GENERAL SCIENCE

FOR COLONIAL SCHOOLS

BY

F. DANIEL, B.Sc.

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PREFACE

This course of ‘ General Science ’ is the result of fifteen years of trial and experiment in two tropical parts of the British Empire. Its first aim is to teach science on a broad basis, as a part of the equipment of the average secondary school pupil for everyday life in a scientific age. I have tried to provide a Course which, although it is ‘ general ’ in touching everyday life at many points, is still ‘ science ’ in that it emphasizes fundamental principles. While the Course is primarily intended for that great majority of secondary school pupils whose schooling stops at the School Certificate stage, it will benefit others who intend to specialize in some corner of the scientific field, by giving an earlier survey of the whole of it.

During the last ten years, every lesson of the Course has been tried out thoroughly with several thousands of pupils (mainly Chinese, Indians, and Malays) in six different Malayan secondary schools. It has been taught by a dozen science teachers of different races, qualifications and experience ; and although the books range over almost the whole field of pure science, the keen teacher who has specialized in one branch of science has proved that he can easily teach himself sufficient of the other branches to handle the Course effectively.[1]

The Course is intended to occupy twenty-four year-periods, and it is recommended that this time-allowance should be allotted as follows: First Year, 5 three-quarter hour periods weekly; Second Year, 5 periods; Third Year, 7 periods; Fourth Year, 7 periods. Some classes have actually completed the Course in only twenty year-periods, but the rate of progress depends largely on the standard of English attained by the pupils. The present division into four one-year parts is that which has proved most suited to the mental outlook and everyday experience of Malayan pupils. After passing through several editions in its earlier, duplicated, loose-leaf form, the Course is now published in book form in the hope that it may also be of use to secondary schools in other tropical and sub-tropical areas.

A special feature of the Course is the very careful attention that has been devoted to language simplification by eliminating non-essential technical words and relating the text to the standard vocabulary of 2,000 words adopted by the Oxford University Press as the basis for colonial school books. This removes one of the main reasons for the pupil's reliance on mere memory—the lowest of the human mental

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[1] An account of the early stages of this experiment appeared in *Oversea Education* for Oct. 1936, Vol. VIII, No. 1, pp. 1-13.

functions—for once he *understands* the ideas involved he can express them in his own words instead of using imposing technical tags as a substitute for real knowledge: merely ‘ labelling his ignorance ’. In order to help the pupil in reading the text and building up his scientific vocabulary, the first occurrence of new words not in the standard vocabulary is marked either by an asterisk or by a dagger. The asterisk refers him to the glossary, where general words are simply defined. The dagger reference enables the pupil by consulting the list of scientific terms whenever necessary to remind himself of their nature or meaning.

In the biological portions of the Course, the physiological approach has been followed throughout, and plant and animal types which are of very wide distribution have been selected with a view to making the Course suitable for use in most tropical and sub-tropical areas. The individual teacher, of course, can always supply additional ‘ local colour ’.

The practical work involves only the simplest and cheapest laboratory equipment for adequate results.

One of the main utilitarian aims of the Course is to provide a background for effective education in health, and since Hygiene (an applied science) can be little more than rule-of-thumb without a proper scientific basis, it is clearly desirable to teach it as a part of a general science course. ‘ Health Science for Tropical Schools ’ is published as Book IIIA of this series.

The Teachers’ Handbook discusses the scope and teaching of General Science, suggests a plan for the construction and equipment of a General Science room and store, and provides running notes on the five books comprising the Course, together with a bibliography and examination papers.

My thanks are due to all those who have helped me in any way with this educational experiment. To my friend and colleague Mr. E. H. Bromley for his many years of conscientious trial and experiment with this Course and for all his suggestions and constructive criticism. To my old pupil, Leong Pak Cheong, who has drawn most of the diagrams from my black-board diagrams and rough sketches. To Cheong San Thau and those senior boys of the Victoria Institution who have helped in working out the biological types, and also to Too Chee Choo, Keshmahinder Singh and A. Krishnappah for their continued interest and help after leaving school. To my assistants, Ahmad bin Haji Osman and Mohamed Basir bin Haji Abdullah, whose patience, skill and industry made it possible to produce the 10,000 loose-leaf text-books needed to carry out the experimental stages of the Course.

I am also grateful to Professor R. Douglas Laurie and to his Senior Lecturer, Dr. E. E. Watkin, for all their help, guidance and encouragement during my two periods of ‘ study-leave ’ in the Zoology Department at Aberystwyth in 1933-4 and 1937-8. To Professor V. H. Mottram (London), Professor H. Munro Fox, F.R.S. (Birmingham), Mr. T. L. Green (Bristol University) and Mr. G. N. Ridley (Wallasey Grammar School), who have given me many new ideas connected with various aspects of animal physiology. To the Board of Education Inspectors and the heads of secondary, central and elementary schools, for all the trouble they have taken to show me new developments in school science teaching during my visits to schools in England and Wales whilst on leave. To my friends in the Malayan Medical and Agricultural Services for giving me the benefit of their local knowledge. To Dr. L. Faucett and Mr. E. C. Parnwell for advice and encouragement in my problem of language-simplification. And especially to my friends, Messrs. W. G. Greaves and J. A. Lauwerys for reading through the manuscript and for many helpful suggestions concerning ‘ General Science ’, a subject on which they are recognized authorities.

The table on p. 310 was very kindly supplied by my friend Mr. F. L. Llewellyn, of the F.M.S. Electrical Department, to whom I am also indebted for reading through the Magnetism and Electricity of this Science Course.

I thank the Local Examinations Syndicate of the University of Cambridge for permission to include a number of questions set in their examinations.

F. DANIEL

Kuala Lumpur
July 1941

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Fig. 108 has been redrawn from a figure in W. H. White: *A Complete Physics*, published by Messrs. George Bell & Sons, Ltd. Owing to the author's absence abroad it had not been possible at the time of going to press to secure his personal permission.

SECTION I

CHEMISTRY

CHAPTER I

CHEMICAL THEORY

THE ATOMIC THEORY

When we studied the general characteristics of solids, liquids and gases (Book II, p. 49), we found that we could explain their physical behaviour by considering matter to be made up of tiny particles called *molecules*. Over 2,000 years ago, some of the Greeks believed that if any lump of matter were divided and sub-divided, again and again, into smaller and smaller pieces, the process of division would finally reach a limit: because matter was made up of exceedingly small, indivisible particles which they called *atoms*. This *theory of atoms* was merely an intelligent guess because the Greeks carried out no experiments to test the correctness of their ideas, and it is only in recent years that scientists have proved the actual existence of atoms and molecules and have measured their size and weight.

This early theory of atoms was of little practical use until about 1808, when *John Dalton* stated his *atomic theory* in such a way that it explained the experimental facts of chemical combination. The chief points in *Dalton's Atomic Theory* were:

(1) *All matter is composed of exceedingly small particles called atoms.*
(2) *Atoms are indivisible during chemical reactions and can be neither created nor destroyed by chemical processes.*
(3) *All the atoms of the same element are exactly alike in every way* (e.g. they have the same weight) *but they are different from the atoms of other elements.*
(4) *Chemical compounds are formed by the combination of two or more atoms of different elements.*
(5) *Chemical combination takes place between simple whole numbers of atoms, e.g.* in the proportions 1 : 1, 1 : 2, 2 : 3, 3 : 4, etc.

Dalton's Atomic Theory is still accepted as the basis of chemical theory, although recent discoveries have shown that some of his statements now need slight modification.

THE LAWS OF CHEMICAL COMBINATION BY WEIGHT

The Laws of Chemical Combination by Weight are:—

(i) *The Law of Conservation of Weight* (or the Law of Indestructibility* of Matter).
(ii) *The Law of Constant Proportions* (or the Law of Constant Composition).
(iii) *The Law of Multiple Proportions.*
(iv) *The Law of Equivalent Proportions* (or the Law of Reciprocal* Proportions).

All these chemical laws, which are based on the results of accurate experiments, can be explained by Dalton's Atomic Theory, and this is why his theory is so important.

THE LAW OF CONSERVATION OF WEIGHT

The French chemist, Lavoisier, from experiments carried out in 1774, concluded that *matter can be neither created nor destroyed during chemical reactions*. Just as the Law of Conservation of Energy is the basis of all quantitative* physics, this Law of Conservation of Weight is the basis of all quantitative chemistry, since it follows that *in all chemical changes the total weight of all the reacting substances is exactly the same as the total weight of all the products of the reaction.*

This law is easily explained by Dalton's suggestion that atoms can be neither created nor destroyed in chemical reactions, and that chemical changes are merely re-groupings of atoms to form different molecules.

THE LAW OF CONSTANT PROPORTIONS

About 1800, the French chemist, Proust, after carrying out a series of quantitative experiments, stated that *the same chemical compound always consists of the same elements combined together in the same proportions by weight.* Proust's results were confirmed later by Stas (1860).

This fact that a pure chemical compound always has the same composition by weight is explained by Dalton's suggestion that all the atoms of the same element are exactly alike in weight. Since a molecule of a chemical compound always consists of the same numbers and kinds of atoms, if the weight of each atom is fixed then the compound must have a constant composition. For example, it is always found that pure water contains 8 parts by weight of oxygen combined with 1 part by weight of hydrogen, no matter what the source of the water may be. It is the same with every other pure substance known to scientists, e.g. whether we make common salt by burning sodium in chlorine, by neutralizing caustic soda with hydrochloric acid, or by dissolving sodium carbonate in hydrochloric acid: the sodium chloride produced will always contain sodium and chlorine in exactly the same proportions by weight. This is because the smallest possible particle of common salt—*one molecule*—contains *one atom* of sodium combined with *one atom* of chlorine; and since all the sodium atoms are of the same weight, and similarly all the chlorine atoms, the proportion of sodium to chlorine must be always the same. (We consider *air* to be a *mixture* and not a *compound* of nitrogen and oxygen, because the proportions of nitrogen and oxygen in the air *are not always exactly the same* at different times and places.)

THE LAW OF MULTIPLE PROPORTIONS

About 1804, Dalton concluded from his Atomic Theory that *when two elements (A and B) combine to form more than one compound, the different weights of one (B) which combine with a fixed weight of the other (A) bear a simple ratio to one another.*

The correctness of this law was afterwards shown by the careful analysis of a large number of compounds. The law agrees with Dalton's suggestion that combination takes place between simple whole numbers of atoms, and since one atom of an element A combines with one or two or three or more atoms of an element B, the different weights of B which combine with a fixed weight of A will bear a simple ratio to each other, e.g. 1 : 2 ; 1 : 3.

The same two elements frequently combine to form more than one compound, e.g. there are two oxides of carbon: (*a*) *carbon monoxide*, formed when carbon burns in a limited supply of air, and (*b*) *carbon dioxide*, formed when carbon burns in a free supply of air. Analysis shows that these two oxides of carbon have this composition:—

	Carbon monoxide	Carbon dioxide
Carbon	1·000	1·000
Oxygen	1·333	2·666

Hence, in *carbon dioxide* there is exactly *twice* as much oxygen combined with one part of carbon as there is in *carbon monoxide*: i.e. in the two oxides of carbon the weights of oxygen that combine with one part of carbon bear a simple ratio to one another, i.e. 1 : 2. This is because a molecule of carbon monoxide contains one atom of carbon combined with one atom of oxygen, while a molecule of carbon dioxide contains one atom of carbon combined with two atoms of oxygen.

To take another example, there are two oxides of sulphur, (*a*) *sulphur* *dioxide*, formed when sulphur burns in air, and (*b*) *sulphur trioxide*,† formed when sulphur dioxide is oxidized (i.e. when it combines with more oxygen). Analysis shows that these two compounds have the following composition:—

	Sulphur dioxide	Sulphur trioxide
Sulphur	2·000	2·000
Oxygen	2·000	3·000

That is, in the two oxides of sulphur, the weights of oxygen that combine with one part by weight of sulphur bear a simple ratio to one another, i.e. 2 : 3. This is because a molecule of sulphur dioxide contains one atom of sulphur combined with two atoms of oxygen, while a molecule of sulphur trioxide contains one atom of sulphur combined with three atoms of oxygen.

Similarly, there are two oxides of copper: *cuprous oxide*† (formed as a red precipitate when testing for sugar with Fehling's Solution), and *cupric oxide*† (the ordinary black oxide). These two oxides have the following composition:—

	Cuprous oxide	Cupric oxide
Copper	1·000	1·000
Oxygen	0·126	0·252

That is, in *cupric* oxide there is *twice* as much oxygen as there is in *cuprous* oxide, or the weights of oxygen which combine with one part of copper bear a simple ratio to one another, i.e. 2 : 1. This is because a molecule of cupric oxide contains one atom of copper combined with one atom of oxygen, while a molecule of cuprous oxide contains two atoms of copper combined with one atom of oxygen. (The ending *-ous* is added to the Latin name of an element to show that the compound contains a smaller proportion of another element. The ending *-ic* shows the higher proportion of another element. For example, cupric oxide contains a higher proportion of oxygen than cuprous oxide. Similarly, sulphuric acid contains a higher proportion of oxygen than sulphurous acid.)

THE LAW OF EQUIVALENT (OR RECIPROCAL) PROPORTIONS

The weights of the elements A and B that combine with a fixed weight of another element C are either the weights in which A and B combine together or simple multiples of these weights. This law follows from the Law of Constant Proportions and the Law of Multiple Proportions. It is explained by Dalton's suggestions (3), (4), and (5).

Some actual examples will make this law clear:—

23 gm. of *sodium* combine with 1 gm. of *hydrogen* to form *sodium hydride.*†

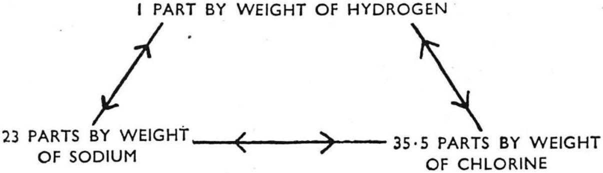


FIG. 1.

35·5 gm. of *chlorine* combine with 1 gm. of *hydrogen* to form *hydrogen chloride.*

23 gm. of *sodium* combine with 35·5 gm. of *chlorine* to form *sodium chloride.*

Hence the weights of sodium and chlorine that combine separately with one part by weight of hydrogen are the weights in which sodium and chlorine combine with each other.

Similarly,

1 gm. of *hydrogen* combines with 8 gm. of *oxygen* to form *water.*

1 gm. of *hydrogen* combines with 3 gm. of *carbon* to form *methane.*†

3 gm. of *carbon* combine with 8 gm. of *oxygen* to form *carbon dioxide.*

Hence the weights of carbon and oxygen that combine separately with one part by weight of hydrogen are the weights in which carbon and oxygen combine with each other.

THE CHEMICAL EQUIVALENTS OF ELEMENTS (COMBINING OR EQUIVALENT WEIGHTS)

The Law of Equivalent Proportions leads to the idea that each element has a definite *chemical equivalent.* In the above examples, it is clear that 1 *part of hydrogen,* 23 *parts of sodium,* 35·5 *parts of chlorine,* 8 *parts of oxygen, and* 3 *parts of carbon are equal to each other in combining power,* i.e. they are *chemical equivalents.* It is convenient to choose a standard weight of a suitable substance in terms of which the *chemical equivalents* (or *equivalent weights*) of all elements (and compounds) can be expressed, and *one unit weight of hydrogen* is taken as this standard.

Hence, when we say that the *equivalent weight of sodium* is 23, we mean that 23 gm. of sodium will combine with 1 gm. of hydrogen, or will displace 1 gm. of hydrogen from its compounds.

The chemical equivalent (or equivalent weight) of an element is that weight of it which combines with, or takes the place of, one part by weight of hydrogen.

Notice that the chemical equivalent of an element is not a weight but merely a *ratio*, i.e. a number.

The equivalent weight of an element can be found experimentally with great accuracy, and this forms the starting-point for all quantitative work in chemistry. You should find the equivalent weight of one or two elements in the laboratory.

VALENCY

The valency† *of an element is the number of hydrogen (or chlorine) atoms it will combine with or displace.*

For example, one atom of sodium can only combine with or displace *one* atom of hydrogen or chlorine, hence sodium is said to have a valency of one. One atom of zinc, however, can combine with *two* chlorine atoms, hence zinc has a valency of two. One atom of aluminium can combine with three chlorine atoms, hence the valency of aluminium is three. In the same way, carbon has a valency of four, phosphorus has a valency of five, and sulphur has a valency of six. No element has a higher valency than eight.

Some elements show more than one valency and so have more than one equivalent weight, e.g. sulphur shows a valency of 4 (and an equivalent weight of 8) in sulphur dioxide, and a valency of 6 (and an equivalent weight of 5·3) in sulphur trioxide. Copper has a valency of 1 (and an equivalent weight of 63·6) in cuprous oxide, and a valency of 2 (and an equivalent weight of 31·8) in cupric oxide.

We shall find that there is a simple relation between the *valency*, the *equivalent weight*, and the *atomic weight* of an element, thus:—

$$\text{VALENCY} = \frac{\text{ATOMIC WEIGHT}}{\text{EQUIVALENT WEIGHT}}$$

This gives us another definition of *valency*:—*The number of chemical equivalents in the atomic weight of the element.*

A knowledge of valencies will be very useful when we come to use chemical formulae.

MOLECULES

An atom is the smallest particle of an element that can take part in a chemical change. Very few atoms, however, can exist singly, by themselves, in a free state, but they usually combine together to form *molecules*.

A molecule is the smallest particle of a substance that can exist in a free state.

For example, one *atom* of sodium combines with one *atom* of chlorine to form one *molecule* of common salt. One atom of oxygen combines with two atoms of hydrogen to form one *molecule* of water. One atom of carbon combines with two atoms of oxygen to form one *molecule* of carbon dioxide. If the atoms in a molecule are of different kinds, then the substance is a *compound*. If the atoms in a molecule are all of the same kind, then the substance is an *element*, e.g. hydrogen and oxygen *molecules* each contain *two atoms*.

CHEMICAL SYMBOLS

Every element has a name, and, for convenience, a *symbol*,* which is usually the first letter of its English or Latin name, e.g. the symbol for *carbon* is *C*, *nitrogen* is *N*, *sulphur* is *S*, *hydrogen* is *H*, *iodine* is *I*, *oxygen* is *O*, *phosphorus* is *P*, *potassium* (Kalium) is *K*.

When the names of several elements all begin with the same letter, two letters are used, e.g. *chlorine—Cl*, *calcium—Ca*, *cobalt—Co*, *magnesium—Mg*, *manganese—Mn*, *nickel—Ni*, *zinc—Zn*, *platinum—Pt*, *Aluminium—Al*, *sodium* (Natrium)—*Na*, *iron* (Ferrum)—*Fe*, *copper* (Cuprum)—*Cu*, *silver* (Argentum)—*Ag*, *gold* (Aurum)—*Au*, *antimony* (Stibium)—*Sb*, *tin* (Stannum)—*Sn*, *mercury* (Hydrargyrum)—*Hg*.

It is to be remembered that these symbols are used *only in chemical equations*. They are *not* shorthand signs for the elements, e.g. Pb does *not* stand for lead in general, any more than $1 means 'money' or 'wealth'. The symbol Pb stands for *one atom of lead*.

CHEMICAL FORMULAE

The next step is to make use of these atomic symbols to represent *molecules*, e.g. a molecule of oxygen contains two atoms of oxygen, and its *formula* is written O_2. Similarly, Cl_2 stands for a molecule of chlorine. (2Cl stands for two atoms of chlorine not combined together. $2Cl_2$ stands for two molecules of chlorine, each consisting of two atoms combined together.) NaCl stands for one molecule of sodium chloride containing an atom of sodium combined with an atom of chlorine. $CaCO_3$ stands for one molecule of calcium carbonate, consisting of one atom of calcium combined with one atom of carbon and three atoms of oxygen. A molecule of *carbon monoxide* contains one atom of carbon combined with one atom of oxygen, therefore its formula is CO. A molecule of *carbon dioxide* contains one atom of carbon combined with two atoms of oxygen, therefore its formula is CO_2. Your teacher will show you how to make up chemical formulae for the chemical compounds you have met in this Course.

ATOMIC WEIGHTS

According to Dalton's Atomic Theory, all the atoms of an element are equal in weight, but differ in weight from the atoms of other elements. A single atom, however, is so exceedingly small that the chemists made no attempt to weigh *one atom*; but they found the *relative atomic weights* of the elements, finding *by how many times an atom of each element was heavier than an atom of hydrogen*, the lightest element. The atomic weight of an element, therefore, like its chemical equivalent, is merely a *ratio*:—

$$\text{ATOMIC WEIGHT} = \frac{\text{WEIGHT OF 1 ATOM OF ELEMENT}}{\text{WEIGHT OF 1 ATOM OF HYDROGEN}}$$

For example, the oxygen atom is 16 times heavier than the hydrogen atom, so that if the *atomic weight of hydrogen* $=1$, then the *atomic weight of oxygen* $=16$. When we say, for example, that the atomic weight of sulphur is 32, we do not mean that we have actually weighed an atom of sulphur, but that an atom of sulphur is 32 times heavier than an atom of hydrogen, or twice as heavy as an atom of oxygen. A list of the atomic weights of some of the commoner elements is given on page 9.

We have learnt that there is a simple relation between the *valency*, the *equivalent weight*, and the *atomic weight* of an element, and that the atomic weight is either the same as the equivalent weight or some simple multiple of it, i.e. *atomic weight* $=$ *equivalent weight* $\times$ *valency*. Atomic weights, therefore, are fixed by first finding the equivalent weight very accurately, and then finding the valency (which is always a simple whole number between 1 and 8) and multiplying the two together.

MOLECULAR WEIGHTS

The *molecular weight* of a compound is simply the sum of the atomic weights of the atoms in one molecule of the compound, e.g. since the atomic weight of sodium is 23 and the atomic weight of chlorine is 35·5, the molecular weight of sodium chloride, NaCl, is $23+35\cdot5=58\cdot5$. To take another example, the formula of sulphuric acid is H_2SO_4: since $H=1$, $S=32$, and $O=16$, the molecular weight of sulphuric acid is $(1\times2)+32+(16\times4)=98$.

English Name	Latin Name	Symbol	Equivalent Weight	Valency	Atomic Weight
ALUMINIUM		Al	9	3	27
CALCIUM		Ca	20	2	40
CARBON		C	3	4	12
COPPER	CUPRUM	Cu	31·8 or 63·6	1 or 2	63·6
CHLORINE		Cl	35·5	1	35·5
HYDROGEN		H	1	1	1
IRON	FERRUM	Fe	28 or 18·67	2 or 3	56
LEAD	PLUMBUM	Pb	207, 103·5 or 51·75	1, 2 or 4	207
MAGNESIUM		Mg	12·15	2	24·3
MERCURY	HYDRARGYRUM	Hg	200·6 or 100·3	1 or 2	200·6
NITROGEN		N	4·67 or 2·8	3 or 5	14
OXYGEN		O	8	2	16
PHOSPHORUS		P	10·33 or 6·2	3 or 5	31
POTASSIUM	KALIUM	K	39	1	39
SILICON		Si	7·075	4	28·3
SODIUM	NATRIUM	Na	23	1	23
SULPHUR		S	16 or 8 or 5·33	2, 4 or 6	32
TIN	STANNUM	Sn	59·5 or 29·75	2 or 4	119
ZINC		Zn	32·7	2	65·4

The molecular weight of a compound is the number of times a molecule of the compound is heavier than one ATOM of hydrogen, i.e. :—

$$\text{MOLECULAR WEIGHT} = \frac{\text{WEIGHT OF ONE MOLECULE OF COMPOUND}}{\text{WEIGHT OF ONE ATOM OF HYDROGEN}}$$

CHEMICAL EQUATIONS

Just as we can use formulae to represent atoms and molecules, we can use *chemical equations* to represent *chemical changes* or *chemical* *reactions.* For example, when zinc dissolves in dilute sulphuric acid, hydrogen is given off, and zinc sulphate remains behind. We can represent this reaction as follows:—

$$Zn + H_2SO_4 \longrightarrow ZnSO_4 + H_2$$

This equation tells us (*a*) *that under certain conditions, not discoverable from the equation, zinc reacts with sulphuric acid so as to form zinc sulphate and hydrogen*;

(*b*) *that one molecule of zinc reacts with one molecule of sulphuric acid to give one molecule of zinc sulphate and one molecule of hydrogen*;

(*c*) *knowing the atomic weights,* $Zn = 65$, $H = 1$, $O = 16$, $S = 32$, *the equation tells us that* 65 *grams, pounds or tons of zinc react with* 98 *grams, etc., of sulphuric acid, to give* 161 *grams, etc., of zinc sulphate and* 2 *grams, etc., of hydrogen*:—

$$\begin{array}{ccccccc} Zn & + & H_2SO_4 & \longrightarrow & ZnSO_4 & + & H_2 \\ & & 2+32+64 & & 65+32+64 & & \\ 65 & & =98 & & =161 & & 2 \end{array}$$

(Hence, if we know the atomic weights, we can work out the quantities of substances taking part in a chemical change.)

This equation does not tell us anything about the conditions under which the reaction takes place, e.g. whether it is necessary to heat or not, whether the acid is concentrated or dilute, what is the speed of the reaction, or whether a catalyst is necessary. Hence, in studying chemical reactions, it is not enough merely to remember equations; they tell us a lot, but not everything.

CHEMICAL EQUATIONS FOR REACTIONS WE HAVE STUDIED ALREADY

(*Copy these into Books I and II.*)

When iron combines with sulphur on heating, the equation is:

$$Fe + S = FeS$$

The following equations represent the burning of various elements in air or oxygen:—

$$S + O_2 = SO_2 \qquad C + O_2 = CO_2 \qquad 2Mg + O_2 = 2MgO \qquad 2Ca + O_2 = 2CaO$$

$$4P + 5O_2 = 2P_2O_5 \qquad 3Fe + 2O_2 = Fe_3O_4 \text{ (magnetic oxide of iron)}$$

The equation which represents the preparation of oxygen from potassium chlorate by heating is $2KClO_3 = 2KCl + 3O_2$. If manganese dioxide is used as a catalyst, it is not necessary to mention it in the equation because it remains unchanged at the end of the reaction, and would therefore appear on both sides of the equation and thus ‘ cancel out ’.*

$2H_2O \longrightarrow 2H_2 + O_2$ represents the *analysis* of water, while $2H_2 + O_2 \longrightarrow 2H_2O$ represents the *synthesis* of water. If we want to show both these reactions in one equation, we write $2H_2 + O_2 \rightleftharpoons 2H_2O$ which means that the change is *reversible*.

The following equations are examples of the *displacement* of one element by another:—

$Fe + CuSO_4 = FeSO_4 + Cu$	(The displacement of copper by iron)
$Zn + CuSO_4 = ZnSO_4 + Cu$	(The displacement of copper by zinc)
$Zn + H_2SO_4 = ZnSO_4 + H_2$	(The displacement of hydrogen by zinc)
$Fe + H_2SO_4 = FeSO_4 + H_2$	(The displacement of hydrogen by iron)
$Mg + H_2SO_4 = MgSO_4 + H_2$	(The displacement of hydrogen by magnesium)
$Zn + 2HCl = ZnCl_2 + H_2$	
$Fe + 2HCl = FeCl_2 + H_2$	
$2Na + 2H_2O = 2NaOH + H_2$	(The action of sodium on water)
$2K + 2H_2O = 2KOH + H_2$	
$Ca + 2H_2O = Ca(OH)_2 + H_2$	
$Mg + H_2O = MgO + H_2$	(The action of magnesium on steam)
$3Fe + 4H_2O = Fe_3O_4 + 4H_2$	(The action of steam on red-hot iron)

Further examples of chemical equations:—

$CaCO_3 = CaO + CO_2$

$CaCO_3 + 2HCl = CaCl_2 + H_2O + CO_2$

$4HCl + MnO_2 = MnCl_2 + 2H_2O + Cl_2$

$KOH + HNO_3 = KNO_3 + H_2O$

$ZnCO_3 + H_2SO_4 = ZnSO_4 + CO_2$

$CaO + H_2O = Ca(OH)_2$

$CaCO_3 + H_2O + CO_2 \rightleftharpoons Ca(HCO_3)_2$

$NaOH + HCl = NaCl + H_2O$

$PbO + 2HNO_3 = Pb(NO_3)_2 + H_2O$

GAY-LUSSAC'S LAW OF COMBINING VOLUMES

When we study chemical reactions between *gases* we find that **gases always combine together in simple proportions by volume.** For example, 2 volumes of hydrogen combine with 1 volume of oxygen to form 2 volumes of steam; 1 volume of hydrogen combines with 1 volume of chlorine to form 2 volumes of hydrogen chloride; 1 volume of nitrogen combines with 3 volumes of hydrogen to form 2 volumes of ammonia gas. All gases behave in a similar way when they combine, one volume of a gas combining with the same, or twice, or three or four times this volume of another gas, but never in odd amounts. The French chemist, Gay-Lussac, studied a large number of reactions between gases, and, about 1808, he came to the conclusion that *when gases combine, they do so in simple proportions by volume.* This is known as *Gay-Lussac's Law of Combining Volumes.*

AVOGADRO'S THEORY

The simple relation which exists between the volumes of combining gases is explained by Avogadro's Theory that **equal volumes of all gases** (under the same conditions of temperature and pressure) **contain the same number of molecules.** This theory, put forward by the Italian chemist, Avogadro, about 1811, at once gives us a clearer idea of how chemical reactions take place. For example, in the combination of hydrogen with chlorine, 1 volume of hydrogen combines with 1 volume of chlorine to form 2 volumes of hydrogen chloride. But, according to

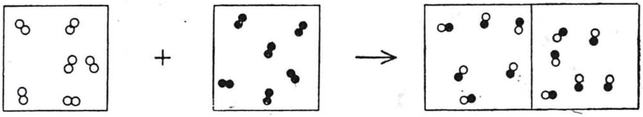


FIG. 2.

1 vol. hydrogen 1 vol. chlorine 2 vols. HCl

Avogadro's Theory, equal volumes of gases contain *the same number of molecules*, therefore *one molecule* of hydrogen combines with *one molecule* of chlorine to form *two molecules* of hydrogen chloride.

Hence, one molecule of hydrogen chloride must contain *half a molecule* of hydrogen and *half a molecule* of chlorine, i.e. the *molecules* of hydrogen and chlorine must each contain at least *two atoms*. This theory connects together Dalton's Atomic Theory and Gay-Lussac's Law of Volumes.

We see, therefore, that our chemical equations also tell us the *volumes* of the gases taking part in chemical changes, for example:

$$\underset{\text{1 vol.}}{N_2} + \underset{\text{3 vols.}}{3H_2} \longrightarrow \underset{\text{2 vols.}}{2NH_3} \quad \Big| \quad \underset{\text{1 vol.}}{H_2} + \underset{\text{1 vol.}}{Cl_2} \longrightarrow \underset{\text{2 vols.}}{2HCl} \quad \Big| \quad \underset{\text{2 vols.}}{2H_2} + \underset{\text{1 vol.}}{O_2} \longrightarrow \underset{\text{2 vols.}}{2H_2O}$$

THE STANDARD VOLUME (OR THE GRAM-MOLECULAR VOLUME)

Since equal volumes of all gases contain equal numbers of molecules, the weights of these equal volumes of gases will be proportional to the *molecular weights* of the gases. Hence, **the molecular weight of any gas, in grams, occupies the same volume** (22·2 litres) **at normal temperature and pressure** (0° C. and 760 mm. of mercury).

(By experiment, a litre of hydrogen weighs 0·09 gm. at N.T.P. But the molecular weight of hydrogen is 2. Hence, 2 gm. of hydrogen occupy $\frac{1\times 2}{0\cdot 09} = 22\cdot 2$ litres, at normal temperature and pressure.

Hence, by Avogadro's Theory, *the gram-molecular volume of any gas will occupy* 22·2 *litres at N.T.P.*)

This fact is very useful when doing chemical calculations. For example, in our reaction $Zn + H_2SO_4 \longrightarrow ZnSO_4 + H_2$, we can now say that 65 grams of zinc react with 98 grams of sulphuric acid to give 161 grams of zinc sulphate and 2 grams *or 22·2 litres of hydrogen (at N.T.P.). Hence a chemical equation also tells us the volumes of gases (at N.T.P.) taking part in the reaction.*

OXIDATION AND REDUCTION

When a substance combines with oxygen, it is said to be *oxidized.* The reverse change, when oxygen is removed from a substance, is called *reduction.*† For example, when lead is heated in air, it forms lead oxide:—

$$2Pb + O_2 \longrightarrow 2PbO$$

This is the *oxidation* of lead to lead oxide. On the other hand, when lead oxide is heated on a charcoal block (with the blowpipe flame), the oxygen is removed from the lead oxide, and metallic lead is left:—

$$PbO + C \longrightarrow Pb + CO$$

This is the *reduction* of lead oxide to lead, and the carbon is called a *reducing agent.* Hydrogen is the commonest reducing agent, e.g.:—

$$CuO + H_2 \longrightarrow Cu + H_2O$$

In some cases the hydrogen merely adds on to the other substance without actually removing oxygen, but the change is still a reduction, e.g. when hydrogen and sulphur are heated together, some hydrogen sulphide (or sulphuretted hydrogen)† is formed:—

$$H_2 + S = H_2S$$

Hence we say that *oxidation is the addition of oxygen to a substance, or the removal of hydrogen from it. Reduction is the addition of hydrogen to a substance or the removal of oxygen from it.*

Most oxidizing agents are substances which are rich in oxygen and which will part with it readily, e.g. potassium nitrate—KNO_3, nitric acid —HNO_3, manganese dioxide—MnO_2, potassium permanganate†—$KMnO_4$, potassium dichromate†—$K_2Cr_2O_7$, hydrogen peroxide—H_2O_2.

Reducing agents are substances with a great liking for oxygen, e.g. hydrogen, carbon, carbon monoxide—CO, sulphur dioxide—SO_2.

TO FIND THE COMPOSITION OF WATER BY WEIGHT

If we heat black copper oxide in a stream of hydrogen, the oxygen of the copper oxide will combine with the hydrogen to form water, leaving metallic copper. In other words, the hydrogen *reduces* the copper oxide to copper, while the oxygen of the copper oxide *oxidizes* the hydrogen to water. This reaction gives us a convenient method for finding the composition of water by weight. It also enables us to find the *chemical equivalent* (or equivalent weight) of oxygen and the equivalent weight of copper in one single experiment:—

$$CuO+H_2 \longrightarrow Cu+H_2O$$

Using the apparatus shown in Fig. 3 (*a*), hydrogen is prepared from zinc and dilute sulphuric acid in the vessel A, and is then dried by passing through the ' U '-tube B, which contains fused calcium chloride. This dry hydrogen then passes through the bent hard-glass tube C and over some weighed copper oxide contained in the porcelain ' boat ' D. When the tube C is heated, the hydrogen reduces the hot copper oxide to metallic copper and the water which is formed is caught in the second ' U '-tube E, which also contains fused calcium chloride. The

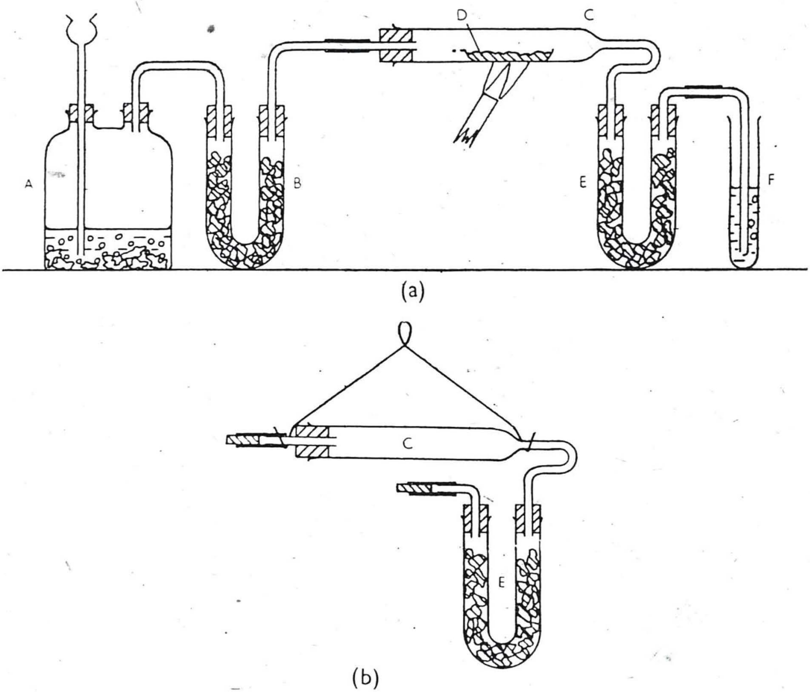


FIG. 3. (*a*) FINDING THE COMPOSITION OF WATER BY WEIGHT
(*b*) WEIGHING THE WATER FORMED

increase in weight of tubes C and E during the experiment gives the amount of *water* formed. The loss in weight of the boat D, containing copper oxide, gives the weight of *oxygen* contained in this weight of water. The difference between the weight of the water and that of the oxygen gives the weight of the *hydrogen* combined with oxygen in this water.

First dry the porcelain boat by heating it in the flame, holding it in the crucible tongs; allow it to cool in the desiccator and weigh it (to the nearest 0·001 gm.), recording the weight below at (*a*). Half fill the boat with black copper oxide, dry it by heating for a few minutes, allow it to cool in the desiccator, weigh and record at (*b*). Also weigh the tubes C and E, with their corks and glass tubes, hanging them on the balance by means of a wire hook (Fig. 3 (*b*)). Record this weight at (*r*). (N.B.—Handle the bent hard-glass tube C and the ‘ U ’-tube E with great care—they are easily broken. Turn the *cork* in the ‘ U ’-tube when the tube C is turned round before weighing.) From the second ‘ U ’-tube E, a glass tube dips into about $\frac{1}{2}$ in. of concentrated sulphuric acid in a test-tube F. This stops water-vapour in the atmosphere from getting into the drying-tube E, and the rate of bubbling also shows the speed at which hydrogen is passing through the apparatus. It also removes the risk of the excess hydrogen catching fire and causing an explosion. (If hydrogen is being given off in A and no bubbles appear in F, then your apparatus is not air-tight.)

Pour about 100 c.cm. of dilute sulphuric acid (6N) on to the zinc in A (after adding a little copper sulphate solution to act as a catalyst) so as to give a steady stream of hydrogen over a long period. (To be on the safe side, cover the vessel A and the ‘ U ’-tube B with a duster in case of an explosion. When you think that all the air has been displaced from the apparatus, collect a small test-tube full of the gas coming through and light it. If it burns quietly, light your burner and heat the hard-glass tube C, gently at first, keeping the flame moving about all the time. When all the black copper oxide appears to have been reduced to reddish brown metallic copper, stop heating, and allow the copper to cool *in a stream of hydrogen.* When the hard-glass tube C is cool, remove the boat (with a wire hook), weigh it and record at (*c*). Then draw air through the tubes C and E and weigh them as before—(*s*).

Weight of empty ‘ boat ’ .. .. =........gm. (*a*)
Weight of boat + copper oxide *before* experiment =........gm. (*b*)
Weight of boat + metallic copper *after* experiment =........gm. (*c*)
Therefore, weight of copper oxide (*b* — *a*) .. =........gm. (*d*)
Therefore, weight of metallic copper (*c* — *a*) .. =........gm. (*e*)

Therefore, weight of oxygen $(d-e)$ or $(b-c)$ .. =........gm. (f)
Weight of tubes C and E *before* experiment .. =........gm. (r)
Weight of tubes C and E *after* experiment .. =........gm. (s)
Therefore, weight of *water* formed $(s-r)$ .. =........gm. (t)
But, weight of *oxygen* in this water $(b-c)$ .. =........gm. (f)
And, weight of *hydrogen* in this water $(t-f)$ .. =........gm.
Therefore, ...gm. of water contain ....gm. of hydrogen and ....gm. of oxygen.
Therefore, 100 gm. of water contain ....gm. of hydrogen and ....gm. of oxygen.[1]

Your experiment also shows that ....gm. of oxygen combines with ....gm. copper. Therefore, 8 gm. of oxygen combine with ....gm. of copper. Since 8 gm. of oxygen combine with 1 gm. of hydrogen, the chemical equivalent of copper is ....

---

[1] The most careful experiments show that water contains 11·1% of hydrogen and 88·9% of oxygen, i.e. 8 gm. of oxygen combine with 1 gm. of hydrogen, or the chemical equivalent of oxygen is 8.

CHAPTER II

SOME COMMON ELEMENTS AND THEIR COMPOUNDS

COMPOUNDS OF NITROGEN

AMMONIA

Ammonia can be prepared by heating any *ammonium salt* with an alkali. Ammonium chloride and calcium hydroxide are generally used for this purpose in the laboratory:—

$$2NH_4Cl + Ca(OH)_2 \longrightarrow CaCl_2 + 2H_2O + 2NH_3$$

Ammonia is a gas which is very soluble in water, hence it cannot be collected over water. As it is much ' lighter ' than air it can be collected by the downward displacement of air, as shown in Fig. 4. If *dry* ammonia is required, it is passed through a drying-tube containing quick-lime.

Mix together 2 gm. of ammonium chloride and 4 gm. of slaked lime and put the mixture in a test-tube. (N.B.—The mixture must not be tightly packed in the test-tube.) Clamp the test-tube in an inclined position and heat gently. Put a few drops of concentrated hydrochloric acid in another test-tube and blow across its mouth towards the delivery-tube.

What happens when the ammonia begins to escape ?......

....

What is formed ?............

....

Write an equation for this reaction........................

Collect the gas in inverted test-tubes as shown in Fig. 4.

First Tube. When the tube is full of ammonia (at which point a piece of red litmus-paper held near the mouth of the tube will turn

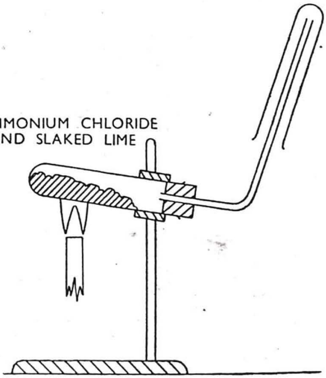


FIG. 4.
PREPARATION OF AMMONIA

blue) close the mouth of the test-tube with the thumb and invert it in a dish of water, removing the thumb when under water.

What happens ? ....

Why ? ....

Test the liquid in the tube with red litmus-paper.

What has been formed ?..................................

Second Tube. Test a tube of the gas with a lighted wooden splinter.

Does ammonia burn ?...................................

Does ammonia support combustion ?.......................

PROPERTIES OF AMMONIA

Ammonia is a colourless gas with a characteristic pungent smell. A dilute solution has a soapy, alkaline taste.[1] It is much ‘ lighter ’ than air. It is the most soluble gas known and its solution in water—*ammonium hydroxide*—is used in the laboratory. Since the solubility of any gas decreases with rise of temperature, the dissolved ammonia is driven off by boiling, and this is an easy way of getting a small quantity of ammonia gas in the laboratory. At ordinary temperatures, ammonia gas can be liquefied by a pressure of 7 atmospheres (100 lb. per sq. in.). When this pressure is reduced to one atmosphere, the liquid ammonia boils at $-34°$ C., and as its latent heat of evaporation is 330 calories per gram, it is used in cooling-machines (see Book III, p. 72).

Ammonia gas does not support combustion, and although it will not burn in air, it will burn *in oxygen*, forming nitrogen and steam:—

$$4NH_3+3O_2 \longrightarrow 2N_2+6H_2O$$

This reaction shows that ammonia is a compound of nitrogen and hydrogen.

If a mixture of ammonia and air is passed over heated *platinum* (which serves as a catalyst) it burns to form oxides of nitrogen which dissolve in water to form nitric acid. The following equation shows the beginning and the end of a complicated series of reactions:—

$$NH_3+2O_2 \longrightarrow HNO_3+H_2O$$

We shall see later that this is the basis of an industrial method for making nitric acid on the large scale.

