinct one from other mosaic viruses (Raychaudhuri and Jha, 1954).

19. *Mosaic of Radish (Raphanus sativus* L.).

A mosaic disease of raidsh (*Raphanus sativus* var. *caudatus*) has been recorded from Bombay (McRae, 1923; Kulkarni, 1924), Poona (Kulkarni, 1924) and Delhi (Raychaudhuri and Pathanian, 1956).

Symptoms :—

The disease is characterized by mosaic mottling of the leaves, accompanied by necrosis and stunting of plants (Raychaudhuri and Pathanian, 1956).

Transmission and Host Range :—

The disease has been shown to be transmissible by sap inoculation and seed transmission does not take place. The host range of the virus is confined to members of Cruciferae, including oil producing species of *Brassica,* in which it produces severe necrosis and death.

Physical properties of the Virus :—

The virus withstands 85°C, for 10 minutes, dilution of 1:10,000,000 and 17 days' storage at 17° to 22°C, and 101 days' storage at 6° to 8°C (Raychaudhuri and Pathanian, 1956).

Affinities of the Virus :—

The virus is considered to be a distinct one from Radish mosaic virus and other viruses affecting cruciferae in its host range and physical properties (Raychaudhuri and Pathanian, 1956).

20. *Yellow Mosaic of Lettuce* (*Lactuca sativa* L.)

During 1946-47, lettuce plants at the Indian Agricultural Research Institute, New Delhi, showed 50 to 80% of Yellow mosaic infection (Vasudeva *et al.,* 1948).

Symptoms :—

The diseased plants show vein-clearing of young leaves, followed by yellow mosaic mottling. Later, pronounced laminar distortion and malformation of inflorescence axis are seen. The flowers are reduced and the market value of the infected leaves is appreciably low.

Transmission and Host Range :—

The virus is shown to be sap-transmissible and 30% seed transmission has been recorded. The virus infects, besides lettuce, Harrison's Special and White Burley tobaccos, and Sutton's Early Market tomato.

Physical properties of the Virus :—

The virus withstands 86°C for 10 minutes, a dilution of 1 : 60,000, and 60 days' storage at 15° to 25°C.

Affinities of the Virus :—

The virus is considered to be a distinct one from other viruses affecting lettuce (Vasudeva *et al.*, 1948).

21. *Distortion Mosaic of* Datura.

The virus :

Datura virus 3, (Capoor and Varma, 1948, n. sp.)
Marmor daturae,
Distortion mosaic virus.

A disease of *Datura alba* Nees. was first observed in Poona in 1939, and later, the disease has been recognised from several places in Bombay Province.

Symptoms :—

The disease is characterized by light and dark green mosaic with blisters of dark green patches on the leaf, and laminar distortion. Diseased plants are pale in appearance. The flowers are severely malformed but dwarfing of plants is rarely seen (Capoor and Varma, 1948b).

Transmission and Host Range :—

The disease is shown to be transmissible by sap, grafting and by insects. *Myzus persicae* Sulz. transmitted the disease from

Datura fastuosa to *D. fastuosa* and *D. alba*, under controlled conditions. The virus is shown to infect tobacco, *Petunia*, and potato producing mosaic symptoms. *Datura stramonium* and *Nicotiana glutinosa* show only local lesions, while it does not affect *Phaseolus vulgaris*, *Vigna sinensis* and *Solanum melongena* (Capoor and Varma, 1948b).

Physical properties of the Virus :—

The virus withstands heating for 10 minutes at 60°C, a dilution of 1 : 10,000 and 13 days' storage in the laboratory. 95% alcohol does not affect its infectivity when treated for 30 hours at 45°C.

The virus is shown to be present in small quantities inside the host plants. The purified virus solution is opalescent and highly birefringent, and shows anisotropy of flow, visible to the naked eye. The virus, iso-electric at pH 4·5, forms long, rod-shaped paracrystals at pH 5·5 and aggregates into small spindles on further concentration. True crystals are not formed, remaining in a liquid crystaline state. The purified preparation, free from pigmentation, gives protein reactions (Capoor, 1950b).

Other species of Datura *affected by Distortion Mosaic virus* :—

(a) *Datura fastuosa* L. is reported to be affected by "Distortion mosaic virus" of *Datura alba*, but the symptom is only a mild mosaic mottle (Capoor and Varma, 1948b).

(b) *Datura inoxia* Mill. shows severe puckering and distortion of leaves and flowers (Capoor and Varma, 1952). The disease is sap-transmissible, and the aphids *Myzus persicae* and *Aphis gossypii* are also capable of transmitting the disease. The host range appears to be confined to solanaceae, except for *Tetragonia expansa* in which the virus causes concentric rings. The virus is inactivated by 6 days' exposure to air at room temperature, and by dessication at room temperature. It withstands a temperature of 45°C for more than 100 days and also complete dessication at this temperature (Capoor and Varma, 1952).

22. *Mosaic of other plants* :—

(a) *Mosaic of Banana* (*Musa* sp.).

Banana is known to suffer from a mosaic disease, which is transmitted by sap as well as by the banana aphid, *Pentalonia*

nigronervosa Coq. (Kamat, 1955). The disease has been successfully transmitted to cucumber (Anonymous, 1954d).

(b) *Mosaic of Maize* (*Zea mays.*).

The mosaic of maize has been reported to be sap-transmissible and by *Aphis miadis, A. gossypii* and *Myzus persicae.* The disease has been transmitted to Sudan grass, *Euchlaena mexicana, Dactyloclenium aegyptiacum* and *Coix-lacryma-jobi;* to maize by *Aphis maidis* and *A. gossypii* and to maize and sorghum by *Myzus persicae* (Anonymous, 1954d, 1955b).

(c) *Mosaic of* Malvastrum coromandelianum *Garcke.*

Vein-clearing of young foliage, and the upward curling of leaves, are the symptoms seen in a mosaic disease of *Malvastrum coromandelianum* Garcke. The condition resembles the vein-clearing mosaic of *Hibiscus esculentus* and tobacco leaf-curl virus. It is suggested that *Malvastrum coromandelianum* might represent a secondary host for the 'Yellow vein mosaic' virus of *Hibiscus esculentus* (Venkatarayan, 1947).

(d) *Enation mosaic of* Hibiscus rosa-sinensis :—

Enation mosaic of *Hibiscus rosa- sinensis* is transmitted by the white-fly, *Bemisia tabaci* to *Hibiscus esculentus,* cotton and *Zinnia elegans* (Anonymous, 1952a).

(e) *Mosaic of* Nicotiana glauca *R. Grah.*

During 1951, a mosaic disease of *Nicotiana glauca* R. Grah. was observed in New Delhi causing 20% infection. The disease increased to 40% during 1952. Dark green mottling, reduction of leaf-size, wavy margins and slight malformation of the leaves are the main symptoms of the disease (Nariani and Singh, 1952).

Transmission and Host Range :—

The virus is readily transmitted by sap-inoculation and better results are obtained by using carborundum. First symptoms appear 15 days after inoculation. The virus has a wide host range of solanaceae and cucurbitaceae. In tomato, fern leaf symptoms are produced. In tobacco varieties Harrison's Special and White Burley, the lamina is reduced to tendrilar structures.

Physical properties of the Virus :—

The virus in standard extract remains infective after an exposure to 57°C for 10 minutes, but loses infectivity at 60°C. Crude juice withstands dilution up to 1 : 100, and at 1 : 200, it is rendered innocuous. In standard extract it remains infective for 7 days at 6° to 8°C and after 8 days at the same temperature, the virus becomes innocuous.

Affinities of the Virus :—

The virus is considered to be closely related to *Cucumis* virus 1. (Nariani and Singh, 1952).

(f) *Mosaic of Mustard* (Sarson) (*Brassica* sp.).

A severe mosaic of Indian mustard (Sarson) is reported from Simla (Anonymous, 1955b).

(g) *Momardica charantia* L. and

(h) *Helianthus annuus* L. have been reported to show mosaic symptoms in Bombay (Uppal, 1933).

STREAK DISEASES

The symptoms of 'streak' being variable, it is often used to indicate necrotic or light green areas or stripes, usually arranged in a definite manner, often parallel to the veins. The incidence of streak diseases in India is very low.

1. *Streak of Sugar cane* (*Saccharum officinarum* L.)

The virus : *Saccharum* virus 4, Smith.