When ammonia dissolves in water, it forms ammonium hydroxide:—

$$NH_3+H_2O \longrightarrow NH_4OH$$

Ammonia neutralizes acids, forming *ammonium salts* and water, e.g. :—

$$NH_4OH+HCl \longrightarrow NH_4Cl+H_2O$$

In doing so it behaves like other bases.

---

[1] Do not taste any chemical unless directed by the teacher.

Ammonium salts can also be made by direct combination between ammonia gas and an acid, without forming water, e.g. :—

$$NH_3 + HCl \longrightarrow NH_4Cl$$

THE NITROGEN CYCLE

The element nitrogen is very inactive, but it is extremely important when combined with other elements, since it forms an essential constituent of all plant and animal food and enters into the composition of all living things. Very few plants or animals can make direct use of the nitrogen of the air, and nearly all green plants can only make use of *combined nitrogen* in the form of *nitrates*. Animals then make use of

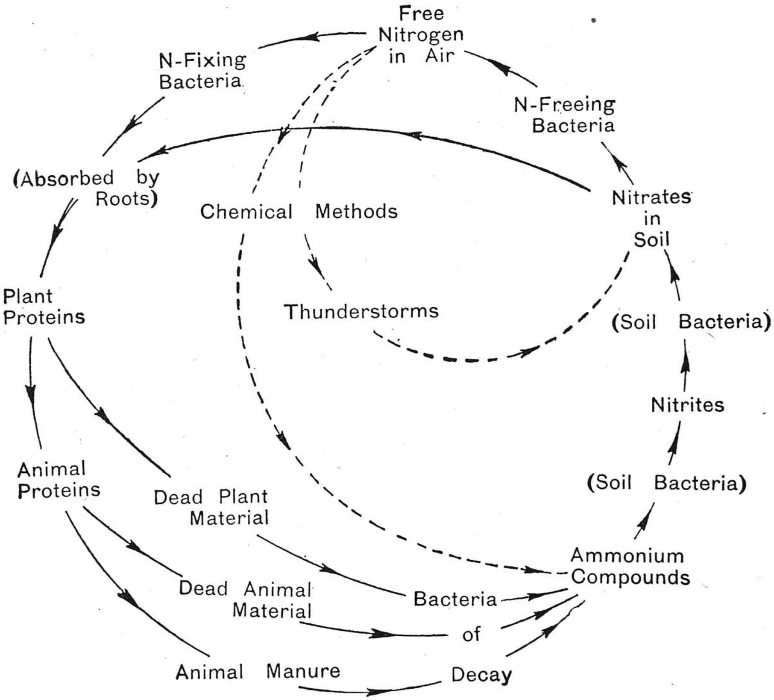


FIG. 5. THE NITROGEN CYCLE

the nitrogen compounds (proteins) built up by green plants, but practically all living things, plants and animals, depend upon soluble nitrates in the first place.

In Nature, most of the nitrogen compounds in plants are returned to the soil when the plants die and decay. *Bacteria* in the soil convert these nitrogen compounds first into *ammonia* and finally into *nitrates* which can be used over again. Hence there is a *Nitrogen Cycle* (see Fig. 5) similar to the Carbon Cycle, and, if there were no losses, all this combined nitrogen could be used over and over again. Actually, much of this combined nitrogen is lost, e.g. nitrates in the soil may be washed deep down out of reach of the roots of plants and, under some conditions, other *nitrogen-freeing bacteria* convert nitrates once more to free, uncombined nitrogen. Hence, in order to make good these losses and to increase the yield of cultivated crops, it is generally necessary to use nitrogenous fertilizers, e.g. ammonium compounds and nitrates.

There is, however, one class of plants which can make use of the free nitrogen of the air. Many *leguminous plants* (i.e. pod-bearing plants) have little swellings on their roots containing *nitrogen-fixing bacteria* which convert free nitrogen from the air into nitrogen compounds. All other plants have to be supplied with nitrates. Besides this, a little *nitric acid* is formed in the air during thunderstorms and is washed down by the rain, but these two natural methods of supplying combined nitrogen can never replace all the nitrogen compounds used up by cultivated crops.

ARTIFICIAL FERTILIZERS (CHEMICAL MANURES)

In order to preserve the balance of the Nitrogen Cycle, combined nitrogen must be supplied to cultivated crops in the form of nitrates or ammonium compounds, and, until methods were found for synthesizing nitric acid and ammonia, the question of the early exhaustion of the world's supplies of soluble nitrogen compounds was a serious problem. For many years the main supply of nitrates came from the rainless parts of South America, where large deposits of *sodium nitrate* are found in Chile. This supply was increased by using the *ammonium sulphate* obtained as a by-product in the manufacture of coal-gas (see p. 51); but even so, in order to produce larger crops to feed the world's rapidly growing population, new sources of soluble nitrogen compounds were necessary, and chemists tried to find methods for converting the nitrogen of the air into nitrogen compounds suitable for the use of plants.

‘FIXING’ ATMOSPHERIC NITROGEN

(1) We have seen that during a thunderstorm small quantities of nitrogen and oxygen combine in the atmosphere. This process has been imitated in Norway (where electricity can be made very cheaply by water-power) by passing air through a large electric arc. Some nitrogen, about 1 per cent, is oxidized to *nitric oxide*[†] which is finally converted into *nitric acid.* This nitric acid is then neutralized with limestone, forming *calcium nitrate* which is used as a fertilizer. This process can only be carried out where there is abundant water-power to supply cheap electricity.

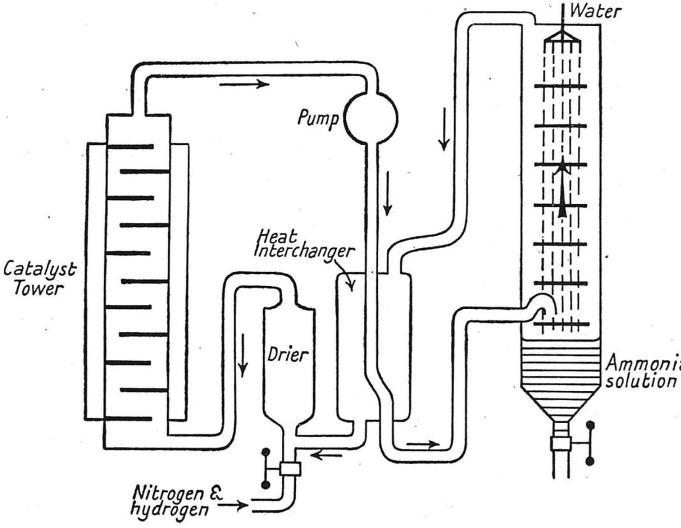


FIG. 6. SYNTHESIS OF AMMONIA

(2) The main modern source of nitrogen compounds is from *synthetic ammonia*, i.e. ammonia made directly from nitrogen and hydrogen. Carefully purified nitrogen (from the air) and hydrogen (from water), under high pressure (200 atmospheres), are passed over a heated catalyst (iron oxide mixed with small quantities of other substances) and they combine together to form *ammonia* (see Fig. 6). Much of this synthetic ammonia is converted into *ammonium sulphate*, which is one of the most widely-used fertilizers. If nitrates are needed, the ammonia can be oxidized to *nitric oxide* by mixing it with air and

passing the mixture over heated platinum, which acts as a catalyst. The nitric oxide then combines slowly with more oxygen to form *nitrogen peroxide* which dissolves in water forming *nitric acid.*

NITRIC ACID

Nitric acid is prepared in the laboratory (and sometimes commercially) by heating *sodium nitrate* (or potassium nitrate) with *concentrated sulphuric acid* :—

$$NaNO_3+H_2SO_4 \longrightarrow NaHSO_4+HNO_3$$

At higher temperatures, the sodium displaces all the hydrogen in the sulphuric acid forming *normal sodium sulphate*:

$$2NaNO_3+H_2SO_4 \longrightarrow Na_2SO_4+2HNO_3$$

but much of the nitric acid is decomposed at this higher temperature, hence only moderate heat is used and the reaction is stopped at the first stage. The sodium hydrogen sulphate (known in the industry as 'nitre-cake') is used for the manufacture of hydrochloric acid and sodium sulphate.

$$NaCl+NaHSO_4 \longrightarrow HCl+Na_2SO_4$$

Much nitric acid is now manufactured by synthetic methods, e.g. by the oxidation of synthetic ammonia.

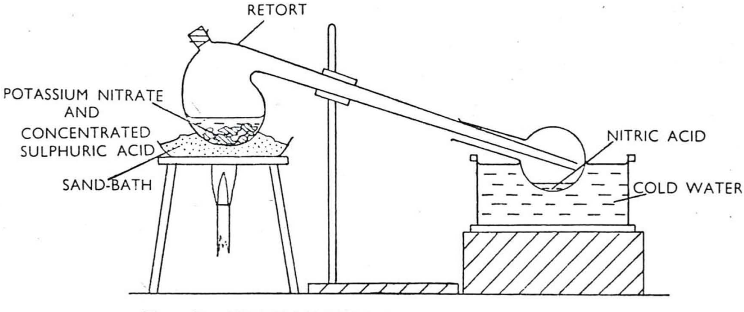


FIG. 7. PREPARATION OF NITRIC ACID

As concentrated nitric acid is one of the most dangerous chemicals used in the laboratory, your teacher will show you how it is prepared, using the apparatus shown in Fig. 7, and putting about 15 gm. of nitrate and 25 c.cm. of concentrated sulphuric acid in the retort.

The properties of nitric acid have already been examined and described in Book II, p. 38.

OXIDES OF NITROGEN

Nitrogen forms at least five oxides, and these five compounds of nitrogen and oxygen provide a good example of the *Law of Multiple Proportions.* (See p. 3). The relative proportions of nitrogen and oxygen in the various oxides are as follows:—

		Nitrogen	Oxygen	
Nitrous oxide† (nitrogen monoxide)†	N_2O	14	8	$=1\times8$
Nitric oxide (nitrogen dioxide)†	NO	14	16	$=2\times8$
Nitrous anhydride (nitrogen trioxide)†	N_2O_3	14	24	$=3\times8$
Nitrogen peroxide† (nitrogen tetroxide)†	N_2O_4	14	32	$=4\times8$
Nitric anhydride (nitrogen pentoxide)†	N_2O_5	14	40	$=5\times8$

Starting with the oxide with least oxygen, we find that *for every* 14 *gm. of nitrogen, the amount of oxygen increases by steps of* 8 *gm.*

Nitrous Oxide. Nitrous oxide is sometimes used by dentists, since it produces unconsciousness and insensibility to pain when breathed. It is prepared by heating ammonium nitrate:—

$$(NH_4).NO_3 \longrightarrow 2H_2O + N_2O$$

It is a colourless gas with a sweetish smell, and, like oxygen, it makes a red-hot wooden splinter burst into flame; but, unlike oxygen, it does not form brown fumes with nitric oxide (see below).

Nitric Oxide. Nitric oxide is obtained when copper is treated with ordinary concentrated nitric acid diluted with an equal volume of water:—

$$3Cu + 8HNO_3 \longrightarrow 3Cu(NO_3)_2 + 2NO + 4H_2O$$

Nitric oxide is a colourless gas which combines readily with oxygen forming brown fumes of nitrogen peroxide:—

$$2NO + O_2 \longrightarrow 2NO_2$$

Nitrogen Peroxide. Nitrogen peroxide is most easily prepared by heating lead nitrate:—

$$2Pb(NO_3)_2 \longrightarrow 2PbO + 4NO_2 + O_2$$

It is separated from the oxygen formed at the same time by passing the mixture through a U-tube immersed in a freezing mixture, when the nitrogen peroxide condenses to a liquid.

EXPLOSIVES

Explosives are substances which, by chemical action, suddenly produce large quantities of heated gases which exert an enormous pressure and hence cause an *explosion.*

The oldest explosive is *gunpowder*, a mixture of *nitre* (potassium nitrate), *sulphur*, and *carbon* (charcoal). In the open it burns fiercely with a bright flash, and only explodes if it is placed in an enclosed space and then fired.

Gunpowder has been replaced by more modern explosives which are all *nitro-compounds* formed by the action of nitric acid on organic substances. A typical example of this type of explosive is formed when *glycerine* is treated with nitric acid, forming *nitro-glycerine*,† an oily liquid which explodes with terrible violence when given a sudden shock. It is much safer when absorbed in porous earth, and the mixture is then called *dynamite*.† *Gun-cotton*,† another very powerful explosive, is made by the action of nitric acid on cotton. *Blasting gelatine*† is a mixture of nitro-glycerine and guncotton. *Cordite*,† a much safer explosive used in rifle cartridges,* is a mixture of gun-cotton, nitro-glycerine and vaseline (petroleum jelly).

The explosives used in big guns are of two different types: (1) *propellents*,* e.g. gunpowder and cordite, which burn comparatively slowly and push the shell or bullet out of the barrel without bursting the gun; and (2) *high explosives*, e.g. T.N.T. (tri-nitro-toluene),† which burn so rapidly that they exert a shattering effect when the shell reaches its objective. These high explosives then require a small explosion to set them off; they do not explode when a light is applied to them.

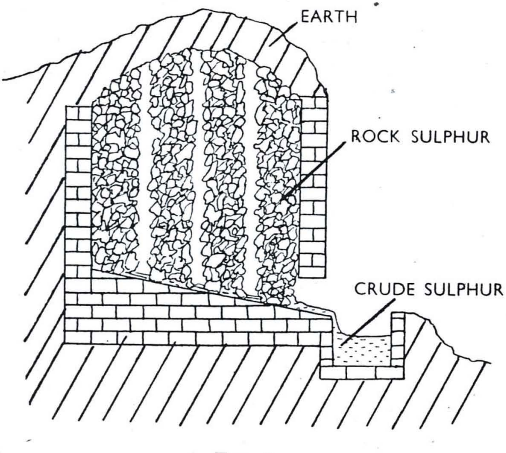


FIG. 8.
EXTRACTION OF SULPHUR (SICILY)

SULPHUR AND ITS COMPOUNDS

Sulphur is found near the surface of the earth in regions of volcanic* activity, e.g. in Sicily, Italy, Japan, and Iceland. There is also a big deposit of sulphur in Louisiana, U.S.A., at a depth of 700 to 900 feet, but this is not thought to have been formed by volcanic action.

THE EXTRACTION OF VOLCANIC SULPHUR

In Sicily, the sulphur is mixed with earth and rock, and the sulphur ore is piled in heaps on a sloping floor, covered with earth, and is then lighted. Part of the sulphur (about one-third) burns away, melting the remainder which flows down the sloping floor into moulds, where it solidifies. Although this may seem a very wasteful method, sulphur is a cheaper fuel in Sicily than coal or wood.

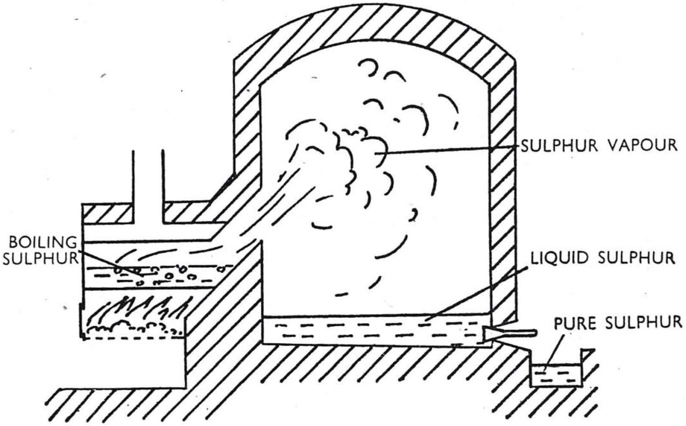


FIG. 9. PURIFICATION OF SULPHUR BY DISTILLATION

This crude sulphur is further purified by *distillation*, usually in a country where fuel is cheaper. The crude sulphur is melted and then boiled in iron retorts, from which the vapour passes into large brick-work chambers where it condenses, first to a fine, pale-yellow powder, called '*flowers of sulphur*', when the chamber is cold, and later to liquid sulphur when the condensing chamber becomes hot. This liquid sulphur is run off into moulds and cast into cylindrical sticks—'*roll sulphur*'.

THE EXTRACTION OF AMERICAN SULPHUR

In Louisiana, where the sulphur is very deep down in the earth, the sulphur is extracted by the *Frasch process* (so called after the inventor).

Four pipes, of diameter 1 in., 3 in., 6 in., and 8 in., are sunk into the earth, one inside the other. Hot water *superheated* to about 170° C. is sent down the two outside tubes to melt the sulphur. Compressed air

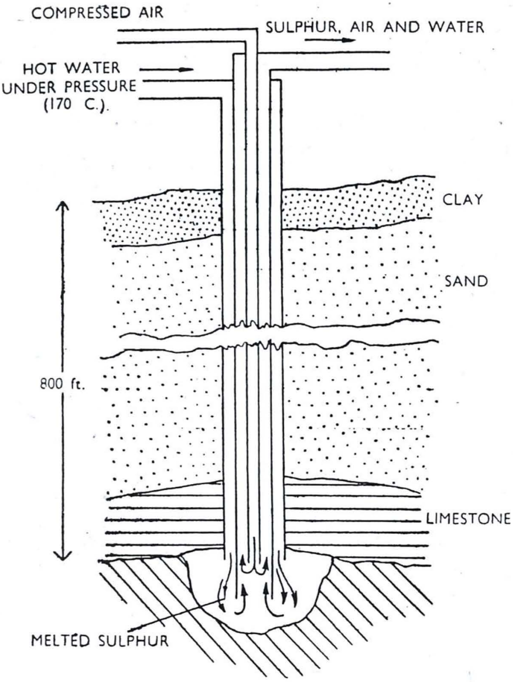


FIG. 10. FRASCH PROCESS FOR EXTRACTING SULPHUR

is then blown down the central 1 in. tube, and the mixture of hot water and liquid sulphur is forced to the surface through the 3 in. pipe. This sulphur, obtained by the Frasch method, is almost pure and does not require any further refining.

ALLOTROPIC FORMS OF SULPHUR

The element sulphur can exist in several different forms called *allotropes*† or *allotropic* forms of sulphur. If roll sulphur is dissolved in carbon disulphide and the solution is allowed to evaporate, crystals of *ordinary sulphur* (or *α-sulphur*) are obtained (Fig. 11 (*a*)). This is one of the allotropic forms of sulphur, and it is the stable form at ordinary temperatures, the other forms of sulphur turning into α-sulphur after standing for some time.

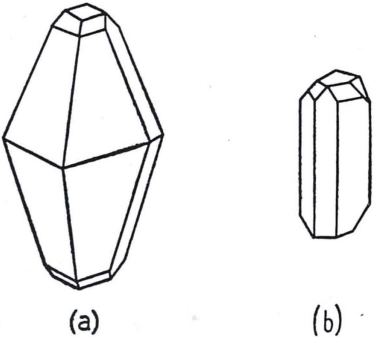


FIG. 11.
(*a*) CRYSTAL OF α-SULPHUR
(*b*) CRYSTAL OF β-SULPHUR

Needle-sulphur, or β-sulphur, a second allotropic form, is made by melting some roll sulphur in a large crucible and then allowing it to cool until a crust just begins to form on the surface. Two holes are pierced in this crust and the still liquid part is then poured off. When the crust is removed, the inside of the crucible is found to be lined with transparent, needle-shaped crystals of β-*sulphur* (Fig. 11 (*b*)). On standing for a few days, the crystals become brittle*, and crumble into little crystals of α-sulphur.

Plastic sulphur*, a third allotropic form of sulphur, is made by heating sulphur to near its boiling point (444° C.) and then pouring the hot, liquid sulphur in a thin stream into cold water. Soft threads of plastic sulphur are formed, which can be moulded between the fingers. This form of sulphur is unstable* and becomes hard and brittle on standing, gradually changing back into α-sulphur.

USES OF SULPHUR

The element sulphur is of very great everyday importance. Since it burns so readily it is used for making fireworks, gunpowder, and matches. Sulphur is used in large quantities for vulcanizing† (or hardening) rubber. Raw rubber and sulphur are heated together, and the higher the proportion of sulphur the harder is the final product. *Vulcanite*,† for example, is rubber which has been hardened in this way with a large quantity of sulphur. Sulphur is also used in large quantities as a *fungicide* for treating plant-diseases (and even human skin-diseases) caused by the growth of a *fungus*.

COMPOUNDS OF SULPHUR

HYDROGEN SULPHIDE (OR SULPHURETTED HYDROGEN) H_2S

Sulphides, when treated with dilute acids, yield hydrogen sulphide, a poisonous gas smelling like bad eggs. For example:—

$$FeS + 2HCl \longrightarrow FeCl_2 + H_2S$$

Hydrogen sulphide is also formed whenever proteins containing sulphur decay.

In the laboratory, hydrogen sulphide is usually prepared by the action of dilute hydrochloric acid on iron sulphide, as shown by the above equation. A *Kipp's apparatus* (Fig. 12) is generally used, so that the

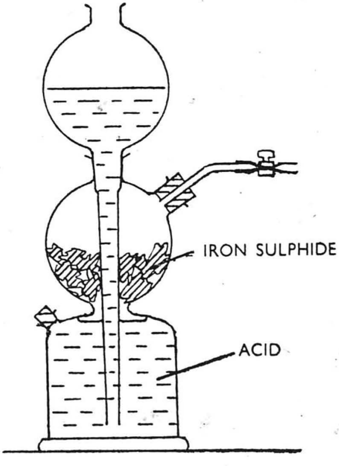


FIG. 12. KIPP'S APPARATUS

supply of this unpleasant and poisonous gas may be controlled. Lumps of iron sulphide are put into the middle bulb, and the acid is poured in at the top until the bottom bulb is full and the acid rises to cover the sulphide. The side tube from the middle bulb is then closed, and the pressure of the gas forces the acid away from the sulphide, thus stopping the action until the side tube is opened again.

PROPERTIES OF HYDROGEN SULPHIDE

Hydrogen sulphide is a colourless gas with a strong and unpleasant smell. It causes headache when breathed in small quantities and death if breathed in large quantities. The gas burns in air with a blue flame. With a limited supply of air, e.g. when the gas is burnt in a tube or jar, sulphur is set free:—

$$2H_2S + O_2 \longrightarrow 2H_2O + 2S$$

With plenty of air or oxygen, e.g. when the gas is burnt at a jet, the sulphur also burns, forming sulphur dioxide:—

$$2H_2S + 3O_2 \longrightarrow 2H_2O + 2SO_2$$

SULPHUR DIOXIDE SO_2

We have seen already that sulphur dioxide is produced when sulphur burns in air or oxygen:—

$$S + O_2 \longrightarrow SO_2$$

In the laboratory, however, sulphur dioxide is usually prepared by the action of metals on warm, *concentrated* sulphuric acid. Copper shavings are generally used, in the apparatus shown in Fig. 13. The flask is heated and the gas is collected by the upward displacement of air. The gas has the taste and smell of burning sulphur. If a tube of the gas is inverted in water, the water rises in the tube, showing that the gas is soluble in water. This solution turns blue litmus red, showing that an *acid* has been formed—*sulphurous*

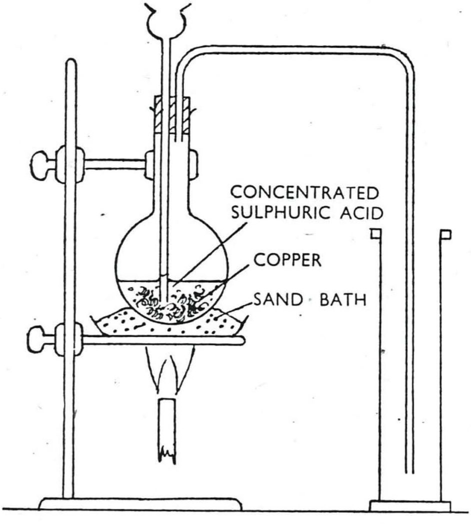


FIG. 13. PREPARATION OF SULPHUR DIOXIDE

acid. A coloured flower, placed in a jar of sulphur dioxide, is bleached.

PROPERTIES OF SULPHUR DIOXIDE

Sulphur dioxide is a colourless gas with a characteristic choking* smell. It is a very heavy gas: about two and a half times heavier than air. The gas is poisonous if breathed in large quantities, and can be used as a *disinfectant** for treating rooms after cases of infectious disease. It dissolves readily in water to form a weak acid—*sulphurous acid*:—

$$H_2O + SO_2 \rightleftharpoons H_2SO_3$$

When this solution is boiled, the sulphur dioxide is driven off again. Sulphurous acid slowly takes up oxygen from the air, forming *sulphuric acid*:—

$$2H_2SO_3 + O_2 \longrightarrow 2H_2SO_4$$

Sulphur dioxide is readily liquefied and is sold commercially in this liquid form in strong glass bottles. Liquid sulphur dioxide (like liquid ammonia) is used in some cooling machines. The gas is used as a mild bleaching agent for delicate materials like silk, wool, straw (for straw hats), bristles* (for tooth brushes), and sponges, which would be damaged if bleached by chlorine. (N.B.—Chlorine bleaches by *oxidation* and the coloured molecules are decomposed, while sulphur dioxide bleaches by *reduction* so that an atom of oxygen is removed which is easily replaced. Hence materials bleached by sulphur dioxide gradually regain their original colour when exposed to the air, e.g. newspapers turn yellow in time because the paper is made from wood-pulp which is bleached with sulphur dioxide.) Sulphur dioxide does not burn itself, and will not allow a lighted candle to burn in it, but a few substances like magnesium, which burn very vigorously, continue to burn in the gas, setting free sulphur:—

$$2Mg + SO_2 \longrightarrow 2MgO + S$$

This is a simple way of proving that sulphur dioxide contains sulphur. It should remind you of a similar experiment, burning magnesium in carbon dioxide:—

$$2Mg + CO_2 \longrightarrow 2MgO + C$$

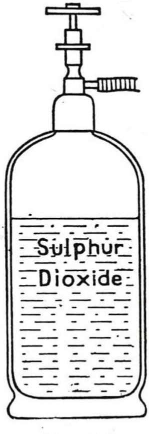


FIG. 14.
LIQUID
SULPHUR
DIOXIDE

SULPHUR TRIOXIDE SO_3

If a mixture of dry *sulphur dioxide* and *oxygen* is passed over finely divided *platinum* in a hot tube, *sulphur trioxide* is formed, the platinum acting as a catalyst:—

$$2SO_2+O_2 \longrightarrow 2SO_3$$

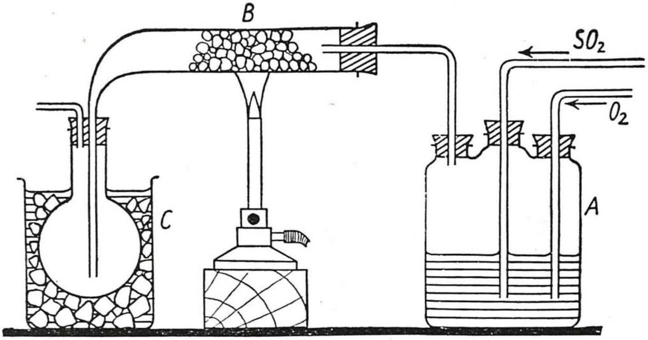


FIG. 15. PREPARATION OF SULPHUR TRIOXIDE

A—vessel containing concentrated sulphuric acid to dry the gases. B—tube containing platinized asbestos. C—freezing mixture surrounding receiver for sulphur trioxide.

Sulphur trioxide is a white solid which crystallizes in long, silky needles. It has a great liking for water, dissolving with a hissing noise to form *sulphuric acid*:—

$$H_2O+SO_3 \longrightarrow H_2SO_4$$

Hence, to make sulphuric acid, the most direct way is to prepare sulphur trioxide and then dissolve it in water.

THE MANUFACTURE OF SULPHURIC ACID

Sulphuric acid is the most important of all commercial chemicals, and millions of tons of it are used each year. It is manufactured commercially by two main methods:—

(i) the older *lead chamber process*; (ii) the more modern *contact process*.

THE LEAD CHAMBER PROCESS

In this process, jets of *water* or *steam* are blown into large leaden chambers along with a mixture of *sulphur dioxide* and *air* containing also *oxides of nitrogen* which act as a catalyst in the oxidation of sulphur *di*oxide to sulphur *tri*oxide. The following equations represent the main changes which take place:—

$$H_2O+SO_2+NO_2 \longrightarrow H_2SO_4+NO$$
$$NO+\text{oxygen from the air} \longrightarrow NO_2$$

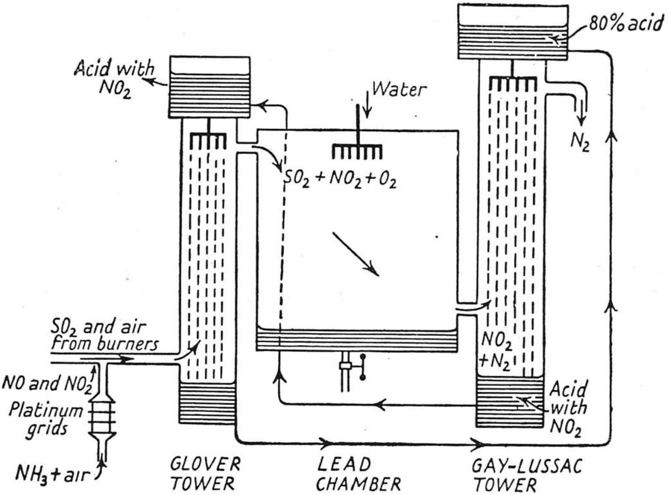


FIG. 16. LEAD CHAMBER PROCESS (diagrammatic)

Nitric oxide (NO) serves as an '*oxygen-carrier*', first taking up oxygen from the air to form *nitrogen peroxide* (NO_2) and then giving it up again to the sulphur dioxide.

The acid which forms in the lead chambers is only of about 65 per cent strength, and, after concentrating by evaporation, it is sold as 'commercial sulphuric acid'. The chamber acid is very cheap but it is also impure, hence it is only used for making 'heavy chemicals' and fertilizers for agricultural purposes.

THE CONTACT PROCESS

In the lead chamber process, nitric oxide is used as the catalyst for making the sulphur dioxide combine with oxygen to form sulphur trioxide. The contact process uses finely divided *platinum* (in the form of *platinized asbestos*) or vanadium pentoxide† instead of nitric oxide and it is essentially the same method as that described above for making sulphur trioxide.

(*Platinized asbestos* is made by dipping asbestos wool into a solution of a platinum salt and afterwards heating it red-hot. This produces a thin layer of very finely-divided metallic platinum on the asbestos.)

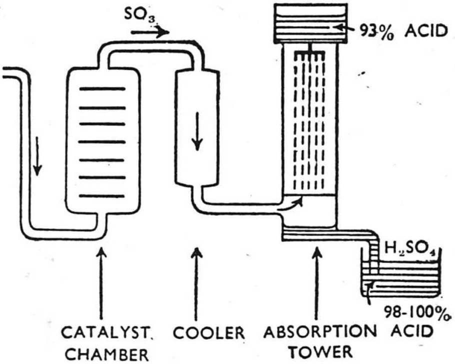


FIG. 17. CONTACT PROCESS (diagrammatic)

Very carefully purified and dried *sulphur dioxide* and *air* (or *oxygen*) are passed over heated *platinum* in a '*contact converter*' and *sulphur trioxide* is formed:—

$$2SO_2+O_2=2SO_3$$

The sulphur trioxide formed is absorbed in 98 per cent sulphuric acid, since unpleasant fumes are formed when water is used for dissolving the sulphur trioxide. Water is added gradually to the acid, combining with the sulphur trioxide as it dissolves and maintaining a 98-100 per cent concentration of acid.

The sulphuric acid prepared by the contact process is much purer than that from the lead chamber process, although it is somewhat more expensive. If the price of the sulphuric acid from the contact process could be reduced, the lead chamber process could not compete with it. Where a concentrated acid is required the contact process is the best source; but for a more dilute acid the chamber process is satisfactory and economical.

The purer acid is used for lead accumulators (batteries), for making 'fine chemicals' and dyestuffs, and in some processes for refining foodstuffs.

We have already examined the properties of sulphuric acid (Book II, p. 35).

The manufacture of *alkalis* is a close second in importance to the manufacture of sulphuric acid. The most important commercial alkalis are *washing soda* (sodium carbonate), *caustic soda* (sodium hydroxide), and *caustic potash*† (potassium hydroxide).

Until the end of the eighteenth century, the only source of alkali for making soap and glass was *plant ashes.* In 1791, Leblanc, a French scientist, invented the first successful method for making washing soda from common salt.

LEBLANC'S PROCESS (THE 'SALT-CAKE' AND 'BLACK-ASH' PROCESS)

(i) The first step in this process was to treat common salt with concentrated sulphuric acid (in A, Fig. 18) forming *sodium sulphate* (or *salt-cake*), together with *hydrogen chloride* as a valuable by-product. *Sodium hydrogen sulphate* ($NaHSO_4$) was first obtained, and this was strongly heated with more salt (in B, Fig. 18) forming *normal sodium sulphate* (Na_2SO_4) and more hydrogen chloride. The hydrogen chloride was dissolved in water and sold as hydrochloric acid:—

(*a*) $NaCl + H_2SO_4 \longrightarrow NaHSO_4 + HCl$ (*b*) $NaHSO_4 + NaCl \longrightarrow Na_2SO_4 + HCl$

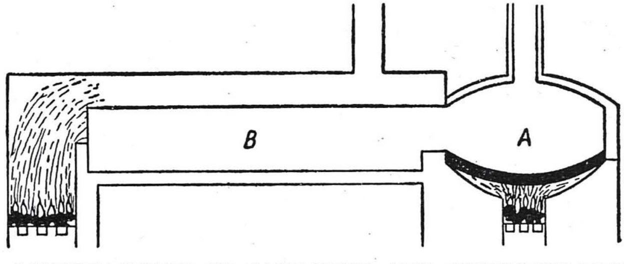


FIG. 18. MANUFACTURE OF SALT-CAKE AND HYDROCHLORIC ACID

(ii) The *salt-cake* was then mixed with powdered *limestone* (calcium carbonate) and *coal-dust* (carbon) and was strongly heated in a revolving furnace, forming '*black ash*'. The coal (carbon) reduced the sodium sulphate to sodium sulphide:

$$Na_2SO_4 + 4C \longrightarrow Na_2S + 4CO$$

The limestone (calcium carbonate) then reacted with this sodium sulphide forming a mixture of *sodium carbonate* and *calcium sulphide*†:—

$$Na_2S + CaCO_3 \longrightarrow Na_2CO_3 + CaS$$

Since sodium carbonate is soluble in water while calcium sulphide is not, the soda was extracted from the ' black ash ' with water. The solution was then evaporated until crystals of *washing soda* were obtained. Leblanc's process is no longer used for making washing soda and is now only of theoretical and historical interest. Reaction (i), however, is still used for manufacturing *hydrochloric acid.* The *salt-cake* (anhydrous sodium sulphate) is used instead of soda-ash in making many kinds of glass.

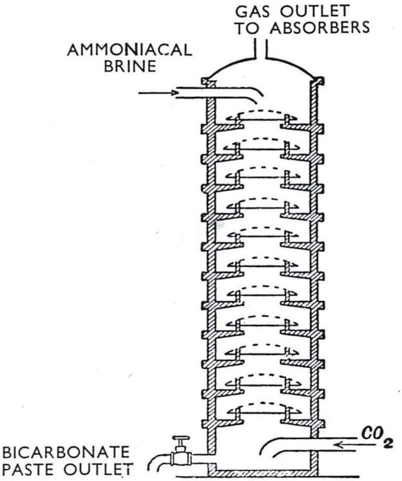


FIG. 19. AMMONIA-SODA PROCESS

SOLVAY'S PROCESS (THE AMMONIA-SODA PROCESS)

This is a more modern and a simpler process than Leblanc's process. *Brine* is first saturated with *ammonia* and then *carbon dioxide* is passed in as shown in Fig. 19. *Ammonium bicarbonate* is first formed from the water, ammonia and carbon dioxide:—

$$H_2O + CO_2 + NH_3 = (NH_4)HCO_3$$

This ammonium bicarbonate then reacts with the sodium chloride in the brine to form *sodium bicarbonate* (baking soda), which, being less soluble in water than the other substances, settles out from the liquid:—

$$NaCl + (NH_4)HCO_3 \longrightarrow (NH_4)Cl + NaHCO_3$$

This sodium bicarbonate is easily converted into *anhydrous sodium carbonate*† (soda-ash) by heating it:—

$$2NaHCO_3 \longrightarrow Na_2CO_3 + CO_2 + H_2O$$

CAUSTIC ALKALIS (CAUSTIC SODA AND CAUSTIC POTASH)

At first sight, it seems an easy matter to obtain *sodium hydroxide* (caustic soda) by electrolyzing a solution of sodium chloride. One would expect metallic sodium to be set free at the cathode† (negative electrode) and to react at once with the water to form caustic soda, while chlorine will be set free at the anode† (positive electrode) and should escape. Unfortunately, some of this chlorine combines with the caustic soda unless special precautions are taken to keep the chlorine away from the alkali. There are many methods for doing this on the large scale, but we shall consider only one of the simplest—the *Castner-Kellner process.*

In this process, the electrolytic apparatus consists of a rectangular tank, about 6 ft. by 4 ft. by 4 ft., divided into three chambers by two vertical walls which do not quite reach to the floor. A shallow layer of mercury closes the chambers and also conducts the electric current from one chamber to another. The two outer chambers are fitted with carbon anodes (+ve) and are filled with saturated brine. The middle chamber is fitted with an iron cathode (—ve) and is filled with water.

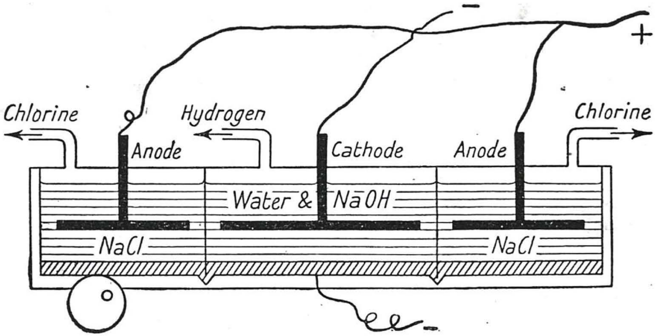


FIG. 20. THE CASTNER-KELLNER PROCESS (diagrammatic)

When the electric current is passed the brine is split up and chlorine is set free at the anodes in the two outer chambers and is led away through pipes in the roof. The layer of mercury on the floor acts as an intermediate negative electrode and the metallic sodium which is set free dissolves in the mercury and is transferred, by means of a gentle rocking motion, into the middle chamber where it dissolves in the water, forming caustic soda and setting free hydrogen. The caustic soda solution is then evaporated.

The chlorine which escapes is a valuable by-product and is often used for making *bleaching-powder* (Book II, p. 30). Sometimes, however, the hydrogen which escapes from the middle chamber is burnt in the chlorine to form hydrogen chloride, which is dissolved in distilled water and sold as chemically pure hydrochloric acid.

It is convenient to divide the elements into the two classes—*metals* and *non-metals*. The most important *chemical property* which distinguishes these two classes is that *metals form basic oxides* while non-metals do not, but *most non-metals form acidic oxides*. The chief *physical properties* of the two classes may be compared as follows:—

Metals	Non-Metallic Elements
(*i*) *Metals can be polished*	*Non-metallic elements cannot be polished*
(*ii*) *Metals can be rolled into thin sheets. For example, gold-leaf is usually* 0·0001 *mm. in thickness*	*Non-metallic elements cannot be rolled into thin sheets*
(*iii*) *Metals can be drawn out into wire. Thus,* 1 *gm. of gold can be drawn out into about* 3,000 *metres of wire*	*Non-metallic elements cannot be drawn out into wire*
(*iv*) *Metals are tough and can resist a strong pulling force without breaking*	*Non-metallic elements are brittle and cannot resist a strong pulling force*
(*v*) *Metals have a high density and a high melting point. With the exception of mercury, all metals are solids at ordinary temperatures*	*Non-metallic elements have a low density and usually have a low melting point. Some of them are liquids or gases at ordinary temperatures*
(*vi*) *Metals are good conductors of heat and of electricity*	*Non-metallic elements are usually either poor conductors or non-conductors*

METALS IN EVERYDAY LIFE

For a metal to be useful for everyday purposes, it must resist the action of air and water. The very active metals *potassium*, *sodium* and *calcium*, which react with cold water, can only be used for chemical reactions in laboratories. *Magnesium*, also, is too active for most everyday purposes, but since it is one of the ‘ lightest ’ metals (S.G. 1·7) it is mixed with aluminium and other metals for making aeroplane parts. Metals in general use which do not rust in air are *aluminium*, *copper*, *nickel*, *tin*, *zinc*, *lead*, *silver*, *gold*, and *mercury*. Gold, silver, and nickel are quite unaffected by air or water. Aluminium, zinc, and lead rust only on the surface, for the film of oxide clings to the surface and so protects the metal underneath from any further action.

Iron is the cheapest commercial metal but it has the great disadvantage that it rusts so quickly in damp air, and that the oxide does not cling to the surface but falls off, exposing a fresh surface of iron until the whole of the iron is converted into rust. Iron articles, therefore, must be protected from the action of air in some way, e.g. by painting, by coating with tin (tin-plate), or zinc (galvanized iron) or silver (silver-plate) or nickel (nickel-plate) or chromium (chromium-plate), and so on. *Copper* is used mainly for electrical work, being a very good conductor of electricity.

ALLOYS

Alloys,† made by mixing melted metals, sometimes have more useful properties than the pure metals alone; e.g. brass, an alloy of copper and zinc, has the most useful properties of both metals. Similarly, pure copper and silver are too soft for making coins, but silver alloyed with a little copper makes much harder coins, while copper is alloyed with a little tin for the same purpose. *Gun-metal*† and *bronze*† are alloys of copper with tin. *Pewter*† and *solder* are alloys of tin and lead. '*German silver*'† is an alloy of copper, zinc and nickel. *Type-metal*, for making printers' type, is an alloy of lead, tin, and antimony. *Bearing metal* (or *anti-friction metal*), for machine bearings, consists of the same three metals mixed in different amounts.

Many '*light alloys*' are used for purposes where weight must be kept as low as possible, e.g. in aeroplane parts and the crank-cases of motor-engines. '*Magnalium*'† and '*Duralumin*'† are mainly aluminium (S.G. $=2\cdot7$) with small quantities of magnesium, copper, nickel and manganese. '*Elektron*'† (90 per cent magnesium with some copper, zinc, manganese and silicon) is one of the 'lightest' of these alloys, with a specific gravity of only $1\cdot8$.

Other alloys we have met with in science lessons are '*Invar*' (65 per cent iron, 35 per cent nickel), and '*Nichrome*' (60 per cent nickel, 15 per cent chromium, and 25 per cent iron).

METALLURGY (THE EXTRACTION OF METALS FROM ORES)

Only a few metals, e.g. gold, platinum, and copper, are found '*native*' in the earth. Other metals are always found combined with other elements, and these compounds are called *ores*. For example, iron and tin are found mainly as oxides, while copper, zinc, lead, and mercury are found as sulphides. *The extraction of metals from ores is called metallurgy,*† and this process, which varies with different ores, is often complicated. In this Course, therefore, we shall only consider the general principles underlying the extraction of metals from their ores.

Ores which contain the metal combined as a carbonate are first heated strongly to drive off carbon dioxide and convert them into oxides. Sulphide ores, too, are first heated strongly in a current of air to convert the sulphur into sulphur dioxide and the metal into its oxide.

In nearly all cases, when the oxide has been obtained, this is reduced to the metal by heating with carbon. (You can do the same thing on a small scale by heating lead oxide with a blowpipe on a charcoal block.) Metallic ores are nearly always mixed with earthy matter, and during the reduction (or *smelting*†) with carbon, limestone, or chalk, is usually added. This lime combines with the earthy matter to form a liquid *slag*† which floats on the top of the liquid metal and can thus be separated.