Very few records of streak disease of sugar cane are available in India. Dastur (1926) reported the disease from Central Provinces, and McRae (1928) recorded the same from Pusa. In both places the disease was found affecting the Uba cane variety.

Symptoms :—

The leaves of the affected plant show numerous stripes of light green areas, running parallel to the veins. In the bleached areas, thin, dark green lines are visible. The affected young leaves assume a stiff, erect position and sometimes show slight crinkling. The streaks do not become confluent; sometimes irregularly outlined designs are seen (Dastur, 1926).

No secondary effects are seen due to the streak disease and no reduction in the yield of molasses is recorded (Dastur, 1926).

2. *Streak of tomato* (*Lycopersicon esculentum* Mill.)

During 1941-'42 streak, associated with leaf-roll of tomatoes has been recorded in Delhi as causing 9·2% infection. The highest incidence of streak as well as other virus diseases of tomato was observed in October, and plants showing streak symptoms were more than the others. Metereological conditions affected the disease incidence probably because, they affected the insect population. Of the virus diseases of tomato, recorded, streak causes the greatest reduction in yield (Vasudeva and Lal, 1947).

CHLOROTIC DISEASES

Pronounced yellowing of the leaf, distinguishes this type of disease from mosaic. As in mosaic, in chlorosis also, other subsidiary symptoms are often seen. A few plants are known to suffer from chlorotic diseases of viral nature in India.

1. *Chlorosis of Jute* (*Chorchorus* sp.)

Jute (*Corchorus capsularis* and *C. olitorius*) is known to suffer from a disease, marked by extreme chlorosis of the leaves, and the cause of the disease was not established for a long time. *Corchorus capsularis* is more susceptible to the disease (Gosh and Kundu, 1954).

The disease sometimes affects only one branch, and the tendency for recovery from the disease is often pronounced. Nitrogenous cattle manure in combination with potash and phosphorous is known to increase the symptoms, while a mixture of calcium, nitrogen, phosphorous and potash, or ammonium sulphate and super-phosphate reduce the disease (Anonymous, 1940).

Transmission :—

That the disease is due to a virus is gathered from the transmission studies. The disease has been found to be graft-transmissible, and to some extent through seed (Gosh and Kundu, 1954).

Pollen from chlorotic plants can transmit the disease during fertilisation. If one of the fusing gametes is from chlorotic tissue, the seed becomes virus-borne (Gosh and Kundu, 1954).

2. *Yellow Chlorosis of* Salvia :—

A mosaic disease of *Salvia,* soon developing into an acute chlorosis, has been recorded from Lucknow (Verma and Bose, 1954).

Symptoms :—

The symptoms start on the young leaves as small yellow specks, later growing bigger, coalescing and giving a chlorotic appearance. Vein-clearing, distortion and curling of leaves, and stunting of the entire plant, are often seen accompanying the chlorotic symptom (Verma and Bose, 1954).

The disease has been transmitted to healthy plants by cleft and bud-grafting.

NECROTIC DISEASES

Pronounced death of tissues either localized or complete, is the main symptom of necrotic diseases. Only potato and tomato are known to suffer from such a disease in India.

1. *Necrosis of Potato (Solanum tuberosum* L.)

Phulwa potato plants show necrotic spots on young and old leaflets, though more of them are seen on the lower leaflets. At its worst, the margins and even the entire leaflets die out. The disease has been shown to be sap- and graft-transmissible (Vasudeva and Gupta, 1946).

The disease has been transmitted to White Burley tobacco, *Capsicum annum, Datura stramonium* and potato variety President and Epicure. The virus is identified as *Solanum* virus 6, a closely related strain of *Solanum* virus 1 (Vasudeva and Gupta, 1946).

Darjeeling Red Round potatoes, raised in insect proof cages, and from virus-free tubers, showed acropetal necrosis, leading to a wilted condition; the leaves were slightly malformed, and gradually they dried, the whole plant later dying. Virus-free potato plants of the varieties President and Craig's Defiance, also showed malformation, browning of leaves, drying but not death of plants. Sap inoculation and other methods of mechanical transmission were unsuccessful. But the disease was successfully grafted to potato and a few other hosts (Vasudeva and Azad, 1948).

Bemisia tabaci was able to transmit the disease from potato to tobacco variety Harrison's Special and tomato. The virus causing necrosis of Darjeeling Red Round potatoes is identified as a mild strain of Tobacco leaf-curl virus (*Nicotiana* virus 10) (Vasudeva and Azad, 1948).

2. *Necrosis of Tomato* (*Lycopersicon esculentum* Mill.)

Sutton's Early Market tomato plants, produced veinal necrosis and necrotic spots when inoculated with the leaf extracts of potato plants showing severe crinkling and mottling of leaves. The first symptom on tomato, 20 to 22 days after inoculation is the slight inward rolling of leaf margins, followed by both veinal and foliar necrosis. Necrotic streaks are also produced on the petioles. The infected plants are dwarfed and tend to shed their leaves (Vasudeva *et al.*, 1949).

The sap from the infected tomato plants has been inoculated to tobacco, *Datura stramonium, Solanum nodiflorum, Petunia hybrida,* and *Nicotiana glutinosa* (Vasudeva *et al.*, 1949; Anonymous, 1955).

From the study of symptoms on these hosts, it is concluded that the virus causing necrosis in tomato is a combination of *Solanum* viruses 1 and 2 (Vasudeva *et al.*, 1949).

LEAF CURL AND LEAF ROLL DISEASES

The main symptom of these diseases is an excessive curling of leaf margins or apices, leading to a gross distortion of the shape of the lamina, often followed by other symptoms like mottling. There are a few reports regarding these diseases, most of them pertaining to crop and ornamental plants.

1. *Leaf-curl of Tobacco* (*Nicotiana tabacum* L.)

The virus: *Nicotiana* virus 10, Smith.

Tobacco suffers from a highly destructive virus disease, chiefly in North India, caused by the tobacco-leaf-curl virus. In some years the incidence is low ranging from 5 to 10%, while in others, as occurred during 1934-'35 in North India, the disease virtually destroys the entire crop (Pal and Tandon, 1937). The disease has been studied mostly on Pusa H 142 variety (Pal and Tandon, 1937; Pruthi and Samuel, 1937). In the Gangetic Delta, comprising

Bihar in east and Baroda in west, the cigarette variety of tobacco is known to be seriously affected (Pruthi, 1945).

Symptoms : –

The general symptoms of the disease are dwarfing of plants, reduction of leaf surface and size, curling of a portion or the entire lamina, vein-banding or thickening, greening of veins of calyx and ovary wall (Pal and Tandon, 1937). The disease does not appear in the seed bed, but is known to occur only after transplanting. The incidence of the disease increases in late October to early November, and declines in mid-January, showing a second rise in March. Monsoon favours infection (Pal and Tandon, 1937).

The symptoms of tobacco leaf-curl are so variable on tobacco, that it has been possible to identify five types of diseases, classified mainly on symptamatology, and supposed to be caused by different strains of the virus (Pal and Tandon, 1936; Pruthi and Samuel, 1937, 1939; 1941; Pruthi, 1945). Types A, B, C, D, and X are recognised as follows : —

Extreme stunting of plants, profuse greening and enations on leaves ——Type A. and B.

Leaves small, much curled and thickened, rugose brittle and dark green ——Type A.

Leaves large, slightly curled, wrinkled, pale or yellowish green with no thickening or brittleness ——Type B.

Stunting less pronounced, vein-clearing seen and enations absent ——Type C. and D.

Vein-clearing pronounced, with "stitches" in old leaves ——Type C.

Vein-clearing less pronounced with no "stitches" in old leaves ——Type D.

Symptoms variable, often a combination of two or more types ——Type X.

That the type X, is a combination of one or more of the other types is also proved by synthetically inducing the symptoms by a mixture of sap of other types (Pal and Tandon, 1937). The various types can be separated by passage through different hosts. Thus, types A and B are separable from C and D, by symptoms

on *Nicotiana glutinosa, Solanum nigrum* and *Petunia* sp., while A is distinguished from B, and C from D, by reactions on *Nicotiana rustica* (Pal and Tandon, 1937).

Transmission and Host Range : —

The disease has been experimentally transmitted by injection of the infective sap into healthy plants, and by grafting. There is now overwhelming evidence to show that in nature the transmission takes place through the agency of the white-fly *Bemisia tabaci* Gen. (*B_3 gossypiperda*) (Pal and Tandon, 1937; Pruthi and Samuel, 1937, 1939, 1941, 1942; Pruthi, 1945). Attempts to transmit the disease by the capsid-bug *Cyrtopeltis* (*Gallobelicus*) *crassicornis* Dist. gave negative results (Pruthi and Samuel, 1937). Seed transmission is shown not to take place (Pal and Tandon, 1937). The disease incidence depends on the white-fly population, which increases in autumn and falls down in winter, with a secondary rise in March (Pruthi and Samuel, 1942).

A number of alternate hosts have been described for the tobacco leaf-curl virus, and their presence as weeds, makes it difficult to eradicate the disease. *Ageratum conyzoides, Althaea rosea, Crotalaria juncea, Euphorbia hirta, Hibiscus rosa-sinensis, Launaea asplenifolia, Petunia* sp., *Schizanthus* sp., *Scoparia dulcis, Sida rhombifolia, Solanum nigrum, Vernonia cinerea* and *Zinnia elegans* have been reported to show leaf-curl symptoms (Pal and Tandon, 1937; Garga, 1939; Pruthi and Samuel, 1939, 1941; 1942). The following types of the tobacco leaf-curl virus have been suspected on the various hosts : —

Euphorbia hirta	——B, rarely C, often mixed with X.
Launaea asplenifolia	——X, sometimes D and X.
Lycopersicon esculentum	——A.
Scorparia dulcis	——X.
Sida rhombifolia	——C, sometimes A.
Solanum nigrum	——B, C and D, and all the 3 combined.
Vernonia cinerea	——A and C, both mixed with X.
Zinnia elegans	——C, D and X. (Pruthi and Samuel, 1941).

Nicotiana plumbaginifolia is suspected to be a symptomless carrier (Pal and Tandon, 1937).

Control Measures:

Constant roguing brought down the infection to 19 %. Selection of resistant varieties, spraying and dusting the leaves at 10-day intervals for two months from seed-bed, and abolition of alternate hosts, are some of the control measures recommended (Pal and Tandon, 1937; Pruthi, 1945).

Experiments on the control of the disease by the use of fertilisers reveal, that plants treated with nitrogen, carried the maximum infection; those treated with phosphorous, none; and those with P N, intermediate type of infection. The P/N ratio of healthy plants was higher than that of the diseased. Addition of nitrogenous fertilisers increased the disease, phosphates reduced it while boron, copper, manganese and zinc slightly reduced the infection (Ramamoorthy *et al.*, 1948). Diseased plants show the absence of a ninhydrin reacting substance, which is present in the healthy, and the presence of 2 new bands corresponding to asparagine and histidine-lysine (Laloraya *et al.*, 1955).

2. *Leaf-curl of Tomato* (*Lycopersicon esculentum* Mill.).

Tomato is known to suffer from a leaf-curl disease for several years in Delhi (Vasudeva and Samraj, 1948). Depending upon the age of infection, the plants show dwarfing, puckering of leaves, vein-clearing, excessive branching mild to severe mosaic, and partial to complete sterility. During 1947, 90 % of plants were affected, and from 1943 to '45, 60 varieties have been found to be susceptible to the disease (Vasudeva and Samraj, 1948).

Transmission is achieved by white-fly *Bemisia tabaci* (*B. gossypiperda*) and 30 to 100% of the plants, allowed to feed by the vector developed symptoms within 15 to 25 days. The disease can be grafted to tobacco varieties Harrison's Special, White Burley, German Samson, Potato variety Craig's Defiance, *Datura stramonium, Nicotiana sylvestris* and *N. glutinosa*. On *Datura stramonium* the reaction is violent (Vasudeva and Samraj, 1948).

The symptoms on various hosts reveal that the disease in tomato is due to both mild and severe strains of the leaf-curl virus, both types of strains producing similar symptoms on tomato. The virulence is shown to be reduced, when the virus is allowed to pass through tomato plants (Vasudeva and Samraj, 1948).

3. *Leaf curl of Papaya* (*Carica papaya* L.)

Leaf-curl disease of papaya has been reported from several parts of India. Thomas (1939) reported the disease from Madras, Thomas and Krishnaswami (1939b) an analogous disease from Coimbatore, Sen *et al.* (1946) from Bihar, and Nariani (1956) from Delhi.

Symptoms : —

The disease is characterized by severe curling, crinkling and distortion of leaves accompanied by vein-clearing and reduction in leaf size. Leaf margins are rolled downwards and inwards in the form of inverted cups, and the veins get thickened and dark green in colour. The petioles are highly twisted. The leaves become leathery and brittle, and the inter-veinal areas become rugose, due to hypertrophy. The affected plants fail to flower, or bear very few fruits, and in the advanced stages, defoliation takes place and the growth of the plant is arrested (Nariani, 1956).

Transmission and Host Range : —

The disease has been transmitted by juice inoculation (Sen *et al.*, 1946), and by grafting (Nariani, 1956). White-fly *Bemisia tabaci* has been found capable of transmitting the disease (Nariani, 1956).

The disease is transmissible to White Burley tobacco and tomato by means of the white-fly. Ranchi, Bombay, Honey Dew and Washington varieties of papaya were tested and shown to be susceptible (Nariani, 1956). The disease is not seed-borne Sen *et al.*, 1946).

Effect of the disease on host physiology : —

Paper chromatographic analysis of the sap of healthy papaya leaves and those infected by the leaf-curl virus reveal a marked increase of asparagine in the latter (Laloraya *et al.*, 1956).

Etiology of the disease :—

There appears to be some confusion regarding the etiology of the disease. Sen *et al.*, (1946) found that the disease could be caused by water-logging in the soil, independent of infection. Khan and Mehta (1956) report the presence of galls in the roots of af-

fected plants induced by *Meloidogyne marioni,* a nematode. The nematode is reported to attack besides Papaya, *Phaseolus mungo, Solanum melongena,* tomato, *Ficus carica,* and *Impatiens balsamina* (Khan and Mehta, 1956). Nariani (1956) identified the causal agent of the leaf-curl of papaya at Delhi as tobacco leaf-curl virus. Thus, it appears that the leaf-curl condition can be brought about by several causes—soil, nematode and virus.

4. *Leaf Curl Mosaic of Sandal (Santalum album* L.)

Venkata Rao (1933) observed leaf-curl symptoms accompanied by mosaic on sandal (*Santalum album* L.) in Bangalore and Mysore.

Symptoms :—

There appears to be two stages for the disease. In the first stage, conspicuous spots develop between the veins, followed by inter-veinal mottling and slight rolling of mature leaves; In the pigmented varieties of sandal, a reddish-brown marginal discolouration is seen. The leaves retain their normal size and the flowering and fruiting is normal. The small branches and the leaves generally show a drooping habit.

During the second advanced stage, the leaves show ruffling at the edges even when quite young, followed by wrinkling, mottling, progressive dwarfing, curling and discolouration, ranging from greenish yellow to reddish brown. In contrast to the 'spike disease', the surface reduction in mosaic leaves is usually greater in length than in breadth. The thickness of the petiole is considerably diminished, the leaves become brittle and fall off early. The twigs and branches die back, and adventitious branches arise, showing highly attenuated leaves. The fruits of the affected trees are small and are shed early (Venkata Rao, 1933).

Transmission :—

The virus is transmissible by ring-bark grafts, but not by sap-inoculation. The cause of natural infection is yet unknown.

5. *Leaf Curl of* Zinnia elegans :—

A leaf-curl disease affecting *Zinnia elegans* is reported from Dehra Dun (Mathur, 1933).

Symptoms :—

Thickening of lower surface of vein-lets, preceded by curling of leaf blades, dwarfing of plants (sometimes less than 1 foot), and few, dwarfed flowers, which are pale and partially sterile. During the warm damp season, the axillary buds give rise to short stunted, rosette-like branches with malformed leaves. Leaf-curl is most prominent from July to September, when the white-fly population is also high. Infection begins at the young leaves near the growing points (Mathur, 1933).

Transmission :—

The disease has been found transmissible to healthy plants by the white-fly *Bemisia tabaci.*

Affinities of the virus :—

The virus causing leaf-curl of *Zinnia elegans* is identified as that causing leaf-curl of cotton at Sudan (Mathur, 1932).

6. *Leaf-curl of other plants* :—

(a) Potato (*Solanum tuberosum* L.)