In a Course of this kind we cannot describe the metallurgy of all the common metals, and so we shall only consider the extraction of the most important metal in everyday life—*iron.*

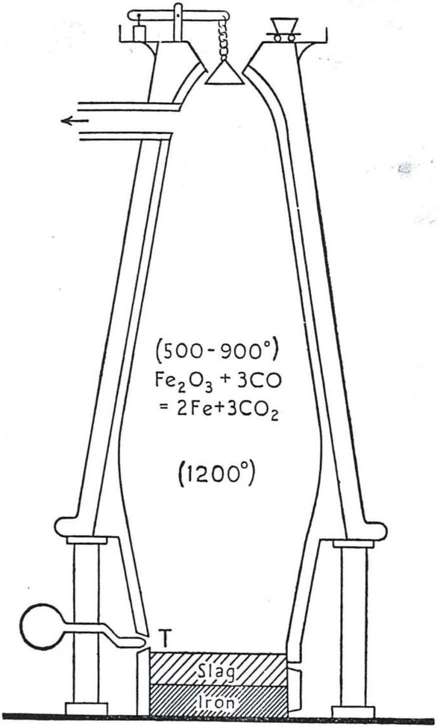


FIG. 21. BLAST FURNACE
(diagrammatic)

IRON AND STEEL

Iron is the most important and useful of all the metals. It is found mainly as iron oxide and iron carbonate ores.

THE METALLURGY OF IRON

The ore is first heated strongly to drive off carbon dioxide and water, leaving ferric oxide†—Fe_2O_3—and this oxide is then *smelted* with carbon (coke†) in a *blast* furnace.* Limestone is added to convert the earthy impurities into a liquid *slag*.

A *blast furnace* is a tall iron tower, 50-100 ft. high, lined with fire-brick. A mixture of iron ore, coke, and limestone is added at the top, and at the bottom is the *hearth** on which the liquid iron and slag collect. Just above the hearth is a ring of jets (T) through which a blast of hot air is blown into the furnace. The coke burns to form *carbon monoxide*, which is the *reducing agent*, and the ferric oxide is reduced to metallic iron:—

$$Fe_2O_3 + 3CO \longrightarrow 3CO_2 + 2Fe$$

The liquid iron and slag collect in the hearth at the base of the blast furnace, the iron below the slag, whence it is run off from time to time through the lower hole, while the slag is run off through the upper hole.

CAST-IRON

The liquid iron from the blast furnace is run into long, narrow channels in sand where it solidifies, forming *cast*-iron*. This form of iron contains not more than 92 per cent of iron, the remainder being mainly carbon and silicon. This impure iron is used for ‘ casting ’ heavy articles which do not have to stand a bending force or a sudden shock, e.g. the bases of laboratory retort stands are made of cast-iron. In making castings, a mould of the required shape is made in sand and liquid iron is poured into the mould, ‘ taking the shape of the containing vessel ’. Owing to its impurities, cast-iron is hard but brittle, breaking readily when hammered or bent.

WROUGHT-IRON

Another important commercial form of iron is *wrought-iron.*† Wrought-iron is tough and ‘ soft ’. This is the purest commercial form of iron, containing over 99 per cent of iron and only about 1 per cent of impurities, of which less than 0·3 per cent is carbon. Wrought-iron is made by melting cast-iron in the presence of air, in a furnace lined with iron oxide. Most of the impurities in the cast-iron either burn away or combine with the iron oxide, leaving wrought-iron. Wrought-iron can be drawn out into wire and rolled out into thin sheets or into rods, e.g. the *rods* of laboratory retort stands are made of wrought-iron. Wrought-iron softens before it melts and it can therefore be shaped and welded† (united) by hammering when red-hot.

STEEL

Steel is the most important iron product. It contains less carbon than cast-iron, but more than wrought-iron. There are several methods of making steel, but the principle is the same in each: either (i) carbon is added to wrought-iron, or (ii) some carbon is removed from cast-iron.

The best steel (which is used for making razors and the best quality cutting tools) is made by the *crucible process*, in which carbon is added to wrought-iron. This ' crucible steel ' is expensive. Ordinary cheap steel is made either by the *Bessemer process* or by the *open-hearth process*, in both of which some of the carbon is removed from cast-iron. The world's yearly production of steel by these processes is about 100,000,000 tons.

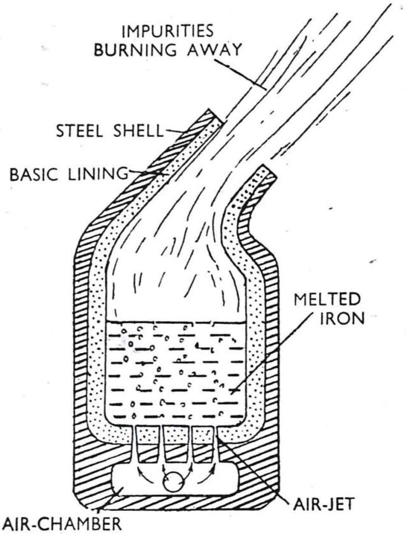


FIG. 22. BESSEMER CONVERTER

STEEL MANUFACTURE

(i) *The Bessemer Process.* In this process, the excess of carbon is removed from liquid cast-iron by a blast of hot air blown through it, in a *Bessemer converter* (Fig. 22). This is a large iron vessel with a fire-proof lining. The converter is half filled with several tons of cast-iron, and hot air is blown through it until most of the impurities in the iron have burned away. This process takes only about half an hour. The necessary amount of carbon is then added to the liquid metal, the converter is turned on horizontal pivots, and the steel is poured out into moulds.

Various kinds of steel can be produced by this method, e.g. *tool-steel* or *carbon-steel* (containing 1·0-1·5 per cent of carbon), *structural steel* for bridges and steel-framed buildings (containing 0·3-0·6 per cent of carbon), and *mild steel*† (containing less than 0·5 per cent of carbon).

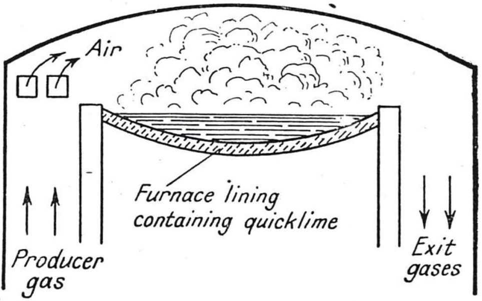


FIG. 23. OPEN-HEARTH FURNACE (diagrammatic)

(ii) *The Open-Hearth Process.* The Bessemer process has been largely replaced in England by the *open-hearth process* which yields a better quality steel. In this process (Fig. 23), a mixture of liquid cast-iron, iron ore, and scrap steel, is heated in an open-hearth furnace. The impurities combine with the lining of the hearth (lime-calcium oxide, or sand-silicon dioxide) forming a slag. The process takes about 10 hours, but the composition and quality of the final product can be more easily controlled than in the Bessemer process.

THE PROPERTIES OF STEEL

The properties of steel depend upon two main factors: (*a*) *the percentage of carbon it contains*, and (*b*) *the heat treatment which it has undergone.* The percentage of carbon may vary from 0·2 to 2·0 per cent. Steel containing little carbon (e.g. mild steel) is comparatively soft, and as the percentage of carbon increases, so also does the hardness of the steel.

If carbon-steel is heated red-hot and then dropped into cold water, it becomes very hard but very brittle. If this steel is re-heated very carefully it becomes tougher and less hard. This *heat treatment* of steel is known as '*tempering*',† and any degree of hardness and toughness can be produced by this method.

SILICON AND ITS COMPOUNDS

Next to oxygen, *silicon* is the commonest element in the Earth's crust, where it is present mainly as *silicon dioxide*† (SiO_2 or *silica*). *Sand*, *flint*,† and *quartz* are common varieties of silica. The *silicates* are also very common constituents of rocks, e.g. *granite* contains from 20 to 30 per cent of silica, while *clay* is mainly aluminium silicate. *Glass*, *pottery*, *bricks*, *mortar* and *cement* are all composed mainly of silicon compounds.

Quartz (or 'rock-crystal') is a hard, transparent form of silica, so hard that it scratches glass but cannot be scratched by steel. It has a very high melting-point (1,500° C.), but it can be melted in the oxy-hydrogen blowpipe flame, and is now used for making tubes, dishes, and flasks for laboratory use. Quartz apparatus has the great advantage over glass that it can be heated and cooled very suddenly without cracking, owing to its very low coefficient of expansion. *Sand* consists of small quartz crystals, white sand being almost pure silica, while yellow and brown sand contain some iron oxide. *Sandstone* consists of sand grains cemented together.

When silica (e.g. sand) is melted with sodium carbonate it forms *sodium silicate*,† a concentrated solution of which is called '*water-glass*'.

This solution can be used for preserving eggs, for hardening concrete floors, for waterproofing cloth, for making asbestos sheets, and for sticking things together.

GLASS

When sand is melted with soda-ash, ‘ water-glass ’ is formed. If sand is melted with soda-ash (anhydrous sodium carbonate) and also lime, ordinary bottle-glass or window-glass is produced. Glass is a mixture of silicates, and common glass is made by melting a mixture of white sand, limestone, and soda-ash. *Hard glass*, or potash glass, to stand higher temperatures without melting, is made in the same way, except that potash is substituted for the soda. *Flint glass*, or *lead glass*, a sparkling glass, of high refractive index, used for making lenses and ‘ cut-glass ’ articles, is made from sand, limestone, soda, and lead oxide. If the silica in a glass is partly replaced by the oxides of *boron*† and *aluminium*, a ‘ boro-silicate glass ’ is obtained which will stand sudden changes of temperature without cracking. Such glass can therefore be made thicker and stronger than ordinary glassware for heating. Coloured glass is made by dissolving small amounts of various metallic oxides in liquid glass, e.g. a trace of cobalt oxide gives a blue glass.

THE PROPERTIES OF GLASS

The most important properties of glass are (*a*) that it is transparent, (*b*) that it resists the action of most chemicals, and (*c*) that it softens gradually before it melts, so that it can be worked while it is in this plastic condition. Most other solids melt sharply, without passing through a plastic or half-solid stage. Glass resists the action of most chemicals, the most important exception being *hydrogen fluoride*† (or hydrofluoric acid†) which dissolves glass, and can therefore be used for marking glass, e.g. in graduating glass apparatus. Strong alkalis attack it slowly when hot.

CLAY, BRICKS AND POTTERY

Clay consists principally of an aluminium silicate containing much water of crystallization. It is formed from igneous rocks by the action of rain-water and carbon dioxide, so usually contains other silicates also, especially that of iron.

Bricks are made by baking impure clay.

Pottery is made from fairly pure clay, while china, or porcelain, is made from ‘ china clay ’. The clay for making pottery is first washed so as to remove the coarser particles, and is then allowed to stand. This purified, plastic clay is then shaped on a potter’s wheel, dried in a warm room, and then fired in a pottery furnace, or ‘ *kiln* ’.† The product is still porous and requires *glazing*. It is coated with a paste of silica and felspar (a potassium-aluminium silicate) and then fired again in the kiln, producing a thin film of glass on the outside. Coloured patterns are produced by painting the design in metallic oxides before glazing.

CEMENT

The Romans found that if a mixture of limestone and clay is heated strongly, the product forms a mortar which will harden even when it is covered with water. Many Roman buildings in which this ‘ hydraulic mortar ’ was used are still standing to this day. Such a mortar, which will harden in water as well as in air, is called *cement*.

The most important kind of cement is called ‘ Portland cement ’, because the earliest cements of this type were said to yield concrete ‘ equal to the best Portland stone ’ for building purposes. This cement is made by strongly heating a powdered mixture of limestone and clay to a temperature of about 1,500° C. in a sloping, revolving furnace. The mixture partly melts and forms a hard mass consisting of a mixture of lime, calcium silicate, and calcium aluminate.† This is ground up into a fine powder and is packed in air-tight containers.

The importance of cement increases steadily year by year. Cement mortar is used largely in place of lime mortar in building with brick and stone. When cement, mixed with sand and gravel, is made into a paste with water, it forms ‘ *concrete* ’, which sets to a solid mass in a short time. Concrete is used in ever-increasing quantities for building, and also for making articles which were formerly made of wood, iron, or stone. Steel rods and wires are often embedded in the concrete to give it greater strength, forming ‘ *reinforced concrete* ’.†

Concrete owes its usefulness to the fact that it can be shaped while in the half-liquid, plastic state. We have learnt that the most characteristic property of *solids* is *rigidity*, and that it is always a matter of some difficulty to change the shape of a solid. A *liquid*, on the other hand, is not rigid, and takes the shape of the containing vessel. Liquid concrete is poured into temporary vessels of the desired shape (moulds), and it takes the shape of this ‘ containing vessel ’, still retaining this shape after it sets to a rigid solid.

PHOSPHORUS

Phosphorus compounds are present in all living matter. Plants take in *phosphates* from the soil, and animals obtain their phosphorus compounds from certain *proteins* synthesized by plants.

The element *phosphorus* was first prepared from animal *bones*, which contain about 50 per cent of *calcium phosphate*.

When bone charcoal is heated in air, all the carbon burns away, leaving *bone-ash*, which is mainly calcium phosphate. When this bone-ash is treated with sulphuric acid, *phosphoric acid*, a thick, sticky liquid, is obtained, and if this liquid is mixed with charcoal and then heated to a white heat in the absence of air, phosphorus distils off as a vapour and can be condensed under water. A more modern method is to heat calcium phosphate rock with sand and coke, in an electric furnace. The silica of the sand forms calcium silicate, displacing *phosphorus pentoxide*, which is then reduced to phosphorus by the carbon.

Phosphorus, like sulphur and carbon, exists in *allotropic forms*, the two commonest being *yellow phosphorus* and *red phosphorus*.

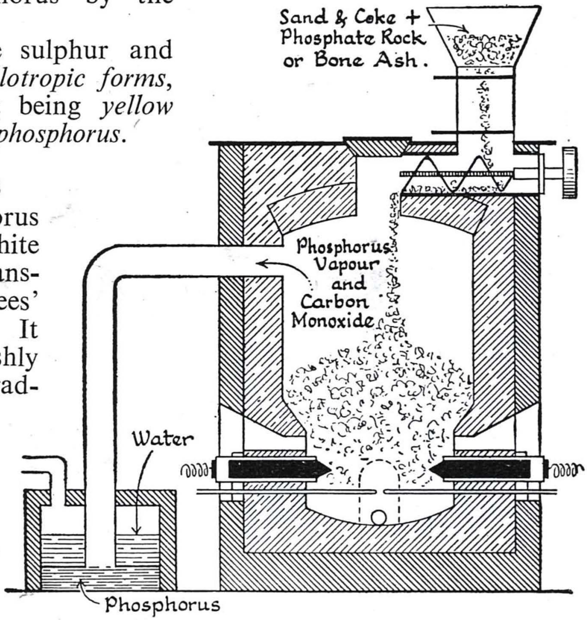


FIG. 24. PHOSPHORUS MANUFACTURE

YELLOW PHOSPHORUS

Yellow phosphorus (sometimes called white phosphorus) is a translucent solid like bees' wax in appearance. It is white when freshly prepared, but it gradually turns yellow on exposure to light. Yellow phosphorus has a very low *ignition temperature*, and will ignite* in air at about 30° C. Hence it must always be handled with the greatest care, particularly in tropical countries. Yellow phosphorus is kept *under water*. This form of phosphorus is a very active *poison*, 0·1 gm. causing death. It always gives off a poisonous vapour, and if small quantities of the vapour are breathed day after day, it causes a disease in which the bones decay, beginning with the jaw-bones. At one time, this disease was very common among workers in match factories, and in most countries the use of yellow phosphorus for making matches is now forbidden by law.

RED PHOSPHORUS

When yellow phosphorus is heated to about 250° C. in the absence of air, it changes to a dark red powder—*red phosphorus*—which is another allotropic form of the element. This variety of phosphorus gives off no poisonous vapour, and it does not catch fire below 240° C.; in fact, it is a much less active form.

The chemical identity* of these two forms of phosphorus can be proved by burning equal weights of yellow and red phosphorus in air or oxygen, when the same weight of the same product (phosphorus pentoxide) is formed in both cases.

MATCHES

For thousands of years, man could only produce fire by friction between pieces of wood, or else by using 'flint and steel'. Matches have only been in use for the last hundred years or so.

Matches were formerly made by dipping splinters of wood into a paste of yellow phosphorus, lead peroxide,† and glue, but when the use of yellow phosphorus was stopped, a sulphide of phosphorus was substituted. Such matches will 'strike' on any rough surface, the friction producing sufficient heat to ignite the mixture.

Modern '*safety matches*' contain no phosphorus in the match-head, but a strip of paper coated with non-poisonous *red phosphorus* is attached to the side of the match-box. The match-head contains an oxidizing agent (e.g. potassium chlorate, potassium nitrate, or red lead) mixed with glue. When the match-head is rubbed on the side of the box, the friction ignites a little of the red phosphorus and thus sets fire to the inflammable match-head, which, in turn, sets fire to the wood of the match-stick.

PHOSPHORUS PENTOXIDE (P_2O_5) AND PHOSPHORIC ACID (H_3PO_4)

When phosphorus burns in dry air or oxygen, a cloud of *phosphorus pentoxide* is formed. On standing, this settles as a white powder. Phosphorus pentoxide has a great liking for water, dissolving with much heat to form *phosphoric acid.* Because of this great liking for water, phosphorus pentoxide is used for drying gases when every trace of water vapour must be removed, since it is the most powerful drying agent known.

Phosphates are the salts of phosphoric acid and we have seen that they are valuable fertilizers, e.g. calcium phosphate.

CHAPTER III

CARBON AND ITS COMPOUNDS

Carbon forms more compounds than any other element, in fact, the study of carbon compounds forms a special branch of chemistry—*organic chemistry*. Although the term '*organic*' suggests that these compounds are obtained from living things, most of them can now be prepared in the laboratory and some of them do not exist in Nature.

The element carbon, like sulphur, exists in *allotropic forms*: (i) *non-crystalline carbon* (e.g. sugar-charcoal), and (ii) *crystalline carbon*, namely *graphite*† and *diamond*. It is surprising to find that diamond, graphite, and charcoal should all be the same chemically, but this is proved to be so by burning the same weight of the three substances in oxygen, when equal weights of the same product, carbon dioxide, are obtained. Hence, although they are so very different in appearance, diamond, graphite, and charcoal are only *different forms of the same element—carbon.*

There are several common kinds of impure non-crystalline carbon: (*a*) *wood-charcoal*, (*b*) *lamp-black* and *soot*, (*c*) *animal* (or bone) *charcoal*, (*d*) *coal* and (*e*) *coke*.

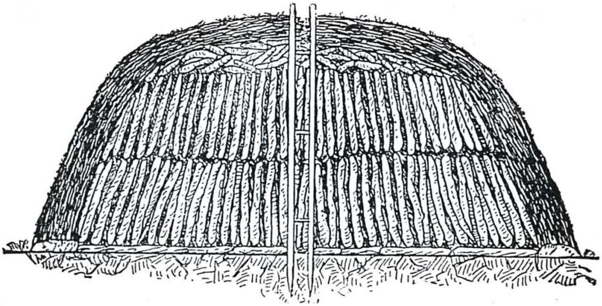


FIG. 25. CHARCOAL BURNING

CHARCOAL

Wood-charcoal is prepared by the *destructive distillation of wood* (i.e. decomposition without flame), either by the old method of burning the wood in heaps covered with earth, or by the modern method of heating the wood in closed ovens or retorts. When the modern method is used, valuable by-products are obtained, e.g. *acetic acid*,† *methyl alcohol*,† *acetone*,† *wood-tar*† and *wood-gas* (mainly methane).

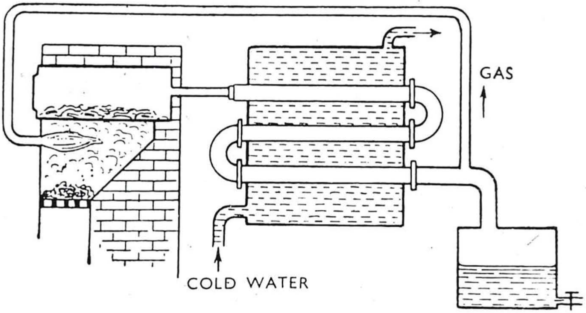


FIG. 26. DISTILLATION OF WOOD IN CLOSED RETORT

When the old-fashioned method is used, these substances escape into the air along with the smoke. Wood-charcoal is a black, porous substance, and owing to its porous structure fresh charcoal can absorb

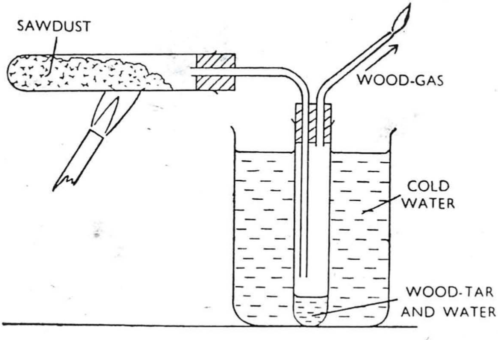


FIG. 27. DISTILLATION OF WOOD

many times its own volume of gas. Fresh charcoal also absorbs solids and liquids, and is therefore used for decolorizing crude brown sugar and similar substances which have to be purified and decolorized for use as food materials.

(i) Set up the apparatus shown in Fig. 27. Heat the sawdust in the hard-glass tube gently at first and then more strongly. Light the *wood-gas* as it escapes from the jet and notice the large volume of gas given off by a small quantity of wood. When there is no further change, allow the apparatus to cool and examine the distillate. Test the watery part with litmus paper. Examine the *wood-tar* and notice its characteristic smell (wood-tar is present in ordinary wood-smoke).

(ii) Half fill a test-tube with the given coloured liquid (a 0·003 per cent solution of methyl-violet†—the dye in ' copying-ink ' pencils and purple typewriter ribbons). Add a few grams of your freshly-prepared *wood-charcoal* and shake vigorously for a few minutes. Filter and notice that the filtrate is practically colourless. After draining, add a few c.cm. of methylated spirit to the charcoal on the filter-paper and notice that a deep violet solution comes through the filter.

LAMP-BLACK

Lamp-black† is very finely divided *soot* obtained by burning substances rich in carbon (e.g. natural gas from oil-wells) in a limited supply of air. It is used in paints, printers' ink, Indian ink, carbon paper, typewriter ribbons, and in gramophone records. It is also mixed with rubber in making rubber tyres. *Soot* from burning wood, coal, or oil, is a substance similar to lamp-black.

ANIMAL CHARCOAL

Animal charcoal is made by distilling bones in iron retorts. Ammonia and bone-oil are given off, leaving a residue of carbon mixed with about 85 per cent of calcium phosphate, the main mineral constituent of bones. If this mineral matter is removed by dissolving in acid, very finely divided charcoal (bone black) remains behind. Animal charcoal is a very good decolorizing agent. Repeat Experiment (ii) above with animal charcoal.

COAL

Coal is a product of the decomposition of plant remains which became buried in the earth millions of years ago. We can trace the formation of coal from vegetable matter by examining the related substances *peat*,† *lignite*,† *coal*, and *anthracite*,† which appear to be successive stages in the decomposition of vegetable matter in the absence of air, under the influence of great pressure and moderate heat. The vegetative origin of coal is also shown by the presence of *fossils* of parts of plants in coal deposits.

Fig. 28. THE STORY OF COAL

Peat consists of partly decomposed plant remains and is clearly of vegetable origin, since parts of stems and roots can be seen in it. Peat is not a very good fuel. It burns very slowly with a great deal of smoke, but gives out very little heat.

Lignite, or *brown coal*, appears to be the next stage in the formation of coal. It is harder and denser than peat, and shows fewer signs of plant remains. Lignite burns with a smoky flame, but it yields more heat than can be got by burning peat. It contains 60-75 per cent of carbon.

Common coal is the next stage in the formation of coal. This is a shiny, black, brittle solid which burns with a bright, hot flame; forming, however, a fair amount of smoke. Common coal often contains fossils of plants. This is the type of coal which is used as the main fuel in cold countries. It contains 75-90 per cent of carbon.

It is also used for making *coal-gas*, since it yields large volumes of gas when distilled.

Anthracite, or *steam coal*, represents the oldest form of coal and consists mainly of carbon (over 90 per cent). It is very hard and burns slowly, without a flame and with very little smoke or ash, but it gives out great heat.

It is thought that *graphite* may represent the final stage in the decomposition of the remains of plants after millions of years.

PEAT ——→ LIGNITE ——→ COMMON COAL ——→ ANTHRACITE

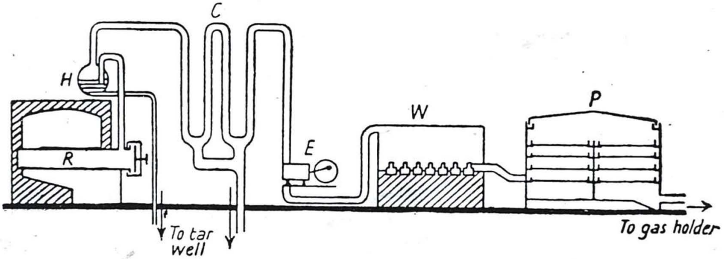


FIG. 29. COAL-GAS WORKS (diagrammatic)
R—retort. H—hydraulic main. C—condenser. E—pump. W—washer. P—purifier.

COAL-GAS

Coal-gas is used very widely in many countries for heating and lighting. It is manufactured in a *gas-works* (Fig. 29) by the *destructive* *distillation of coal* in fire-clay retorts (R). The coal-gas, mixed with many impurities, passes through pipes into the *hydraulic main** (H), a large channel containing water which cools the gas and removes some of the impurities besides serving as a valve to stop gas escaping when a retort is opened. *Coke*, a porous form of carbon, is left in the retorts, and the crude* gas, containing *coal-tar*, *ammonia*, and *sulphur compounds*, passes through several stages of purification. It first passes through the *condensers* (C), where the *tar* and water containing dissolved *ammonia* are removed from the gas. The coal-gas then passes through a *washer* (W) where the remainder of the ammonia dissolves in water. Sulphur compounds are then removed by passing the gas through the *purifier* (P), over layers of iron oxide or slaked lime which combine with the sulphur compounds. The purified coal-gas then passes into the *gas-holder* ready for delivery through pipes to the users.

BY-PRODUCTS FROM COAL-GAS MANUFACTURE

(i) *Coke*. The coke left behind in the retorts is used as a cheap fuel, for making 'producer-gas' and 'water-gas'† and also for reducing metals from their ores, particularly in the manufacture of iron in the blast furnace.

(ii) *Coal-tar*. Coal-tar is a thick black liquid which is a complicated mixture of at least 200 different substances, some of which are of great commercial value. The more valuable constituents are separated from the tar by *distillation*, condensing and collecting the vapour which comes off at different temperatures. The first part of the tar to distil over is known as the 'light oil', containing *benzene*,† *toluene*,† and *xylene*,† which form the raw materials for the preparation of *synthetic dyestuffs* and *fine organic chemicals*. The second fraction to distil over is known as the 'middle oil', and this consists mainly of *phenol* (or *carbolic acid*†), and *naphthalene*. Carbolic acid is very poisonous to bacteria and is used as a disinfectant, while one use of naphthalene is as 'moth-balls' to drive away insects. The next fraction is the 'heavy oil', used for preserving wood. The last fraction to distil over is the 'anthracene oil' which yields *anthracene*,† another raw material for synthetic dyes, and the residue left behind in the retort is *pitch*,† which is used for road-making.

(iii) *Ammonia*. The ammonia in the liquid from the washers is set free and then absorbed in sulphuric acid, forming ammonium sulphate, which is used as a fertilizer for supplying crops with combined nitrogen. We have seen how important a supply of combined nitrogen is for modern agriculture .

Set up the apparatus shown in Fig. 30, using a 3 in. × $\frac{1}{2}$ in. hard-glass test-tube, half full of powdered coal. Heat strongly and first set fire to the crude gas before connecting up the condenser and purifier.

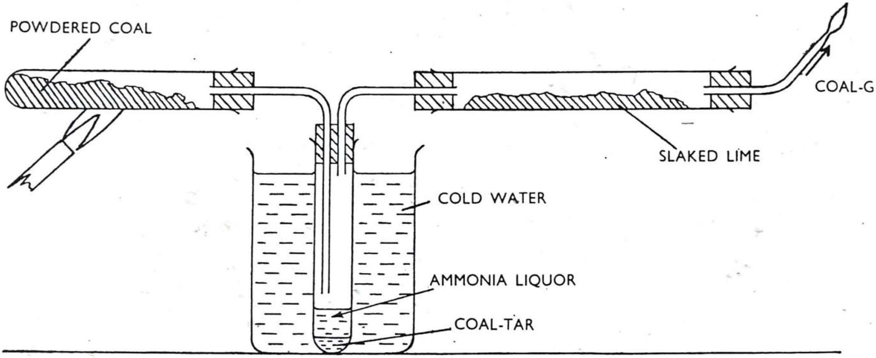


FIG. 30. DISTILLATION OF COAL

Notice the smoky flame. Then connect the apparatus to the condenser and purifier and set fire to the purified gas. Stop heating when no more gas is given off and allow the apparatus to cool. When cool, test the watery distillate with litmus paper. Shake out the residue from the hard-glass tube and examine the *coke*.

CRYSTALLINE FORMS OF CARBON

Graphite, which is used in making ' lead ' pencils, is also known as ' black lead ', because it was formerly supposed to contain lead. Natural

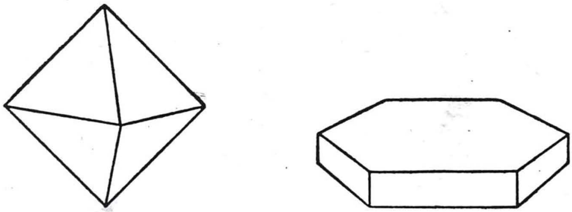

Diamond
Graphite

FIG. 31. CARBON CRYSTALS

graphite is found in Ceylon, Siberia, Poland, and North America. A very soft and pure artificial graphite is made at the Niagara Falls, where powdered coke or anthracite is mixed with sand and heated strongly in an electric furnace.

Graphite crystallizes in soft, flat plates which feel slippery, and it is therefore used for *lubricating* when a liquid lubricant (e.g. oil) cannot be used. The filling for lead pencils is made by mixing graphite and clay in suitable proportions and then heating the mixture. A ‘ soft ’ pencil has a larger proportion of graphite, while a ‘ hard ’ pencil has a larger proportion of clay.

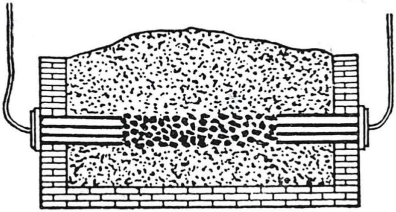


Fig. 32. MANUFACTURE OF GRAPHITE

Diamonds, which are valuable mainly because they are so rare and beautiful, are found in many parts of the world, but chiefly in South Africa and South America. For ornamental purposes, in jewellery, the natural diamonds are cut and polished so as to reflect as much light as possible. Diamond is the hardest substance known, and the splinters produced in cutting diamonds for jewellery are used for cutting glass. Black or other discoloured diamonds, which are useless for jewels, are mounted in the faces of rock-drills used in boring for oil and water.

Artificial diamonds are said to have been made by dissolving carbon in liquid iron at a temperature of about 3,000° C. in an electric furnace, and then cooling suddenly. After cooling, the iron was dissolved away with acid, leaving very small diamonds behind, the largest being about the size of a pin’s head. This process has no commercial value.

COMPOUNDS OF CARBON

CARBON MONOXIDE

In our earlier lessons we spent some time studying carbon dioxide (CO_2). This, however, is not the only oxide of carbon; there is another important compound of carbon and oxygen—*carbon monoxide*† (CO). This compound, which is also a gas, is like carbon dioxide in having neither colour nor smell, but it differs in being very inflammable and very poisonous.

Carbon monoxide is formed when carbon burns in a limited supply of air. It can be prepared by passing carbon dioxide through a tube containing red-hot carbon, as shown in Fig. 33:—

$$C + CO_2 \longrightarrow 2CO$$

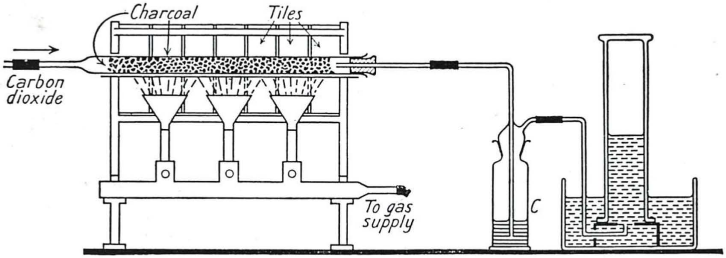
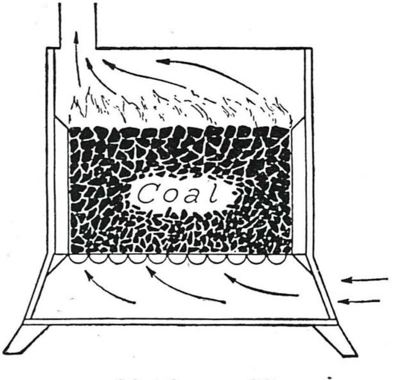


FIG. 33. LABORATORY PREPARATION OF CARBON MONOXIDE

After passing over the red-hot carbon, the gas is passed through a wash-bottle C containing sodium hydroxide solution, to remove any traces of carbon dioxide which have not been reduced by the carbon. The gas is collected over water.



$$2CO + O_2 \longrightarrow 2CO_2$$
$$CO_2 + O \longrightarrow 2CO$$
$$2C + 2O_2 \longrightarrow 2CO_2$$

FIG. 34. REACTIONS IN A CHARCOAL FIRE

PROPERTIES OF CARBON MONOXIDE

Carbon monoxide is a gas without colour, taste, or smell. It is slightly ‘lighter’ than air and is almost insoluble in water. Carbon monoxide burns readily in air with a characteristic blue flame, forming carbon dioxide:—

$$2CO + O_2 \longrightarrow 2CO_2$$

(This blue flame is often seen above a charcoal fire.) Because of the readiness with which it combines with oxygen, carbon monoxide is often used as a *reducing agent*, e.g. in the blast furnace:—

$$Fe_2O_3 + 3CO \longrightarrow 2Fe + 3CO_2$$

Carbon monoxide is a very poisonous gas. As little as one volume of carbon monoxide mixed with 100,000 volumes of air will make people seriously ill, while 1 per cent of carbon monoxide in air causes death in a few minutes, because the haemoglobin in the red blood-cells takes up carbon monoxide in preference to oxygen. Since carbon monoxide is formed by motor-engines running on a rich mixture (e.g. when starting up), it is very dangerous to run an engine inside a closed building and breathe the exhaust gases.

PRODUCER-GAS AND WATER-GAS

Since carbon monoxide burns with a very hot flame it is often used as a fuel for industrial purposes. When air is passed through red-hot coke, a mixture of carbon monoxide and nitrogen is formed. This mixture is known as '*producer-gas*' and is used for heating furnaces because of its hot flame.

If *steam* is blown through red-hot coke, a mixture of *carbon monoxide* and *hydrogen* is formed:—

$$C+H_2O \longrightarrow CO+H_2$$

and this mixture is known as '*water-gas*'. Water-gas burns with about three times as much heat as producer-gas. Producer-gas and water-gas are often made together. Air is blown through coke until it is white-hot, forming producer-gas; steam is then blown through instead of air, forming water-gas, until the coke begins to cool down, when it is made white-hot once more by blowing air through it (or a mixture of steam and air in the correct proportions is passed continuously). In this way a very cheap fuel-gas is obtained. Producer-gas and water-gas are sometimes mixed with coal-gas to produce a cheaper fuel.

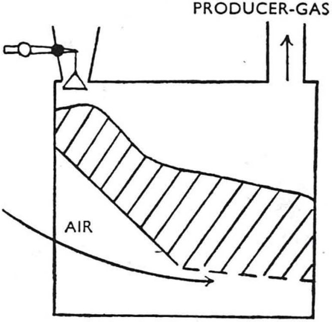


FIG. 35. GAS-PRODUCER (diagrammatic)

HYDROCARBONS

There are many compounds containing only *carbon* and *hydrogen*, called *hydrocarbons*, and, as we might expect, they all burn readily, giving out light and heat.

METHANE

Methane (or *marsh-gas*†) is the simplest hydrocarbon and is formed in Nature by the slow decay of vegetable matter in the absence of air.

Bubbles of methane can sometimes be seen rising to the surface of marshy pools. It also escapes from cracks in coal-mines where it is known to the miners as '*fire-damp*'†, causing dangerous explosions, since a mixture of methane and air explodes violently when lighted:—

$$CH_4 + 2O_2 \longrightarrow CO_2 + 2H_2O$$

Since one-fifth of the air is oxygen, one volume of methane uses up all the oxygen from 10 volumes of air, forming carbon dioxide in its place. Hence there is a serious danger of suffocation by carbon dioxide (called '*after-damp*'† by miners) after a 'fire-damp' explosion in a coal-mine.

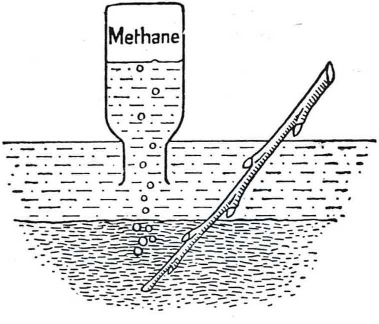


FIG. 36. MARSH-GAS

THE PARAFFIN SERIES OF HYDROCARBONS

There is a whole series of hydrocarbons similar to methane, e.g. *methane*—CH_4, *ethane*—C_2H_6, *propane*[1]—C_3H_8, *butane*[1]—C_4H_{10}, *pentane*[1]—C_5H_{12}, *hexane*[1]—C_6H_{14}, and so on. This series of hydrocarbons is called the *paraffin series*, and has the general formula C_nH_{2n+2}. The lowest members of the series (CH_4 to C_4H_{10}) are gases at ordinary temperatures, from C_5H_{12} to $C_{17}H_{36}$ are liquids or half-solids, while the paraffins from $C_{18}H_{38}$ upwards are solids; e.g. paraffin-wax is a mixture of solid paraffins.

THE ETHYLENE SERIES OF HYDROCARBONS

Ethylene,† C_2H_4, is the simplest member of another series of hydrocarbons called the *ethylene series*. Ethylene is present in coal-gas, burning with a luminous, smoky flame. The general formula of this series is C_nH_{2n}. The hydrocarbons of the ethylene series all burn with a more luminous and more smoky flame than the corresponding paraffins, since they contain a greater proportion of carbon.

THE ACETYLENE SERIES OF HYDROCARBONS

Acetylene, C_2H_2, is the first member of another series of hydrocarbons called the *acetylene series*, the general formula of the series

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[1] There is no need to remember these names.

being C_nH_{2n-2}. Acetylene is easily prepared by dropping water on calcium carbide, a method used in some lamps:—

$$CaC_2 + 2H_2O \longrightarrow C_2H_2 + Ca(OH)_2$$

Acetylene is a colourless gas with an unpleasant smell. It burns in air with a luminous and very smoky flame. Special burners are made, however, which will burn acetylene with a bright white light and without smoke. When mixed with air or oxygen, acetylene explodes with great violence when a light is applied to the mixture: in fact, it is not safe for you to try the experiment. The *oxy-acetylene*† *blowpipe*, burning acetylene in oxygen, provides the hottest flame known (over 3,000° C.). It will melt and cut through steel quite easily.

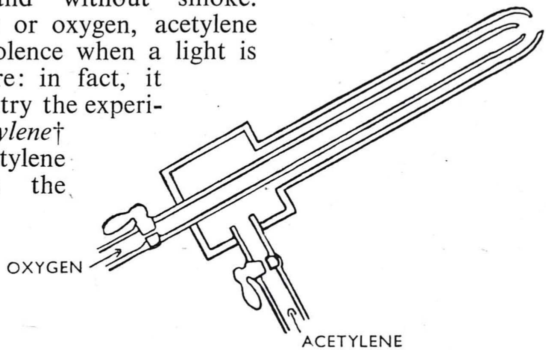


FIG. 37. OXY-ACETYLENE BLOW-PIPE

THE BENZENE SERIES OF HYDROCARBONS

Benzene—C_6H_6—is the first member of another important series of hydrocarbons, called the *benzene series*. Higher members of this series are *toluene* and *xylene*, which are present in coal-tar along with benzene.

You will realize from this outline that the number of different compounds of carbon and hydrogen is very large indeed, but the study of all these hydrocarbons is simplified by the fact that *the members of the same series all have very similar properties*.

PETROLEUM

Petroleum (*mineral-oil*, or *rock-oil*) is a natural mixture of various hydrocarbons. It is found in many parts of the world, e.g. North America, Russia, Rumania, Persia, Trinidad, and the Dutch East Indies. American petroleum consists mainly of *paraffins*, i.e. hydrocarbons of the *methane series*, while Iraqi petroleum contains hydrocarbons of the *benzene series*.

Petroleum is obtained by drilling a bore-hole down through the earth until the petroleum is reached. At first, the oil usually rushes out under its own pressure, but the pressure decreases slowly until, after a time, the oil has to be pumped to the surface. The crude petroleum is a thick, heavy liquid which varies in colour from greenish-brown to black, consisting of a mixture of several hundred different hydrocarbons. The purification, or *refining*,* of this crude petroleum and its products is now a great industry, yielding over 200 commercial products.

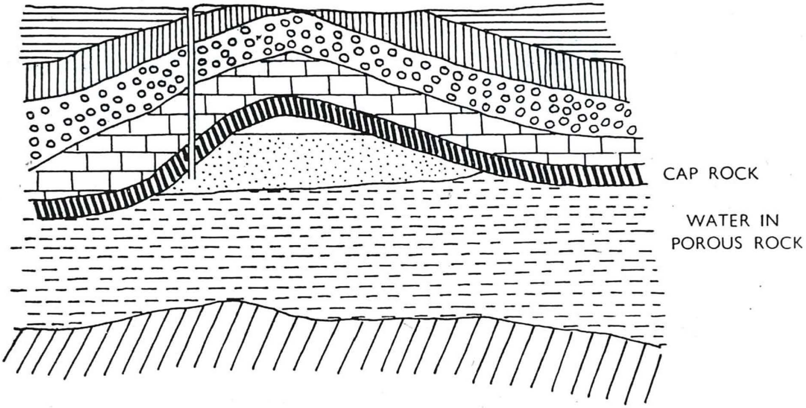


FIG. 38. SECTION OF OIL-WELL (diagrammatic)

PETROLEUM REFINING

The constituents of petroleum are separated by *fractional distillation*, each constituent having a different boiling-point and thus distilling over at a definite temperature. The crude petroleum is run into large *stills** and is heated gradually. The portion with the lowest boiling-point, which distils over first, is the main source of *petrol* (or gasoline†). The next product to distil over is *kerosene* (or paraffin-oil) which has a higher boiling-point and catches fire less readily than petrol. At a still higher temperature, large quantities of a thicker and heavier oil distil over, known as *gas-oil*,† which is used as a fuel for oil-engines (e.g. Diesel-engines), and which can also be converted into petrol by a process known as ‘ *cracking* ’.† (In this process, the gas-oil is passed through a red-hot tube at a high pressure, and the ‘ heavy ’ hydrocarbons of high molecular weight and high boiling-point split up into ‘ light ’ hydrocarbons of lower molecular weight and lower boiling-point. About half the world's supply of petrol is now made by ‘ cracking ’ heavier oils.) At a still higher temperature, a very thick oil distils over which yields *lubricating oils*, and also *vaseline* (petroleum-jelly) and *paraffin-wax*. *Pitch*, used for surfacing roads, is left behind in the still, together with some *coke*. A typical American petroleum yields about 4 per cent of gas, 25 per cent petrol, 15 per cent kerosene, 40 per cent gas-oil, 12 per cent lubricating oil, and 4 per cent coke.

PETROL SUBSTITUTES OF THE FUTURE

Geologists* consider that the natural deposits of petroleum are very limited, and that, at the present rate of oil-production, they may be exhausted comparatively soon. Most geologists calculate that the world's petroleum supplies will not last for more than 100 years; some calculations are as low as 50 years. Scientists have therefore begun to search for petrol substitutes. One idea is to use alcohol instead of petrol, and another idea is to make petrol from *coal*.

POWER ALCOHOL

Ethyl alcohol† (which we shall study shortly) can be produced very cheaply by the fermentation† of carbohydrates, but it is not a very satisfactory motor-fuel by itself, although when mixed with petrol it forms a very good motor-fuel. *Methyl alcohol* is also a satisfactory motor-fuel, and besides being formed during the destructive distillation of wood, it can be synthesized very cheaply from water-gas. When this mixture of carbon monoxide and hydrogen is passed over a suitable catalyst at a high temperature and pressure, methyl alcohol is formed:

$$CO + 2H_2 \longrightarrow CH_3.OH$$

The fermentation of carbohydrates and this synthetic process for making methyl alcohol offer possibilities for the world's future supplies of motor-fuel.

THE ' HYDROGENATION ' OF COAL

When powdered coal is heated with hydrogen under the enormous pressure of 200 atmospheres (about 3,000 lb. per square inch) a thick, black, oily liquid is produced.

This oily liquid formed by the ' hydrogenation ' of coal yields a number of valuable products on distillation. By this process, a ton of coal yields about 40 gallons of motor-fuel, about 50 gallons of fuel-oil for Diesel-engines, and over 10,000 cubic feet of gas for lighting and heating. It is therefore probable that as the world's supplies of petroleum become exhausted, some such method of ' hydrogenating ' coal will supply motor-spirit. One great advantage of this process is that it can be used with peat, lignite, and coal of poor quality. This process, of course, depends on coal, but the world's coal deposits are more extensive than the oil deposits; in fact, it has been calculated that the coal deposits of the world will last for about 5,000 years if mined at the present rate.

COMBUSTION AND FLAME

A chemical change which is accompanied by the production of heat and light is called combustion, as when we speak of the combustion of coal in a railway-engine; the combustion of petrol inside the cylinder of a motor-engine (internal combustion); the combustion of magnesium in oxygen; the combustion of copper in chlorine. Some combustion takes place without *flame*; in fact, flame is only an accidental result of combustion. In the combination of iron and sulphur, for example, there is no flame, the mixture merely becomes white-hot. *Substances that yield gases on heating, e.g. burning wood or oil, burn with a flame.*

We can distinguish between two types of flames, (i) *luminous flames* which give out light, e.g. the candle flame, and (ii) *non-luminous flames* which produce little or no light, e.g. burning hydrogen or alcohol, or a well-regulated Bunsen flame.

THE CANDLE FLAME

The wax of which the candle is made cannot burn until it is changed into gas, and this is the purpose of the *wick*. The heat of the flame changes some solid wax into liquid wax, and this liquid wax rises up the wick by *capillarity*, until it reaches the inside of the flame, where it changes into gas.

The heat of the flame then decomposes the hydrocarbons of which the wax is composed, setting free tiny particles of carbon which are heated white-hot by the hydrogen as it burns to form steam. If there is a good supply of air, the carbon near the outside of the flame burns in the oxygen of the air, forming carbon dioxide, but if the air supply is insufficient, some of the carbon escapes, unburnt, as black smoke (or soot).

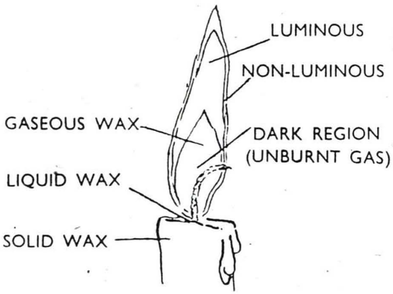


FIG. 39. CANDLE FLAME

A candle flame consists of (i) a dark, inner region of unburnt gas, (ii) a luminous region containing white-hot carbon particles, and (iii) a non-luminous outer region.

Hold a sheet of paper horizontally across a candle flame, withdrawing it just before it catches fire. Notice the charred ring, showing that the middle of the flame is cool, consisting of unburnt gas. Make a ' vertical section ' of the flame in the same way.

THE BUNSEN FLAME

For laboratory purposes we require a hot flame which will not deposit soot, and the Bunsen burner provides such a flame. When the air-holes are closed, we get a luminous, smoky flame. When the air-holes are opened, the rush of gas through the jet at the bottom of the burner-tube draws in air through the air-holes, and this air mixes with the gas in the burner-tube. With the right amount of air, all the carbon in the gas burns at once to carbon dioxide, instead of separating as white-hot particles as it does when the air-holes are closed. Hence, when the air-regulator admits the correct amount of air, the coal-gas burns with a clean greenish-blue flame called the *Bunsen flame*. If the proportion of air is increased, an explosive mixture is formed and the flame ' strikes back ' down the burner-tube, burning at the jet itself.[1] This ' striking back ' can be prevented by covering the top of the burner-tube with wire gauze. The principle involved is the same as in the miners' safety lamp which we shall discuss shortly.

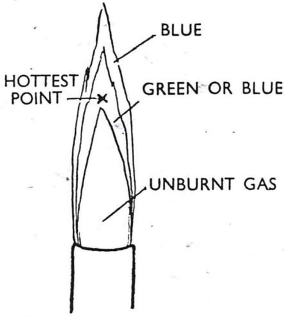


FIG. 40. BUNSEN FLAME

The Bunsen flame consists of two regions of burning gas surrounding an inner cone of unburnt gas. If a piece of paper is held for a moment horizontally across the flame, the paper is charred in a ring, showing that the inside of the flame is quite cool, unburnt gas. In the same way, if a wide tube is connected to the gas supply and the top is covered with wire gauze, a small heap of gunpowder on the centre of the gauze does not catch fire when the gas is lighted. This shows clearly that the flame is *hollow*.

---

[1] With petrol-gas, however, the flame usually ' floats off ' when too much air is admitted.

The inner region of the Bunsen flame is green or blue in colour, and in this part of the flame the hydrocarbons in the gas are partly burnt to form carbon monoxide and steam. The outer region is blue in colour, and in this part of the flame the carbon monoxide formed in the inner cone burns to form carbon dioxide.

The Bunsen flame gives out much heat but little light. For lighting purposes, this heat energy can be converted into light energy by allowing a Bunsen flame to heat a ' gas-*mantle* '† which contains the oxides of *thorium*† (99 per cent) and of *cerium*† (1 per cent). These oxides when heated become white-hot, and the mantle gives out a bright, white light.

THE SAFETY LAMP

When candles were used in coal-mines, ' fire-damp ' (methane) explosions were very common and many lives were lost. In 1815, Sir Humphry Davy studied the causes and possible means of preventing these accidents. Davy found that if a flame is cooled, it is extinguished, because inflammable gases all have a definite temperature below which they cannot be ignited. This temperature is called the ' *ignition temperature* '. The ignition temperature of methane is fairly high, so that it cannot be set alight by a red-hot object, but only by something at a higher temperature. Davy also noticed that if a piece of wire gauze is held over a flame, the flame burns on the underside of the gauze only. The flame does not pass through the gauze because the metal conducts away the heat and does not get hot enough to ignite the gas passing through the gauze. If, however, the gauze is left over the flame for a long time, it may become hot enough to ignite the gas above the gauze. Conversely, if a piece of wire gauze is held over an unlighted burner, the gas can be lighted *above* the gauze, but the flame does not pass down through the gauze and ignite the gas at the top of the burner-tube.

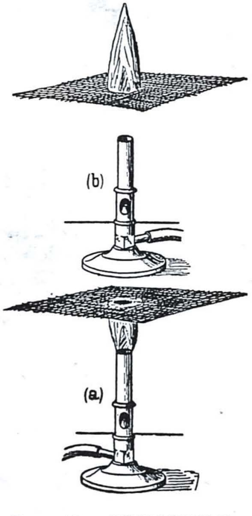
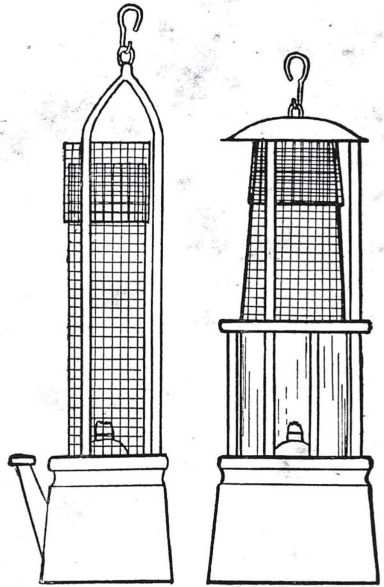


FIG. 41. PRINCIPLE OF SAFETY LAMP

Experiments like these led Davy to the idea of using a wire gauze as a ' *flame sieve* ', which, while allowing gas, air, and light to pass, would stop the passage of a flame. The ' *Davy safety lamp* ' consists of an oil-lamp surrounded by a cylinder of wire gauze. When placed in air containing ' fire-damp ', the methane passes through the gauze and burns *on the inside* of the gauze without igniting the explosive mixture outside.

Later and improved forms of the Davy lamp have a strong glass cylinder round the lower part of the lamp, and gauze above. This allows more light to escape and also prevents the flame from being blown through the gauze by a wind, since, in modern coal-mines, the ventilation current is sometimes very strong.



Original form Modern form

FIG. 42. SAFETY LAMPS

CARBOHYDRATES

Carbohydrates are compounds of carbon, hydrogen and oxygen, with the hydrogen and oxygen present in the same relative proportions as in water, i.e. 2 atoms of hydrogen to every atom of oxygen. The carbohydrates, which play a very important part in plant and animal life, can be classified as (i) *sugars*, (ii) *starches*, (iii) *celluloses*; or for chemical purposes as (*a*) *simple sugars* (mono-saccharides,† e.g. glucose and fructose†), (*b*) *complex sugars* (di-saccharides,† e.g. sucrose,† lactose† and maltose†), and (*c*) *complex carbohydrates* (e.g. starch, dextrin,† glycogen† and cellulose).

COMPLEX SUGARS—DI-SACCHARIDES

Ordinary household ' sugar ' is *sucrose* (or *cane-sugar*) and it is obtained from the sugar-cane,† the sugar-beet,† or the sugar-palm.† Sugar-cane is crushed between iron rollers to extract the cane-juice. This juice, which contains about 20 per cent of sugar, is then purified, filtered and evaporated, and the crystals of ' raw sugar ' are removed from the liquid residue (treacle† or molasses†). This crude brown sugar is refined by melting and decolorizing with charcoal, and it is then evaporated at a low temperature, to produce the familiar white crystals used in the household. Sugar-beets are crushed to a pulp and the sugar is extracted with water, the solution being purified and evaporated by processes similar to those used in the manufacture of sugar from the sugar-cane. Sugar is obtained from the sugar-palm by tapping the flower-shoots and evaporating the sap which flows out. The product, in each case, is *sucrose.*

Sucrose, $C_{12}H_{22}O_{11}$, is a complex sugar (or a di-saccharide) because it can be split up into two simple sugars by boiling with very dilute acids or by the action of certain *enzymes*† (see p. 98), one molecule of sucrose yielding one molecule of glucose and one of fructose:—

$$C_{12}H_{22}O_{11} + H_2O \longrightarrow C_6H_{12}O_6 \ (\textit{glucose}) + C_6H_{12}O_6 \ (\textit{fructose})$$

This mixture of glucose and fructose, produced by boiling sucrose with dilute acid, is called ' *invert sugar* ',† and it is more readily digested than sucrose. *Honey* is a natural form of invert sugar. *Jam** (particularly ' home-made ' jam) contains a large proportion of invert sugar formed when the original sucrose is boiled with the acid fruit-juice.

Lactose (or *milk-sugar*), which is present in cows' milk to the extent of about 5 per cent, is another complex sugar (or di-saccharide). Like sucrose, lactose can be split up into two simple sugars, one molecule of lactose yielding one molecule of glucose and one molecule of galactose.† Lactose is not sweet like sucrose, but is almost tasteless.

Maltose (or *malt-sugar*), another complex sugar, is formed in germinating seeds by the action of the enzyme *diastase*† on the starch stored as a reserve food material. (The enzyme *ptyalin*† in the saliva also converts starch into maltose.) Commercially, *barley*† is made into *malt*† by allowing the grains to germinate until some diastase has been formed. Germination is then stopped by heating, and the malted barley is extracted with water, yielding ' malt extract ', which contains maltose and diastase. Maltose, like sucrose and lactose, is readily split up by dilute acids or by a suitable enzyme (e.g. *maltase*,† which is present in the human small intestine), yielding two molecules of glucose.

SIMPLE SUGARS—MONO-SACCHARIDES

Glucose (*grape-sugar* or *dextrose*†) is the commonest and most important of the simple sugars. It is the first carbohydrate formed during photo-synthesis in green plants. It is present in some fruits, e.g. grapes† (raisins†—dried grapes—contain crystals of glucose). Glucose is also the final product of the digestion of carbohydrates in the animal body.

Glucose is always present in small quantities in the blood, being carried to the body-cells where it is oxidized to supply *energy*. Commercially, glucose is made by boiling starch with very dilute acids under pressure. The product provides a cheap sweetening agent used for making jams and sweets.

Fructose (*fruit-sugar* or *levulose*†), another simple sugar, is present in most sweet fruits and in the nectar of flowers. We have seen that it is formed, together with glucose, when sucrose is split up by heating with dilute acids or by the action of enzymes, e.g. *sucrase*,† or *invertase*,† which is present in the human small intestine.

Galactose, another simple sugar, is uncommon in the free state in Nature, but it is formed, along with glucose, when lactose is split up by heating with dilute acids or by the action of the enzyme *lactase*† (which is also present in the human small intestine).

All the simple sugars are sweet, crystalline substances, soluble in water. They cannot be split up into simpler carbohydrates, but as they are readily oxidized they are all good *reducing agents*.

EXPERIMENTS WITH SIMPLE SUGARS—MONO-SACCHARIDES

The chemical reactions of all the simple sugars are very similar and may be illustrated with a 1 per cent solution of glucose. They are all '*reducing sugars*' and will reduce copper salts to red cuprous oxide and silver solutions to metallic silver. Copper sulphate is most commonly used in alkaline solutions. (N.B.—All the common sugars we have discussed—except sucrose—are 'reducing sugars', but the reducing power of the simple sugars is greater than that of the complex sugars.)

Fehling's Test.—Take 2-3 c.cm. of the test solution and add an equal volume of diluted (1/5) Fehling's solution (an alkaline solution of copper sulphate containing also Rochelle salt (sodium hydrogen tartrate)). Place in a boiling water-bath. A red precipitate of cuprous oxide shows the presence of a reducing sugar.

Silver Mirror Test.—Take 2-3 c.cm. of silver nitrate solution and add a very weak solution of ammonia (e.g. 1 c.cm. of 2N ammonia in a test-tube full of water) *drop by drop*, until the precipitate first formed is just re-dissolved. Then add an equal volume of 1 per cent glucose solution and warm on a water-bath. A mirror of metallic silver forms on the sides of the tube.

EXPERIMENTS WITH COMPLEX SUGARS—DI-SACCHARIDES

There are two kinds of di-saccharides, (*a*) those that reduce copper solutions, e.g. maltose and lactose, and (*b*) those that do not reduce, e.g. sucrose. The reducing power of glucose and fructose, however, is greater than that of maltose and lactose.

Repeat Fehling's test with maltose, lactose and sucrose, and compare their reducing power with that of glucose as follows:—Take five test-tubes, A, B, C, D, and E, each containing 2-3 c.cm. of dilute Fehling's solution (1/5). To A add 1 c.cm. of 1 per cent glucose solution, to B add 1 c.cm. of 1 per cent fructose solution, to C add 1 c.cm. of 1 per cent maltose solution, to D add 1 c.cm. of 1 per cent lactose, and to E add 1 c.cm. of 1 per cent sucrose. Place all five test-tubes in a boiling water-bath and notice when a precipitate of cuprous oxide appears: first in A (glucose) and B (fructose) and a little later in C (maltose) and D (lactose). Notice that *no precipitate forms in* E (sucrose). *Sucrose has no reducing action*, and the reducing power of glucose and fructose is greater than that of maltose and lactose.

To convert Complex Sugars into Simple Sugars.—Take 2-3 c.cm. of (*a*) 1 per cent sucrose solution, (*b*) 1 per cent maltose solution, and (*c*) 1 per cent lactose solution, and add to each tube 1 c.cm. of dilute hydrochloric acid (2N). Boil for 2 minutes. Then cool, and neutralize with 1 c.cm. of dilute sodium hydroxide solution (2N). Take four test-tubes, A, B, C, and D, and put 2-3 c.cm. of diluted Fehling's solution in each. To A, add 2 c.cm. of the boiled sucrose solution; to B, add 2 c.cm. of the boiled maltose solution; to C, add 2 c.cm. of the boiled lactose solution; and to D, add 2 c.cm. of 1 per cent glucose solution, unboiled. Place all four tubes in a boiling water-bath and notice the rapid reduction to red cuprous oxide in every case, showing that the complex sugars sucrose, lactose and maltose have been split up into simple sugars with greater reducing power.[1]

COMPLEX CARBOHYDRATES—POLY-SACCHARIDES†

Starch is one of the commonest and most important of the complex carbohydrates. It is the commonest reserve food material in plants, where carbohydrates are *transported* in the form of *soluble glucose* and are *stored* in the form of *insoluble starch.* (There is a very similar arrangement in the animal body—we shall refer to this later.) In plants, starch is in the form of grains, and as each starch-grain is surrounded by a layer of cellulose, this raw starch is insoluble in water. On boiling, however, the starch-grains swell up and burst, setting free the 'true' starch which dissolves in the water forming a translucent solution which becomes a jelly on cooling. We have seen that starch can be changed

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[1] Only a selection of experiments is given in this book to illustrate the physiology of animals. A further group of experiments is given in the Teacher's Handbook together with some details of the preparation and composition of solutions.

into glucose by heating with dilute acids or by the action of suitable enzymes.

Dextrin is another complex carbohydrate formed when starch is heated to about 200° C. Dextrin is also formed as an intermediate* product when starch is heated with acids or treated with enzymes, e.g. ptyalin. It is a yellow or brown powder which dissolves in water. Dextrin is used commercially for making gum, e.g. for postage stamps.

FIG. 43. STARCH GRAINS (highly magnified)
A—starch grains being attacked by diastase. B—single starch grain.

Glycogen (sometimes called '*animal starch*') is the commonest carbohydrate reserve in animals (just as starch is the commonest carbohydrate reserve in plants) where it is found mainly in the liver and the muscles.

Glycogen is difficult to extract from liver and muscle since it is rapidly changed into sugar after the animal dies. It is most readily obtained from oysters and other molluscs. Glycogen is soluble in water, forming a translucent solution.

Cellulose is the name given to a number of complex carbohydrates which form the cell-walls of plants. Cotton, linen and filter-paper are nearly pure cellulose. Cellulose, in its various forms, is of very great importance to man. It is one of the main constituents of wood, which is used in enormous quantities for a great variety of purposes. Paper, too, is mainly cellulose. The cheapest quality of paper (' newsprint ') is made from wood-pulp which is heated with chemicals, leaving cellulose. Better quality paper is made by mixing the wood-pulp with china-clay or some similar ' filler ' and then rolling the mixture into sheets. The best quality paper is made from old linen, the cellulose fibres in linen being longer than those in wood-pulp, thus producing a tougher paper. Paper passed quickly through fairly concentrated sulphuric acid becomes water-proof.

EXPERIMENTS WITH COMPLEX CARBOHYDRATES—POLY-SACCHARIDES

Starch. (i) Take about 1 c.cm. of 1 per cent starch solution in a test-tube and add one drop of iodine solution (N/20 in potassium iodide). Notice the deep blue-black colour. Fill up the tube with water and notice that the colour is still deep blue.

(ii) Take about 1 c.cm. of 1 per cent starch solution and one drop of iodine solution, add about 5 c.cm. of water and boil. Notice that the blue colour disappears on heating. Cool under the tap. Notice that the blue colour returns on cooling. (N.B.—Remember that the iodine test does not work in hot solutions.)

(iii) Take 1 c.cm. of 1 per cent starch solution and make it alkaline to litmus with sodium carbonate solution. Add a drop of iodine solution. There is no blue colour. Repeat this with 1 c.cm. of starch solution which has been acidified with dilute hydrochloric acid: iodine gives the usual blue colour. (N.B.—Remember that the iodine test does not work in alkaline solutions, but only in neutral or acid solutions.)

(iv) Take a small quantity of dry rice-starch in a test-tube and shake vigorously with about 5 c.cm. of cold water. Filter half the mixture. Add a drop of iodine solution to the filtrate and notice that no blue colour develops, showing that starch is insoluble in cold water. Boil the remainder of the suspension and then filter. Add one drop of iodine solution to the cold filtrate. Notice that a blue colour is produced at once, showing that starch is soluble in hot water.

Dextrin. Take 2-3 c.cm. of a 1 per cent solution of dextrin and add one drop of iodine solution. Notice the reddish-brown colour. (Compare this with the colour produced by adding one drop of iodine to 2-3 c.cm. of water.) Heat the mixture of dextrin and iodine—the reddish-brown colour disappears. Cool under the tap and notice that the colour returns.

Glycogen. Take two test-tubes, A and B. In A put about 5 c.cm. of the given glycogen solution, and in B put about 5 c.cm. of water. Add iodine solution to each tube drop by drop and compare the colours. The reddish-brown colour in A is due to the presence of glycogen. (N.B.—The iodine test gives similar results with dextrin and glycogen. *Dextrin* solutions, however, are clear and *transparent,* while *glycogen* solutions are *translucent.*)

Cellulose. Put one drop of iodine on a piece of filter-paper. No blue colour is formed. Add one drop of strong sulphuric acid, 50 per cent: it turns blue.

TO CONVERT STARCH INTO SIMPLER CARBOHYDRATES

(i) First show that starch has no reducing action. Take 2-3 c.cm. of 1 per cent starch solution and add an equal volume of dilute Fehling's solution. Place in a boiling water-bath. No precipitate of cuprous oxide is formed, showing that *starch has no reducing action* on Fehling's solution.

(ii) Take about 10 c.cm. of 1 per cent starch solution in a test-tube and add 1 c.cm. of concentrated hydrochloric acid. Place in a boiling water-bath, noting the time. On a white tile* (or a glass plate over white paper) put 12 drops of iodine solution, and at the end of every minute, quickly stir the starch solution with a clean glass rod and then remove a drop of the solution and test it with a drop of iodine on the tile. Notice the gradual change of colour from blue (starch) to purple and then reddish-brown (dextrin); finally the solution becomes colourless. Pour 2-3 c.cm. of the remaining solution into a test-tube and neutralize with about 1 c.cm. of sodium hydroxide solution (2N). Then add an equal volume of diluted Fehling's solution and heat in a boiling water-bath. A red precipitate of cuprous oxide shows that sugar (glucose in this case) has been formed.

(iii) It is interesting to compare the splitting-up of starch by acids with that brought about by the enzyme *ptyalin.* Put 2-3 c.cm. of 1 per cent starch solution in a test-tube and add a little saliva. Shake up and stand it in the 37° C. water-bath for a few minutes. Then add an equal volume of dilute Fehling's solution and heat in a boiling water-bath. A red precipitate of cuprous oxide shows the presence of a reducing sugar.

GLYCOGEN

(i) Take about 5 c.cm. of a glycogen solution and add a few drops of concentrated hydrochloric acid. Heat in a boiling water-bath for 10 minutes, then cool under the tap and divide into two parts A and B. To A add two or three drops of iodine solution. No reddish-brown colour develops, showing that the glycogen has disappeared. Neutralize B with sodium hydroxide solution, add an equal volume of diluted Fehling's solution, and heat in a boiling water-bath. A red (or yellow) precipitate shows that the glycogen has been converted into a reducing sugar (glucose in this case). (ii) If time permits, try the action of saliva (see (iii) above).

Pure glycogen is very expensive to buy in the solid form, but a solution is readily prepared from oysters, mussels, or similar molluscs, as follows:—Boil 200 c.cm. of water in a beaker and then drop in several oysters or mussels. Boil gently for half an hour to extract the glycogen and to clot most of the proteins. Then add dilute acetic acid, drop by drop, until the liquid just reddens litmus, to complete the clotting of proteins. Filter, and carry out the experiments with the translucent liquid. Tinned oysters can be used, in which case most of the glycogen is in the *liquid*, which should be heated in a boiling water-bath for half an hour, acidified with acetic acid, and then filtered, as above.

ALCOHOLIC FERMENTATION

It has been known for centuries that when grape-juice or any other dilute solution of sugar is left exposed to the air, it loses its sweet taste, acquires a burning taste, and becomes intoxicating.* The sugar is said to have *fermented* and has been converted into *ethyl alcohol*, giving off carbon dioxide in the process:—

$$C_6H_{12}O_6 \longrightarrow 2C_2H_5OH + 2CO_2$$

The fermentation is brought about by *yeast*, a microscopic, non-green plant which consists of a single cell. Instead of seeds, the yeast-plant (like other *fungi*) forms *spores*, which are small enough and light

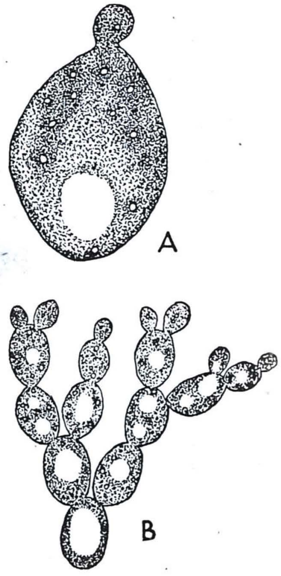

FIG. 44. YEAST
A—single yeast plant (×2,500).
B—yeast plants reproducing by 'budding'.

enough to float about in the air, so that whenever sugar solutions are exposed to the air, yeast spores soon enter, grow, and multiply, setting up fermentation. If this ' living dust ' is kept out, sugar solutions cannot ferment. The actual chemical change involved in fermentation is brought about by an *enzyme* (called *zymase*†) contained in the yeast-cells, which acts as an organic catalyst. This fermentation is a special form of *respiration* which takes place *in the absence of oxygen*, setting free energy for life and growth.

Fermentation stops when about 14 per cent of alcohol has been formed, and stronger alcohol is obtained from this dilute alcohol by *fractional distillation*; the alcohol, having a lower boiling-point than water, distilling off first. The strongest and purest alcohol is called *absolute alcohol* and contains over 99 per cent of ethyl alcohol. In order to discourage people from drinking alcohol sold for industrial purposes, crude wood-alcohol (methyl alcohol), paraffin, bone-oil and dyes are added, the mixture being known as *methylated spirit*. Ethyl alcohol is used in large quantities as a solvent for making varnish and it is also used, mixed with petrol, as a motor fuel.

Just as there is a *paraffin series* of hydrocarbons (methane—CH_4, ethane—C_2H_6, etc.), there is a corresponding series of *simple alcohols*, e.g. *methyl alcohol*—CH_3OH, *ethyl alcohol*—C_2H_5OH, and so on. In everyday use, when we speak of ' alcohol ' we mean ' ethyl alcohol ', as this is the commonest alcohol. ' Wood-alcohol ' is ' methyl alcohol ' and is a very poisonous substance.

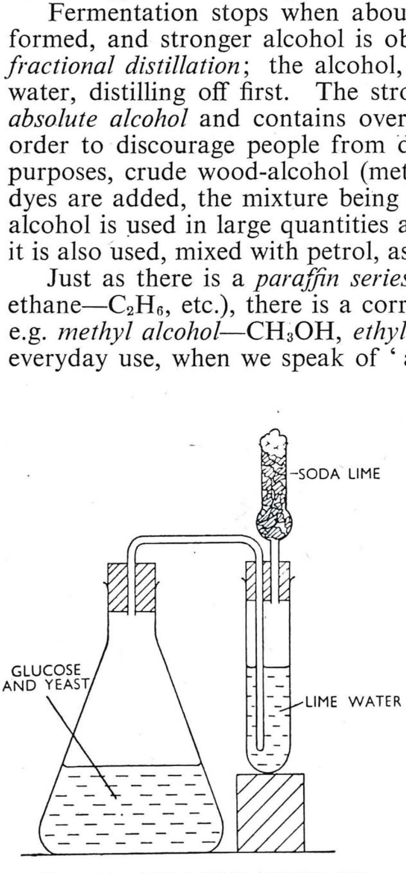

FIG. 45. FERMENTATION OF GLUCOSE

TO SHOW ALCOHOLIC FERMENTATION OF GLUCOSE

(*Demonstration*.) In a 500 c.cm. boiling-flask put 250 c.cm. of a 10 per cent solution of glucose. Add 5 gm. of dried yeast and fit the flask with a cork carrying a glass tube leading to a tube containing clear lime-water, as shown in Fig. 45. This tube is closed by a two-holed cork carrying a soda-lime tube to prevent any carbon dioxide entering from the air. Keep the flask at about 37° C. for 24 hours (or if this is not possible, first warm the flask to this temperature and afterwards keep in a warm room for at least 24 hours). Notice that large volumes of gas escape through the lime-water, turning it chalky at first, but it afterwards becomes clear owing to the large volume of carbon dioxide which is given off (see Book II, p. 13).[1] After a day or two, filter the contents of the flask into a distilling-flask and distil over a small flame, collecting the first 25 c.cm. to come over. Take about 5 c.cm. of this distillate in a test-tube and add a few drops of sodium hydroxide solution (2N). Then add iodine solution, drop by drop, until the mixture is faintly brown. Warm gently, and notice the peculiar smell. Repeat this test with a few drops of ethyl alcohol. The same smell is produced. This is due to the compound *iodoform*,† produced from alcohol by the action of the alkali and the iodine. Repeat the test with a few c.cm. of 10 per cent glucose solution—there is no smell of iodoform. Hence the yeast has converted some of the sugar into alcohol, setting free carbon dioxide during the process.

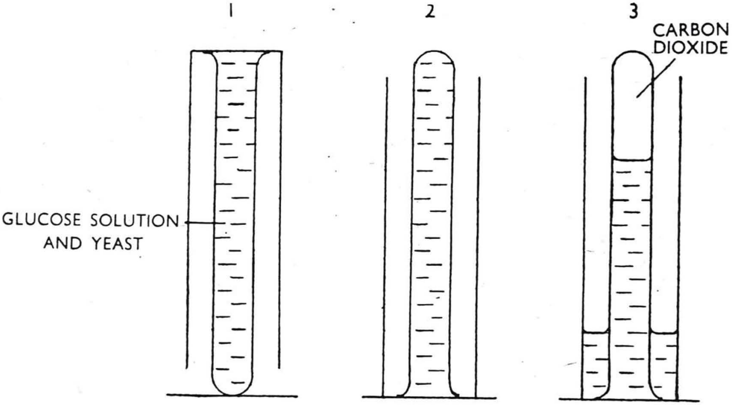

FIG. 46. FERMENTATION OF GLUCOSE

A SMALL-SCALE EXPERIMENT TO SHOW ALCOHOLIC FERMENTATION

Take 1·5 gm. of dry yeast (e.g. zymin†) in a mortar, and add to it, a little at a time, 15 c.cm. of a 20 per cent solution of glucose, grinding thoroughly after each addition. Then quite fill a 6-in. by $\frac{1}{2}$-in. test-tube

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[1] Baryta water—barium hydroxide† solution—may be used instead of lime-water as it does not become clear again when carbon dioxide is passed for a long time.

with the mixture and carefully push this tube up into an inverted 6-in. by 1-in. flat-bottomed tube (a specimen-tube) until the flat bottom of this tube closes the mouth of the test-tube, as shown in Fig. 46 (1). The apparatus can then be inverted without any solution escaping from the test-tube or any air entering (Fig. 46 (2)). (A small air-bubble, however, may be neglected.) The apparatus is then put in the 37° C. water-bath and, after a time, bubbles of gas begin to rise to the top of the test-tube, displacing the liquid into the outer tube (Fig. 46 (3)). After about half an hour the test-tube becomes full of gas. When this stage is reached, remove the apparatus from the water-bath, and, keeping the mouth of the test-tube closed by the bottom of the specimen-tube, pour off the displaced liquid into another test-tube and set aside to test for alcohol later. Into the outer specimen-tube, pour about 15 c.cm. of sodium hydroxide solution (2N) and then rock the inner tube gently. Notice that the alkali rises inside the test-tube as the carbon dioxide is absorbed, and continues rising until it fills the test-tube, showing that the tube was full of carbon dioxide.

Filter the mixture of yeast, sugar and alcohol poured off at the end of the experiment, and to the filtrate add a few drops of sodium hydroxide solution (2N) and then add iodine solution, drop by drop, until the liquid is faintly brown. Warm gently, and notice the characteristic smell of *iodoform* (CHI_3), showing the presence of *ethyl alcohol.*

Examine a little of the residue from the filter-paper under the high power of the microscope.

If you examine, under the high power of the microscope, a drop of the liquid from the flask, where the yeast has had a good supply of food for over 24 hours, you may see the yeast cells ‘ *budding* ’, i.e. reproducing by division.

VINEGAR AND ACETIC ACID

Fermented sugar solutions turn ‘ sour ’ when they are exposed to the air, owing to the formation of *acetic acid.* This is formed by another fermentation process brought about by ‘ acetic bacteria ’, the spores of which are always floating about in the air. The ethyl alcohol is oxidized to acetic acid:—

$$C_2H_5OH + O_2 \longrightarrow CH_3COOH + H_2O$$

Acetic acid—CH_3COOH—is one of the simplest members of another important series of organic compounds—the *fatty acids*—so called because the higher members of the series are found, combined with *glycerine*, in many fats. It is present in *vinegar*,† formed by allowing fermented sugar solutions to ferment still further; but concentrated acetic acid is prepared from the watery liquid which distils off when wood is heated in closed retorts (see p. 47). The *acetates*† are salts of acetic acid.

‘ ORGANIC SALTS ’—ESTERS

The alcohols, in organic chemistry, are rather like the bases we have studied in inorganic chemistry, and just as *acid + base*——→*salt + water* in inorganic chemistry, so *fatty acid + alcohol*——→‘ *organic salt* ’ + *water* in organic chemistry. Such ‘ organic salts ’ are called *esters.*† A simple example of *ester formation* is the reaction between acetic acid and ethyl alcohol to form *ethyl acetate,*† and water:—

$$CH_3COOH + C_2H_5OH \longrightarrow CH_3COOC_2H_5 + H_2O$$

Mix together in a test-tube about 1 c.cm. each of acetic acid (concentrated), ethyl alcohol, and concentrated sulphuric acid (to remove the water from the reaction), and warm the mixture gently. Notice the sweet smell of ethyl acetate.

OILS AND FATS

Many plants store *oil* as a reserve food material instead of starch, e.g. coconut-oil, palm-oil, soya-bean-oil, olive-oil,† etc. Fishes and other ‘ cold-blooded ’ sea animals also produce oils, e.g. cod-liver-oil,† while ‘ warm-blooded ’ animals store similar substances with higher melting-points, which are solids at ordinary atmospheric temperatures. These solid oils we call *fats.* (Coconut-oil and palm-oil are *liquid oils* in the tropics but are *solid fats* elsewhere.)

Such oils and fats are *not* hydrocarbons like petroleum, but they contain the elements *carbon, hydrogen and oxygen.* They are *not,* however, carbohydrates, since the hydrogen and oxygen are not present in the same proportions as in water. There is in fats much less oxygen in proportion to hydrogen than in carbohydrates. Fats are ‘ organic salts ’, or *esters,* being compounds of the ‘ organic base ’ *glycerine* (or *glycerol,*† since it is an alcoh*ol*) and *fatty acids* of high molecular weight: one molecule of glycerine combines with three molecules of a fatty acid (since glycerine contains three (OH) groups). For example, olive-oil is mainly composed of the glyceride of *oleic acid*†; palm-oil is mainly the glyceride of *palmitic acid*†; mutton-fat† contains the glyceride of *stearic acid,*† and butter is mainly the glyceride of *butyric acid.*†

TESTS FOR FATS

When a drop of oil is placed on a piece of filter-paper, a permanent translucent *grease-spot* is formed. This grease-spot, unlike the translucent spot produced by water or any liquid that evaporates, does not disappear on standing, and it cannot be washed away with water. This is a useful test for fats when present in fairly large quantities.

(i) Place one drop of coconut-oil on a strip of filter-paper and notice that a permanent translucent grease-spot is formed. (Keep this strip of filter-paper for Experiment v.)

(ii) Place single drops of (*a*) water, (*b*) alcohol, (*c*) petrol, (*d*) glycerine, on strips of filter-paper. Notice that although a translucent spot is first formed in each case, the water, alcohol, and petrol soon evaporate and leave no trace behind. The glycerine, however, does not evaporate. Wash the fourth (glycerine) paper under the tap and then pin it up to dry. Notice that no translucent spot is left after drying the paper.

(iii) If the substance under test contains very little fat, we can make use of the fact that oils and fats are soluble in petrol. Take about 1 gm. of crushed soya-bean in a *dry* test-tube and add about 5 c.cm. of petrol (away from flames). Close the mouth of the tube with the thumb and shake vigorously for several minutes. Then allow the solid to settle. With a narrow glass tube, remove a little petrol and allow it to fall—a drop at a time—on a strip of filter-paper, allowing one drop to evaporate before adding the next drop. Notice that a permanent grease-spot is formed as the petrol evaporates. (Keep this filter-paper for Experiment v.)

(iv) If the substance contains much fat, it will form a grease-spot if it is placed on a filter-paper and gently warmed. Try this test with some oil-yielding seeds.

(v) Take the strips of filter-paper from Experiments (i), (ii *d*) and (iii), place them in a test-tube and add a few c.cm. of Sudan III solution. Shake and then allow to stand for several minutes. Then remove the strips and wash with water, under the tap. Notice that the grease-spots are now stained a bright orange-red. This staining test enables us to test for fat in substances which do not give the grease-spot test readily.

SOAPS

There is a whole series of *fatty acids* corresponding to the paraffin series of hydrocarbons, one of the simplest members being acetic acid. When fatty acids are neutralized by alkalies, salts are formed, e.g. when acetic acid is neutralized with sodium hydroxide, sodium acetate† is formed:—

$$CH_3COOH + NaOH = CH_3COONa + H_2O$$

Similar salts are formed when any fatty acid is neutralized by a base; but the sodium and potassium salts of the higher fatty acids have peculiar properties and are called *soaps*. The *sodium salts of the higher fatty acids are 'hard soaps'*, while the corresponding *potassium* compounds are '*soft soaps*'.

Soaps are not made commercially by direct neutralization of fatty acids with alkalies, but by boiling fats with caustic alkalies, forming a mixture of soap and glycerine. *Fat + alkali = soap + glycerine.* Fats or oils are mixed with caustic soda and the mixture is heated, either by steam-coils or by blowing steam through the liquid. After the mixture has been heated for some time, all the fat is converted into soap. At this point salt is added. The soap, being insoluble in salt solution, separates out as a solid on the surface, the glycerine remaining behind in solution. The soap is purified and is then pressed into moulds where it cools and hardens. The glycerine, being a very valuable by-product, is recovered by distillation. It is used mainly for making explosives and for medicinal purposes.

Only sodium and potassium soaps are soluble in water. The stearates, oleates and palmitates of other metals are all *insoluble soaps*. Hence, when soap is added to *hard water* (containing calcium and magnesium salts) no lather is produced at first, but only a surface layer consisting of the calcium and magnesium soaps, which are insoluble in water. For example:—

Sodium stearate† + calcium sulphate = calcium stearate† + sodium sulphate
(*soluble soap*) (*insoluble soap*)

Hence the amount of soap required to produce a permanent lather measures the *hardness* of a sample of water (see Book II, p. 15).

It is possible that the formation of soluble soaps plays a part in the digestion of fats. The pancreatic juice contains an enzyme—*lipase*†—which splits up fats into fatty acids and glycerine. It is thought that these fatty acids may combine with the sodium carbonate in the pancreatic juice to form soaps, which are then absorbed, along with the glycerine, both being soluble in water. After absorption, the fatty acid is again set free and combines with the glycerine, producing fat once more in the lymph vessels (*lacteals*†). The sodium carbonate set free from the absorbed soap is returned to the intestine to combine with more fatty acid.

TO PREPARE SOME HARD SOAP

(*Demonstration.*) Take a 250 c.cm. flask fitted with a cork carrying a long glass tube (4-5 ft. long). Put 25 c.cm. of coconut-oil in the flask, and add 100 c.cm. of alcoholic sodium hydroxide solution (N =4 gm. NaOH in 100 c.cm. of alcohol or methylated spirit). Replace the cork and heat over a water-bath. By using alcoholic caustic soda, we obtain a reaction very quickly, since the oil dissolves in the alcohol where it can react with the alkali. Any alcohol which boils away condenses in the long glass tube and is returned to the flask. After 10 minutes, quickly pour out a few drops of the clear solution into some water in a test-tube and replace the flask on the water-bath. If drops of oil separate in the water in the test-tube, the reaction is not yet complete. If a clear solution is obtained, all the oil has been converted to soluble soap. When the reaction is complete, pour the hot solution into a basin and allow it to cool. The alcohol soon evaporates, leaving a solid mass of soap (which still contains some glycerine).

(*Class Experiments.*) (*a*) Take a small, stoppered bottle and put in it 25 c.cm. of coconut oil. Take 20 c.cm. of 20 per cent sodium hydroxide solution in a measuring cylinder, and add about 1 c.cm. of this to the oil. Shake the bottle vigorously for 2 minutes (so as to break up the oil into tiny droplets, thus increasing the surface area of oil and alkali in contact with each other). Then add 5 c.cm. of caustic soda and shake for 1 minute, and so on until all the alkali has been added. Then pour out about half this mixture into a 100 c.cm. beaker and allow to stand while you do the following experiments with the remainder of the soap.

The success of this method depends upon adding the alkali very slowly at first and shaking very vigorously. It also works better with some samples of coconut oil than with others. The following experiment gives an alternative method.

(*b*) Weigh out 4 gm. of solid sodium hydroxide in an evaporating-dish and dissolve it in a mixture of 10 c.cm. of water and 5 c.cm. of alcohol. Then add a mixture of 25 c.cm. of coconut oil and 5 c.cm. of alcohol slowly, stirring all the time with a glass rod. Heat on a boiling water-bath with constant stirring until the mixture becomes stiff (10-15 minutes).

EXPERIMENTS WITH SOAP

(i) Half fill the stoppered bottle with warm water and shake vigorously to dissolve the soap. Notice the characteristic lather—a mass of air-bubbles, each bubble surrounded by a film of soap. Keep this solution for the following experiments.

(ii) Put about 25 c.cm. of saturated brine into a 100 c.cm. beaker and pour into it an equal volume of the soap solution. Notice that the soap is ‘ salted out ’ and rises to the surface.

(iii) Take about 10 c.cm. of the warm soap solution in a test-tube, add to it about 1 c.cm. of concentrated hydrochloric acid and shake. The fatty acids are set free, forming a milky suspension. On standing, the fatty acids rise to the surface to form a separate oily layer which solidifies on cooling.

(iv) *Emulsions.* Take three test-tubes, A, B, and C. Put 10 drops of coconut oil in each tube. To A add about 10 c.cm. of water; to B add 10 c.cm. of water and a few drops of sodium hydroxide solution; to C add 5 c.cm. of water and 5 c.cm. of your soap solution. Shake all three tubes vigorously and then allow them to stand in the test-tube stand. In each case, vigorous shaking breaks up the oil into very fine drops which are suspended in the liquid to form an *emulsion.*† On standing, however, the oil-drops in A soon run together and rise to the surface to form a separate layer, i.e. the emulsion is unstable. In B, the emulsion is permanent, because the alkali combines with the oil to form a little soap which forms a protective film round each tiny oil-drop. In C, the emulsion is also stable, for the same reason.

In the same way, when greasy or oily articles are washed with soap, the soap solution ‘ *emulsifies* ’ the grease, and this emulsion can easily be washed away with water.

(v) (*a*) Put a little lamp-black (finely-divided carbon) in a test-tube and add about 10 c.cm. of water. Shake vigorously. Filter half the suspension and allow the remainder to stand. The filtrate is colourless, all the solid particles being retained by the filter. On standing, the solid particles gradually settle out from the suspension.