Bismarck variety potato has been known to suffer from a leaf-curl disease, characterized by stunting of the apical shoots, downward curling of leaf-tips and outward or inward curling of leaf-margins. The cause of the disease has been attributed to the tobacco leaf-curl virus (Anonymous, 1951b).

(b) *Hibiscus rosa-sinensis* L.

The plant when affected with the leaf-curl virus shows faint mottling, thickening of veins and reduction of leaves; subsequently enations develop on the veins on the lower side. Puckering is common, and the leaf margins usually curl upwards (Anonymous, 1951b). The disease has been found transmissible by grafting and by white-fly *Bemisia tabaci. Hibiscus esculentus,* H. *tuberculatus, H. manihot,* hollyhock and cotton are the alternate hosts for the virus (Anonymous, 1954d).

(c) *Saponaria vaccaria* L.

The disease was first observed on *Saponaria vaccaria* L. in Delhi during 1952.

The main symptoms of the disease are stunting of plants, curled and reduced leaves, profuse irregular granular outgrowths on the veins of lower side of leaves, scanty flowering, and in some cases of late infection, production of purple leaves (Azad, 1953). The disease has been transmitted to healthy plants by grafting. The disease is shown to be due to a virus different from the tobacco leaf-curl virus (Azad, 1953; Anonymous, 1954-d).

(d) *Physalis peruviana* L.

The disease was first observed in 1953, in New Delhi. The main symptoms of the disease are stunting, dwarfing and curling of leaves, and enations on the lower veins of the leaves. The virus has been identified as tobacco leaf-curl virus (Nariani and Pathanian, 1954).

(e) *Sann hemp* (*Crotalaria juncea* L.)

Production of enations, curling of leaves and dwarfing of plants are the main symptoms of the leaf-curl disease of sann hemp (*Crotalaria juncea* L.). The disease is graft-transmissible (Raychaudhuri, 1948).

7. *Leaf roll of potato.* (*Solanum tuberosum* L.)

Leaf-roll of potatoes is one of the highly destructive diseases of potatoes, and the disease appears to be fairly common in North India. During 1937-'38, 1,000 maunds of seed potatoes of Dunbar, Cavalier and Majestic varieties were obtained from abroad, to replace the mosaic and leaf-roll affected seed potatoes of Agra (Pushkarnath, 1943). Imperia and Sattoo variety are reported to be resistant to the leaf-roll virus (Pal and Pushkarnath, 1951). The disease causes the highest yield reduction in potatoes compared to other potato virus diseases, and in India, the loss is estimated as 72 to 90% (Vasudeva and Azad, 1952). High temperature (summer temperature) reduces the disease incidence in seed potatoes (Thirumalachar, 1954).

Venkata Rao (1930) inoculated the infective juice of potato leaves suffering from leaf-roll, by means of hypodermic needles and fine pin punctures into healthy tomato plants. Symptoms of leaf-roll were observed 50 days after inoculation (Venkata Rao, 1930).

8. *Leaf roll or 'Bobri' of Chilli* (*Capsicum fruitescens* L.)

Leaf-roll or 'bobri', a virus disease of chilli, has been reported from Rajasthan, causing serious havoc on chilli plants (Anonymous, 1951-c). The affected plants are small, and stunted; the leaves exhibit intense rolling, fruiting is retarded and the fruits are small. During 1950-'51, the disease caused 100% damage in Kapasin district (Rajasthan).

9. *Leaf-roll of Tomato* (*Lycopersicon esculentum* Mill.)

Tomato plants are known to suffer from a virus disease, characterized by leaf-roll, associated with streak and mosaic. Typical leaf-roll has however, not been observed in India (Vasudeva and Lal, 1947).

SPIKE DISEASES

In the foregoing diseases, the main symptoms are foliar, while in "spike", the disease is manifested by the entire plant, and particularly by the branches, which appear like a spike, with attenuated leaves and branchlets, often similating a condition described as "witches broom" or "rosette".

1 *Spike of Sandal* (*Santalum album* L.).

The virus :

Sandal spike virus, Coleman.
Santalum virus 1, Smith.
Sandal spike rosette virus, Rao and Iyengar, 1934.
Chlorogenus santali Holmes.

Spike disease of sandal (*Santalum album* L.) is one of the earliest known virus diseases in India. The disease was first observed in Coorg in 1899, by McCarthy. By 1903, the disease extended to other parts, chiefly Mysore, Madras and North Coimbatore. The disease soon spread towards south and south-east India. Salem experienced heavy losses due to the disease and during 1917-'25, 98,734 sapwood trees were removed in North Salem, due to the disease, which increased from 30 to 20,000 acres (Hart and Rengaswamy, 1926) and the ultimate loss in controlling the disease amounted to Rs. 47 lakhs. The annual loss due to spike disease of sandal, in South India, is estimated as Rs. 6 to 7 lakhs (£ 45,000 to £ 52,000) (Narasimhan, 1933).

Symptoms :—

The characteristic symptom of the disease, is the extreme reduction of leaf size and the attenuation of branches into stiff pointed structures resembling a spike. The internodes are shortened, phyllody of flowers is common and the root ends and haustoria die, which gradually leads to the death of the tree (Narasimhan, 1928). Sreenivasaya (1934) has recorded two types of disease, causing almost similar symptoms; one, a non-infectious stunting in South India, due to adverse environmental conditions, not transmissible by grafting and curable by the supply of good soil and a favourable host, the other, the infectious and highly destructive spike disease, transmissible by grafting. The girth of the stem, affected by the spike disease is adversely reduced, and the minimum girth observed for a fairly old tree is ½″ (Hart and Rengaswamy, 1926). The ratio of length and breadth of the affected leaves (L/B) is significantly higher, and can be used as a test to identify the disease (Varadaraja Iyengar, 1938a).

In some cases, the leaves become pale in colour, and in later stages, the spike leaves attain a reddish tinge. The average period from the appearance of the symptoms to the death of the tree is about 15 months, the time varying with the girth of the tree (Rangaswami Iyengar and Griffith, 1940).

Masking of symptoms is sometimes observed in sandal, and hence external appearance sometimes proves useless in detecting the disease (Sreenivasaya, 1930b). Symptomless sandal however, develops symptoms when mechanically injured as when pruning or making holes in stem. Apparently healthy trees can be tested for the virus by the leaf or bark grafting. The masking is influenced by sunshine and temperature. During the masked period the virus appears to be localized in the phloem, where it multiplies and exists in a highly virulent form (Sreenivasaya, 1930b).

Transmission :—

Earlier observations on the cause of sandal spike revealed that the disease is caused by a difficulty in obtaining water to the plant, due to unfavourable selection of the host plant (Jivanna Rao, 1921a, 1921b and 1925). Coleman (1923) showed that the disease could be transmitted by grafting, and through haustoria. Bud-grafting, patch-grafting and leaf-insertion methods gave very good results of transmission (Sreenivasaya, 1930c; Sreenivasaya and Naidu,

1928). Transmission did not take place by patch-grafting of root bark on root, leaf mutilation, wood grafting or injection of tissue fluids. According to Sreenivasaya (1933c), best results of transmission are obtained by the leaf-insertion method. According to Venkata Rao and Iyengar, (1934b) twig grafting gave 100% infection, budding 63%, patch-grafting 18% and leaf insertion 17%. The average incubation period for both bud and twig grafting is three to four months. Callous formation has been shown to occur 10 to 15 days after the insertion of infected leaf into the healthy plant. Transmission of the disease by pollen and seed does not take place (Venkata Rao, 1935).

As early as 1923, Coleman suspected the role of an insect vector in the natural spread of the disease. Dover and Appanna (1934) obtained indirect evidence that a species of *Macrosiphum* from *Lantana camara* can cause infection. Similar results were obtained for studies on transmission by *Moonia albimaculata*. Rangaswami and Sreenivasaya (1935) reported that the vector of the disease is to be found among three types of pentatomidae, two of jassidae, and three of fulgoridae, and that the vectors are active during night. Rangaswami and Griffith (1941) showed that *Jassus indicus* Walk. can effectively transmit the disease, the maximum and minimum incubation period for the virus inside the vector, being 122 and 102 days respectively. The vector of the spike disease has been shown to be active between the months of April and June, with a secondary climax in October to December.

Histopathology :—

In the spiked leaf, six to seven lines of mesophyll cells are packed very closely without any intercellular space, and the cells show a high accumulation of starch (Jivanna Rao, 1921a). The starch thus accumulated disintegrates in the later stages of infection (Narasimhan, 1933). Necrosis of phloem due to collapse of sieve tubes is often seen (Narasimhan, 1934).