(*b*) Put about the same quantity of lamp-black in another test-tube and add 5 c.cm. of water and 5 c.cm. of the soap solution. Shake vigorously. Filter half the mixture, as before, and notice that the filtrate is black, each separate particle of lamp-black having been surrounded by a film of soap and thus enabled to pass through the filter. Allow the remainder of the mixture to stand—the lamp-black remains suspended. This experiment explains how soap removes dirt, each particle of dirt being surrounded by a film of soap so that it is readily washed away with water.

PROTEINS

Proteins are the third great class of organic compounds present in plants and animals along with carbohydrates and fats. Seeds are always rich in proteins, and small quantities of proteins are found in roots, stems and leaves. The bodies of animals always contain a larger proportion of protein than is found in plants. These animal proteins are all obtained, in the first place, from plant proteins. Plants synthesize their proteins from simple materials like water, carbon dioxide, nitrates and ammonium compounds.

Proteins always contain carbon, hydrogen and oxygen *together with nitrogen*. Nearly all the proteins also contain *sulphur*, while many of them contain *phosphorus* as well. Proteins are some of the most complex of all chemical compounds, their molecules being very large and containing enormous numbers of atoms ; for example, $C_{146}H_{226}N_{44}SO_{50}$ is the sort of formula a protein molecule may have ! In a General Science Course of this kind we cannot attempt more than an outline of the chemistry of the proteins, but in view of their great everyday importance as essential constituents of all living things and as food-materials we cannot omit the proteins.

AMINO-ACIDS

When proteins are boiled with acids for a long time, they are split up into simpler compounds called *amino-acids*,† in fact, a protein molecule appears to consist of a long chain of amino-acids. There may be 50 to 500 links in this chain, and each link is an amino-acid molecule. These amino-acid ‘ links ’ can be ‘ unlocked ’ by heating with acids or by the action of *protein-splitting enzymes* (just as the poly-saccharides and di-saccharides can be split up into mono-saccharides). Or, to put it another way, the amino-acids are the units, or ‘ bricks ’, from which proteins are built.

About 20 of these amino-acids are found in natural proteins, and one protein differs from another in the number and arrangement of its constituent amino-acids.

The simplest amino-acid is *glycine*.† The name means ‘ sweet glue ’, since glycine has a sweet taste and was first prepared by boiling glue or gelatine with acids. The chemical name for glycine is *amino-acetic acid*, as it is derived from acetic acid, CH_3COOH, one hydrogen atom of the (CH_3) group being replaced by the *amino* group (NH_2), giving $CH_2(NH_2)COOH$. All the amino-acids can be regarded similarly as derivatives of fatty acids and ammonia.

This has an important bearing on the use of proteins as foods. The proteins used as food by animals are very seldom the same as those in the animal's body-tissues, and the sole object of the digestion of proteins is to split up the food-proteins into their constituent amino-acids before they enter the blood-stream.

The body-cells then select from these amino-acids the ones which can be built up into the kind of body-protein they require for growth and repair.

During digestion, the very complex proteins are broken down by stages:—

PROTEINS——→PROTEOSES†——→PEPTONES†——→POLY-PEPTIDES†——→AMINO-ACIDS

We shall learn more about these changes later.

All proteins are not of equal food-value, because they do not all contain all the essential amino-acids in suitable proportions. A protein which contains all the essential amino-acids in the proportions necessary to support human life is called a ‘ complete ’ (or ‘ first-class ’) protein, or it is said to be a protein of good biological value. A protein which does not contain all the essential amino-acids, or which contains them in unsuitable proportions, is called an ‘ incomplete ’ (or ‘ second-class ’) protein, or it is said to be a protein of poor biological value. All animal proteins (except gelatine) are ‘ complete ’ proteins, while plant proteins are all ‘ incomplete ’. Examples of animal protein are lean meat (beef, mutton, fish and birds), eggs, milk and cheese. These are all expensive foods. Examples of foodstuffs supplying vegetable proteins are the various kinds of legumes (pod-bearing plants—peas, beans, pulses, lentils, soya-beans). These plant-proteins are cheaper than animal proteins.

The amino-acids with which we are most concerned in protein foods are tryptophane,† tyrosine,† cystine,† arginine,† lysine,† histidine† and phenyl-alanine.†

TESTS FOR PROTEINS

Nearly all the tests for proteins depend upon reactions given by the constituent amino-acids, hence *all* proteins do *not* give *all* the colour reactions, so that it is necessary to carry out several tests before the presence of protein is verified. Try the following tests on a solution of egg-white.

BIURET REACTION

This test is given by all *proteins* (in which the individual amino-acids are linked together) but *not by the individual amino-acids.* Take 2-3 c.cm. of the test-solution in a test-tube and add an equal volume of sodium hydroxide solution (2N—8 per cent). Then add two drops of 1 per cent copper sulphate solution, and shake gently to mix. Proteins give a *violet* colour at once, without heating. (The less complex proteoses and peptones give a reddish or pink colour.) In doubtful cases, a control tube should be used, with 2-3 c.cm. of water, an equal volume of sodium hydroxide solution, and two drops of 1 per cent copper sulphate solution, comparing the colour of this liquid with that given by the test solution.

XANTHO-PROTEIC REACTION

This is given by the amino-acids *tyrosine*, *tryptophane* and *phenylalanine* (either the individual amino-acids or the amino-acids when present in combination in a protein). Take 2-3 c.cm. of the test-solution and add about 1 c.cm. of concentrated nitric acid. If one of the above amino-acids is present, a white precipitate is formed which turns *yellow* and partly dissolves on warming. *Cool* under the tap and then add excess of ammonia (*care* !). The yellow colour turns to orange.

SECTION II

PHYSIOLOGY

CHAPTER IV

THE STRUCTURE AND FUNCTIONS OF LIVING MATTER

CELLS

In 1665, Robert Hooke, examining thin slices of bottle-cork (cork-oak† bark) under the newly invented microscope, noticed that the cork, when magnified, looked very much like a honey-comb, being built up of a large number of closely-fitting little boxes to which he gave the name *cells*. The cork-cells examined by Hooke were actually *dead cells*, and it was not until nearly 200 years later that there was any clear idea of the true nature of cells and that *the most important part of a cell is its living contents and not its dead cell-walls.*

All living matter, whether plant or animal, is alike in being built up of these tiny units called *cells*. Some of the simplest living things consist of one single cell, e.g. Amoeba, but a higher plant or a higher animal consists of millions of these cell-units. All cells, however, are not alike, for they have different functions to perform. For example, the cells in the skin of a leaf are different from those in the middle of a leaf; the cells of animal muscle are different from nerve-cells, and so on. Collections of similar cells, all doing similar work, are called *tissues*. In higher plants, for example, there is one type of cell that forms the *strengthening tissue* in the stem, and another kind that forms the *conducting tissue* through which manufactured food-materials travel from the leaves to other parts of the plant. In higher animals there is skin-tissue, nerve-tissue, muscle-tissue, bone-tissue, gland-tissue, etc., all carrying out different functions and built up of different kinds of cells. **Cells, therefore, are the tiny living units from which plants and animals are built up.**

CELL STRUCTURE

A young *plant cell* consists of a tiny ‘ box ’ with a thin *cell-wall* composed of the carbohydrate *cellulose*. This non-living cell-wall encloses the actual living matter—*protoplasm*—which is the basis of all *life*. Protoplasm is a colourless, almost transparent, jelly-like substance, somewhat like the ‘ white ’ of a raw egg, and of very complex chemical composition. The protoplasm of almost every cell contains a denser, rounded portion called the *nucleus*, which seems to control the working of the cell. Fine threads of protoplasm pass through tiny holes in the cell-wall and join up the protoplasm of neighbouring cells, so that the protoplasm is continuous from end to end of the plant. In very young plant-cells, e.g. those at the growing point of the stem or root, the protoplasm fills the whole of the cell, but as the cell grows bigger, *cell-spaces* appear in the protoplasm, and these spaces are filled with *cell-sap*, a watery solution of sugars and mineral salts. As the cell grows still larger, the protoplasm merely forms a living lining round the inside of the cell-wall, and the middle of the cell is filled with sap. In still older cells, e.g. the *heart-wood* of an old tree, the protoplasm disappears altogether and the cells are no longer living (although this dead tissue is still useful to the tree in providing strength).

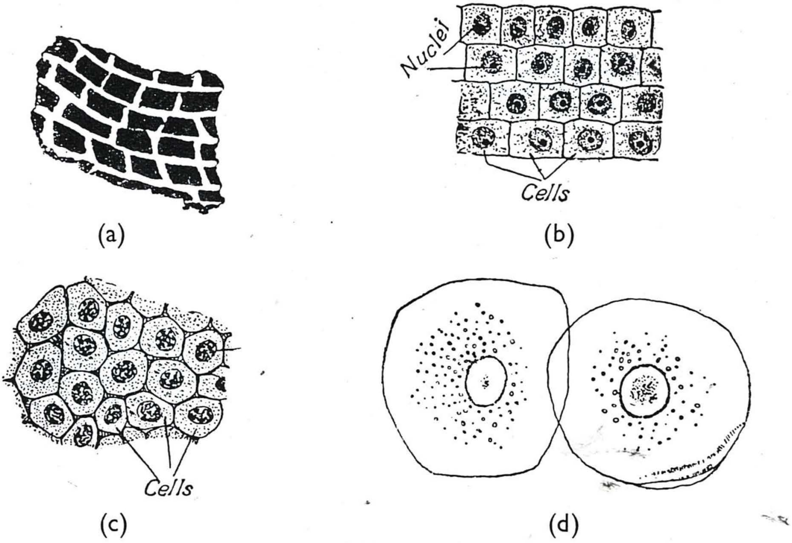

FIG. 47 (highly magnified).
(*a*) Hooke’s drawing of cork cells. (*b*) Plant-cells from young root.
(*c*) Animal skin-cells. (*d*) Cells from inside of cheek.

An animal-cell also consists of protoplasm with a nucleus, but it is surrounded by a very thin, *living*, cell-membrane (composed of a firmer layer of *protoplasm*) and is not enclosed by a dead *cell-wall* (composed of *cellulose*) as in plant-cells. This absence of a non-living external layer of cellulose is one of the most characteristic differences between the cells of plants and animals, and for this reason plant-cells are more definite in shape and show a clearer outline when examined under the microscope. Another characteristic difference between plant- and animal-cells is that fully-grown plant-cells usually have cell-spaces (filled with cell-sap) occupying a large part of the cell, while animal-cells never have these large cell-spaces but are nearly always full of protoplasm.

EXAMINATION OF CELLS

Plant-cells. (i) Peel off, with forceps, a small piece of the transparent outer skin from the scale-leaf of a bulb, e.g. onion or lily, or from a green leaf, e.g. Allamanda. Mount it in a drop of water, under a cover-glass, and examine it with the microscope or micro-projector. Notice (*a*) the *cell-wall*, and (*b*) the *nucleus* in the middle of the protoplasm which fills the cell.

Animal-cells. (ii) Examine some animal-cells by scraping the inside of your cheek with a clean finger-nail and mounting the scraping in a drop of saliva under a cover-glass. Notice that these cells, unlike plant-cells, have *no cell-wall.* This is one of the main differences between plant- and animal-cells—*plant-cells have a non-living cellulose cell-wall, while animal-cells have no such non-living cell-wall but a living membrane of protoplasm which is more solid than the remainder.* Another difference is that plant-cells usually have cell-spaces in them, filled with cell-sap; animal-cells as a general rule do not have such cell-spaces inside the protoplasm.

Every living cell, however, plant or animal, consists of protoplasm and nucleus.

(iii) Examine under the microscope a little of the soft pulp from inside a tomato. Notice the rounded cells.

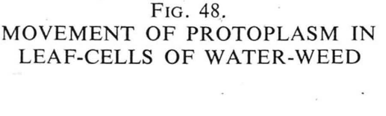

FIG. 48.
MOVEMENT OF PROTOPLASM IN LEAF-CELLS OF WATER-WEED

(iv) Mount a young leaf of a moss or a water-weed in a drop of water under a cover-glass and examine with the microscope. Under the high power, a continuous current of protoplasm can usually be seen, moving slowly round the walls of the cell and carrying the green bodies with it. (This movement is usually most rapid in the centre of the leaf, in cells either in or adjoining the mid-rib.)

(v) Examine a few threads of Spirogyra under the microscope and notice the cell structure. Mount some more Spirogyra in 10 per cent sugar solution and examine under the microscope. The solution diffuses through the cell-wall but cannot penetrate the lining layer of protoplasm, which is thus forced away from the cell-wall and is more easily seen.

THE FUNCTIONS OF LIVING PROTOPLASM

A one-celled animal like Amoeba, or a single cell of a simple green plant like Spirogyra, can carry out all the following everyday functions: (*a*) *nutrition* (or feeding), (*b*) *respiration*, (*c*) *excretion* (or getting rid of waste products), (*d*) *moving*, (*e*) *feeling* (or response to external stimuli), (*f*) *growing*, and (*g*) *reproduction*. *These are the fundamental functions of living protoplasm.* Since all living things are built up of one or more cells of protoplasm, these are also the fundamental functions of *all living things*. In the 'lowest', one-celled, plants and animals, one single cell can perform all these functions for itself. In the 'higher' plants and animals, consisting of many cells, there is *specialization*, and some cells are specially good at performing one special function. Thus some cells (muscle-cells) are specially good at movement, some (nerve-cells) are specially good at feeling, while others (wood- and bone-cells) are specially strengthened, and so on.

We shall now study these functions of living protoplasm in turn, examining the different ways in which different living things carry out the functions. This branch of science is called *physiology*.

CHAPTER V

NUTRITION IN PLANTS AND ANIMALS

All living things, plants and animals, require *food* for three purposes: (*a*) *to provide energy*, (*b*) *for growth and repair*, and (*c*) *to protect the organism from disease* and to regulate its healthy working. Since animals are more active than plants, the greater part of their food is used to supply energy, while plants use most of their food for growth (i.e. for making new cells and tissues). **The essential living material in both plants and animals is protoplasm, and this requires the same food substances whether it is part of a plant or part of an animal, namely, carbohydrates, fats, proteins, mineral salts, and water.**

There is a very important difference, however, in the way that plants and animals obtain these essential foodstuffs. Animals have to get their carbohydrates, fats and proteins ' ready-made ', but green plants can make these essential foodstuffs for themselves from simple raw materials—carbon dioxide from the air together with water and dissolved mineral salts from the soil.

NUTRITION IN GREEN PLANTS

We have seen that the green plant builds up carbohydrates from carbon dioxide and water, the necessary energy being obtained from sunlight. Some of the carbohydrates formed by photo-synthesis are then changed into oils. Plant-proteins are also formed from carbohydrates: sugars reacting with nitrates from the soil to form amino-acids which are finally built up into proteins. Some plant-proteins contain sulphur and phosphorus, and these elements are obtained from sulphates and phosphates absorbed from the soil. The plant makes these proteins into new protoplasm for new cells, the cell-walls being composed of cellulose which the plant makes from the carbohydrates formed during photo-synthesis. Under favourable conditions, plants make more food than they need for immediate use, and this excess is stored. Reserve carbohydrates are usually stored as insoluble starch in seeds, storage roots and underground stems. Reserve oils and fats are usually stored in fruits and seeds. Reserve proteins are usually stored in seeds, leaves and growing tissues.

In order to make use of these reserve carbohydrates, fats and proteins, the plant has to *digest* them, i.e. change them into simpler, soluble substances which will pass through the cell-walls. Although a plant, unlike most animals, has no special digestive system, similar chemical changes are brought about by similar *enzymes* in both plants and animals. The enzyme *diastase* changes starch into the di-saccharide *maltose*, and this is changed into the mono-saccharide *glucose* by the enzyme *maltase*. Protein-splitting enzymes split up the reserve proteins into amino-acids. Reserve oils and fats in seeds are changed into simple sugars by enzymes during germination. These soluble food-materials—simple sugars and amino-acids—are then carried to all parts of the plant and are used by the plant-cells to build up new protoplasm or to supply energy.

NUTRITION IN NON-GREEN PLANTS

Our earlier studies of plants were all based on *green* plants. Some plants, however, are not green, e.g. *fungi* and *bacteria*, hence they cannot carry out photo-synthesis and build up their own foodstuffs

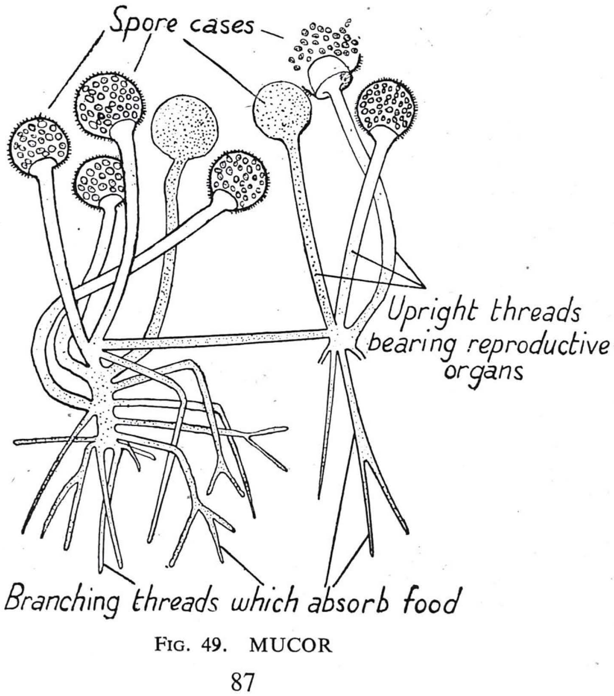

FIG. 49. MUCOR

from simple raw materials. They have to get their manufactured food ‘ ready-made ’, either from living green plants or from the dead remains of plants and animals.

FUNGI

This group of non-green plants includes *moulds, mildews,*† *mushrooms,*† *toadstools,*† *puff-balls*† and *yeasts.*

A common example is *mucor*† (Fig. 49), a mould which soon appears on damp bread if it is exposed to the air for a few days. The Mucor *spores* are present in the air. The Mucor plant consists of a network of fine white threads growing in the bread and on its surface. After a time, upright threads arise, each one capped with a small black head—the reproductive organ—which is full of microscopic spores. The branching threads of Mucor produce enzymes which split up (or digest) the solid food material into simpler, soluble substances which can be absorbed by the plant, the carbohydrates being changed into sugars and the proteins into amino-acids.

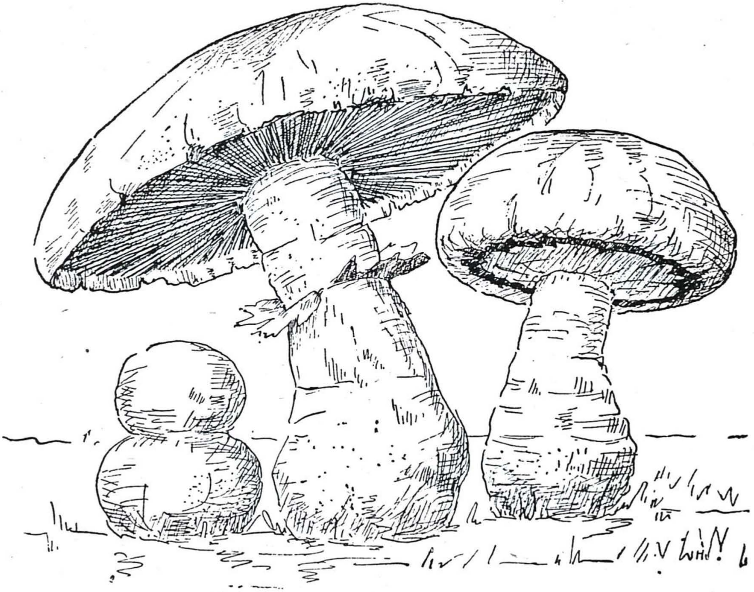

FIG. 50. MUSHROOMS
(Reproductive organs producing spores.)

The *mushroom* plant (Fig. 50) also consists of a branching network of fine threads feeding on decaying organic matter in the soil. The most familiar part of the mushroom plant, which appears above the ground, is only the reproductive organ, producing enormous numbers of spores. (One mushroom may produce over 1,000,000,000 spores.)

A *yeast* plant consists of a single oval cell. With the help of an enzyme (zymase) it oxidizes sugar to alcohol and carbon dioxide, setting free energy. This is a special type of respiration which goes on *in the absence of air*. Since oxidation is incomplete under such conditions, much less energy is set free by this type of respiration than by respiration in presence of air when the sugar is completely oxidized to carbon dioxide and water. (' Chinese yeast '† is a mixture of yeast cells and Mucor spores. When added to boiled rice, the Mucor changes the starch of the rice-grains into sugar, which is then converted into alcohol by the yeast.)

NUTRITION IN ANIMALS

Animals have to get their organic food—carbohydrates, fats and proteins—' ready-made ', and we have seen that the original source of all these materials is the green plant. We have also seen that the structure and method of life of any animal depends mainly on its method of feeding. This accounts for the much greater variety of animal types than of plant types. A green plant is usually surrounded by suitable raw food materials—carbon dioxide in the air and water in the soil; and it can take in these raw materials most efficiently by remaining fixed to one spot, spreading out its flattened leaves in the air where they will receive as much light as possible, and absorbing water by means of a much-branched root-system in the soil. But an animal has to move about in search of suitable food. The food must then be got inside the animal's body to be digested and absorbed. The animal's problems of ' food-capture ' and how to escape from its enemies results in highly specialized arrangements in different types of animals. For example, among the insects we find complicated mouth parts adapted for biting or sucking, according to the insect's diet; among the birds we find many different types of beaks and feet; among the mammals we find several different arrangements of cutting-, tearing- and grinding-teeth in flesh-eaters and plant-eaters. Hence, although nutrition is the same essential function in all living things, the details of the process are more varied and more complicated in animals than in plants.

For convenience, we shall consider animal nutrition in four stages: (i) *ingestion*†—the capture or intake of food, (ii) *digestion*—the process of breaking down complex food materials into simpler, soluble and diffusible substances which can enter the body-cells, (iii) *absorption*—the intake of soluble and diffusible food-substances by the true interior of the body, i.e. the body-cells, and (iv) *assimilation*—the building up of these food-substances into the chemical compounds which the body requires.

NUTRITION IN MAMMALS (e.g. Man)

Ingestion. Regarding Man as our example, food is taken in at the mouth with the aid of the teeth and lips. If the solid food is in large lumps, smaller pieces are bitten off by the cutting-teeth (incisors) at the front of the mouth. This food is then chewed, or ground up, by the grinding-teeth (molars†) at the back of the jaws. While the teeth are tearing and grinding the food, it is moistened* and softened by *saliva*, and worked into a round ball by the tongue in readiness for swallowing.

By thorough chewing, the solid food is broken up into small pieces which will offer a larger surface area to the action of the digestive juices in the food-canal. (N.B.—A liquid can only react with an insoluble solid *at its surface*.)

Digestion. The aim of digestion is to prepare the food for absorption. After ingestion, the food passes through a long tube—the food-canal—leading from the mouth to the anus, where the complex carbohydrates, fats, and proteins are broken down into simpler, soluble substances whose molecules are small enough to diffuse through the walls of the food-canal into the true interior of the body. Digestion is brought about by the action of various *enzymes* which are poured into the food-canal by a series of *digestive glands. These enzymes are organic catalysts produced by living cells.* In chemistry lessons we have seen that there are many chemical changes which take place very slowly when left to themselves, but which take place rapidly when a small quantity of a suitable *catalyst* is added. **Every living cell produces similar organic catalysts, or enzymes, which enable the cell to carry on its life processes at a suitable speed.** Each digestive gland is associated with a different part of the food-canal and produces a digestive juice containing its own special enzymes for acting on part of the food as it passes through the food-canal. The enzymes which are concerned with digestion may be classified as (i) *carbohydrate-splitting enzymes*, (ii) *fat-splitting enzymes* and (iii) *protein-splitting enzymes*. All these enzymes are destroyed by boiling.

In Man, digestion begins in the mouth, where the food is mixed with *saliva*, a slightly alkaline juice produced by the *salivary glands*. Man has three pairs of salivary glands: (*a*) a pair by the angles of the jaws, just beneath the ears, (*b*) another pair between the two halves of the lower jaw, and (*c*) a third pair beneath the tongue. Saliva softens the food and helps in swallowing, but it is also the first digestive juice, containing the enzyme *ptyalin* (or salivary diastase) which changes cooked starch into sugar. In practice, the food is not kept in the mouth long enough to change much starch into sugar, hence the digestion of starch only begins in the mouth; it is completed later on, in the stomach and the small intestine. Ptyalin acts on starch (cooked), glycogen and dextrin, converting them into the di-saccharide *maltose*. It has no action on di-saccharides (e.g. sucrose), uncooked starch, fats or proteins. Its activity is increased when sodium chloride is present.

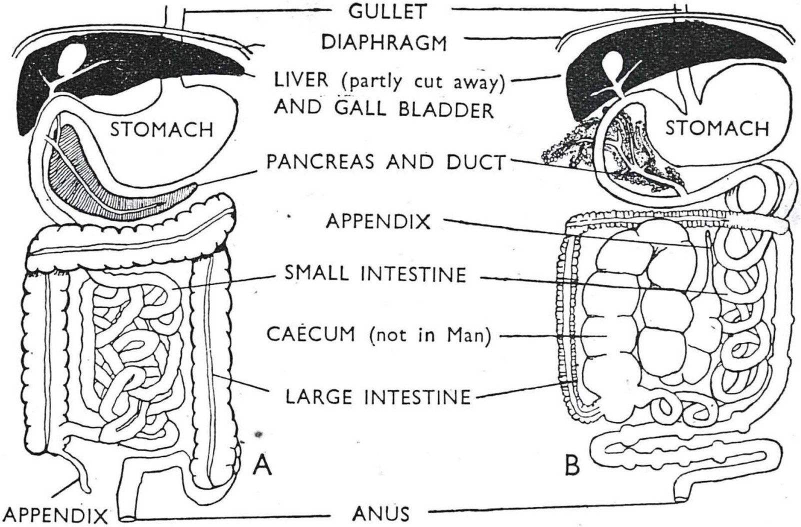

FIG. 51. FOOD-CANAL OF MAN (*a*), AND RABBIT (*b*) (diagrammatic

EXPERIMENTS ILLUSTRATING DIGESTION IN THE MOUTH—ACTION OF SALIVA

(i) Place small pieces of red and blue litmus paper (or neutral, purple litmus paper) on the tip of the tongue. Notice that saliva is very slightly alkaline.

(ii) Prepare a solution of saliva as follows:—Warm 50 c.cm. of water in a beaker to about 37° C. and then wash out the mouth thoroughly with about 20 c.cm. of the warm water. Then take another 20 c.cm.

of warm water in the mouth and keep it moving with the tongue for at least one minute. Collect this saliva solution in a clean beaker and filter it. Place test-tubes of saliva solution and 1 per cent starch solution (separately) in a water-bath at about 37° C. to warm up to a suitable temperature for your experiments. On a white tile (or a glass plate over a piece of white paper) place drops of iodine solution (N/20) at regular intervals.

(iii) Put about 3 c.cm. (1-in. in a $\frac{5}{8}$-in. test-tube) of the warm 1 per cent starch solution in a test-tube and add an equal volume of the warm saliva solution. Note the time. Shake up and place in the 37° C. water-bath.

After 30 seconds, remove a drop of the mixture with a glass rod and add it to one of the drops of iodine solution on the white tile. A blue colour shows the presence of starch, a purple or reddish colour shows the presence of dextrin, while a negative result shows that the starch is completely changed. Remove drops of the mixture every 30 seconds until there is no colour change with iodine. If this point is reached in less than 2 minutes, dilute some of your saliva solution with an equal volume of water and repeat the experiment.

(iv) When you have reached the colourless point with iodine, add an equal volume of diluted Fehling's solution (1/5) and heat in a boiling water-bath. A red precipitate shows the presence of a reducing sugar. This shows that the enzyme *ptyalin* in saliva only begins the digestion of starch, converting it to the di-saccharide maltose, which is further split up into glucose in the small intestine.

(v) Boil the remainder of the raw starch suspension from Experiment (iii) and cool under the tap. Take 2-3 c.cm. of this suspension, add an equal volume of a saliva solution and test at intervals with iodine. Notice that saliva digests cooked starch.

DIGESTION IN THE STOMACH

After being chewed, the ball of food is passed down the throat and gullet to the stomach, where it remains for several hours. Here it meets the *gastric juice*, produced by many tiny glands in the lining of the stomach-wall. This gastric juice is strongly acid, owing to the presence in it of about 0·2 per cent of hydrochloric acid, hence it gradually stops the action of the alkaline saliva as the food becomes mixed with the gastric juice. In practice, this takes some time, since the last food to be swallowed goes to the middle of the stomach and the alkaline saliva is not neutralized by the acid gastric juice until later. (See Fig. 52.) Hence the digestion of starch by saliva continues in the stomach. The hydrochloric acid in the gastric juice also kills bacteria and stops the food from decay in the stomach. Gastric juice contains two enzymes, (*a*) *pepsin*,† which begins the digestion of proteins, splitting them up in an acid solution into *proteoses* and *peptones* (it has no action on carbohydrates or fats), and (*b*) *rennin*,† which clots milk so that it is retained in the stomach while the pepsin acts on it.

In the stomach, the food and the gastric juice are gradually mixed up into a soft liquid mixture; and, at intervals, a ring of muscle between the stomach and the small intestine relaxes and the partly-digested food is passed, a little at a time, into the small intestine, where it meets more digestive juices.

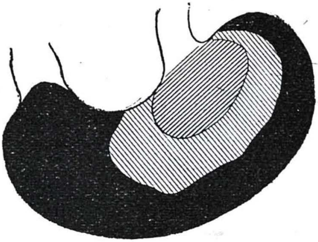

FIG. 52. SECTION OF FROZEN STOMACH OF RAT AFTER THREE SEPARATE MEALS OF DIFFERENTLY COLOURED FOOD

EXPERIMENT ILLUSTRATING DIGESTION IN THE STOMACH—ACTION OF PEPSIN

To show the digestion of Protein by Pepsin. Put 4 c.cm. of 1 per cent pepsin solution in a test-tube and add 1 c.cm. of 1 per cent hydrochloric acid. (1 c.cm. of 1 per cent hydrochloric acid has now been diluted to 5 c.cm., so that the mixture contains 0·2 per cent of hydrochloric acid, i.e. the same percentage as in human gastric juice.) Add a *small* piece of washed white fish-muscle[1] and place in the 37° C. water-bath. Shake occasionally (to imitate the movements of the stomach) and notice that the protein gradually dissolves, i.e. it is *digested* by the pepsin. Filter the final solution and divide into two. To one half add 2 c.cm. of 40 per cent sodium hydroxide solution containing 2 drops of 1 per cent copper sulphate solution. The characteristic purple Biuret reaction shows the presence of dissolved proteins. To the other half, add about 1 c.cm. of 1 per cent *bile-salt* (or ox-gall) solution. A white precipitate of partly-digested protein is formed. (Bile-salts precipitate *proteoses* in the small intestine, thus giving the pancreatic juice a better chance to complete the digestion of the proteins in the food.)

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[1] If a *small* piece of fish is used it will be completely dissolved and there can be no doubt about the action of the enzyme on the protein. If a large piece of fish is used it will not *all* dissolve and it will be difficult to be sure from direct inspection whether *any* has dissolved.

THE CLOTTING OF MILK BY GASTRIC JUICE

Gastric juice contains pepsin, which begins the digestion of proteins. In addition, it contains another enzyme, *rennin*, which causes milk to clot.

(*a*) *To show the action of Rennin on Milk.* Take about 5 c.cm. of fresh cows' milk in a test-tube and add about 1 c.cm. of rennin solution (e.g. 'essence of rennet'). Shake, and keep the tube in the 37° C. water-bath for 5 minutes. Notice that a *clot* is formed, and that the tube can be turned upside-down without any milk flowing out. The clot consists of the *casein* of the milk together with the *milk-fat*, which is carried down by the casein. Put back the tube in the 37° C. water-bath for an hour, and notice how the clot gradually contracts, squeezing out a clear liquid, the '*whey*'.† (This is the first stage in the process by which milk is made into cheese.)

(*b*) *To show that Rennin is destroyed by boiling.* Repeat the last experiment, using 1 c.cm. of *boiled* rennin solution. No clot is formed, showing that rennin (like any other enzyme) *is destroyed by boiling.*

EXPERIMENTS WITH BILE

To show how Bile lowers the surface tension of Water. The following experiments may be done with diluted ox-bile (1 volume of bile to 9 volumes of water), or with a 1 per cent solution of bile-salts.

(*a*) Take two small glass beakers. In one put 10 c.cm. of distilled water and in the other put 10 c.cm. of bile solution. Shake a little 'flowers of sulphur' on the liquid surface in each beaker and then rock the vessels gently. The fine particles of sulphur continue to float on the surface of the pure water owing to the high surface tension, but the sulphur sinks in the bile solution because the bile-salts lower the surface tension of the water.

(*b*) Put two similar folded filter-papers in two filter-funnels. Wet one with pure water and the other with bile solution. Then pour about 1 c.cm. of vegetable oil, e.g. coconut oil, on to each filter and allow to stand for some hours. Notice that the oil will not pass through the filter-paper moistened with pure water, owing to the high surface tension of the oil-water surface. The oil gradually passes through the filter-paper wetted with bile solution, however, showing that the bile-salts lower the surface tension. This fact is of great importance in breaking up oils into very tiny drops so as to form an *emulsion*, and this is the main action of bile in digestion.

EXPERIMENTS WITH EMULSIONS

Take four test-tubes, A, B, C, and D. Put 10 drops of vegetable oil, e.g. coconut oil, in each tube. To A add 10 c.cm. of distilled water. To B add 5 c.cm. of distilled water and 5 c.cm. of sodium carbonate solution (1 per cent) giving about the same amount of alkali as in human pancreatic juice. To C add 10 c.cm. of 1 per cent bile-salt solution. To D add 5 c.cm. of 1 per cent sodium carbonate solution and 5 c.cm. of 1 per cent bile-salt solution. Shake all the tubes vigorously and allow them to stand in the test-tube rack. In each case, shaking breaks up the oil into very fine drops which are suspended in the liquid to form an *emulsion.* On standing, however, the oil-drops in A soon run together and rise to the surface to form a separate layer, i.e. the emulsion is only *temporary.* In B the emulsion is *permanent,* because the alkali forms a trace of soap which produces a protective film round the oil-drops.

The pancreatic juice is alkaline, and thus helps to break up oils and fats into small droplets. In C, a permanent emulsion is formed because the bile-salts lower the surface-tension between the oil and the water. In D, a very stable emulsion is formed by the combined action of the sodium carbonate and the bile-salts (both of which are present in the small intestine).

DIGESTION IN THE SMALL INTESTINE

Action of Bile and Pancreatic Juice. When the acid contents of the stomach enter the small intestine, they are neutralized by the alkaline *bile* (from the liver) and the alkaline *pancreatic juice* (from the pancreas). In this first part of the small intestine, starches, proteins and fats are all split up by the action of three enzymes: *amylopsin*† acts on the starch, *trypsin*† acts on the proteins, and *lipase* acts on the fats.

Bile is a greenish, alkaline fluid made by the liver and stored in the *gall-bladder.* In Man, a small *bile-duct* introduces the bile into the first part of the small intestine. The bile contains no digestive enzymes, but, being alkaline, it neutralizes the acid of the gastric juice and thus stops the action of the pepsin. It contains *bile-salts* which break up fats and oils into tiny droplets so that they are more readily attacked by *lipase,* the fat-splitting enzyme of pancreatic juice. The bile-salts also precipitate proteoses, thus delaying their passage through the small intestine and giving more time for the pancreatic juice to act on them.

PANCREATIC JUICE

The pancreatic juice, like bile, is alkaline, owing to the presence of *sodium carbonate.* It contains three enzymes—*amylopsin* (starch-splitting), *trypsin* (protein-splitting), and *lipase* (fat-splitting). The starch-splitting enzyme—amylopsin—acts on any starch which has not been split up by the ptyalin in the saliva. Amylopsin, however, can act on uncooked starch, while ptyalin only acts on starch after the starch-grains have been burst by cooking. The protein-splitting enzyme—trypsin—carries the digestion of proteins a stage further than pepsin, some amino-acids being formed. The fat-splitting enzyme—lipase—splits up oils and fats into fatty acids and glycerine. After the combined action of bile and pancreatic juice, the carbohydrates of the original food are converted into maltose, the fats to fatty acids and glycerine, and the proteins into proteoses, peptones and some amino-acids. Digestion is then completed by the enzymes of the *intestinal juice*, produced by cells in the walls of the small intestine.

EXPERIMENTS ILLUSTRATING PANCREATIC DIGESTION

To show the action of Pancreatic juice on Starch. Amylopsin, the starch-splitting enzyme of pancreatic juice, does not act properly except in the presence of certain salts, hence in making up the ‘ pancreas solution ’ for these experiments on the digestion of starch, add 0·8 gm. of sodium chloride and 1·2 gm. of sodium phosphate ($Na_2HPO_4.12H_2O$) to every 100 c.cm. of ‘ pancreas solution ’. In a test-tube put 3 c.cm. of the ‘ pancreas solution ’, 3 c.cm. of 1 per cent starch solution and 2 c.cm. of 2 per cent sodium carbonate solution. Shake and place the tube in the 37° C. water-bath. Remove drops of the mixture at one-minute intervals and test for starch with drops of iodine solution on a white tile as in your experiments on salivary digestion. (N.B.—Use N/20 iodine solution acidified with dilute hydrochloric acid, since starch gives no blue colour with iodine in alkaline solutions.) After 20 minutes, add a little Fehling’s solution and heat in a boiling water-bath. A yellow or orange precipitate shows that the starch has been converted into a reducing sugar.

To show the action of Pancreatic Juice on Protein. Mix together 3 c.cm. of 1 per cent ‘ pancreas solution ’ (without sodium chloride or phosphate) and 3 c.cm. of 1 per cent sodium carbonate solution in a test-tube, and place in the 37° C. water-bath. Add a *small* piece of washed white fish-muscle and shake from time to time (to imitate the movements of the small intestine). Notice that the protein gradually dissolves, i.e. it is digested by the trypsin of the pancreatic juice.

Filter the solution and divide it into two. To one half, add 2 c.cm. of 40 per cent sodium hydroxide solution containing 2 drops of 1 per cent copper sulphate solution. A *pink* colour shows that trypsin carries the digestion of proteins a stage further than pepsin. (Pepsin changes proteins into *proteoses* and *peptones*, which give a *purple* colour with the Biuret test, while trypsin converts proteins to *poly-peptides* and *amino-acids*, and the poly-peptides give a *pink* colour with the Biuret test.) To the other half of the filtrate add a few drops of 1 per cent bile-salt solution. There is no precipitate, showing the absence of proteoses. (In the corresponding experiment on the filtrate from pepsin digestion there *was* a precipitate, showing that pepsin only digests proteins as far as proteoses.)

To show the action of Lipase on Fats. Take about 15 c.cm. of fresh cows' milk (which contains fat in a very finely divided condition) in a test-tube and add a few drops of 1 per cent phenol-phthalein† solution. Then add 1 per cent sodium carbonate solution, drop by drop, shaking after adding each drop, until a permanent pink colour is obtained, showing that the mixture is slightly alkaline. Pour 5 c.cm. of the mixture into each of three test-tubes—A, B, and C—and place the tubes in the 37° C. water-bath. Take about 5 c.cm. of 1 per cent ' pancreas solution ' and make it slightly alkaline by adding 1 per cent sodium carbonate solution, drop by drop, until just pink. To the alkaline milk in A, add 2 c.cm. of water. To B, add 2 c.cm. of the alkaline ' pancreas solution ' and shake. To C, add 2 c.cm. of alkaline ' pancreas solution ' which has been boiled. Tube B soon loses its pink colour, showing that the lipase has begun to split up the milk-fat into glycerine and fatty acid. This free fatty acid neutralizes the sodium carbonate in the mixture, which thus becomes acid. Add more 1 per cent sodium carbonate solution to B, drop by drop, until it again turns pink, and then replace in the water-bath. The pink colour soon disappears again as more fatty acid is set free by the action of the lipase on the milk-fat. Tube C remains pink, showing that the fat-splitting enzyme is destroyed by boiling.

Milk was used in the above experiment because it contains fat in very tiny particles (i.e. in the form of an *emulsion*) which are readily attacked by the enzyme. Other fats and oils may be used, but digestion is much slower unless the fat is finely divided. Instead of milk, an emulsion of egg-yolk can be used—1 volume of egg-yolk shaken up with 9 volumes of water. To carry out a similar experiment with a vegetable oil, e.g. coconut oil, take 5 c.cm. of absolute alcohol in a dry test-tube, add 10 drops of the oil, and warm gently until a clear solution is obtained. Pour this alcoholic solution of oil into 10 c.cm. of water. The oil is thrown out of solution as very tiny particles which form a milky emulsion.

DIGESTION IN THE SMALL INTESTINE. ACTION OF THE INTESTINAL JUICE

The cells of the inner lining of the small intestine produce *intestinal juice.* This digestive juice contains the enzymes *erepsin,*† *sucrase, lactase* and *maltase.* The erepsin acts on proteoses and peptones, finally changing them into amino-acids (although erepsin will not act on the original, unchanged proteins of the food). Sucrase (or invertase) acts on the di-saccharide sucrose, splitting it up into glucose and fructose. Lactase acts on the di-saccharide lactose (milk-sugar), splitting it up into glucose and galactose. Maltase acts on the di-saccharide maltose, splitting it up into glucose. Experiments to illustrate intestinal digestion are too difficult for this Course. Sucrase, however, is easily obtained from plant sources, and this may be used in the laboratory to split up sucrose into glucose and fructose.

THE FINAL PRODUCTS OF DIGESTION

As a result of digestion in the mouth, the stomach, and the small intestine: (i) *the proteins are finally split up into amino-acids,* (ii) *the carbohydrates are finally split up into simple sugars,* (iii) *the fats are split up into fatty acids and glycerine, the fatty acids combining with the sodium carbonate in the intestine to form soluble soaps.* All these substances are now ready for *absorption.*

ABSORPTION

In Man, the walls of the small intestine are covered with large numbers of tiny, finger-shaped projections called *villi,*† each villus being about 0·05 in. long, and containing a network of capillary blood-vessels running just beneath the covering layer of cells (Fig. 53). In the centre of each villus is a blind-ended tube called a *lacteal,* which is the beginning of the *lymphatic system* of the body. These villi present a very large surface area to the digested food in the small intestine, and it is through the cells covering the villi that the digested food enters the true interior of the body. The *simple sugars* (e.g. glucose) and the *amino-acids* pass directly into the capillary blood-vessels running through the villi. All the veins from the villi join together to form one large vein (the *portal vein*) which runs to the *liver* and then divides again into capillaries. The *soaps* and *glycerine* which result from the digestion of fats do not pass into the blood-vessels of the villi, but enter the lymph-vessels (or lacteals) where the glycerine re-combines with the fatty acids of the soaps to form tiny droplets of fat once more. If these droplets of fat from the food-canal entered the blood-vessels and were carried to the liver, they might run together into larger drops which would block up the capillaries of the liver. Instead of this, the droplets of fat enter the lymph-vessels in the villi, and the latter join together to form larger lymph-vessels which finally discharge the lymph into the large vein from the left fore-limb, at the base of the neck. The droplets of fat are then carried round the body in the blood-stream.

By the time that the food has travelled the whole length of the small intestine, practically all the soluble, digested food has been absorbed, and the indigestible residue, still mixed with much water, is passed on to the large intestine. In Man, the large intestine plays no important part in digestion, but a lot of water is re-absorbed from the indigestible food-residues, which are reduced to about one-tenth of their former volume, and are then got rid of as semi-solid *faeces* from the anus. The large intestine also gets rid of some waste products from the blood in the form of salts which are not sufficiently soluble to be got rid of through the kidneys.

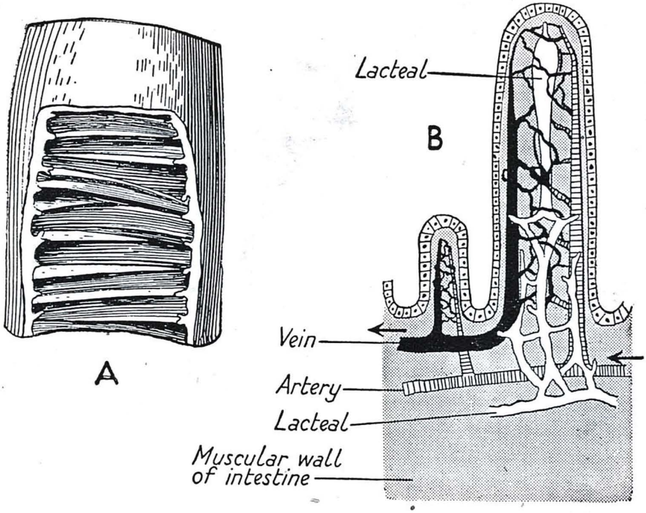

FIG. 53.
A—portion of human small intestine cut open to show transverse folds.
B—a villus, very highly magnified, to show the blood-capillaries and lacteals.

In plant-eating mammals, e.g. rabbit and cavy, the small intestine enters the large intestine in the middle of one side, so that one end of the large intestine forms a very large, blind-ended sac—the *caecum.*

A very long intestine and a large caecum are characteristic of plant-eating animals, since the food has to be retained for a long time to enable *bacteria* to convert the cellulose into simple sugars. (In Man, the food-canal is about 20 ft. long. The food-canal of a giraffe is about 270 ft. long !)

DIFFUSION, OSMOSIS AND DIALYSIS

In order to understand how the final products of digestion pass through the walls of the villi, it is necessary to know more about *diffusion* and similar processes. If a jar of oxygen is placed mouth-to-mouth with a jar of nitrogen, the two gases soon form a uniform mixture by *diffusion*, the gas molecules moving about in all directions until the two gases are thoroughly mixed. If a test-tube is half filled with copper sulphate solution and is then slowly filled up with water, without shaking, the blue colour gradually spreads throughout the tube until the colour is uniform. The copper sulphate molecules have *diffused* throughout the water in the tube. Diffusion in liquids, however, is much slower than diffusion in gases. If a solution of egg-white is used instead of copper sulphate in the last experiment, diffusion is still slower, because the protein molecules of the egg-white are very much bigger than the copper sulphate molecules.

If a solution of sugar in water is enclosed in a membrane (e.g. parchment*-paper) surrounded by water, some sugar molecules diffuse through the membrane. This process is called *dialysis*.† At the same time, some water molecules diffuse through the membrane into the sugar solution. This process is called *osmosis*. When sugar solution is put inside the bulb of a thistle funnel with its wide end closed by a membrane and standing in water, water diffuses inwards across the membrane (*osmosis*) and sugar diffuses outwards (*dialysis*). The explanation of this is that the moving molecules of water and sugar are constantly hitting against the membrane, which behaves like a sieve with very tiny holes. The water molecules, being much smaller and moving more rapidly than those of sugar, find their way through the sieve much more easily, so that more water passes through the membrane into the sugar solution than sugar molecules pass outwards through the membrane into the water. Hence the level of the liquid inside the thistle funnel rises.

Thomas Graham (1861-4), while experimenting with this diffusion of dissolved substances through membranes, found that some substances, e.g. salt, sugar, urea, and glycerine, *dialysed* (or passed through membranes); while others, e.g. egg-white, starch, gelatine, and glue, did not dialyse. As the first class of substances crystallize readily, Graham called them *crystalloids*.† The second class of substances, whose molecules are too big to pass through membranes, he called *colloids*.† Animal and plant material consists of both crystalloids and colloids, hence it is important to know something about the two classes of substances. For example, the proteins and starches are *colloidal* and so cannot pass through the walls of the small intestine until their large molecules have been broken down into simpler, smaller, diffusible molecules of *crystalloids* (the amino-acids and the simple sugars).

EXPERIMENTS ON OSMOSIS AND DIALYSIS

One-third fill a ' sausage-skin ' paper sac with 1 per cent starch solution and then fill up to two-thirds with 5 per cent glucose solution. Stand the sac in a narrow jar or tall beaker of water, being very careful not to spill any of its contents into the beaker. After 20 minutes, pour a few c.cm. of water from the beaker into a test-tube and add a drop of iodine solution. No blue colour develops, showing that no starch (a *colloid*) has *dialysed* out through the membrane. Pour out another 2-3 c.cm. of water from the beaker and test for glucose with Benedict's solution. A red precipitate of cuprous oxide shows that some glucose (a *crystalloid*) has dialysed out through the membrane. Add a few drops of iodine solution to the remainder of the water in the beaker and set aside for half an hour. A blue colour develops *inside* the membrane, showing that iodine has *dialysed* through the membrane. No blue colour develops in the surrounding water, showing that no colloidal starch has dialysed outwards through the membrane.

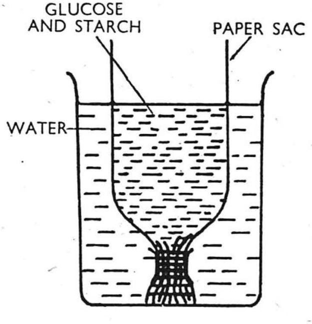

FIG. 54. DIALYSIS

Digestion, therefore, prepares the food for absorption by (i) *converting insoluble materials into soluble substances* (e.g. starch to sugar), (ii) *converting the large molecules of colloids into smaller molecules of diffusible crystalloids* (e.g. proteins to amino-acids), (iii) *converting soluble substances into substances acceptable to the tissues* (e.g. maltose to glucose).

ASSIMILATION

All the veins from the food-canal join together to form one large vein (the *portal vein*) which runs to the liver and then divides up into capillaries again. The liver is the largest *gland* in the body and it is situated in the belly-cavity adjoining the diaphragm and partly overlapping the stomach. We have seen that one function of the liver is to make *bile*, but it has some other equally important functions: (i) *it regulates the amount of sugar in the blood and stores up any excess*, and (ii) *it neutralizes harmful waste substances in the blood and prepares them for excretion by the kidneys.*

All the simple sugars formed by the digestion of carbohydrates are carried by the portal vein from the small intestine to the liver. If all this sugar is not needed immediately, any excess is converted into *glycogen* and is stored in the liver-cells until required. When the muscles require sugar for producing energy, they take it from the blood-stream, and the liver at once converts some of its reserve glycogen back again into sugar and passes it into the blood-stream. In this way, the liver controls the amount of sugar in the blood.

All the amino-acids absorbed from the small intestine are also carried to the liver by the portal vein. When they reach the liver, one of two things may happen: If the amount of amino-acids is small, they may be passed on directly to the rest of the body in the blood-stream; but if there is more amino-acid in the blood from the small intestine than the body needs, this excess is *never stored up* but is converted into glycogen and ammonia. The ammonia formed in this way would be a poison if it were allowed to circulate in the blood-stream, and it is converted by the liver into *urea*, a harmless, neutral substance which is returned to the blood-stream for removal later by the kidneys.

The glucose, which is carried to the tissues by the blood and the lymph, is oxidized to set free *energy* (the greater part of this energy appearing as *heat*). The amino-acids are removed by the body-cells and are built up into proteins and protoplasm. The excess fat is removed from the blood-stream and stored under the skin and around the organs in the belly-cavity. When the animal is short of food for a long time, it makes use of this reserve fat as a source of energy.

Having dealt with the process of *nutrition* in some detail in the case of Man (the most complicated and specialized animal) we can now discuss nutrition in other types of animals much more shortly.

NUTRITION IN BIRDS (e.g. PIGEON)

Since a bird has no teeth it cannot chew its food, but swallows it at once. Hence a bird's salivary glands are of little importance for digesting carbohydrates, although the saliva softens the food and makes it easy to swallow. The food is passed down the gullet, the posterior part of which is enlarged to form a large, thin-walled *crop* in which the food is stored. The food then passes, a little at a time, into the *stomach* where it receives a digestive juice and then passes into the *gizzard*, which has very strong, muscular walls and contains small stones which help to grind up the food. The small intestine, which is long and much looped, is connected by ducts to the liver and pancreas, receiving bile and pancreatic juice from these glands. The large intestine is very short and ends in a *cloaca*, which also receives the urinary and reproductive ducts. The details of digestion, absorption and assimilation are much the same in both birds and mammals. Birds have very efficient digestive systems, and the amount of waste matter passed out as faeces is only small.

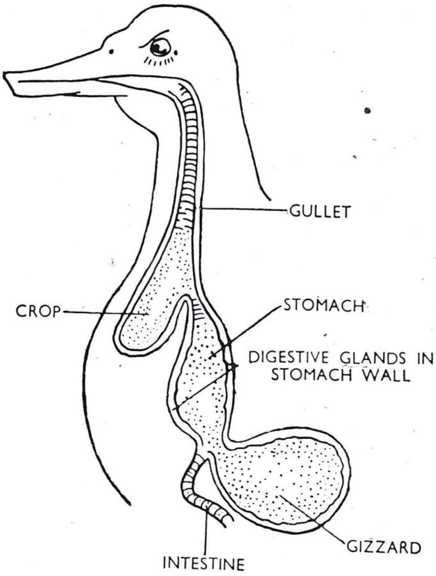

FIG. 55. FOOD-CANAL OF BIRD (diagrammatic)

NUTRITION IN AMPHIBIANS (e.g. FROG)

The frog has a long tongue attached to the front of the mouth-cavity. The animal shoots this tongue out of its mouth suddenly, catching small animals (e.g. insects, worms and snails) with the sticky end and then draws the tongue back into the mouth. The food is usually swallowed whole, and there are no salivary glands. After being forced down the gullet into the stomach, the food is acted on by the gastric juice, forming a soft pulp which passes through a ring of muscle into the small intestine. The bile-duct, leading from the gall-bladder, opens into the beginning of the small intestine. The pancreatic ducts also join the bile-duct, so that both bile and pancreatic juice enter the small intestine through the same duct. The small intestine is a long, narrow tube, much looped. It provides a large area for absorption. It widens to form the large intestine, which is comparatively short, and the faeces are passed out at the cloaca, which also receives the urinary and reproductive ducts. The chemical processes involved in digestion, absorption and assimilation are very similar to those in mammals and birds.

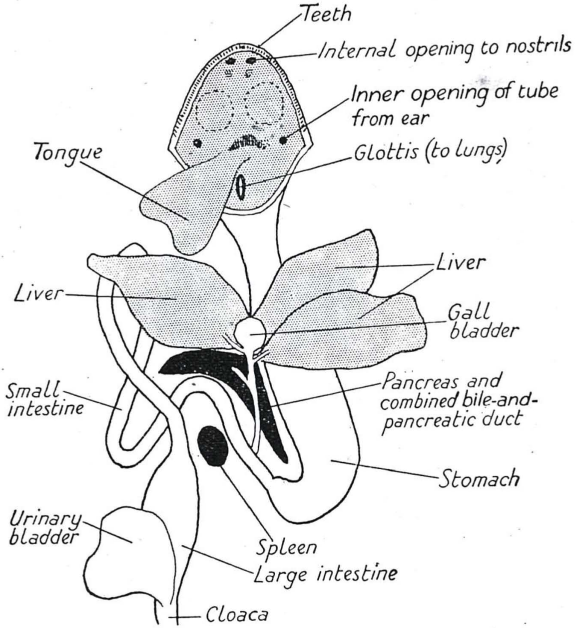

FIG. 56. THE FOOD-CANAL OF A FROG (diagrammatic)

NUTRITION IN FISHES

Digestion in fishes is very similar to digestion in the frog. Fishes are provided with teeth and movable jaws. The food is taken into the mouth and passed through a ring of muscle into the gullet, comb-like *gill-rakers* preventing food from passing through the gill-slits on the way. The general arrangement of the digestive organs and the processes of digestion and absorption show no fundamental differences from those of the frog.

NUTRITION IN MOLLUSCS (e.g. SNAIL)

The snail has a horny jaw-plate in the roof of its mouth and a file-like ' tongue ' in the floor of its mouth. The ' tongue ' is covered with thousands of tiny, backward-pointing ‘ teeth ’, and this flexible ‘ file ’ is drawn backwards and forwards over the plants on which the animal feeds. The finely-divided food is mixed with saliva in the mouth and is passed down the gullet into the large, thin-walled crop, and from there into the ‘ stomach ’ and intestine (which are embedded in the ‘ liver ’). The undigested food-remains are then passed through the large intestine to the anus.

NUTRITION IN INSECTS (e.g. COCKROACH)

Insects show great variety in their feeding habits, and their digestive organs are modified to suit the insect's food. We shall consider only one example—the cockroach. The cockroach crushes its food with its three pairs of mouth parts, mixes it with saliva, and forces it into its mouth and down a short gullet into a large, thin-walled *crop*, where the food is mixed with digestive juices so that the digestion of carbohydrates, fats and proteins begins in the crop. Behind the crop is the *gizzard*, a round, muscular organ in which the food, mixed with digestive juices, is ground up by means of a ring of six horny ‘ teeth ’. The gizzard leads to the ‘ stomach ’ (or *mid gut*) where the food is mixed with a digestive juice supplied by seven or eight small finger-like tubes which are extensions of the food-canal. Digestion and absorption are completed in the mid gut and its side tubes (the only part of the food-canal which is not lined with horny material like that on the outer surface of the animal) and the undigested food-remains pass through the *hind gut* to the anus. The digestive enzymes of the cockroach enable it to digest almost any animal matter or vegetable matter, e.g. paper, cotton, wool, silk, leather, hair and horn. The cockroach usually stores large reserves of fat in the body-cavity. The absorbed food reaches the tissues by way of the blood which fills the body-cavity.

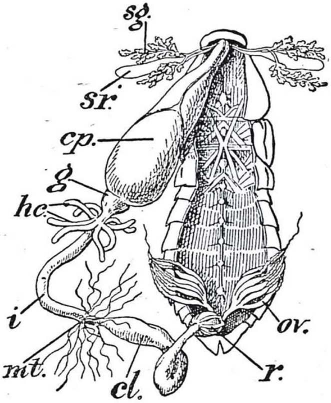

FIG. 57. DIGESTIVE AND REPRODUCTIVE SYSTEMS OF FEMALE COCKROACH

sg—salivary gland. sr—salivary receptacle. cp—crop. g—gizzard. hc—digestive glands. i—mid gut. mt—excretory organs. cl—hind gut. r—anus. ov—ovary.

NUTRITION IN WORMS (e.g. EARTH-WORM)

The food-canal of the earth-worm consists of a straight tube, running through the middle of the body-cavity from the mouth to the anus: the simplest possible food-*canal*. We have seen that earth-worms eat their way through the soil, sucking in and swallowing soil containing particles of animal and vegetable matter. The soil is ground up by the *gizzard* and is then passed through the intestine, where digestion and absorption take place.

NUTRITION IN STINGING ANIMALS (e.g. HYDRA)

Hydra has no food-*canal* but only a food-*cavity* surrounded by a double layer of cells. (There is no *body*-cavity between the food-cavity and the body-wall as there is in all the higher animals.) Living food-organisms are stunned by the stinging-cells and are swept into the ' mouth ' by the hollow threads (or *tentacles*†) which surround the upper, open end of the animal. Some of the cells of the inner layer of the food-cavity are *gland-cells* which pour out digestive juices into the food-cavity, where digestion of the larger food-particles begins. After these larger particles have been broken up by the enzymes, the smaller particles are *ingested* by Amoeba-like cells, and digestion is completed by these cells. Undigested food-remains are got rid of through the only opening to the food-cavity—the ' mouth '. Hydra is an excellent example of the beginnings of ' *division of labour* ' between different kinds of specialized cells. For example, there are several different kinds of cells taking part in the one function of nutrition: gland-cells produce enzymes, Amoeba-like cells ingest food-particles and complete the process of digestion, stinging-cells stun the prey, muscle-cells move the tentacles, and nerve-cells control the other kinds of cells.

NUTRITION IN ONE-CELLED ANIMALS (e.g. PARAMECIUM AND AMOEBA)

Paramecium shows the beginnings of a definite path for the food through the body during digestion. There is a groove on one side, leading to a funnel-shaped opening into the softer protoplasm underlying the firm outer layer. Food-particles, e.g. bacteria, are swept into this groove or gullet, and enter the protoplasm enclosed in a drop of water. Owing to the streaming of the protoplasm, these food-particles circulate round the body, following a definite path, while

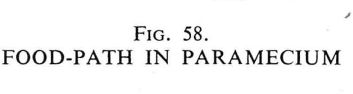

FIG. 58.
FOOD-PATH IN PARAMECIUM

digestion and absorption take place. The undigested residue is finally got rid of, always at the same point in the surface. Although no permanent opening is visible at this spot, this can be regarded as a ‘ temporary anus ’.

Nutrition in Amoeba is even simpler than in Paramecium. Food can be ingested *at any point* in the animal’s surface, there is *no definite path* along which the food moves during digestion and absorption, and the undigested food-residues are got rid of *at any point* in the surface —there is no one spot serving as a temporary anus. In spite of the extreme simplicity of its method of nutrition, however, Amoeba carries out the same fundamental processes as the higher animals. After *ingestion*, the food-particle is acted on by an acid digestive juice followed by an alkaline digestive juice so that *digestion* of the food takes place. *Absorption* and *assimilation* by the protoplasm complete the process of nutrition.

FIG. 59. AN AMOEBA TAKING IN A ONE-CELLED PLANT
(highly magnified)

SUMMARY OF NUTRITION IN ANIMALS

It should now be quite clear that *nutrition* consists of the same fundamental processes of *ingestion*, *digestion*, *absorption* and *assimilation* in all animals. They all feed on ‘ ready-made ’ complex organic food in many different forms, but the carbohydrates, fats and proteins of the food are split up into simpler, soluble and diffusible substances which can be absorbed by the body and used for growth and repair or for supplying energy.

CHAPTER VI

RESPIRATION IN PLANTS AND ANIMALS

We have seen that one of the functions of food is *to provide energy*. A heat engine (e.g. a steam-engine) gets its energy by burning (oxidizing) fuel. A living thing gets its energy in a very similar way—by oxidizing part of its food and converting some of its chemical energy into mechanical energy and heat. This process—*respiration*—is a fundamental activity of all living protoplasm. We have defined *respiration* as *the process of setting free energy by breaking down complex chemical compounds into simpler compounds possessing less chemical energy*.

RESPIRATION IN ANIMALS

In studying nutrition, we started with the "highest' plants and animals and worked downwards to the 'lowest'. We shall begin our study of respiration with the lowest one-celled animals and work upwards to Man.

RESPIRATION IN AMOEBA AND PARAMECIUM

In a tiny, one-celled animal like Amoeba or Paramecium, there is no special respiratory system. The animal is surrounded by water containing dissolved oxygen which is absorbed *over the whole surface*, and every part of the protoplasm can get the oxygen it requires *by diffusion*.

RESPIRATION IN HYDRA

Although Hydra shows the beginnings of 'division of labour' between different kinds of specialized cells, there are *no special respiratory organs*. As Hydra is never more than two cells thick, oxygen can be absorbed from the surrounding water over the whole surface and can reach every part of the body *by diffusion*, as in the simplest one-celled organisms.

RESPIRATION IN THE EARTH-WORM

The earth-worm has *no special breathing organs*, but absorbs oxygen *over its whole surface* through its *damp skin*. But an earth-worm is too ‘ thick ’ for oxygen to reach every cell quickly enough by diffusion, hence a *transport system* is necessary. The earth-worm has a simple blood-system for this purpose. Underneath the skin is a network of capillary vessels, containing blood. This blood is coloured red by its dissolved haemoglobin (a protein substance containing iron) which is the ‘ oxygen-carrier ’ of the blood. The oxygen absorbed through the skin combines with this haemoglobin, forming oxy-haemoglobin.† This loosely-combined oxygen must now be distributed to every part of the worm, and the carbon dioxide produced during respiration must be got rid of through the skin, hence the blood must *circulate*. In the earth-worm there is a closed system of blood-vessels, some of which contract and expand regularly and thus cause the blood to circulate round the body. Part of the blood filters through the walls of the capillaries and bathes every cell as *lymph*.

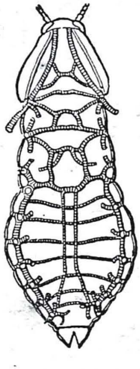

FIG. 60. MAIN AIR-TUBES OF FEMALE COCKROACH

RESPIRATION IN INSECTS

Insects transport dry air from the atmosphere to their tissues quite differently from other animals—*by a system of branching air-tubes* which open at the surface of the body. The finest branches of these air-tubes reach near to every cell in the insect's body, and the blind end of each air-tube contains a little water in which oxygen can dissolve. The air-tubes carry oxygen to the tissues and remove the carbon dioxide formed during respiration. In insects, since air is carried to all parts of the body by the air-tubes, the blood system can be very simple since the blood is only concerned with carrying dissolved food and not with carrying oxygen, as it does in most other animals. We have seen that the size of insects is limited owing to the difficulties of changing the external skeleton as the animal grows. The method of breathing by branching air-tubes also limits the size, as this system would be inefficient in larger animals.

RESPIRATION IN THE SNAIL

The snail changes the air in its mantle-cavity from time to time. The mantle-wall is richly supplied with a network of capillary blood-vessels, so that exchange of oxygen and carbon dioxide takes place between the blood and the air (as in the *lung* of the higher animals). The snail's blood, like that of some of the other lower animals, does not contain haemoglobin (an iron-protein substance) but *haemocyanin* (a copper-protein substance). Haemocyanin is colourless, but after combining with oxygen the oxy-haemocyanin is slightly blue. The snail's heart is very simple, consisting of *two* chambers only, a thin-walled auricle and a thick-walled, muscular ventricle. The oxygenated blood from the capillaries of the mantle is collected into a large vein leading to the auricle. The ventricle then pumps the blood along the main artery to all parts of the body.

RESPIRATION IN FISHES

Most fishes use the oxygen which is dissolved in the surrounding water. Water, containing dissolved oxygen, is taken into the mouth, passed over the *gills* and out through the gill-slits. The gills consist of numerous folds of very thin membrane, richly supplied with capillary blood-vessels. Exchange of oxygen and carbon dioxide takes place between the water and the blood in the gill-capillaries, the oxygen combining with the haemoglobin in the red blood-cells. This oxygenated blood from the gills is distributed by arteries to every part of the body and it becomes de-oxygenated in the body-capillaries. This de-oxygenated blood is then collected up by the veins and returned to the heart, from which it is pumped to the gills to complete the circulation.

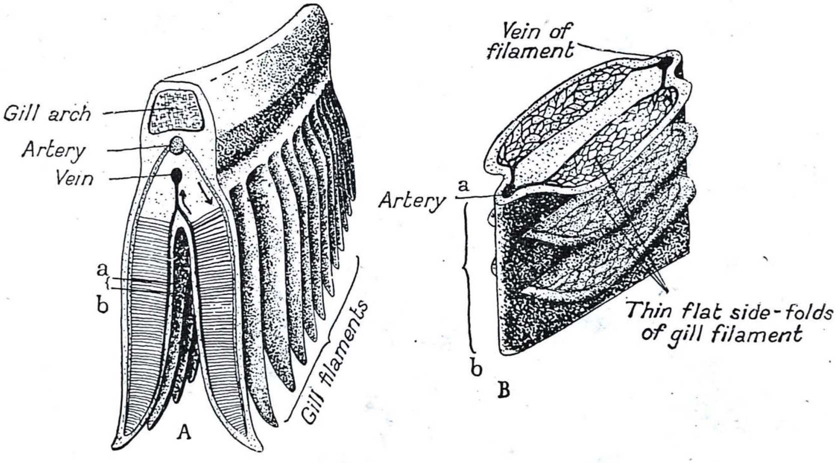

FIG. 61. GILL STRUCTURE OF BONY FISH
B is a more highly-magnified view of the portion *ab* in A.

The heart of a fish, therefore, is very simple, consisting of *two* chambers only, an auricle and a ventricle, with valves between the two to stop ' back-flow '. It contains de-oxygenated blood only.

RESPIRATION IN THE FROG

The frog is of special interest since it has several different methods of respiration. An adult frog has no less than three ways of absorbing oxygen: (*a*) *through its lungs*, (*b*) *through its damp skin*, and (*c*) *through the lining of its mouth.* Throughout its whole life-history, the frog can absorb oxygen through its skin, which has a network of capillary blood-vessels. In its earliest *tadpole* stage, however, immediately after hatching, it has three pairs of external gills on the sides of the body, just posterior to the head. A little later, gill-slits appear and the external gills wither away, being replaced by a new set of gills like those of a fish. Its blood-system, also, is like that of a fish, consisting of a simple *two*-chambered heart circulating blood through the gill-capillaries and round the body. Up to this stage in its development the tadpole is practically a *fish.* After this stage, the frog loses its fish-like habits and structure. Fore and hind limbs appear and the tail is absorbed. The gills wither and the gill-slits close up, being replaced by *lungs.* The blood-system also changes to suit this new method of respiration, the auricle dividing into two, producing a *three*-chambered heart. The frog's lungs are simple, hollow sacs, the thin walls of which are folded inwards to increase the internal absorbing area. The walls of the lungs are richly supplied with blood-capillaries, and gases are exchanged between the blood and the air in the lungs. We have already discussed the way in which a frog breathes (Book III, p. 166). In the frog, *lymph* plays an important part in respiration, carrying oxygen from the red blood-cells through the walls of the capillaries to every cell. After exchanging oxygen for carbon dioxide and other waste products of the cells, the lymph is collected up by thin-walled lymph-vessels.

In the frog there are large *lymph-sacs* between the skin and the muscular body-wall, and these are connected with the lymph-vessels. Two pairs of *lymph-hearts* return this lymph to the veins.

The heart of an adult frog is somewhat conical in shape and it is divided into *three* chambers—a right auricle, a left auricle, and a single thick-walled ventricle. *De-oxygenated blood* from the *body-circulation* enters the *right auricle*, while *oxygenated blood* from the *lung-circulation* enters the *left auricle.* As the two auricles contract, both kinds of blood are forced into the ventricle, but there is very little mixing of the different kinds of blood owing to the spongy structure of the ventricle. Hence, when the ventricle is full, its left side contains oxygenated blood, its right side contains de-oxygenated blood, and there is a little mixed blood in the middle. When the ventricle contracts, the blood is forced along the main artery, which branches and distributes the blood to the lungs and to all other parts of the body. The three branch-arteries on each side, however, are arranged in such a way that (*a*) *the de-oxygenated blood goes to the lungs and skin*, (*b*) *the mixed blood goes to the hinder part of the body*, and (*c*) *the oxygenated blood goes to the head and brain.*

RESPIRATION IN BIRDS (e.g. PIGEON)

Birds are very active animals and have a high body-temperature, several degrees warmer than that of mammals. This calls for a great deal of energy which can only be supplied by a very efficient method of respiration. The bird's lungs, which are comparatively small, are attached to the ribs, hence they do not expand and contract during breathing. But the bird's lungs are not blind sacs, they open into a system of large, thin-walled *air-sacs* lying in the body-cavity. Every time air is breathed in, it rushes *through the lungs* into the air-sacs, and the same half-used air is later forced out from the air-sacs through the lungs again. In this way, the blood in the lung-capillaries has two opportunities of absorbing oxygen from the same air during each breathing movement. During flight, the action of the flying-muscles helps the breathing movements. This very efficient breathing system is one of the main factors that provide a bird with its high body-temperature and enable it to be one of the most active animals.

The bird's blood-system, too, is very efficient. The red blood-cells are large and more in number (per c.cm.) than in any other animal. As in the mammal, there is a *four*-chambered heart, and a *double circulation* keeps the oxygenated blood quite separate from the de-oxygenated blood. The right side of a bird's heart contains only de-oxygenated blood returning from the body-circulation, while the left side contains only oxygenated blood returning from the lung-circulation.

RESPIRATION IN MAMMALS (e.g. MAN)

We have already discussed the mechanism of breathing (or *external respiration*) in Man (Book III, p. 89). You should note that mammals are the only animals with a *diaphragm* forming the air-tight cavity necessary for this type of breathing. The lungs are covered with a thin membrane and they hang freely in the chest-cavity, which is also lined by a similar membrane. A film of liquid separates the two layers of membrane so that they slide easily over each other during breathing. The system of branching air-tubes ending in air-sacs inside the lungs presents an enormous area for the absorption of oxygen by the blood in the lung-capillaries. We have already described the blood-system of the mammal (Book III, p. 88).

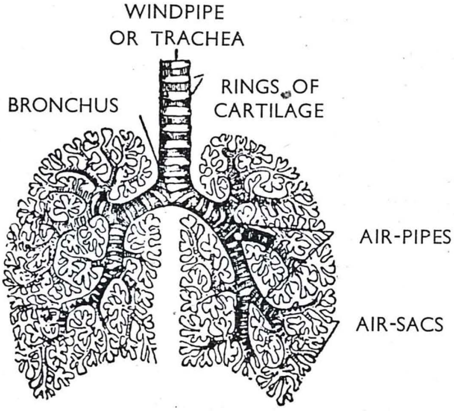

FIG. 62. HUMAN LUNGS

Oxygen diffuses into the blood-capillaries through the thin membrane lining the air-sacs, and combines with the haemoglobin of the red blood-cells.

This oxygenated blood finally reaches the capillary blood-vessels of the body-circulation. In these thin-walled capillaries, the oxy-haemoglobin gives up its loosely-combined oxygen to the lymph, which escapes through the walls of the capillaries and bathes the individual body-cells. *The essential feature of respiration—the actual setting free of energy—takes place in the individual body-cells.* (This is called *internal respiration*, or *tissue-respiration.*) In the living cell, the food-substance which is oxidized to set free energy is always a simple sugar (usually *glucose*) and the oxidation is brought about by *enzymes* present in the cell. Besides setting free energy, this oxidation also produces carbon dioxide and water, which diffuse out of the cell into the lymph and from there to the

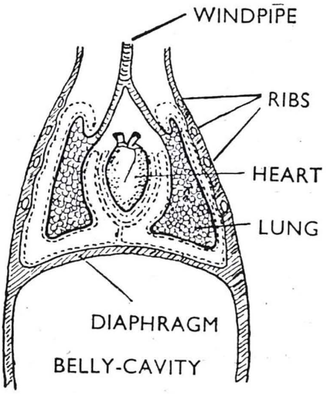

FIG. 63. DIAGRAM OF CHEST-CAVITY

blood-capillaries. On reaching the lungs, the carbon dioxide and water diffuses into the air-sacs and is got rid of during breathing.

You should notice that there is no essential difference between respiration in a one-celled organism like Amoeba and respiration in the individual body-cells of the higher animals.

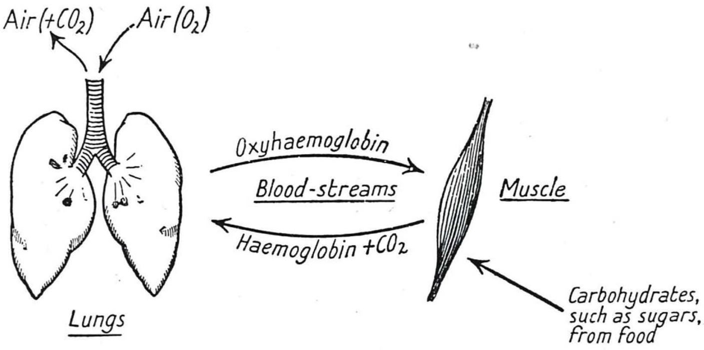

FIG. 64. DIAGRAM ILLUSTRATING RESPIRATION

BODY-TEMPERATURE (ANIMAL HEAT)

Animals are sometimes classified into two groups: (*a*) the *cold-blooded animals* and (*b*) the *warm-blooded animals*. The temperature of the blood of fishes, frogs and snakes, for example, changes with the temperature of their surroundings. Such animals are said to be 'cold-blooded', although this is a very misleading term since the blood of a snake lying in the hot sun may rise to about the same temperature as a mammal's. '*Variable-temperature animals*' would be a better term for such animals. Birds and mammals, however, are able to keep their bodies at a constant temperature, no matter what the temperature of their surroundings may be. A man's blood, for example, is always between 98° and 99° F. when he is healthy, whether he is living near the Equator or near the Arctic Circle. Birds and mammals are called 'warm-blooded animals', although '*constant-temperature animals*' would be a better term.

This constant high temperature is a great advantage to birds and mammals, since it enables the animal to be equally active under different external conditions. We have seen that an animal gets its supply of energy by oxidizing sugar, and that this chemical change takes place in the body-cells. Now all chemical changes are more rapid at higher temperatures; roughly, a rise in temperature of 10° C. *doubles* the speed of a chemical reaction. For example, if limestone is treated with dilute hydrochloric acid at a temperature of 30° C., the rate at which carbon dioxide is given off can be approximately doubled by warming the mixture to 40° C., while a further rise from 40° to 50° C. would almost double the flow of gas once more, i.e. about four times as much gas will be given off at 50° C. as at 30° C. The same principle applies to all the living processes taking place in the animal body, thus muscular action is more rapid when the body-temperature is reasonably high. This is one of the main reasons why birds are the most successful animals in the air, and also mammals on the land, because their activity does not depend on the temperature of the surrounding air. When the air-temperature is 30° C., we should expect a bird, with its body at 40° C., to be twice as active as a frog or snake of the same size. With the air at 20° C., the bird should be about four times as active, and at 10° C., eight times as active as the ' cold-blooded animal '.

REGULATION OF BODY-TEMPERATURE

We have seen that the heat-energy which keeps an animal's body warm is produced by the oxidation of sugar in the living cells. Since heat-energy is produced by every part of the body as long as it is alive, if it were allowed to accumulate in the body the body-temperature would get higher and higher, hence the excess heat must be got rid of. The heat set free by oxidation in the body-cells during respiration is distributed* to all parts of the body by the blood-stream, and, in time, it escapes into the surrounding atmosphere. Some of this heat (but seldom more than one-fifth) is got rid of during breathing. The air we breathe out is at the body-temperature (98·4° F. or 37° C.), and this is usually warmer than the air we breathe in; thus the body loses some heat. It is clear that in cold countries much more heat will be given up with the breath than in hot countries; and also that this loss of heat cannot be controlled, since we must always go on breathing.

In Man, the greater part of the excess heat produced during respiration is lost from the *skin*, and the amount of heat given off by the skin is regulated in such a way that the temperature of the body is kept very nearly constant. The skin gives up heat to the surrounding atmosphere in two different ways: (i) *by the ordinary physical processes of heat-transfer*, i.e. by conduction, convection and radiation, and if sufficient heat is not lost in this way, (ii) *by the production and evaporation of sweat.*

The loss of heat by conduction, convection and radiation is increased by an increased flow of blood through the capillary blood-vessels of the skin. When it becomes necessary to lose more heat, these capillaries become wider, more blood flows through them, and more heat is given off from the skin. (This method of heat-regulation resembles the way in which the engine of a motor-car gives up its excess heat through the ' radiator '.) If the outside temperature falls, the blood-vessels in the skin contract and allow only a small flow of blood, thus the loss of heat is reduced.

SWEAT

When the temperature of the surrounding air is high, or when the muscles are in very active use, the skin cannot get rid of the excess heat quickly enough by conduction, convection and radiation, and further cooling is brought about by *sweating*. Even under cool conditions, the skin is always giving off a little water-vapour, but when hard muscular work is done under hot conditions, large quantities of water are lost as *sweat*, sometimes as much as a litre in an hour. This sweat escapes from *sweat-pores* in the skin. It is formed in knots of coiled tubes—*sweat-glands*—deep down in the skin, and passes along the *sweat-ducts* to the *sweat-pores*.

Sweat is mainly water and the sweat-glands remove this water from the blood flowing through the skin. From our work on the latent heat of steam we know that the evaporation of 1 gm. of water from the skin will require 536 calories of heat. Most of this latent heat will be taken from the skin itself, thus cooling it. Hence the more sweat evaporated from the skin the more heat given up by the body. It is important to realize, however, that sweat can have no cooling effect unless it evaporates, and when drops of liquid sweat remain on the skin without evaporating there is great discomfort owing to the lack of any cooling effect. This is why a hot, damp atmosphere, almost saturated with water-vapour, is so much more uncomfortable than a dry atmosphere *at exactly the same temperature*. In air at 105° F. which has a relative humidity of 50 per cent a man can keep ' cool ', but in air at 105° F. which is saturated with water-vapour he will experience no cooling effect and his temperature will rise, finally causing ' heat-stroke '.

Since evaporation is more rapid in a current of air than in still air, moving air feels ' cooler ' than still air of the same temperature and relative humidity.

To sum up, the body can lose heat by (*a*) *evaporation of sweat from the skin*, (*b*) *evaporation of water from the lungs during breathing*, (*c*) *warming up the air during breathing*, and (*d*) *radiation, convection and conduction from the skin*.

PREVENTING LOSS OF HEAT

When the temperature of the surrounding air is low, or when the muscles are not being used, or when the animal is short of food, birds and mammals have to avoid loss of heat. This is done by having a non-conducting layer at the surface of the body. A bird's body is covered with feathers which cause the body to be surrounded with a layer of air, which is a very bad conductor of heat. The covering of hair on the skin of most mammals also encloses some air and serves a similar purpose. Man has to provide himself with clothes which also serve to keep a layer of non-conducting air near the skin. A layer of fat under the skin serves a similar purpose, and this is particularly important in mammals which live in water, e.g. whales and seals, which have a very thick layer of fat just beneath the skin.

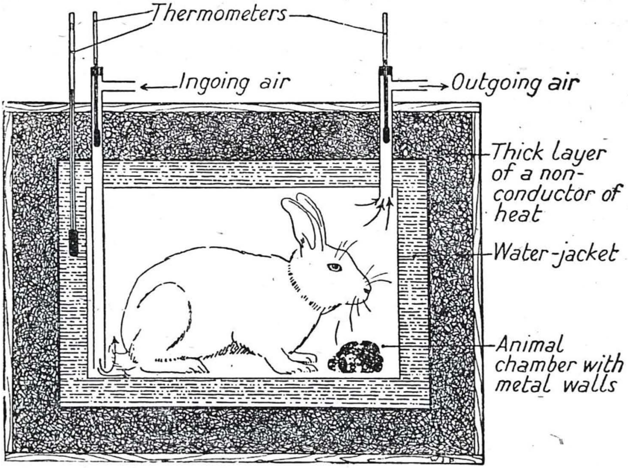

FIG. 65. ANIMAL CALORIMETER

FOOD AND ENERGY

In Chemistry lessons we learnt that the general principle underlying all chemical changes is the *Law of the Conservation of Matter*, i.e. *matter can be neither created nor destroyed.* In other words, the weight of everything we start with in a chemical reaction is exactly the same as the weight of all the products formed during the reaction. In Physics lessons we learn that the general principle running through the whole of Physics is the *Law of the Conservation of Energy*, i.e. *energy can be neither created nor destroyed.* In other words, energy can be *transformed* from one form into another, but the total amount of energy remains unchanged throughout the transformation.

These two laws apply equally to all the living processes taking place in plants and animals. In the case of animals, the truth of this has been verified experimentally by comparing the energy an animal produces in a given period with the amount of energy in the animal's food over the same period. Animals have been given a measured amount of food and kept in specially designed calorimeters for several hours, and the amount of heat-energy produced has been carefully measured. The same amount of the same kind of food has then been burnt in oxygen in another type of calorimeter and its *heat of combustion* has been found. In such experiments (provided that the animal neither gains nor loses weight while it is in the calorimeter) it is always found that *the same amount of food produces the same amount of heat-energy whether it is burnt rapidly in oxygen or oxidized slowly in an animal's body.* An animal, therefore, is just as much a *heat-engine* as a petrol-engine is. A motor-engine gets its energy from the rapid oxidation of a hydrocarbon (petrol), while the animal gets its energy from the slow oxidation of food substances in its body-cells with the aid of enzymes.

ENERGY OUTPUT AND FOOD REQUIREMENTS

Since the Law of Conservation of Energy applies to animals, if we know how much energy an animal produces (as heat and work) in a day, we can find out how much food it must eat in order to supply this energy.

The energy output of an animal, e.g. a man, will depend on the kind of work he is doing. It is important to realize, however, that some muscles are always working even when a man is ' resting ', e.g. his heart-muscle and the muscles concerned with breathing-movements are *doing work* even during sleep. A man resting completely for 24 hours produces about 1,700 *kilo-calories* of heat. (The ' small calorie ' of our Physics lessons is inconveniently small when we deal with the heat-production of large animals, hence the ' large calorie ' or ' Kilo-calorie ' is used. 1 Kilo-calorie is equal to 1,000 ' small calories ': or, the Kilo-calorie is the quantity of heat required to raise the temperature of 1 Kilogram of water through 1° C.) These 1,700 Kilo-calories per day (about 70 Kilo-calories per hour) will be the *minimum energy* required. Any muscular effort will require additional energy, e.g. merely standing up increases the heat-production from 70 to about

100 Kilo-calories per hour; walking at an ordinary speed will increase it to about 300 Kilo-calories per hour, and very hard muscular work to about 750 Kilo-calories per hour. The ' average man's ' output of energy in 24 hours is usually taken as 3,000 Kilo-calories.

If the body is neither to gain nor lose weight, it must get the same amount of energy from food (plus an extra allowance—about 10 per cent—for that part of the food which is not digested). Hence an ' average man's ' food must supply him with *about* 3,300 Kilo-calories of energy per day. (The energy requirement of people living in a hot, tropical climate—where there is no difficulty in ' keeping warm '—is usually taken as about 10 per cent less than they would need in a cool, temperate climate.) We shall now consider the ' energy-values ' of the various kinds of food-materials.

ENERGY VALUES OF FOODS

Of the six classes of food-materials—carbohydrates, fats, proteins, water, salts, and vitamins—only the first three will yield energy on oxidation. The approximate energy-values of carbohydrates, fats, and proteins are:—

One gram of a carbohydrate yields about 4 *Kilo-calories on oxidation.*
One gram of a fat yields about 9 *Kilo-calories on oxidation.*
One gram of a protein yields about 4 *Kilo-calories when partly oxidized.*

The ' average man's ' energy-requirement of 3,300 Kilo-calories can be made up from many different proportions of carbohydrate, fat, and protein, but a good proportion would be:—

500 grams of *carbohydrate*, giving	2,000 Kilo-calories
100 grams of *fat*, giving	900 Kilo-calories
100 grams of *protein*, giving	400 Kilo-calories
Total	3,300 Kilo-calories

Here is a tropical diet suggested by Professor Rosedale:—

16 ounces Rice	1,547 Kilo-calories
*4 ounces Mutton	197 Kilo-calories
1½ ounces Fat	406 Kilo-calories
4 ounces Lentils	398 Kilo-calories
1 ounce Peanuts	178 Kilo-calories
8 ounces Vegetables	150 Kilo-calories
Total	2,876 Kilo-calories

He suggests, however, that the animal protein (*) is rather low and that a few ounces of *fish* might well be added. The relative proportions of the various foodstuffs should be such that *about two-thirds of the energy requirement is provided by carbohydrates, one-sixth by fats, and one-sixth by proteins* (*half this protein being animal protein and half vegetable protein*). These proportions can be varied considerably according to individual needs, as long as the following principles are borne in mind: (*a*) *there must always be sufficient protein for growth and repair*, (*b*) *hard muscular work requires an increased proportion of carbohydrates*, (*c*) *under cold conditions, an increased proportion of fat will provide extra heat.*

RESPIRATION IN GREEN PLANTS

Although plants do not show such active movements as animals, since they are usually rooted in one spot all their lives, they carry out slow movements which require energy. For example, a tall tree has had to do work against the force of gravity in order to raise its topmost branches from ground level; roots have to be forced through the soil and growing parts need energy to build up new protoplasm.

We have seen that *respiration* is a chemical process carried out by all living protoplasm, and the details of *internal* or *tissue respiration* are essentially the same in plants and animals. The differences in *external respiration* in plants and animals are mainly differences in the arrangements for transporting oxygen from the atmosphere to the tissues. Also, since plants are less ' active ' than animals, their energy-requirements are less and their respiration is much less rapid. Green water-plants can absorb dissolved oxygen over the whole of their under-water surface. The under-water leaves of water-plants are usually very thin and small: the surface-layer is usually only one cell thick, and since no part of the leaf is far from the surface no special transport system is necessary.