A characteristic symptom of the infected leaf is the presence of round or oval vacuolate intra-cellular bodies, in close association with the nucleus. Generally, only a single such body is present in a cell, though rarely, two or three have been observed. These bodies, stain readily with iron-alum haematoxylin or Fleming's triple stain and reacts with Good pastur's Carbol anilin-fuchsin stain, similar to those found in animal virus diseases like

rabies (Narasimhan, 1933). The inclusion bodies have a diameter of 4·3 to 8·7 μ. They have a reticulate matrix containing usually 1 to 11 vacuoles of different sizes. As many as seven proliferations have been observed arising from these bodies.

Effect of environment on the incidence of the disease :—

Many workers have shown that environment exerts a considerable influence on the incidence of spike disease. Hart and Rengaswamy (1926) showed that the spread of the disease is from southwest to northeast, i.e., in the direction of the prevailing monsoon winds at the flowering time. This is in keeping with the view of Sreenivasaya and Rangaswami (1934) that the viruliferous insects, acting as vectors spread along the direction of the wind. The highest infection is in June, while the lowest is in October (Venkata Rao and Iyengar, 1934b). Norris (1930) showed that the nature of host parasitized by the sandal has a considerable bearing on the incidence of infection. The disease is reported to be very high in areas where *Lantana camara* dominates the vegetation, which not only depletes the soil moisture and nutrients but also might harbour insect vectors (Anonymous, 1931). That the role of undergrowth has a pronounced effect on the disease is stressed by many workers. *Zizyphus oenoplia, Dodonaea viscosa, Stachytarpheta indica* and *Vinca rosea* are known to suffer from spike-like diseases and these are considered to act like alternate hosts for sandal spike. Sandal trees parasitic on *Cajanus indicus, Cassia montana, Acacia suma* and *Pongamia glabra* become susceptible to the disease, while those parasitic on *Azadirachta indica, Murraya koenigii, Eugenia jambolana, Toddalia aculeata* and *Dodonaea viscosa* seem to resist infection (Mathew, 1955). Shade, rainfall followed by excessive heat, dense stands of sandal trees, and an annual rainfall of 15 to 20 inches, contribute to the disease (Anonymous, 1931; Mathew, 1955).

Physiology of the spiked sandal :—

Considerable work has been done on the various physiological aspects of the spiked sandal and only the salient points are mentioned here.

The infected tissues show increased diastatic activity, the increase being sometimes up to four times that in the healthy plant (Sreenivasaya and Sastri, 1928; Sastri and Sreenivasaya 1929). Higher enzyme concentration, osomotic concentration, depression

in the freezing point, (due to an increased soluble matter in the sap) more reducing sugars, lower specific conductivity of sap, and lower ratio of specific conductivity to depression of freezing point of the sap characterize the diseased sandal (Sreenivasaya and Sastri, 1928, 1929; Varadaraja Iyengar, 1928; Rao and Sreenivasaya, 1928). Infected leaves show higher total nitrogen, increased water-soluble nitrogen and basic nitrogen, and decreased nitrate nitrogen (with reference to dry weight and total nitrogen content) (Varadaraja Iyengar, 1928; Rao and Sreenivasaya, 1928; Narasimhamurthy and Sreenivasaya, 1929). The accumulation of total organic nitrogen takes place in the diseased foliage, and this is increased by an increase in the proteoclastic activity. The higher peptonase activity of the diseased tissues and tissue fluids accounts for the presence of a higher concentration of amino nitrogen (Subrahmanyam and Sreenivasaya, 1941). Histidine is found in large quantities in the infected leaves, which is found in negligible amounts in the healthy, and its conversion into histamine is believed to explain the necrosis of root ends of spiked plants (Rao, 1933). Mannitol has been found in the spiked plants but not in the healthy (Sreenivasaya 1930a). Increased production of ammonia due to an increased oxidative deamination and increased carbon-dioxide are observed in the spiked plants. While the healthy plants show pyrocatechin group of tannins, the infected ones show the pyrogallol group. The buffering capacity of tissue fluids is greatly disturbed consequent on infection. (Varadaraja Iyengar, 1933a 1935a and 1937). Oxalic, malic and succinic acids are seen to accumulate in the spiked leaves, and is explained as due to respiratory changes, the oxygen uptake being half and the oxidase and peroxidase activities about the double of the healthy leaves. Lower concentration of calcium either hinders their neutralization or removal by precipitation (Sastri, 1946).

The spiked plants contain lower amounts of calcium and potassium, while phosphorous and manganese are found in greater quantities than in healthy plants (Sreenivasaya and Sastri, 1928; Varadaraja Iyengar, 1928, Rao and Sreenivasaya, 1928; Sreenivasaya and Sastri, 1929b Badami and Iyengar, 1932). The ratio of calcium to nitrogen is low in the infected plant (Varadaraja Iyengar, 1929). The inability of plants to assimilate calcium is the immediate reaction of sandal stems and leaves to spike infection (Varadaraja Iyengar, 1937). There is a high gradient of calcium concentration from stem towards leaves, which tends to flatten out

as the disease progresses. Due to the decreased concentration of calcium, irregularities in the translocation of starch are seen. The root of the infected plant often contains a high concentration of calcium, and the translocation of calcium from the roots to the aerial parts is checked. Nitrogen on the other hand, rapidly reaches the aerial parts, causing proliferation of the vegetative parts at the expense of the reproductive. The reduced size of the affected leaves is probably due to calcium deficiency. In the affected areas, the soil beneath the spiked plants contain twice as much calcium as in healthy plants. It is observed that the immediate cause of spike is poor intake of calcium by roots and growth, due to some toxin secreted after infection. The importance of this view is that it serves to connect together the virus and the physiological theories of the nature of the disease (Varadaraja Iyengar, 1938b).

The higher manganese in the spiked plants has little effect on protein synthesis (Badami and Iyengar, 1932).

Movement of the Virus in the Host :—

That the sandal spike virus moves through the phloem and parenchyma of the cortex has been shown by ringing experiments. The fact that the virus does not pass through the xylem is indirectly shown by the failure of disease transmission by grafting the wood of the infected plant. The most rapid movement recorded is twelve inches in 89 days after inoculation (Venkata Rao and Iyengar, 1934a).

Resistance to the disease :—

Sandal possesses (i) an autogenic resistance, which is largely independent of the nature of the host it parasitises, many sandal varieties remaining healthy in severely infected areas and (ii) an acquired resistance, exhibited when specific hosts are provided under certain natural environmental conditions, the disease incidence and spread being dependent upon the floristic composition of the area. Based on this, two areas have been established in North Salem using resistant strains of sandal, reinforced by hosts imparting immunity like *Melia indica, Dodonaea viscosa* and *Ruta graveolens* (Sreenivasaya, 1948).

In pot culture studies, sandals growing in large pots were more resistant than those in small ones. Deep-rooted hosts are bene-

ficial to sandal. Infected sandals succumb to disease soon, in the absence of a host (Sreenivasaya, 1948; Sreenivasaya and Rangaswami, 1938).

The identification of sandal plants resistant to spike is difficult. Profuse root development, and a strong tendency to haustorial formation are taken to be characters of such varieties. Some resistant varieties are said to have ovate leaf, stem with rich lenticels, and a root with high haustorial capacity (Sreenivasaya, 1930c; Anonymous, 1932). Tissue fluids of sandals, grown in combination with *Ruta graveolens, Murraya koenigii, Azadirachta indica,* and *Toddalia aculeata,* are more buffered than those grown with *Acacia farnesiana.* Thus, a significant correlation would appear to exist between disease resistance and the buffering capacity of the tissue fluids of sandal grown with different hosts (Srinivasan and Sreenivasaya, 1935).

Control measures :—

In certain areas, prompt removal of infected plants appears to have controlled the disease (Hart and Rengaswamy, 1926; Rangaswamy and Griffith, 1939a). Lopping or ringing the affected part is found to be useful in some cases, but proves unsatisfactory when the disease is in the latent form (Anonymous, 1932). Abolition of *Lantana camara,* and other unfavourable hosts of sandal, has given excellent results.