Although green land plants are surrounded by air, they cannot absorb oxygen freely over their whole surface, but do it mainly through *stomates* in their green parts and through *lenticels* in the bark of their stems and roots (and also through their root-hairs which absorb soil-water containing dissolved oxygen). The leaf-skin usually has a waterproof layer on its outer surface which limits the inward passage of oxygen (and, of course, the outward passage of water). We have seen that plant-cells are not tightly packed and that air can be drawn through the air-spaces in stems and leaves (Book II, pp. 112 and 126). These air-spaces extend throughout the living parts of leaves, stems and roots, and every living plant-cell is near one of these air-spaces; so that oxygen is absorbed through the damp cell-walls, and carbon dioxide and water are got rid of in the same way. In sunlight, however, the oxygen set free during photo-synthesis supplies all the oxygen needed for respiration in the green parts of plants.

RESPIRATION IN NON-GREEN PLANTS

In fungi (e.g. Mucor and the Mushroom), oxygen diffuses through the surface layer and internal respiration is the same as in green plants. In yeast, if oxygen is available, the same type of respiration takes place; but, in the absence of oxygen, yeast frees energy from glucose by a different chemical change—*fermentation.* In this change alcohol and carbon dioxide are the waste products instead of carbon dioxide and water.

EXPERIMENTS ON RESPIRATION

To compare Inspired Air, Expired Air, and Air from the Air-sacs. (i) Take three similar gas-jars, A, B, and C. Fill B and C with water and invert in a gas-trough. Using a piece of rubber tube, fill B with *expired air* by breathing out normally through the tube. Fill C with air from the *air-sacs* of the lungs by holding the nose and breathing out through the tube *at the end of a forced expiration.* Invert the three jars, in turn, over a lighted candle, noting the time the candle continues to burn in each case. Repeat the experiment.

	(*a*) Fresh Air	(*b*) Expired Air	(*c*) Air-sac Air
Experiment I			
Flame extinguished in	........sec.	........sec.	........sec.
Experiment II			
Flame extinguished in	........sec.	........sec.	........sec.
Experiment III			
Flame extinguished in	........sec.	........sec.	........sec.
Average	........sec.	........sec.	........sec.

What do your results suggest? ....

....

FURTHER EXPERIMENTS ON RESPIRATION

To compare Inspired Air, Expired Air, and Air from the Air-sacs. (ii) Take a burette fitted with a rubber tube and a spring clip; pour 50 c.cm. of water into it and record the level. Then fill to the top with water, close the clip and invert in a trough of water.

Open the tap and admit 50 c.cm. of fresh (inspired) air, taking care that the water levels inside and outside the burette are approximately the same. Close the clip and put a well-fitting rubber cork in the open end of the burette—under water. Holding the burette the right way up, quickly remove the cork and drop in 5 little balls of sodium hydroxide; then replace the cork. Shake the burette until all the sodium hydroxide has dissolved, then invert the burette in water and remove the cork. Cool the air to its original temperature by pouring water over the outside of the burette, and then adjust the water-levels as before, making sure before reading off the volume of the air that all the liquid has drained down from the narrow tube just below the clip. Record the new volume below. (Notice that there is little or no change in volume, since fresh atmospheric air contains so very little carbon dioxide—0·03 per cent.) Cork the burette again, under water, and then, holding it the right way up, quickly take out the cork and put in about 0·1 gm. of alkaline *pyrogallol.*† Replace the cork and shake the burette so as to absorb all the oxygen. (The 'alkaline pyrogallol' turns brown as it absorbs the oxygen.) Then invert the burette and remove the cork under water. Notice how the water rushes in to take the place of the oxygen which has been absorbed. After cooling the gas to its original temperature and equalizing the pressure, read off the new volume and record below. The gas which remains is nitrogen.

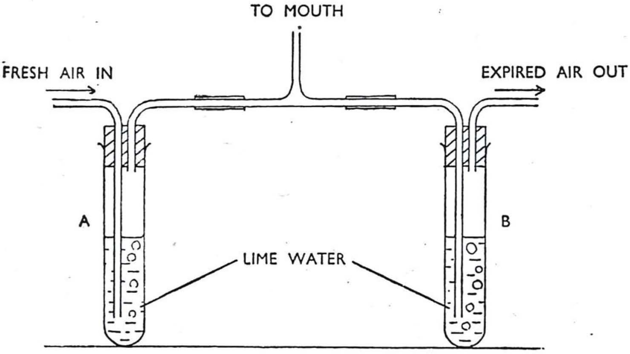

FIG. 66. CHANGES PRODUCED IN AIR BY BREATHING

Repeat the experiment, first putting 50 c.cm. of *expired air* into the burette by breathing out *normally* through a short rubber tube attached to the burette. (Put in more than 50 c.cm. at first, cool the air down to room temperature by pouring water down the outside of the burette, then allow any excess of air to escape by opening the clip and adjusting the water-level.)

Again repeat the experiment, first putting 50 c.cm. of air from the *air-sacs* into the burette by holding the nose and breathing out *at the end of a forced expiration.*

	Fresh Air	Expired Air	Air-sac Air
Original volume of air	50·0 c.cm.	50·0 c.cm.	50·0 c.cm.
Burette reading at start of experiment			
Burette reading after shaking with NaOH			
Therefore, volume of carbon dioxide absorbed =			
Therefore, percentage of carbon dioxide =			
Burette reading after shaking with alk. pyro.			
Therefore volume of oxygen absorbed =			
Therefore, percentage of oxygen =			
Therefore, percentage of nitrogen =			

Notice that the percentage of carbon dioxide in expired air is greater than that in fresh air, while the percentage of oxygen is less. The percentage of nitrogen, however, remains about the same. Notice, also, that air from the air-sacs contains still more carbon dioxide and still less oxygen, though the percentage of nitrogen remains almost unchanged.

(iii) Fit up the apparatus shown in Fig. 66. Half-fill both test-tubes with clear lime-water. Apply the mouth to the 'T'-tube T, hold the nose, and breathe gently in and out through the mouth. The air enters by way of A and leaves by way of B.

Notice that the lime-water in B turns chalky long before that in A, showing that expired air contains much more carbon dioxide than inspired (fresh) air.

(iv) To 25 c.cm. of very dilute sodium hydroxide solution (N/500 NaOH) add 10 drops of an indicator solution, such as Phenol Red, Bromo-Thymol Blue, Neutral Red or Phenol-phthalein, and repeat the above experiment, using equal volumes of the indicator solution in the tubes A and B. Notice that the colour of the liquid in B changes long before that in A, owing to the greater proportion of carbon dioxide in expired air.

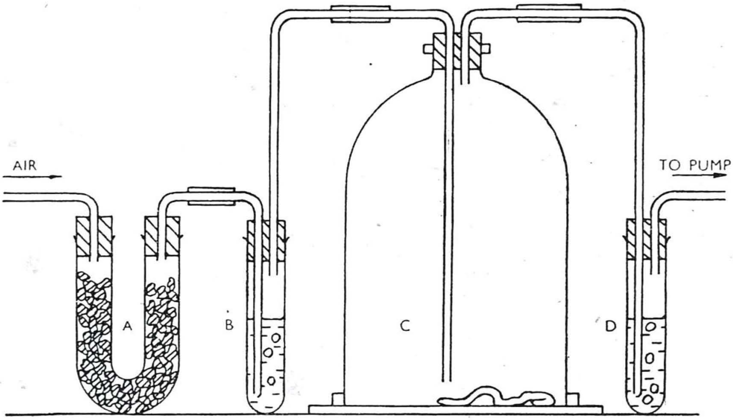

FIG. 67. RESPIRATION EXPERIMENT

All these indicators are sensitive to carbon dioxide and can be used for detecting and comparing small quantities of carbon dioxide in respiration experiments.

To show that a small animal gives out Carbon Dioxide. Fit up the apparatus as shown in Fig. 67. Air is drawn through the apparatus from left to right, first through soda-lime in the 'U'-tube A so as to free the air from carbon dioxide before it passes over the animal (e.g. a snail, earth-worm, frog or cockroach) in C. (The test-tube B, between A and C, containing lime-water, shows that the air is free from carbon dioxide.) After passing through C, the air passes through the lime-water in tube D. Although the lime-water in B remains clear, that in D turns chalky, showing that the animal is giving out carbon dioxide.

To measure the Oxygen used by a small animal during Respiration. If a small animal is kept in a closed vessel connected to a manometer, small changes in pressure or volume can be observed. If the carbon dioxide produced by the animal is absorbed by sodium hydroxide, the total volume of air in the apparatus will diminish as the animal uses up oxygen, and this change in volume is shown by the movement of the liquid in the manometer. In practice, however, such a simple arrangement as that shown in Fig. 68 is not satisfactory, since changes in the temperature and pressure of the air in the apparatus may produce bigger changes in volume than those produced by the animal's respiration.

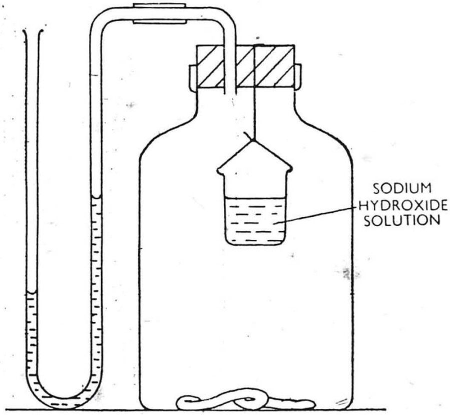

FIG. 68. SIMPLE RESPIROMETER

The modified Haldane apparatus shown in Fig. 69 overcomes this difficulty and enables us to measure the rate of respiration accurately under different conditions, e.g. at different temperatures. The two vessels, A and B (which may be either bottles, flasks, or wide tubes) are of the same size *so that any changes in temperature and pressure have exactly the same effect on the air in A and B.* Equal numbers of sodium hydroxide balls (or equal volumes of the same sodium hydroxide solution) are placed in A and B. The animal (e.g. a small frog, snail, earth-worm or insect) is placed in A, while B serves as a *control.* Both A and B are connected by thick-walled rubber tubing to either side of the water-manometer M, whose level can be adjusted by the levelling-tube X. The vessel A is also connected to the graduated tube T (a 1 c.cm. pipette graduated in 0·01 c.cm.) containing coloured water whose level can be adjusted by moving the other levelling-tube Y.

At the start of the experiment, the clips R and S are left open and both the vessels A and B are placed in a large vessel of water so as to keep their temperature steady. After allowing sufficient time for A and B to come to the temperature of the water, the water-level in T is set at the lowest mark and the water-levels in the manometer M are set at some convenient point, where the level is marked by stretching a piece of thread CD between two lumps of soft wax (or plasticene) stuck on the board to which the apparatus is attached. The two clips R and S are then closed (at the same instant). As the animal uses up oxygen from A and gives out carbon dioxide (which is absorbed by the sodium hydroxide) the pressure of the air in A becomes less than that in B. Hence the water in the manometer M stands at a different level in the two limbs, rising on the side C and falling on the side D. The volume of oxygen used up by the animal is measured by raising

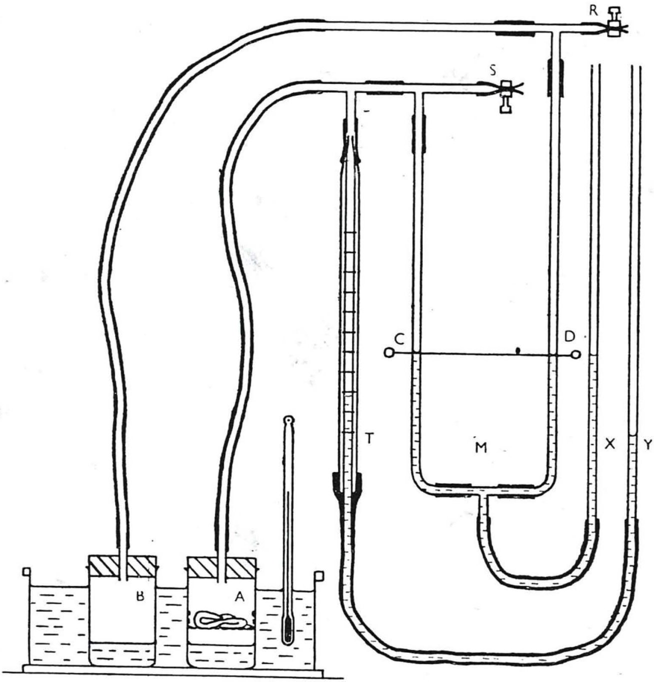

FIG. 69. RESPIROMETER

the levelling-tube Y and displacing air from T until the water in the manometer M returns to its original level (also moving the levelling-tube X if the water-surfaces are above or below CD when levelled). A and B are then both at the original atmospheric pressure, and the rise of liquid in the graduated tube T gives the volume of oxygen used up.[1]

Carry out the first experiment at 10° C. Before starting the experiment, allow the temperature of the animal-vessel to become steady and allow the animal to settle down; then place at C a tube containing about 25 c.cm. of water, turn on the pump, and adjust the rate of bubbling by adjusting the screw-clip. Draw air through the apparatus for several minutes so as to sweep out any accumulation of carbon dioxide from the animal-vessel. Then place a tube of dilute alkali and indicator at C (prepared as at (iv), p. 124), noting the time. Take the time again when the liquid in C is exactly the same colour as that in the control-tube B. Repeat the experiment until two successive results are in close agreement, showing that the animal has settled down to its normal rate of respiration at this temperature. (N.B.—Be very

Temperature	10° C.		20° C.		30° C.	
Experiment No.	C.C.T.	R.R.	C.C.T.	R.R.	C.C.T.	R.R.
I	sec. /	min.	sec. /	min.	sec. /	min.
II	,,	,,	,,	,,	,,	,,
III	,,	,,	,,	,,	,,	,,
IV	,,	,,	,,	,,	,,	,,
V	,,	,,	,,	,,	,,	,,
Average under steady conditions	)____	)____	)____	)____	)____	)____

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[1] This apparatus can also be used for measuring the rate of respiration of plant material. It is interesting to compare the rate of respiration of equal weights of plant and animal material, e.g. calculate the oxygen used up in unit time by 100 gm. of germinating seeds and 100 gm. of frog.

careful to sweep out any accumulations of carbon dioxide from the animal vessel and the tubes of the apparatus before putting a fresh tube of alkali and indicator at C.)

Repeat the experiment (*a*) at 20° C. and (*b*) at 30° C., recording all your results in the table on p. 127. (C.C.T. = 'colour-change-time'.)

If a frog is used, it is interesting to measure the 'respiratory rate' at the same time, counting the number of movements of the floor of the mouth per minute. If the 'respiratory rate' and the 'colour-change-time' are plotted against the temperature, the two graphs are seen to be very similar. (R.R. = 'respiratory rate'.)

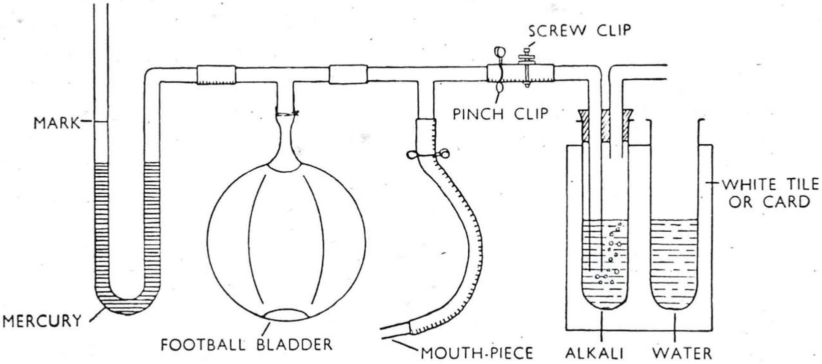

FIG. 70. RESPIROMETER

To show the effect of exercise on Carbon Dioxide Production in Man.[1]

The apparatus shown in Fig. 70 consists of a mercury manometer, a No. 4 football bladder, a piece of rubber tube fitted with a *spring*-clip and a tube containing 25 c.cm. of N/500 NaOH solution with phenolphthalein as an indicator.

The horizontal piece of pressure-tubing is fitted with a *spring*-clip and a *screw*-clip, the latter being first adjusted to give a suitable rate of bubbling and left in this position during all the experiments. The bladder is pressed out flat between experiments.

Before taking any additional exercise, open the clip on the mouth-tube and blow up the bladder *with one breath.* Then close the clip on the mouth tube. The surface of the mercury in the open limb of the manometer will then stand higher than that of the other limb. Mark this level in the open limb by means of a rubber band. Then

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[1] This experiment was devised by G. N. Ridley and was described in *The School Science Review* for March, 1936.

determine your pulse-rate, i.e. the number of times your pulse ' beats ' per minute, and record this below. Then open the spring-clip on the horizontal tube for a few seconds before attaching the test-tube containing the alkali so as to sweep out the air from the tubes of the apparatus. Then close the clip and put the test-tube in position as shown in Fig. 70. Open the spring-clip on the horizontal tube, noting the time, and allow the expired air from the bladder to bubble through the dilute alkali and indicator, noting the time when the last trace of colour disappears.[1] Record your result below.

Squeeze all the air out of the bladder and then blow it up again after 30 seconds of vigorous exercise, e.g. ' knees up ', taking care that the mercury in the manometer stands at the same level as before. Find your pulse-rate. Sweep out the air from the tubes of the apparatus as before, and then place another 25 c.cm. of the coloured alkali solution in the test-tube and take the time required to decolourize the liquid as before. Record your result below.

Repeat the experiment after 1 minute of vigorous exercise and record the ' colour-change-time ' and your pulse-rate below.

Time of Exercise	Colour-change-Time	Pulse-rate
0 sec.	sec.	per min.
30 sec.	sec.	per min.
60 sec.	sec.	per min.
90 sec.	sec.	per min.

Draw graphs, (*a*) plotting ' colour-change-time ' against time of exercise, (*b*) plotting pulse-rate against time of exercise. Notice the relation between the two graphs, showing that heart-beat (pulse-rate) and carbon dioxide production are both affected in the same way by exercise.

BLOOD

THE CLOTTING OF BLOOD

Although blood is a liquid when in the blood-vessels, after escaping it soon sets to a jelly-like *clot*. Thus, if blood is collected in a beaker

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[1] A control tube containing water will help in judging the colour change.

and is then allowed to stand for 5-10 minutes, the beaker can be inverted without any blood running out. The blood clots, owing to the formation of a network of fine threads of *fibrin*,† an insoluble protein, formed from *fibrinogen*,† a soluble protein contained in the blood, when blood is exposed to the air. This net of fibrin threads works itself round the blood-cells. After standing for some time, the fibrin threads contract so that the clot contracts also, squeezing out a clear yellow liquid called *serum*.†

If freshly-collected blood is stirred or shaken, the fibrin can be removed as it is formed, and the *de-fibrinated blood* which remains will not clot.

Since the calcium salts in the blood are necessary for clotting, if the calcium compounds are removed the blood will not clot. Potassium oxalate† or ammonium oxalate† will precipitate the calcium from the blood as insoluble calcium oxalate.† This *oxalated blood* can be made to clot by adding a calcium salt.

(i) Place strips of red and blue litmus paper on a clean white tile and moisten them with 10 per cent sodium chloride solution. Place one drop of blood (de-fibrinated, oxalated, or fresh—your own, for example) on each strip, and after about 30 seconds wash off the blood with more salt solution. Notice that the blood is slightly alkaline to litmus.[1]

(ii) Examine a drop of blood under the microscope, first spreading the drop with a second glass slide so as to form a thin film of blood. Using the high power objective, identify the red blood-cells (reddish-yellow under the microscope) surrounded by a colourless liquid—*serum*.

(iii) Take about 25 c.cm. of oxalated blood which has been warmed to about 37° C. and add 1 c.cm. (25 drops) of 2 per cent calcium chloride solution. Mix and then allow to stand. Note the formation of a *blood-clot*, after standing for about an hour. Notice how the clot begins to contract and becomes surrounded by *serum*. Examine the blood-clot. Notice the bright red colour on the surface which is in contact with the air. Cut the clot in half and notice that the part which has not been in contact with the air is a darker, purple colour, and that it turns bright red on standing in contact with the air.

(iv) Dilute 10 c.cm. of blood to 100 c.cm. with water. Fill two 5-in. by $\frac{5}{8}$-in. test-tubes with this diluted blood (20 c.cm.) and to

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[1] Drops of your own blood can best be obtained as follows:—Sterilize* the skin at the tip of the left fore-finger by wiping with cotton-wool soaked in alcohol. Then wrap the corner of your handkerchief, or a piece of string, several times round the end joint of this finger (but do not *tie* it) and bend the finger so that the finger-tip becomes swollen with blood. With a sterile needle (which has been kept in alcohol) pierce the skin sharply and squeeze out a drop of blood. Wipe the spot with alcohol after removing the blood. This method is safe and practically painless.

one tube add 1 c.cm. of 1 per cent *sodium hydro-sulphite solution* ($Na_2S_2O_4$)—a weak *reducing agent.* (This solution must be freshly prepared.)

Close the mouth of the tube with the thumb and mix gently. Notice that the bright red colour darkens to purple because the *oxy-haemoglobin* (characteristic of *oxygenated* blood) has been reduced to *haemoglobin* (characteristic of *de-oxygenated* blood). Transfer this purple liquid to a boiling-tube and shake vigorously in contact with air. The oxygen of the air oxidizes the haemoglobin back to oxy-haemoglobin and the bright red colour returns. Transfer to the original test-tube and compare the colour with that of the ' control ' tube of diluted blood.

EXAMINATION OF HEART AND LUNGS

Examine the heart and lungs of a large mammal, e.g. sheep or goat, obtained from the abattoir* with the blood-vessels cut off some distance from the heart and with the wind-pipe and voice-box complete. Slit open the *heart-sac* to expose the heart and identify (*a*) the ventral surface of the heart with its diagonal groove filled with fat, (*b*) the right ventricle (thin-walled and rather soft to the touch), (*c*) the left ventricle (thick-walled and therefore firm and hard), (*d*) the right and left auricles (thin-walled and divided from the ventricles by a horizontal groove filled with fat). Draw the heart and lungs as seen from the ventral surface.

Cut horizontal slices across the apex of the heart until you expose the cavities of both ventricles. Notice (*a*) the shape of the cross-section of each ventricle and (*b*) the relative thickness of their muscular walls. Draw what you see. Wash out the heart with water and then slit up the ventral wall of each ventricle and notice (*a*) the valve-strings and valve-flaps of the valves which separate the ventricles from the auricles (there are 3 valve-flaps in the right side of the heart and 2 in the left), (*b*) the opening of *the artery to the lungs* in the *right* ventricle, (*c*) the opening of *the main artery to the body* in the *left* ventricle. Push a thick, round-ended glass rod into these openings and follow the course of these two arteries.

Slit up the auricles and open up the right and left sides of the heart. Notice (*a*) the openings of the two *main veins* (anterior and posterior) into the *right* auricle, and (*b*) the openings of the two *veins from the lungs* into the *left* auricle. Follow the course of these veins by pushing a glass rod along them.

Finally, slit open the main artery just above the left ventricle to expose the three *pocket-shaped valves* which prevent back-flow of blood

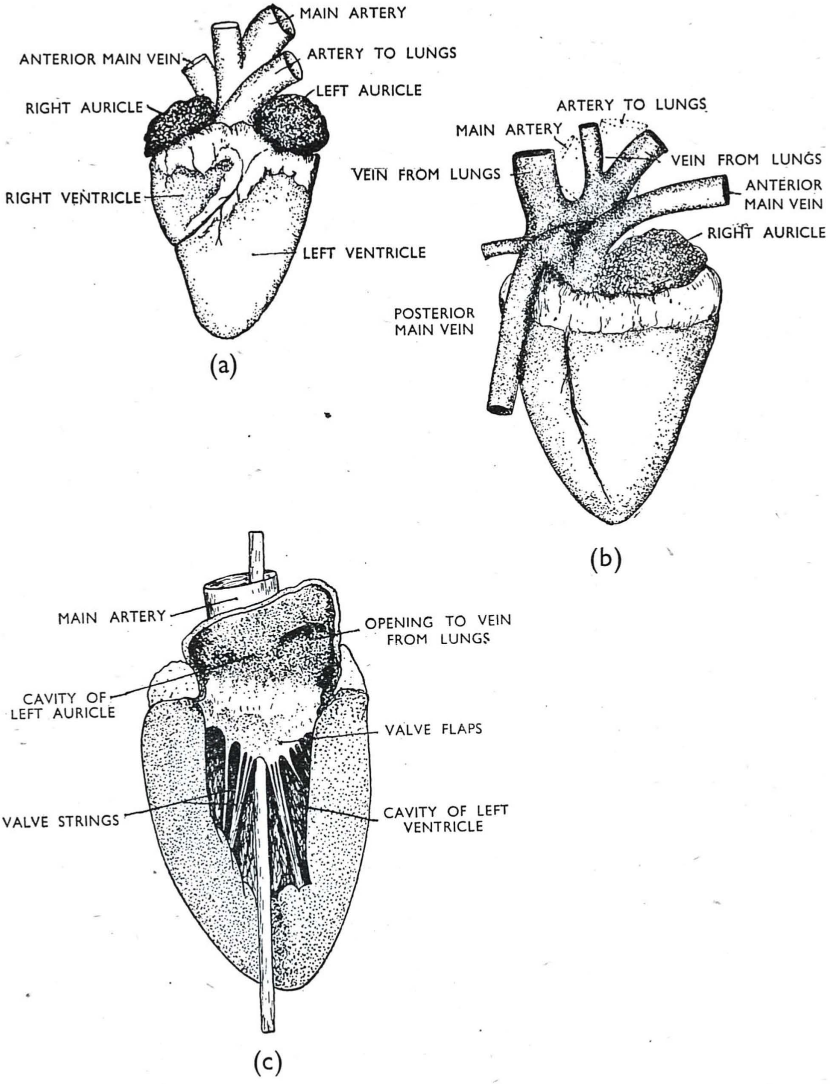


FIG. 71. VIEW OF SHEEP'S HEART
(*a*)—From dorsal surface. (*b*)—From ventral surface.
(*c*)—From left side, with glass rod through main artery

into the ventricle. Pour some water on to these valves and notice how they expand to meet each other and close the opening.

EXAMINATION OF LUNGS

Examine the lungs and wind-pipe of a large mammal and draw their external appearance from the ventral view. Identify (*a*) the voice-box, (*b*) the *epiglottis*,† a lid-like structure which closes the opening to the wind-pipe when swallowing. Look down inside the voice-box and notice the *glottis*, a narrow slit, with the *voice-cords* on either side. Slit open the voice-box and find the voice-cords.

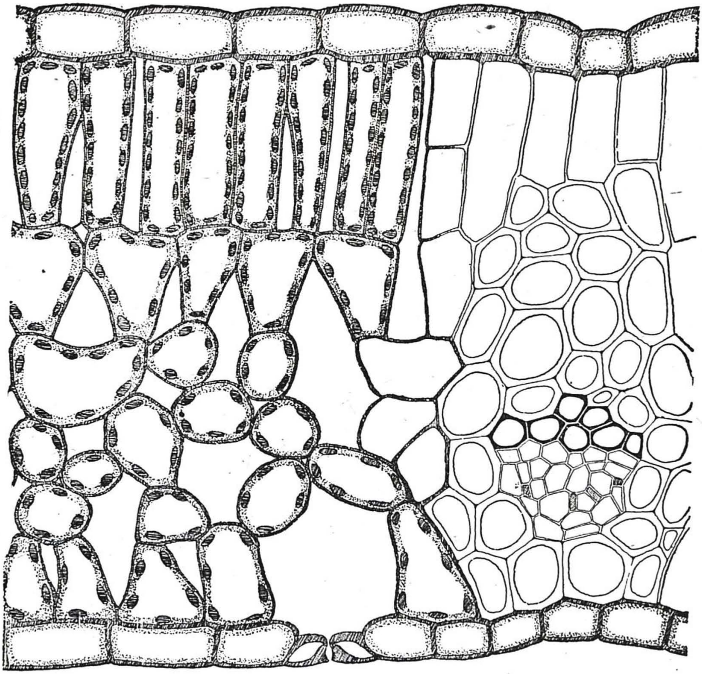


FIG. 72. TRANSVERSE SECTION OF LEAF (highly magnified)

Examine the wind-pipe and notice how it is stiffened by ' C '-shaped rings of gristle so that it is always kept open. Slit open the wind-pipe as far as its posterior end where it divides into two branches. Slit open one of these branches and follow its course as it divides again and again into smaller and smaller air-tubes. At the same time, identify the arteries and veins running through the tissues of the lungs.

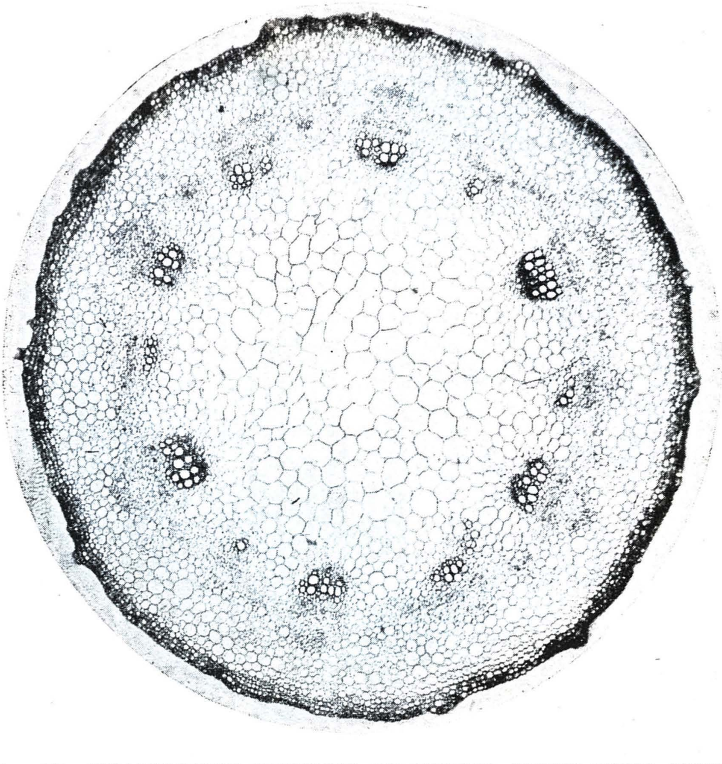


FIG. 73. TRANSVERSE SECTION OF YOUNG SUNFLOWER STEM (a di-cotyledon)

PLANT STRUCTURE

MORE ABOUT THE INTERNAL STRUCTURE OF PLANTS

In order to understand clearly the processes of nutrition and respiration in plants it is necessary to learn more about their internal structure as seen through the microscope.

We have already seen that a leaf has a ‘ skeleton ’ of *veins*; these veins are continuations of the *conducting bundles* (or *vascular bundles*) of the stem and its branches. Each vein consists of two parts, the *wood* on the upper side of the leaf-stalk and the *bast* on the under side. If a thin cross-section of a leaf is examined under the microscope, it is seen (Fig. 72) to have a *leaf-skin* on both its upper and its under side. This skin consists of a single layer of flattened, transparent cells. In the under-surface can be seen *stomates*, each with its two *guard-cells* which can change their shape and so regulate the size of the pore of the stomate.

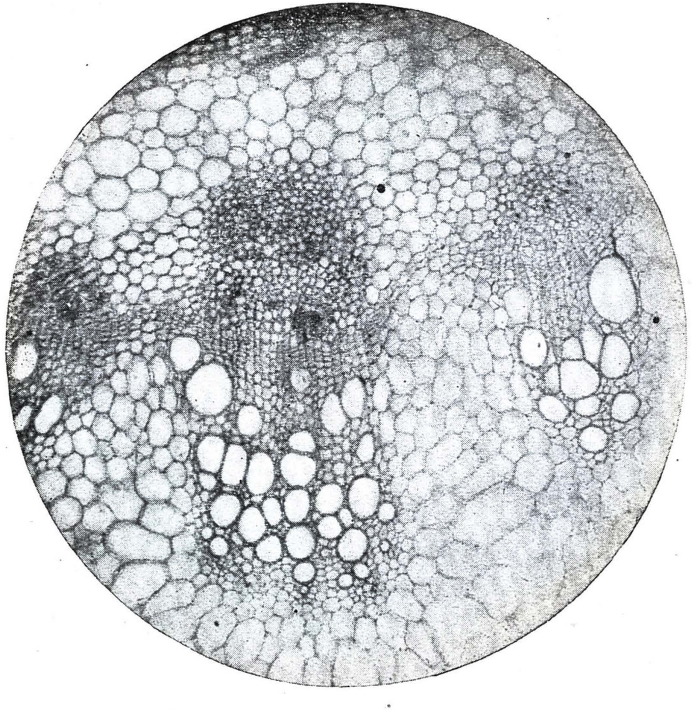


FIG. 74. MORE HIGHLY MAGNIFIED TRANSVERSE SECTION OF PART OF SUNFLOWER STEM SHOWING VASCULAR BUNDLES AND CAMBIUM RING

Just beneath the upper leaf-skin is a layer of long cells, packed closely together perpendicular to the leaf-surface, called *palisade cells.*† These palisade cells contain most of the green colouring matter of the leaf, and since this part of the leaf gets most light, it is in these cells that most of the carbohydrates are synthesized. The cells in the lower part of the leaf, below the palisade cells, are usually loosely packed, forming a *spongy tissue*, with plenty of *air-spaces* between the cells. It is through the air-spaces in the spongy tissue that the *exchange of gases* takes place during photo-synthesis, respiration and transpiration.

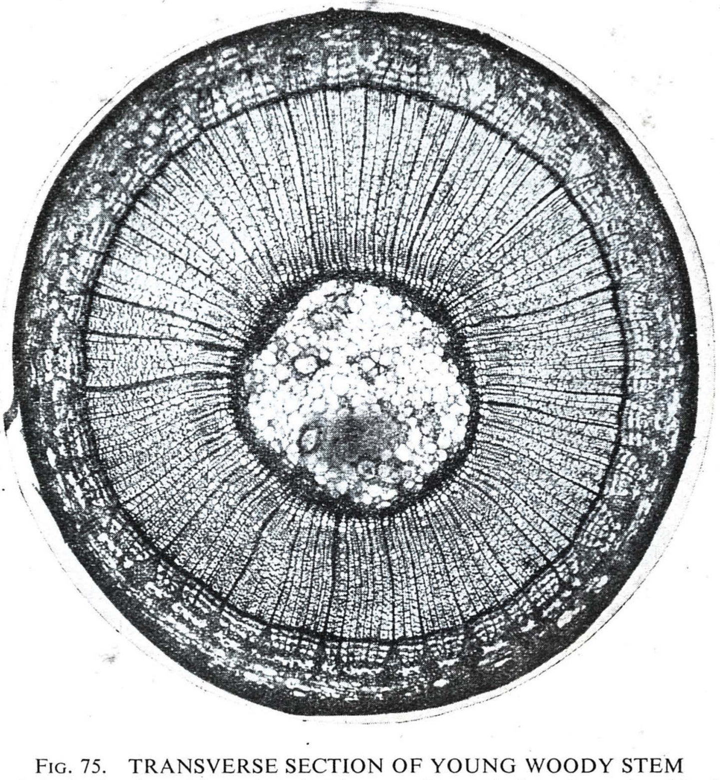


FIG. 75. TRANSVERSE SECTION OF YOUNG WOODY STEM (di-cotyledon) showing pith surrounded by complete cylinder of wood, with ring of cambium separating wood from bast. Notice the outer layer of bark and also the rays running through the wood.

The greater part of the stem of a young non-woody plant, e.g. Bean, Sunflower, Cucumber, consists of simple, thin-walled cells forming the *ground-tissue*. Running lengthwise through such a stem are a number of tough threads called *vascular bundles*. In di-cotyledons, the vascular bundles are arranged more or less in a circle, near the outside of the stem, while in mono-cotyledons they are scattered throughout the ground-tissue but are more in number near the outside of the stem.

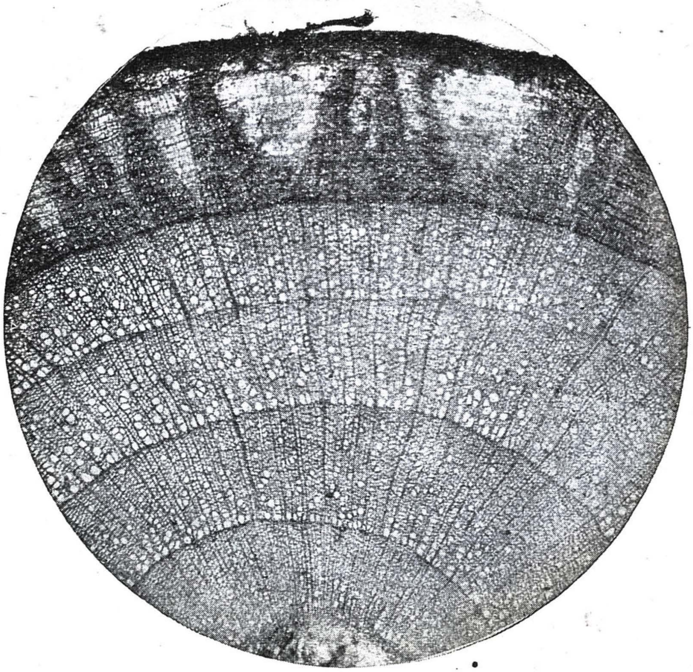


FIG. 76. TRANSVERSE SECTION OF OLDER WOODY STEM (di-cotyledon), showing growth-rings and rays.

If a cross-section of the stem of a young di-cotyledon is examined under the microscope, it is seen (Figs. 73, 74) that each vascular bundle consists of two distinct parts, the inner part having wider and thicker-walled cells (the *wood*-vessels) and the outer part of the bundle having narrow, thin-walled cells (the *bast*-vessels). These vascular bundles form the *conducting tissue* of the stem and conduct water and food-materials from one part of the plant to another. The wood-vessels carry water upwards from the roots, while the bast-vessels carry manufactured food from the leaves to the roots and to the growing-points of the shoot.

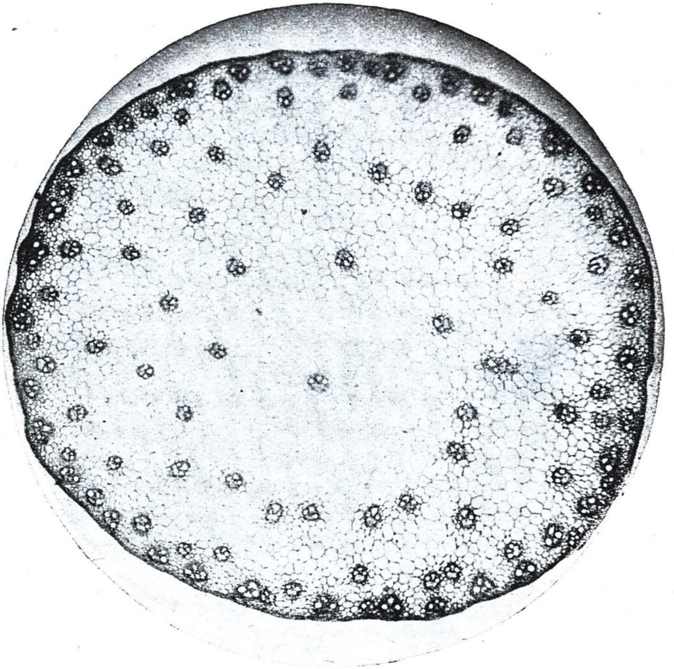


FIG. 77. TRANSVERSE SECTION OF MAIZE STEM (a mono-cotyledon), showing scattered vascular bundles.

Separating the wood from the bast is the *cambium*† layer, an actively-growing tissue. In woody di-cotyledons, as the stem grows in thickness, the *cambium* forms new wood-cells on the inside of each vascular bundle, and new bast-cells on the outside until a continuous layer of bast is formed, surrounding a complete layer of wood: i.e. the wood and bast form two tubes, one inside the other, with *pith* filling up the region inside the wood (Fig. 75).

The young wood-vessels, known as *sap-wood*, continue to conduct water, while the older wood-cells die and form the *heart-wood*, which gives the strength to a woody stem. This growth of the wood inwards squeezes up the *pith* more and more until, in an old tree, the pith can no longer be seen when the stem is cut across.

Owing to this addition of layer after layer of wood, it is sometimes possible to estimate the age of a tree by counting the *growth rings* on the cut surface of a tree-trunk (Fig. 76). In the wet season, the wood grows rapidly and the vessels are thin-walled. In the dry season, the wood grows very slowly and the cell-walls are thicker. In this way, *growth-rings* are often seen on a cross-section of a tree-trunk, each ring representing the growth which took place in one wet season and one dry season.

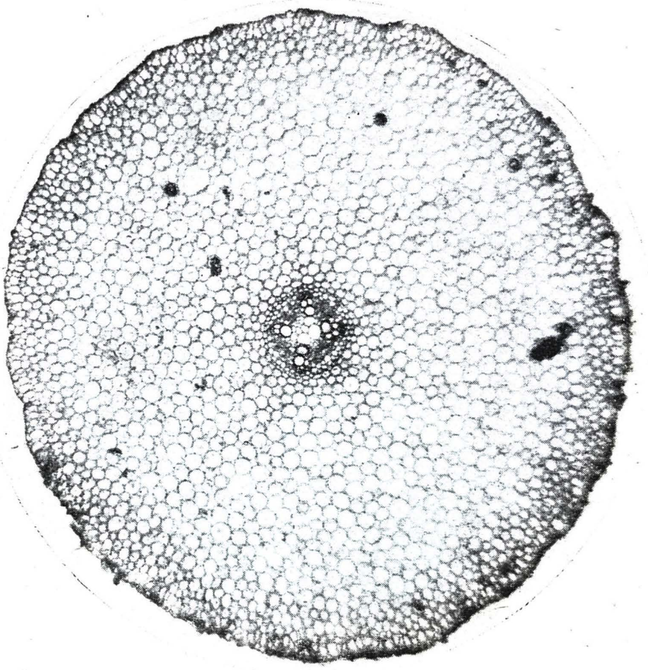


FIG. 78. TRANSVERSE SECTION OF YOUNG ROOT OF DI-COTYLEDON, showing central cylinder.

In temperate climates, each ring usually represents one year's growth, but in tropical climates with more than one well-marked wet season each year, there may be more than one ring formed each year, or growth may be so uniform that no distinct rings are visible.

In the stems of mono-cotyledons, no complete cylinders of wood and bast are formed, the vascular bundles remaining separate, so that the stems of mono-cotyledon trees are usually fibrous, and the timber* is of little use for wood-work. The vascular bundles of mono-cotyledon stems have *no cambium*, hence there is no growth in thickness;

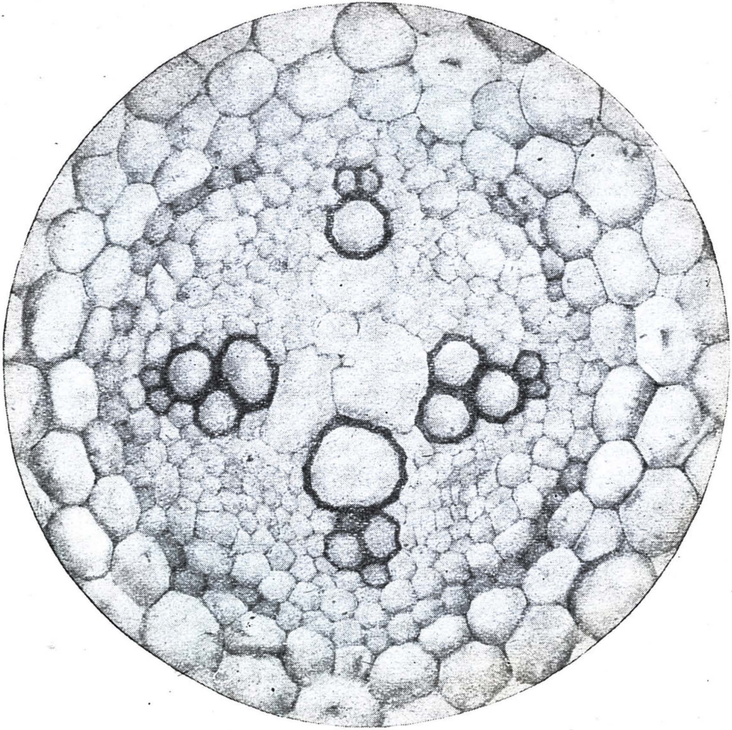


FIG. 79. MORE HIGHLY MAGNIFIED TRANSVERSE SECTION OF CENTRAL CYLINDER FROM FIG. 78,
showing bundles of wood-vessels (thick-walled) alternating with bundles of bast-vessels (thin-walled).

and the stems of such plants, e.g. Palms, are of practically the same diameter throughout their length (Fig. 77).