Several methods have been adopted to destroy the affected plants. The older method was to burn down the plant after removing the heartwood, but recently certain chemicals are shown to do the job better. Thus the patent arsenical chemical "Atlas", has become popular (Varadaraja Iyengar, 1935c). Potassium and sodium chlorates, potassium arsenate, (2½ lbs. per gallon) arsenic oxide (4 lb. in 1 lb. of NaoH per gallon) and sodium arsenate also give good results (Varadaraja Iyengar and Rangaswamy, 1935).

2. *Pendulous Spike of Sandal* (*Santalum album* L.)

Venkata Rao and Iyengar (1934a) described a disease, distinct from the spike disease, and which they designate "Pendulous spike".

Symptoms :—

The individual infected shoots show continuous apical growth usually attaining a height of one or three feet. The leaves are

confined to the upper portions of such branches. The characteristic feature of the disease is the drooping habit of the branchlets, these bearing leaf-blades, which are usually large and broad. Extensive branching is rare, due to the non-development of the axillary buds.

The flowers of the diseased plants are invariably abnormal, the pedicels often elongating to three or four times the length of those in healthy plants. Anthers show a reddish tinge, the pistil is usually seen as a thick cylindrical body, the apex of which is bent on itself in the bud condition. Sterility being very pronounced, fruits are seldom produced. Death of root ends and haustoria does not occur in this disease.

This disease is transmissible by the same grafting methods as applicable to spike disease, and inclusion bodies similar to those found in spike disease are reported in sandals affected with pendulous spike.

3. *Other plants affected with the spike disease* :—

(a) *Dodonaea viscosa* L.
(b) *Vinca rosea* L.
(c) *Zizyphus oenoplia* Mill.

These plants are often seen to suffer from spike-like diseases, producing symptoms similar to those of spiked sandal, like stunted growth, excessive branching, diminutive leaves and internodes and phyllodic flowers. In the leaves of *Vinca rosea* affected with spike, inclusion bodies similar to those found in the leaves of spiked sandal, have been described (Narasimhan, 1928).

In *Vinca rosea,* the diseased parts usually contain 1½ times as much starch as in healthy parts. The ash content and nitrogen show an increase during infection (Varadaraja Iyengar, 1933b). Calcium is seen in greater concentration in roots, less in the leaves of the infected plants, while nitrogen is seen in greater concentration in leaves and less in roots, suggesting a rapid mobilisation of nitrogen to the aerial parts. The protein content of the different parts of the infected plant are generally lower than similar parts in healthy plant, but the reverse condition is seen in the case of ammonia. The infected leaves are further characterized by starch accumulation and increased diastatic activity (Varadaraja Iyengar, 1935b).

In *Dodonaea viscosa* L., non-protein nitrogen constituents of spiked trees show a higher value for nitrate, nitrite, ammonia and

amide nitrogen. Sugars and starch accumulate in the spiked tissues and the diastatic activity of such tissues is always higher than in healthy ones. Low quantities of calcium is yet another feature of infected parts (Sastri and Narayana, 1931).

In *Zizyphus oenoplia* Mill., the affected leaves show a higher accumulation of starch and sugars, with an increased diastatic activity and low calcium content (Sastri and Narayana, 1931, Varadaraja Iyengar, 1935b).

It is obvious from the above observations, that spike disease causes similar disturbances in the physiology of the host in addition to the display of similar symptoms. Muniyappa *et al.* (1941) have shown that the diseased leaf tissue from spiked *Vinca rosea* when grafted to healthy sandal, causes the production of spike symptoms in the latter. These facts definitely show that the sandal spike virus has allernate hosts in the above-said plants.

4. *Rosette of Groundnut* (*Arachis hypogea* L.)

A disease, similar to spike, called the "rosette" disease, has been known to affect groundnut (*Arachis hypogea* L.). The main symptoms of the disease are reduction of leaf and internodes, phyllodic flowers and a bushy appearance of the plant (Anonymous, 1945).

PHYLLODY AND BUNCHY-TOP DISEASES

Phyllody is characterized by the conversion of the floral parts into foliaceous structures, resulting in the sterility of the flower.

1. *Phyllody of Sann hemp* (*Crotalaria juncea* L.)

Bose and Misra (1938) reported twelve different types of phyllody observed from 0·05 to 6·61% in seven varieties of *Crotalaria juncea* at Pusa. The floral parts in the affected plants are transformed into greenish leaf-like structures. Proliferation of floral parts, adherence of vegetative shoots formed by the converted stamen and carpel and fasciation are other symptoms of the disease. Sterility is pronounced in the affected plants. Occurrence of similar disturbance in *Crotalaria striata*, and *C. saltiana* is recorded from Bombay.

Transmission :—

The disease has been successfully transmitted to healthy plants by grafting (Bose and Misra, 1938). Sap-inoculation, the white

fly *Bemisia tabaci* and the jassid *Deltocephauls* sp. also caused transmission. (Anonymous, 1952a; 1955d).

Physical properties of the Virus :—

Sap from *Crotalaria juncea* infected with the phyllody virus is shown to remain infective for 370 days at 6° to 8°C. (Anonymous, 1953b).

2. *Phyllody of Sesame* (*Til*) (*Sesamum* sp.)

Dey (1948) recorded the phyllody disease of sesame for the first time in India. Mehta (1952) reported the occurrence of the disease from Uttar Pradesh, and it is now known that the disease is fairly widespread throughout North India.

Symptoms :—

The disease in some areas is called the 'green flowering' disease. The infected plants of *Sesamum indicum* show flowers in which all floral parts except the stamen, are modified into leaf-like structures. The anthers seldom contain pollen, and the plants are often sterile. Fasciation is a common feature. The leaves are small, and near the floral axis they become pale. Phylloid flowers are radially symmetrical, with polysepalous calyx, and the veins of the sepals heavily thickened. The fifth anterior stamen is usually developed unlike in normal flowers, the style is reduced and the carpellary wall is reduced to a leaf-like structure (Pal and Nath, 1935).

Twenty varieties of *Sesamum orientale* as well as *S. occidentale, S. indicum* and *S. radiatum* proved susceptible to the disease.

Transmission :—

Sap transmission is not successful; but the disease has been successfully transmitted by grafting (Pal and Nath, 1935). The jassid *Deltocephalus* sp. has been shown to effectively transmit the disease, the symptoms of which appeared 33 to 59 days after feeding by the vector (Vasudeva and Shambi, 1956).

3. *Bunchy-top of Banana* (*Musa* sp.).

Bunchy-top of banana is a very destructive virus disease, recorded from various parts of India. The disease is fairly widespread in Assam, Bihar, Bombay, Hyderabad, Khandesh, Orissa, Poona, Surat, and Travancore (Kottayam) (Nandhi, 1941; Verghese, 1945; Kamat and Patel, 1951; Varasi, 1954). The high incidence of the disease prevented the export or transport of

banana plants or any articles used in packing them, from the states of Orissa and Travancore-Cochin (Anonymous, 1953a). The disease appears to have been first observed in India during 1943, and is presumed to have come from Ceylon (Varasi, 1954).

Symptoms :—

The disease is characterized by stunted growth of the shoot, dwarfing of leaves (lamina and leaf-stalk), failure of the leaf to emerge from the overlapping leaf-sheaths, chlorotic leaf-margin and lamina, and failure to flower. The disease gets the name, due to the bunchy appearance of the entire plant after infection, and is known to affect all varieties of banana and at all stages of growth (Verghese, 1945).

Transmission :—

The disease is transmitted by the aphid *Pentalonia nigronervosa* Coq.

Control measures :—

Avoiding the use of infected suckers, pouring boiling water on diseased stumps, addition of lime to the underground portions of the diseased plants, destroying the affected plants by cutting them into bits and deeply burying them in the ground and the eradication of insects are some measures of control recommended (Verghese, 1945; Varasi, 1954).

STENOSIS AND LITTLE LEAF DISEASES

1. *Stenosis of Cotton* (*Gossypium* sp.).

The virus :

Cotton small-leaf virus.

Kottur and Patel (1920) described a disease of cotton plants, in all parts of West India, characterized by a malformation of leaves leading to sterility, and of unknown cause, Afzal (1933a, b) reported a similar disease affecting American varieties of cotton grown in the canal waters of Punjab and ascribed symptoms of "stenosis" of cotton. Gokhale (1936) reported a small leaf disease of cotton, the cause of which was not known, and hence he considered it to be a teratological case. It is now clear, that the small leaf disease or stenosis of cotton, is a virus disease, (Uppal *et al.*, 1944). These authors report the occurrence of the disease in Bombay, Madras and Punjab.

Symptoms :—

The diseased plants show extreme stunting of the aerial parts, crinkled deformed and highly reduced leaves, sometimes with yellow patches, reduction of the number of leaf-lobes, normal sized petioles light in colour, discoloured epicalyx, and reduction of root system. A large proportion of developing bolls and young buds on the diseased plants are shed, and the few bolls that are allowed to ripen, contain a few small seeds. The fibre obtained from the diseased plants is weak in quality. (Afzal, 1933a, b).

Starch accumulation takes place in the tissues of leaves, vascular bundles of petioles, stems, and roots of stenosised plants. (Likhite and Desai, 1934).

The disease affects plants at all stages of growth, especially after the plants become six weeks old, and is very virulent during August-September.

Transmission :—

The disease is not seed-borne, nor is it sap-transmissible. While the possibility of a vector relationship is strongly suspected (Afzal, 1933b) the disease at present is shown to be graft-transmissible (Uppal *et al.*, 1944).

Susceptible varieties :—

The most susceptible variety of cotton, to stenosis is reported to be the American variety, while indigenous cottons are shown to be fairly immune to the disease. The common susceptible varieties are Rozi, Jarilla, American, Gaorani and Mungari (Uppal *et al.*, 1944).

2. *Little leaf diseases* :—

Synonyms : 'Small leaf' disease; 'Rosette' disease.

In this group, the characteristic symptom of the disease is manifested mainly on the leaf, which undergoes a striking reduction in size, simulating a "little" or "small" leaf. It should be noted however, that this symptom occurs in other diseases like spike and phyllody. Since no better name for the disease is available and since many have used this foliar symptom to identify the disease, the convention is followed here.

(a) *Little leaf of egg plant (Brinjal) (Solanum melongena* L.).

The virus :

Cranberry False-blossom virus, Dobroscky.
Vaccinium virus 1, Smith.
Chlorogenus vaccinii Holmes (Smith, 1957).

Symptoms :—

The main symptom of the disease is the reduction in leaf size, the new leaves becoming smaller and smaller and sessile. The lamina becomes thin, soft, glabrous and pale green. In thorny varieties the thorns are attenuated or absent. The stimulation of the axillary buds to grow, results in a bushy appearance of the affected plants. Phyllody is very common, and the plants are sterile. The disease affects the plants at all stages of growth (Thomas and Krishnaswami, 1939a) and is widespread in South India.

Transmission and Host range :—

The virus is not sap-transmissible, but is transmitted by grafting. Two species of jassids *Empoasca devastans* and *Eutettix phycitis* are shown to be vectors of the disease. The virus occurs naturally on *Datura fastuosa,* and it has been experimentally transmitted both by grafting and by leaf-hoppers, to tomato, tobacco, potato, *Solanum trilobatum, S. pubescens, S. xantho-carpum, S. seaforthianum* and *Withania somnifera* (Thomas and Krishnaswami, 1939a).

(b) *Little leaf of Vinca rosea* L.

The little leaf of *Vinca rosea* L. has already been described under the spike diseases. The virus has been shown to be transmissible by grafting, and by *Eutettix phycitis* (Anonymous, 1954b).

Histopathological investigations of "little leaves" of Brinjal, *Datura fastuosa,* and *Vinca rosea* reveal that a "little leaf" can be identified by the peculiar anatomical features, from a similar-sized young healthy leaf and a mature healthy leaf. Furthermore, the stomata per unit area is shown to be highest in a small young leaf and lowest in a "little leaf". The nuclei in the "little" leaf are often irregular and sometimes crescent-shaped (John and Malini, *unpublished*). Intra-cellular inclusion bodies have been described in the spiked leaves of *Vinca rosea* and are shown to be similar to those described in the spiked leaves of sandal (Narasimhan, 1928).

(c) *Small leaf of Sann hemp* (*Crotalaria juncea* L.).

Sterility is shown to be the most important symptom of the small leaf disease of sann hemp which is shown to be due to a sap-transmissible virus. The physical properties of the virus are thermal death point of 85°-90°C, longevity *in vitro* 41-50 days and dilution end point of 1:25:000 to 1:30,000 (Raychaudhuri, 1948).

OTHER VIRUS DISEASES

Under this group a few virus diseases, for which no distinct names are available, are included.

(a) *Big bud of tomato* :—(*Lycopersicon esculentum* Mill.).

The virus :

Cranberry False Blossom virus, Dobroscky.
Vaccinium virus 1, Smith
Chlorogenus vaccinii Holmes (Smith, 1957).

During 1943, Sutton's Early Market tomato plants at the Imperial Agricultural Research Institute, New Delhi, showed symptoms of tomato big bud virus, not previously known in India. Erect growth habit, floral anomalies such as adhesion, certain types of overgrowth, virescence and thickening of the pedicels due to the formation of an abnormal tissue in the phloem, were the chief symptoms of the disease.

Transmission :—

The disease has been found to be transmissible by bud grafting and inarching. The symptoms appear 45 days after bud grafting and 105 days after inarching in healthy plants (Vasudeva and Lal, 1946).

(b) *A latent virus disease of tomato* (*Lycopersicon esculentum* Mill).

The cause of a faint mottling of tomato plants suffering from "smalling", has been ascribed to a latent virus, which is known to produce a very faint mottling even in the healthy tomato plants, variety Sutton's Early Market.

Transmission :—

The disease is transmissible by sap, using carborundum. In addition to tomato, the virus affects White Burley tobacco.

Inclusion bodies :—

X-bodies have been described from the epidermal cells of the diseased plants. Usually only one X-body is seen in each cell, and the nuclei in such cells are enlarged and deformed. Prior to the formation of the X-body, small granules appear in the cells, which gradually aggregate, and finally become the X-body. They are usually round, with an average diameter of 25 μ. The crystalline inclusions are cubical, and their number and size are variable. The maximum diameter of one such block is 6 μ.

Physical properties of the Virus :—

The virus has a thermal death point of 59°C. for 10 minutes, a dilution end point of 1:10,000 and longevity *in vitro* of 24 hours at room temperature (Vasudeva and Samraj, 1947).

(c) *Grassy shoot of sugarcane* (*Saccharum officinarum* L.).

A new virus disease has been described as, "grassy shoot" affecting sugarcane, from Lucknow (Anonymous, 1956), and Bombay (Vasudeva, 1956). In Lucknow, the affected variety is Co 453, while in Bombay Co 419 showed the disease.

Symptoms :—

On Co 453, the symptoms are stunted grassy shoots, with narrow chlorotic yellowish white leaves, thin stalks, with the leaves of the spindle chlorotic, and premature sprouting of side shoots, with yellowish white leaves (Anonymous, 1956). In general, the affected plants show a bushy appearance, and the growth of these plants is adversely arrested (Vasudeva, 1956).

Transmission :—

The disease appears through infected cane setts, and *Aphis sacchari* is suspected to be the vector for the disease (Vasudeva, 1956).

Control measures :—

Treatment of cuttings in hot water (50°C.) for two hours appreciably reduced the disease (Vasudeva, 1956).

(d) *Line Pattern of Plums* (*Prunus* sp.).

A new virus disease of plums is reported by Bhargava and Bist (1957), which they call "Line pattern". The disease was first noticed in Ranikhet and Ramgarh orchards in Kumaon during 1955.

Symptoms :—

The leaves of the infected plant show yellow or creamy white chlorotic lines often leading to a net-work. Vein-banding is a common feature. Sometimes "oak-leaf" like symptoms are also seen.

Transmission :—

The disease has been shown to be bud-graft transmissible.

(e) *Virus disease of* Solanum jasminoides *Paxt.*

A virus disease of *Solanum jasminoides* has been described from Simla (Pushkarnath, 1952).

Symptoms :—

The disease produces usually a fleeting type of mottle. On White Burley tobacco, complete collapse of the lower leaves occur within 4 to 5 days after inoculation. Vein-banding and necrotic banding along the veins are also observed. On *Nicotiana glutinosa* mottling is the only symptom. On *Solanum nigrum,* floral anomalies are seen. The disease is also transmissible to *Hyocyamus niger, Petunia* and tomato.

Transmission :—

The virus is sap-transmissible.

Physical properties of the Virus :—

The dilution end point of the virus lies between 1:1,900 and 1:2,000. It can withstand 24 to 48 hours' storage at room temperature, and the thermal death point is 60°C. Fifty per cent alcohol inactivates the virus.

Affinities of the Virus :—

From studies on cross-immunity and host range, the virus is found to behave like Potato virus Y in many respects, yet differing from it appreciably. It is believed that the virus occurring on *Solanum jasminoides* plants at Simla has come from South America, the original home of the host plant.

(f) *Virus disease of* Thevetia neriifolia *Juss* :—

A virus disease affecting *Thevetia neriifolia* Juss. is reported from Delhi (Garga, 1953).

Symptoms :—

The leaves of the affected plant show reduction in leaf-size and extreme curling of margins. Fine chlorotic streaks are seen

along the veinlets, giving a unique pattern. During 1948, there was 100% incidence of the disease.

Transmission and Host Range :—

The disease has been grafted by wedge-grafting. In addition to *Thevetia neriifolia*, the disease has been transmitted to *Datura stramonium*, tobacco varieties Harrison's Special, and White Burley, and Sutton's Early Market tomato. In all these hosts, leaf rolling and smalling are seen.

(g) *Yellowing of* Sorghum (*Sorghum* sp.).

Thomas (1937) has reported a virus disease affecting *Sorghum*, characterized by yellowing of leaves. The disease is not sap-transmissible, but is shown to have a vector in *Pundaluoya simplicia* Distant.

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Zea mays	Leaf-curl (413)
Zinnia elegans	Spike (423)
Zizyphus oenoplia Mill.	

TABLE I

Comparison of the physical properties of the sap-transmissible viruses

No.	Virus	Thermal Death Point	Dilution End Point	Longevity *in vitro*	Reference
1.	Bottle gourd mosaic	86-88°C for 10′	1 : 10,000	90 days at 33-35° C	Vasudeva *et al.*, 1954.
2.	„ „ „ (*Cucumis* Virus 2B)	95° for 10′	1.: 1,000,000	308 days	Capoor and Varma, 1948e.
3.	„ „ „ (*Cucumis* virus 2C)	86-88°C for 10′	1 : 1,000 (crude) 1 : 10,000 (pure)	90 days	Vasudeva *et al.*, 1950.
4.	„ „ „ (*Cucumis* virus 3)	60°C for 10′	1 : 500	6 hours	Vasudeva and Lal, 1943.
5.	Chilli mosaic	60°C for 10′	1 : 30,000	22 days	Raychaudhuri and Jha, 1954.
6.	Cowpea mosaic	85°C for 10′	1 : 50,000 to 1 : 10,000	9 days at 24°C	Capoor *et al.*, 1947.
7.	„ „	——	1 : 50,000 to 1 : 100,000	19 days at 24°C	Capoor and Varma, 1956.
8.	*Datura* Distortion mosaic	60°C for 10′	1 : 10,000	13 days	Capoor and Varma, 1948b.
9.	*Dolichos* enation mosaic	90°C for 10′	1 : 5,000,000 to 1 : 3,000,000	——	Capoor and Varma, 1948a.
10.	Egg plant mosaic	60-70°C	1 : 1,000	15 days at 35°F	Anonymous, 1954c.
11.	Lettuce Yellow mosaic	86°C for 10′	1 : 60,000 (crude)	60 days at 15-25°C	Vasudeva *et al.*, 1948.

12. *Nicotiana glauca* virus (*Cucumis* virus 1)	57°C for 10′	1 : 100	7 days at 6-8°C	Nariani and Singh, 1952.
13. Papaya mosaic	53°C for 10′	1 : 1,000	28 hours	Capoor and Varma, 1948d.
14. *Phaseolus mungo* mosaic	60°C for 10′	1 : 2,000	26 hours at 15-37°C	Anonymous, 1952a.
15. Potato X	75°C for 10′	1 : 100,000	234 days	Vasudeva and Lal, 1945a.
16. Potato Y	50-54°C	1 : 1,000	24 hours	Vasudeva and Lal, 1945b.
17. Radish mosaic	85°C for 10′	1 : 10,000,000	17 days at 17-22°C 101 days at 6-8°C	Raychaudhuri and Pathanian, 1956.
18. Sann hemp mosaic (*Crotalaria juncea*)	68-70°C	1 : 1,000 to 1 : 5,000	71-76 days	Raychaudhuri, 1947a.
19. Sann hemp mosaic (*C. mucronata*)	80-85° C	1 : 20,000 to 1 : 30,000	134-142 days at 12-15°C	Raychaudhuri and Pathanian, 1950.
20. (Southern) Sann hemp mosaic (*C. juncea*)	90° C for 10′	1 : 100,000	557 days	Capoor, 1950.
21. Sann hemp small leaf	85-90°C	1 : 25,000 to 1 : 30,000	41-50 days	Raychaudhuri, 1948.
22. Sugarcane mosaic virus X	65°C	1 : 100	6 hours	Chona, 1944.
23. „ „ „ Y	55°C	1 : 100	6 hours	
24. „ „ „ Z	45°C	1 : 50	2 hours	
25. *Solanum jasminoides* virus	60°C	1 : 1,900 to 1 : 2,000	24 to 48 hours	Pushkarnath, 1952.
26. Tomato latent virus	59°C for 10′	1 : 10,000	24 hours	Vasudeva and Samraj, 1947.

TABLE II

Insect vectors and the diseases transmitted by them.

Insect vector	Disease transmitted	Reference
Aphis gossypii Glover	Chilli mosaic	Raychaudhuri and Jha, 1954.
	Cowpea mosaic	Capoor and Varma, 1956.
	Datura inoxia Distortion mosaic.	Capoor and Varma, 1952.
	Maize mosaic	Anonymous, 1954d.
	Papaya mosaic	Capoor and Varma, 1948d.
A. maidis Fitch.	Maize mosaic	Anonymous, 1954d.
	? Sugarcane mosaic	Sundararaman, 1928a.
A. malvae Koch.	Papaya mosaic	Capoor and Varma, 1948d
A. medicaginis Koch. (*A. craccivora* Koch)	Cowpea mosaic	Capoor *et al.*, 1947.
		Capoor and Varma, 1956.
A. sacchari	? Sugarcane grassy shoot	Vasudeva, 1956.
Bemisia tabaci Gen. (*B. gossypiperda* Misra & Lamba)	Bhendi Yellow Vein mosaic	Varma, 1952.
	Dolichos Yellow mosaic	Capoor and Varma, 1950a.
	Double bean Yellow mosaic	Capoor and Varma, 1948c.
	Hibiscus rosa-sinensis Enation mosaic and Leaf-curl	Anonymous, 1952a; 1954d.
	Papaya leaf-curl	Nariani, 1956.
	Pumpkin Yellow Vein mosaic	Varma, 1955b.
	Potato necrosis	Vasudeva and Azad, 1948.
	Sann hemp phyllody	Anonymous, 1952a.
	Tobacco leaf-curl	Pruthi and Samuel, 1937.
	Tomato leaf-curl	Vasudeva and Samraj, 1948.
	Zinnia leaf-curl	Mathur, 1932.
Deltocephalus sp.	Sann hemp phyllody	Anonymous, 1955d.
	Sesame phyllody	Vasudeva and Shambi, 1956.
Empoasca devastans Distant.	Egg plant little leaf	Thomas and Krishnaswamy, 1939a.
	Egg plant mosaic	Raychaudhuri, 1947b.

Insect vector	Disease transmitted	Reference
Eutettix phycitis	Egg plant little leaf	Thomas and Krishnaswamy, 1939a.
	Vinca rosea little leaf	Anonymous, 1954b.
Jassus indicus Walk.	Sandal spike	Rangaswami and Griffith, 1941.
Macrosiphum sp.	? Sandal spike	Dover and Appanna, 1934.
Moonia albimaculata	? Sandal spike	Dover and Appanna 1934.
Myzus persicae Sulz.	Cowpea mosaic	Capoor and Varma, 1956.
	Datura inoxia Distortion mosaic	Capoor and Varma, 1952.
	Egg plant mosaic	Anonymous, 1954c.
	Maize mosaic	Anonymous, 1954d.
	Papaya mosaic	Capoor and Varma, 1948d.
Pentalonia nigronervosa Coq.	*Amomum* mosaic	Anonymous, 1955b.
	Banana bunchy top	Varasi, 1954.
	Banana mosaic	Kamat, 1955.
	Cardamom mosaic	Varma and Capoor, 1953.
	Foorkey disease	Anonymous, 1955e.
Pundaluoya simplicia Distant.	*Sorghum* yellow	Thomas, 1937.

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