ROOT STRUCTURE

Although the structure of an old, thickened root is like that of an old, thickened stem, a young root has a different internal structure from that of a young stem (Fig. 78). In young roots there are separate bundles of wood- and bast-fibres, arranged alternately to form a circle; but in the root these fibrous bundles are nearer the centre than they are in the di-cotyledon stem, thus they form a tough, *central cylinder* (Fig. 79). This is the best arrangement of the strengthening tissue for standing against the *pulling strain** which a root has to meet. (Stems usually have to stand a *bending* strain, and against this the best arrangement is to have the strengthening tissue nearer the *outside* of the stem.)

Where the root joins the stem, the wood-vessels of the root join up with those of the stem, while the bast-vessels of root and stem are also continuous with each other. (Examine leaf, stem, and root sections under the microscope.)

CHAPTER VII

EXCRETION IN PLANTS AND ANIMALS

When a motor-engine is working, it converts the chemical energy of the fuel (petrol) into heat energy and energy of movement. There are also waste products—a little unburnt fuel, carbon dioxide and steam in the exhaust gases, and also the tiny particles of steel which gradually wear away from the moving parts as they rub against each other. There are similar waste products in living things. In higher animals, a small proportion of the food always passes right through the food-canal without being digested and absorbed, and this is passed out as faeces. There are also the waste-products—mainly compounds of nitrogen—formed by the wear and tear of the body-cells, together with the carbon dioxide and water formed in the living cells during respiration. We have seen that the term *excretion* is applied only to the getting rid of these *internal waste-products from the body-cells*. The solid, undigested food-remains, passed out as faeces, have never been inside the body-cells but have been *rejected** by the body and not ' excreted '.

In plants, respiration yields the same waste products, carbon dioxide and water, but there is very little nitrogenous waste in plants. Plants also get rid of excess oxygen (formed as a result of photo-synthesis) and excess water (during transpiration).

EXCRETION IN ANIMALS

In tiny, one-celled animals like Amoeba and Paramecium, waste products can be excreted simply *by diffusion* through the surface layer, or they can be squeezed out by contractions of the protoplasm. An earth-worm has special excretory organs in each segment of its body—fine, coiled tubes opening through the body-wall or leading into the food-canal. Body-fluid, containing waste products, can pass along these ' excretory tubes ' from the body-cavity to the outside, either through the skin or by way of the food-canal. Part of each ' excretory tube ' is surrounded by a network of capillary blood-vessels, so that waste products are absorbed from the blood and are passed out through the ' excretory tube '.

In insects (e.g. the Cockroach), the hind gut bears many yellow thread-like excretory tubes which remove waste nitrogenous matter from the blood and discharge it into the hinder part of the food-canal. Insects use some of their excreted nitrogen compounds to build up their horny, external skeleton, which is composed of *chitin*,† a protein material.

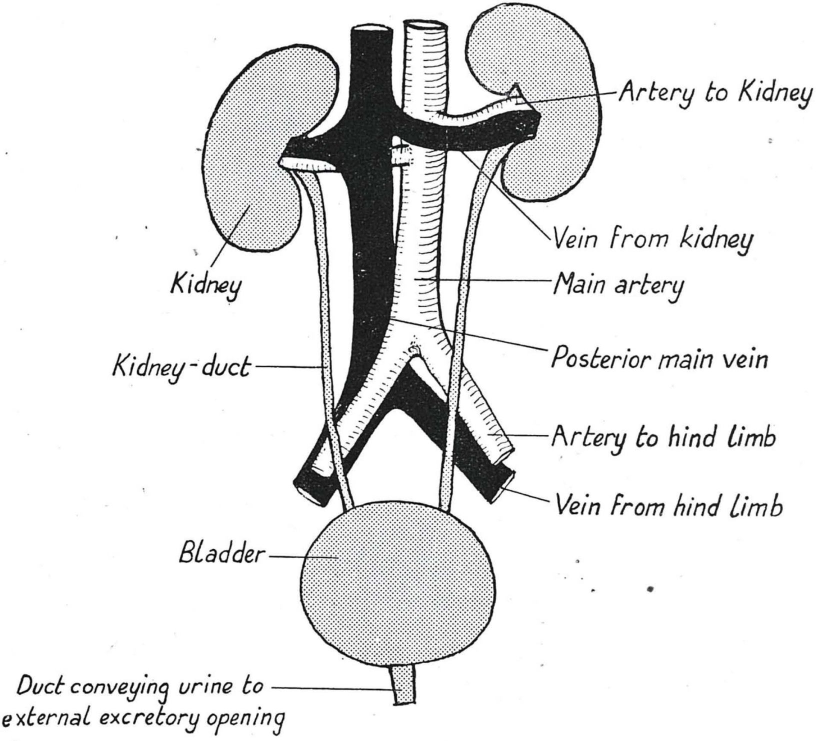


FIG. 80. HUMAN KIDNEYS AND THEIR BLOOD SUPPLY (diagrammatic)

In the Snail, the excretory organ or ‘ kidney ’ is a triangular body in the roof of the mantle-cavity, near the heart. It consists of a hollow sac with spongy walls which are well supplied with capillary blood-vessels from which nitrogenous waste products are absorbed and discharged to the outside by a narrow ‘ kidney-duct ’ running alongside the large intestine.

In all backboned animals, e.g. amphibians, fishes, birds and mammals, the excretory organs are the *kidneys*, consisting of large numbers of ' kidney-tubes ' packed closely together and surrounded by a network of capillary blood-vessels. We shall take the human kidney as our example (a sheep's or goat's kidney will serve for dissection). The kidney of a mammal is a bean-shaped gland attached to the dorsal wall of the belly-cavity. It is richly supplied with blood by a large artery leading straight from the main dorsal artery, and a large vein returns the blood to the main posterior vein. A narrow kidney-duct conveys urine from each kidney to the bladder, where it is stored and discharged at intervals. Inside the kidney the arteries branch and branch again until they form a network of capillaries surrounding the kidney-tubes. The kidney-tubes absorb *urea* (nitrogenous waste prepared for excretion by the liver) and some mineral salts dissolved in water. This solution is known as *urine*.

Besides getting rid of nitrogenous waste, the kidneys also regulate the composition of the blood. Soon after we have drunk a lot of water, the blood becomes more dilute, but the kidneys at once adjust this by removing the excess of water and excreting a larger quantity of more dilute urine. Conversely, after we have eaten a lot of salt, the percentage of salt in the blood rises, but the kidneys at once begin removing more salt and less water, i.e. excreting a more concentrated urine, until the composition of the blood returns to normal again. Similarly, when sweating is rapid and the blood is giving up water through the skin, less water is absorbed by the kidneys and a smaller volume of more concentrated urine is excreted. Under cold conditions, when very little water is being lost as sweat, the kidneys absorb more water from the blood and excrete a larger volume of more dilute urine.

In reptiles and birds, water is absorbed from the urine on its way from the kidneys, and these animals produce a semi-solid, white excretion, which is mainly *uric acid*.†

In all backboned animals the *liver* is an important excretory organ, in addition to its other functions. When proteins are used as a source of energy, ammonia is produced by the decomposition of amino-acids in the liver. This poisonous ammonia combines at once with carbon dioxide and water to form *urea*, a harmless, neutral substance which dissolves in the blood and is carried to the kidneys for excretion.

EXAMINATION OF KIDNEY

Examine the kidney of a large mammal (e.g. sheep, ox, or goat). Identify (*a*) the artery bringing blood to the kidney, (*b*) the vein taking blood from the kidney, (*c*) the kidney-duct carrying urine from the kidney. With scissors, cut open the kidney-duct and trace it up to where it enlarges into a funnel-shaped cavity inside the kidney. Note the conical projections inside this cavity—this is where the kidney-tubes discharge their urine into the cavity. Cut open the kidney longitudinally into two halves and notice that there are two different layers. The outer, dark-brown layer consists of the coiled ends of the kidney-tubes and the closed, outer end of each tube is in close contact with a bunch of capillary blood-vessels. The inner layer is of a lighter colour and appears striped. This layer consists of the straight portions of the kidney-tubes which collect the urine and discharge it into the central cavity from which it is drained to the bladder by the kidney-duct.

In Man and some other mammals, the skin plays a small part in excretion, although the function of sweating is not primarily the getting rid of waste matter but the regulation of body-temperature. Sweat, however, is not pure water, it contains small quantities of mineral salts (as shown by its taste), acids (as shown by its action on litmus) and urea. Although there is only a very small quantity of urea in sweat, it gives a smell to sweaty clothes when it decomposes.

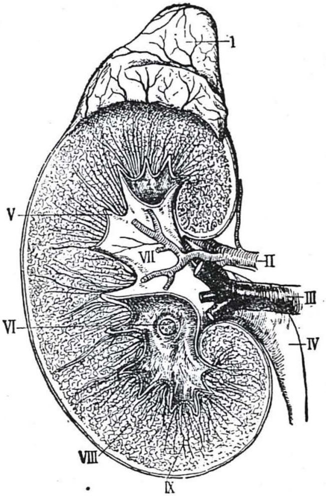


FIG. 81. SECTION OF KIDNEY. I—adrenal body. II—artery to kidney. III—vein for kidney. IV—kidney-duct. V—cavity of kidney. VI — conical projection. VII — branches of kidney-artery. VIII—outer layer. IX—kidney-tubes.

EXCRETION IN GREEN PLANTS

Plants have to get rid of the waste products of respiration (carbon dioxide and water), photo-synthesis (oxygen) and excess water absorbed by the roots. Plants are much less active than animals and they produce so little nitrogenous waste that we need hardly mention this side of excretion which we have seen to be so important in animals.

The carbon dioxide produced during respiration can be got rid of, from all parts of a plant, through the lenticels and stomates. In green parts of plants, in the light, the carbon dioxide formed during respiration is used for photo-synthesis. In the same way, some of the oxygen produced during photo-synthesis is used for respiration, and only the excess is excreted through the stomates.

Since roots can only absorb the essential mineral salts in very dilute solution, a plant has to take in much more water than it requires to keep its cells turgid and for synthesizing carbohydrates. This excess water is given off as water-vapour from the leaves during transpiration.

Excretion in plants, therefore, is a much simpler subject than excretion in animals.

CHAPTER VIII

MOVEMENT IN PLANTS AND ANIMALS

Movement is a characteristic of all living things although, since most plant movements are very slow, it is much more noticeable in animals than in plants.

MOVEMENT IN PLANTS

We have seen that most animals have to move rapidly in order to capture their food and to escape from their enemies. Plants do not need to move from one place to another to get food. Their leaves are spread out to receive as much light as possible, and they are always surrounded by air containing carbon dioxide. Their roots spread through the soil and absorb water containing dissolved mineral salts, so that the ‘ higher ’ plant is firmly rooted in one spot. (Some of the simplest, one-celled plants, however, swim about freely in water.)

In some of our earliest lessons on plants we learnt that plants respond to the stimuli of *gravity*, *light* and *water*.

MOVEMENTS IN RESPONSE TO GRAVITY

We have seen that when a seedling is placed in damp air, with its main root and shoot horizontal, the tip of the root turns vertically

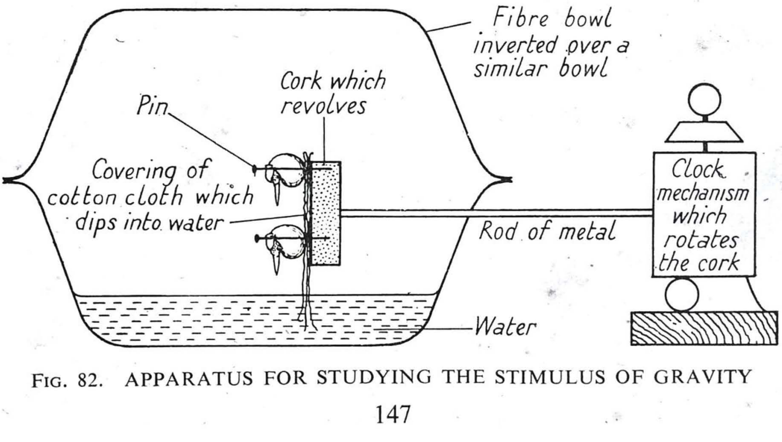


FIG. 82. APPARATUS FOR STUDYING THE STIMULUS OF GRAVITY

downwards and the whole of the shoot bends upwards. If the experiment is repeated with seedlings from which the tips of the root and shoot have been cut off, there is no bending. This shows that the stimulus of gravity affects only the tips of the root and shoot.

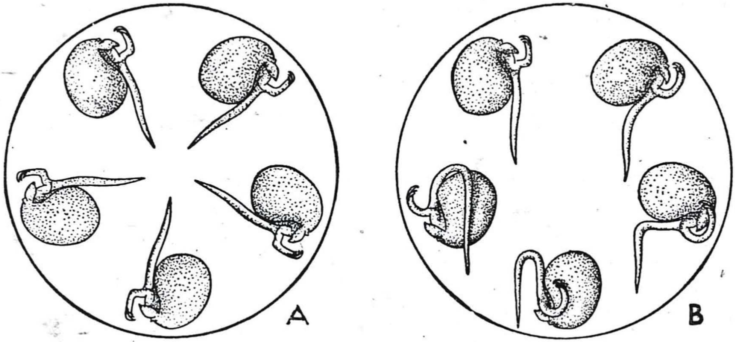


FIG. 83. RESULTS OF EXPERIMENTS USING APPARATUS SHOWN IN FIG. 82

We cannot set up a ' control ' experiment in which there is no force of gravity acting on the plant, but we can make gravity act equally in every direction by fastening seedlings to a vertical disc which can be

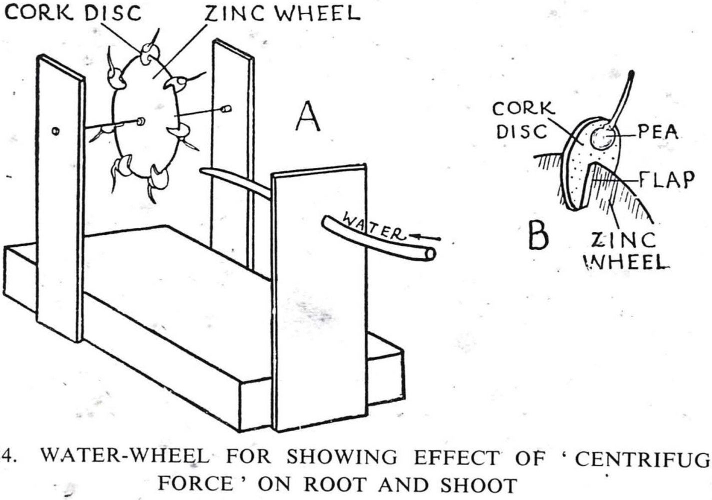


FIG. 84. WATER-WHEEL FOR SHOWING EFFECT OF ' CENTRIFUGAL FORCE ' ON ROOT AND SHOOT

turned by a clock, as shown in Fig. 82. When the disc revolves slowly (e.g. once every hour) the direction of the force of gravity relative to the seedlings is continually changing and is thus neutralized, as it were, and the seedlings continue to grow in the direction in which they were placed at the beginning of the experiment (Fig. 83 (*a*)). In a control experiment with seedlings fastened to a similar vertical disc which is *not turned*, it is found that, in response to the force of gravity, the root turns downwards and the shoot turns upwards (Fig. 83 (*b*)).

If a similar experiment is carried out with seedlings fastened to a *rapidly revolving* vertical disc (turned by a small water-turbine or electric motor), a *centrifugal*† *force* (see p. 148) is set up which acts in the same way as the force of gravity, and the roots grow outwards and the shoots inwards towards the centre of the disc (Fig. 84). If the experiment is repeated with the rapidly revolving disc in the horizontal plane, the centrifugal force acts outwards and the force of gravity downwards. The seedlings respond to both forces, so that the roots grow outwards and downwards and the shoots inwards and upwards.

FIG. 85. LEAF-MOSAICS

PLANT MOVEMENTS IN RESPONSE TO LIGHT

In earlier lessons we saw that *stems grow towards the light*, that roots usually grow away from the light, and that leaves usually arrange themselves at right angles to the direction of the light so as to catch as much light as possible. These movements are clearly to the plant's advantage: an exposed root, in growing away from the light, buries itself in the soil; while the stem, in growing towards the light, brings the leaves into the most favourable position for photo-synthesis. In addition, the leaves usually arrange themselves so that they do not overlap and shade one another. This arrangement of the leaves is called a *leaf-mosaic*. When a plant receives light from one side only, the directive influence of light on the stem may be as strong as that of gravity.

We have seen that it is only the tips of the root and shoot which are sensitive to gravity. In the same way, it is only the tip of the shoot which is sensitive to light. For example, if seedlings are placed in a box with light entering through a small hole in one side of the box, their stems bend towards the light, but *if the tip of the stem has been cut off there is no bending*. Similarly, *if the tip of the stem has been shaded* with a tiny light-proof cap *there is no bending towards the light*.

FIG. 86. LEAF-MOVEMENTS OF SENSITIVE PLANT

It is interesting to note that although only the tip of the stem is sensitive to light, the actual bending takes place some distance from the tip. A *growth-controlling substance* or *plant hormone*† is produced by the tip of the stem and passes down the stem, causing more rapid growth on the side of the stem away from the light. This results in the bending of the stem. We shall discuss *hormones*, or *chemical messengers*, in more detail later.

PLANT MOVEMENTS IN RESPONSE TO WATER

In our earlier lessons we have seen that roots grow towards water and that the influence of water may be greater than that of gravity, so that the roots always grow towards the damper parts of the soil.

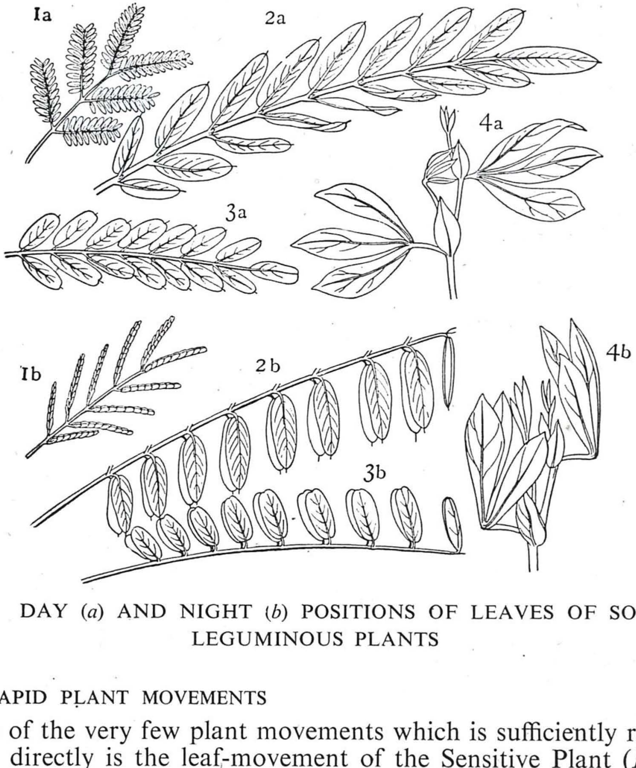


FIG. 87. DAY (*a*) AND NIGHT (*b*) POSITIONS OF LEAVES OF SOME LEGUMINOUS PLANTS

MORE RAPID PLANT MOVEMENTS

One of the very few plant movements which is sufficiently rapid to be seen directly is the leaf-movement of the Sensitive Plant (*Mimosa pudica*). If one leaflet is touched, the neighbouring leaflets close up in pairs, the movement travelling down the leaf-stalk until the whole leaf closes up and hangs down. Similar ' sleep-movements ' take place in Mimosa and some other leguminous plants, e.g. the Rain Tree† (*Saman*), as darkness falls.

Many flowers show fairly rapid movements, opening in the morning and closing up later in the day, e.g. Morning Glory and ' Nine-o'Clock Flower '† (*Turnera*).†

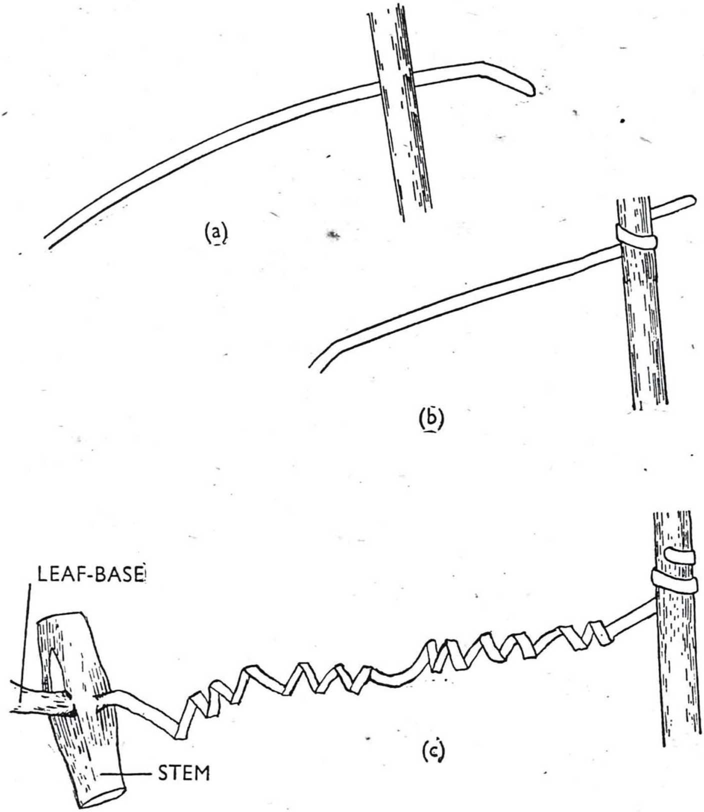


FIG. 88. STAGES IN THE TWINING OF A TENDRIL

(*a*) Tendril stimulated by contact.
(*b*) Tendril beginning to react to the stimulus.
(*c*) The reaction completed. The whole tendril has responded.

Plants which climb by twining or by tendrils show a fairly rapid response to the stimulus of *touch.* Tendrils and the tips of twining stems circle slowly round in search of a support. When they touch a *solid object*, rapid growth takes place on the side *away* from the point of contact, so that the stem or tendril twines round the support. After a tendril has grasped a support, the whole tendril coils up from either end, forming a double spring. This pulls the stem nearer to the support and also reduces the danger of breakage.

MOVEMENT IN ANIMALS

Just as there are some free-swimming plants, there are some exceptional animals which remain fastened to one spot, e.g. Sponges and Corals. By far the greater number of animals, however, can move from place to place in search of food and to escape from their enemies. This movement is brought about in various ways.

In *Amoeba*, part of the protoplasm flows out to form a rounded projection in the direction in which the animal is moving. The rest of the protoplasm then flows in the same direction, so that the cell moves forwards with a rolling motion over the surface on which it rests.

In *Paramecium*, another one-celled animal, the whole surface is covered with tiny protoplasmic threads, called *cilia*, which strike the water something like the oars of a boat and so drive the animal through the water. As the cilia are arranged in spiral rows, they give Paramecium its characteristic spiral movement.

In *Hydra*, which consists of a large number of cells arranged in two layers to form a hollow sac, some of these cells have ' muscle-tails ' which can contract. The ' muscle-tails ' of the outer layer of cells run lengthwise up and down the animal, while the ' muscle-tails ' of the inner layer run round the body. When the longitudinal ' muscle-tails ' of the outer layer contract, the animal shortens in length, and when the circular ' muscle-tails ' of the inner layer contract the animal becomes long and thin. Similar contractions produce the movements of the hollow threads (tentacles) surrounding the mouth.

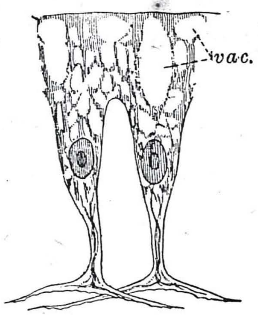


FIG. 89. TWO HYDRA CELLS WITH ' MUSCLE-TAILS '
vac.—cell-spaces

Although Hydra spends most of its life attached to one spot, e.g.

to a piece of water-weed, it is able to move from place to place. It can change its position in different ways. Slight changes of position are carried out by slowly *sliding* along on its base. More rapid movements are carried out by ‘*looping*’ as shown in Fig. 90 (*a*): the animal bends over, fastens its ‘mouth’ to a spot a little distance away from its base, then draws its base towards its ‘mouth’. The animal then repeats the process. Still more rapid movements are carried out by *turning* ‘*head-over-heels*’ as shown in Fig. 90 (*b*).

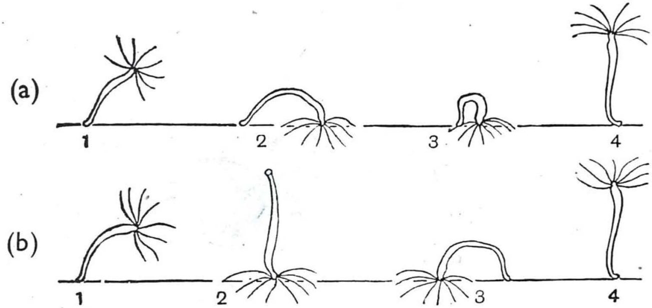


FIG. 90. MOVEMENTS OF HYDRA
(*a*) ‘looping’. (*b*) turning ‘head-over-heels’.

In the *earth-worm*, the body-wall consists mainly of two layers of muscles—an outer layer of *circular muscle-fibres* running round the animal and an inner layer of *longitudinal muscle-fibres* running up and down the animal from head to tail. The earth-worm's body-cavity is filled with liquid, and the pressure of this liquid keeps the body-wall stretched and firm (like the turgid plant-cell). Rings of tiny bristles project from each segment. The earth-worm moves by alternate

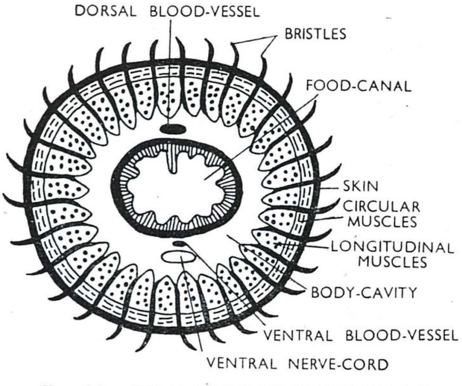


FIG. 91. TRANSVERSE SECTION OF EARTH-WORM (diagrammatic)

contractions of its two layers of muscles. When the circular muscles of the hind part contract, the diameter of this part of the body is reduced and the liquid in the body-cavity is compressed, thus pushing the front part of the animal forwards. The bristles at the head end are then pressed into the soil, while the longitudinal muscles contract and draw the hind part forward. The bristles at the tail end are then pressed into the soil and the whole process is repeated.

The *snail* has a remarkable gliding* motion. The 'foot' is a mass of soft muscle, and alternate waves of contraction and relaxation run forward through these muscles, carrying the animal forwards along its slimy track.

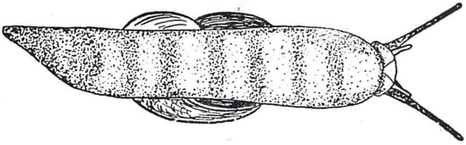


FIG. 92. VENTRAL VIEW OF SNAIL MOVING ON A SHEET OF GLASS, SHOWING UNDER-SURFACE OF 'FOOT' AND WAVES OF MUSCULAR CONTRACTION

SKELETONS AND MOVEMENT

The animals whose movements we have studied so far have been comparatively small, 'soft-bodied' animals, and their movements are comparatively slow. For more efficient and more rapid movements the muscles must be attached to rigid supports, and *muscle-attachment* is one of the main functions of a *skeleton.* In larger animals, too, the soft parts of the body need support, and the skeleton, besides serving to protect delicate parts of the body, provides a rigid framework. Since the study of movement in the higher animals requires some knowledge of the skeleton to which the muscles of movement are attached, this is a convenient point at which to discuss skeletons in general—in plants as well as in animals.

PLANT SKELETONS

We shall deal first with plant skeletons because they are much simpler than those of animals—plants do not have to move about. We have seen that in the young green parts of plants, the cells are kept *turgid* by the outward pressure of the cell-sap on the cell-wall. If these young cells lose water they become soft, and the green parts of the plant *wilt.*

In the older parts of plants, turgidity is no longer the only means of support, and special *strengthening tissue* is formed. The cellulose cell-wall is converted into *wood*, and such cells finally lose their living contents (although they may still serve to carry water). Some of these cells are long and pointed so that they interlock and form tough *fibres*. We have seen such strengthening tissue in the *leaf-skeleton* left behind after the soft parts of a dead leaf have decayed. We have also seen how this strengthening tissue—wood—is arranged in stems and roots. In young di-cotyledon stems, the wood is near the outside of the stem, where it will offer most resistance to a *bending strain*. As di-cotyledon stems get older, the wood first forms a hollow cylinder with soft *pith* inside, and then the wood gradually extends inwards until the greater part of the stem is solid ' heart-wood '. In mono-cotyledon stems, the strengthening fibres are scattered throughout the stem, but they are much more in number near the outside, where their strengthening effect is greatest. Most grasses, in fact, e.g. Bamboo, have hollow stems, making the most efficient use of their strengthening tissue to resist a bending strain. It is interesting to note that engineers, when they want to combine maximum strength with minimum weight of material, use exactly the same arrangement in the hollow posts they use to support heavy loads. (We shall see later that the larger bones of vertebrates are built on the same plan.)

In roots, which are not subjected to a bending strain but only to a pulling strain, the strengthening tissue is arranged as a *central cylinder*. The same principle is adopted in strengthening a thick rope with a wire cable—the stronger material is put in the centre.

ANIMAL SKELETONS

We have seen that the most primitive type of supporting tissue in animals is the *external skeleton* found in many invertebrates. Its simplest form is an unjointed shell. In chalk-forming animals, this shell serves only for protection and support, playing no part in movement. The shell of a snail serves a similar purpose, and the animal escapes from its enemies by drawing itself back into the shell. In the insects, we find an entirely different type of external skeleton—a *jointed skeleton*—where the muscles of movement can pull on the jointed limbs and thus enable the animal to move from place to place. This hollow, jointed, tubular skeleton is very efficient for support and protection, although it is somewhat less efficient for movement; and some of the insects are among the most successful and efficient of all animals. But a dead, inelastic, external skeleton has one very serious disadvantage—*it hinders growth*. We have seen that all insects have to cast off their external skeleton several times during their life history. This skin-casting is often very difficult and dangerous. It is often difficult for the soft-bodied animal to get out of its horny shell, and it is helpless and defenceless until the new shell grows and hardens. We have seen that many of the higher insects, e.g. butterflies, moths and mosquitoes, undergo *complete metamorphosis* during their complicated life history, so that growing and skin-casting take place in the larval stage, before the development of any complicated wings, jointed walking-legs and jointed mouth parts. The perfect winged insect cannot grow in size because it can no longer cast its external skeleton. For this reason, the insects are all relatively small animals.

An insect's external skeleton is composed of *chitin*, a horny, protein substance similar to the material forming human finger-nails and hair.

In crabs, the external skeleton is thickened and hardened by the addition of *calcium carbonate* to the horny *chitin*.

In vertebrates, the supporting tissue is arranged on an entirely different plan: there is an *internal skeleton*. This bony framework is built on the same essential ground-plan in all vertebrates, e.g. fish, frog, bird and mammal, consisting of a *main axis* (*skull*, *backbone*, *ribs*, *breastbone*) and two pairs of *attachments* (*limbs and limb-girdles*) which may be either legs, or arms, or fins, or wings. Such an internal skeleton does not give such complete protection to the soft parts of the body as the external skeleton of an insect, but it offers no hindrance to growth and is better adapted to free body-movements. We have already discussed the general arrangement of the bones in the vertebrate skeleton (Book III, p. 99), but we have still to discuss how movements of bones are brought about by muscles.

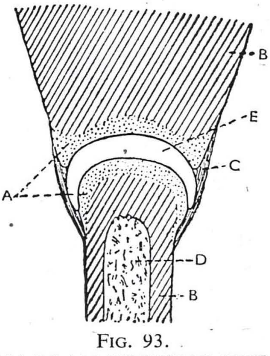


FIG. 93.

DIAGRAM SHOWING THE STRUCTURE OF A JOINT

A—layer of gristle. B—bone. C—tendon. D—bone-marrow. E—sac of gristle enclosing slimy fluid.

Where neighbouring bones, e.g. most of the skull-bones and the bones of the hip-girdle, do not have to move independently they often grow together. Many bones, however, must be able to move freely, and some kind of *joint* is necessary between neighbouring bones. Such a joint must allow the two bones to move freely and without friction, but the bones must be bound together in some way so that they are not readily pulled ‘ out of joint ’ (or *dislocated*†). In joints like the *hip* and the *shoulder*, the rubbing surfaces are polished and rounded, and they are also covered with a thin layer of smooth *gristle* (cartilage†). Besides this, such a joint is enclosed in a close-fitting sac containing a slimy fluid which lubricates the joint. This slimy fluid is formed from the gristle as it wears away and is replaced by a new growth. The two jointed bones are also bound together by strong *tendons* (ligaments† or sinews†). The type of joint which allows the freest movement in all directions is the ‘ *ball-and-socket* ’ type which joins the limbs to the limb-girdles (hip- and shoulder-joints). ‘ *Hinge-joints* ’ (e.g. those of the elbow, knee, jaw and finger) allow movement in one plane only.

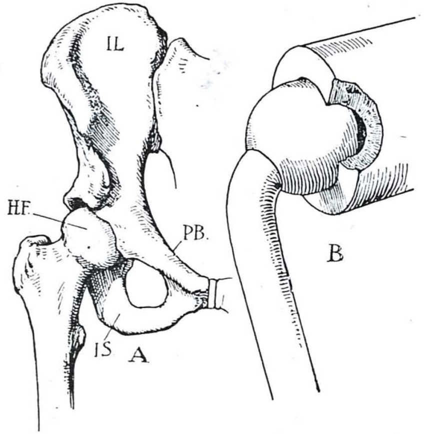


FIG. 94. (*a*) THE HUMAN HIP-JOINT. (*b*) A SIMPLE BALL-AND-SOCKET JOINT
HF—rounded head of thigh-bone.
IL, PB, IS—fused bones of hip-girdle.

The joints between the individual bones (*vertebrae*) of the backbone are so firmly bound together by tendons that very little movement can take place, since the main function of the backbone is to provide strength and support.

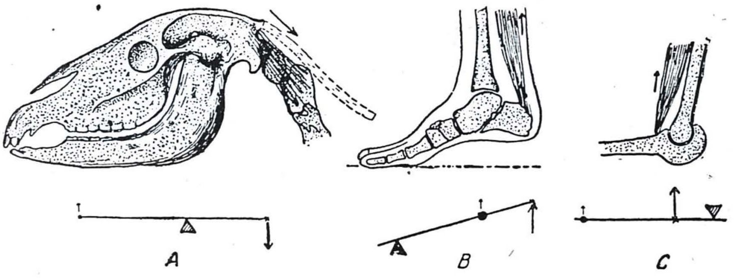


FIG. 95. THREE TYPES OF ANIMAL LEVERS
A—head of horse (Class I). B—human foot (Class II). C—human elbow-joint (Class III).

Every muscle of movement is attached to at least two jointed bones, and nearly all these muscles are arranged in pairs—one muscle bringing the jointed bones closer together when it contracts and the other muscle pulling the bones back to their original position once more. When we discuss the *principle of the lever* in other lessons we shall discuss the mechanical principles of the various systems of bony levers found in vertebrate skeletons.

We have seen that the internal skeleton of vertebrates gives less effective protection than the external skeleton of insects. In most vertebrates, however, additional protection is given by a ‘ *skin skeleton* ’ covering the body. We have seen that fishes and reptiles have their skin covered with *scales*; birds have *feathers*; mammals have *hair*. All these are external protective coverings. In primitive, gristly fishes, e.g. sharks, the skin is covered with thousands of tiny scales. Each scale arises from a flat, bony base embedded in the skin and consists of a backward-pointing spine of *dentine* (ivory), covered with a layer of *enamel* (the hardest of all tissues). There is a *pulp-cavity* in the centre of the spine. These scales of primitive fishes are of exactly the same structure as the animals' *teeth*. They are very much like our own teeth.

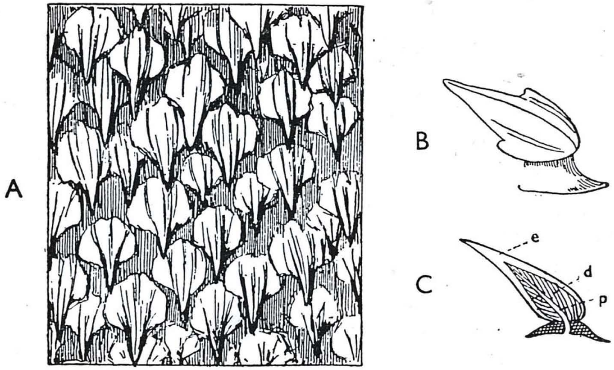


FIG. 96. SCALES OF SHARK
A—portion of skin as seen with hand-lens. B—single scale (magnified). C—section of scale: d—dentine; e—enamel; p—pulp-cavity.

The overlapping scales of bony fishes, e.g. carp, are of different structure and consist of thin, flexible, translucent plates covered by the outer layer of skin. They often show *growth-rings* which can be used to tell the age of the fish.

The horny scales of reptiles and the scales on the legs of birds are different from those of either gristly or bony fishes.

We have already discussed *feathers* and *hair*, two other coverings which develop from the ‘ skin skeleton ’.

TEETH

In vertebrates, the lining of the mouth is essentially an in-folding of the skin. In primitive gristly fishes, the lining of the mouth is covered with the same spiny scales as the outer surface of the body, and in the region of the jaws these spiny scales grow very big and become *teeth.* In gristly fishes, as one row of these teeth wears out, another row grows forward over the jaw to take the place of the old teeth. Such sharp, backward-pointing teeth are primarily for seizing and holding the food and not for chewing, and this is their main function in amphibians and reptiles. Teeth are never present in birds.

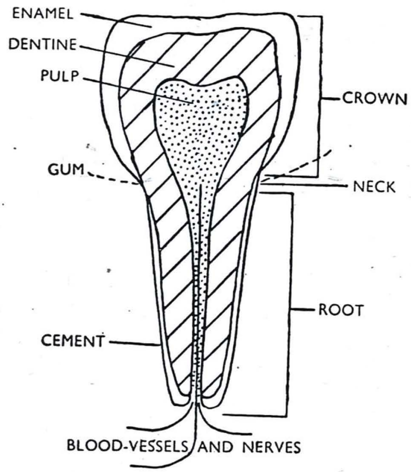


FIG. 97. SECTION OF HUMAN TOOTH

We have seen that one important characteristic of mammals is their *specialized teeth.* The teeth of all mammals are built on the same ground-plan, although they may differ widely in size and shape according to the animal’s feeding habits. That part of the tooth which projects into the mouth is called the *crown*, while the part embedded in the jaw is called the *root.* The region level with the *gum*, where the crown joins the root, is called the *neck* of the tooth. The greater part of a tooth consists of *dentine* (or ivory†). Dentine, although it is very strong, is *elastic*, so that it can withstand sudden pressure without breaking. Dentine, however, is too soft to stand continuous wear, and the crown of the tooth is covered with a layer of very hard *enamel.* Enamel is the hardest of all tissues, and in fully-grown teeth it consists almost entirely of mineral (or inorganic) matter. Although the enamel is brittle, it is saved from breakage because it rests on the elastic dentine

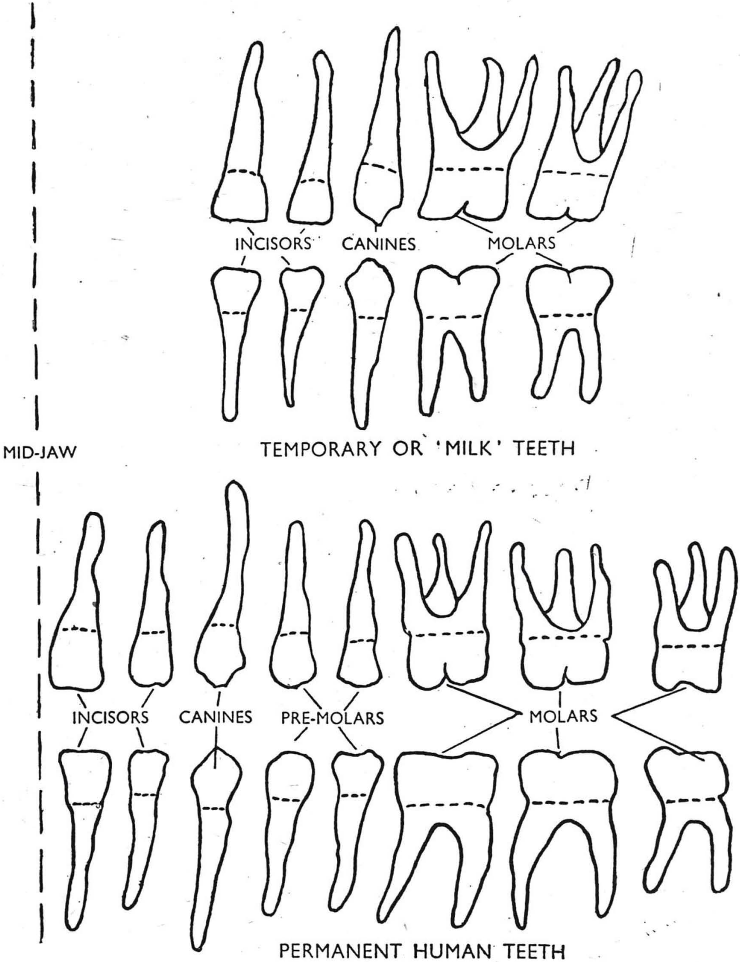


FIG. 98.
TEMPORARY OR 'MILK' TEETH
PERMANENT HUMAN TEETH

beneath. The enamel extends down to the neck of the tooth. The dentine of the root is covered with a thin layer of bone-like *cement* which helps to fix the tooth in its socket in the jaw. In the middle of the dentine is the *pulp-cavity*, containing a soft mass of blood-vessels and nerve-fibres, which enter through a tiny hole at the tip of the root.

We have seen that mammals have different kinds of teeth, specially adapted for different work. In *Man*, there are four kinds of teeth: (*a*) *incisor teeth*—flattened, cutting teeth for biting. These are the four front teeth in each jaw. The incisors are well developed in gnawing mammals, e.g. rats, mice, cavies, squirrels. (*b*) *canine*[†] *teeth*—more pointed and stronger, two in each jaw, on either side of the incisors. These canine teeth are best developed in flesh-eating mammals, e.g. cats, dogs, tigers, being used for tearing up the food. (*c*) *pre-molar*[†] teeth—four small, grinding teeth, placed in pairs behind each canine tooth. The pre-molars are present only in the permanent human teeth and do not occur in the temporary ' milk teeth '. (*d*) *molar teeth*—six large teeth for grinding up the food and mixing it with saliva: three on either side of each jaw. (In the temporary ' milk teeth ' there are only four molars in each jaw.) The molars and pre-molars (or ' double ' teeth) are well developed in grass-eating mammals, e.g. sheep, goats, cattle.

Man grows two sets of teeth—a *temporary set* during childhood and a *permanent set* later on. When a baby is about 7 months old, the first front teeth appear. Other teeth appear gradually until by the time the child is about 2 years of age, the *temporary set* is complete, consisting of 4 incisors, 2 canines and 4 molars in each jaw. The ' milk teeth ' serve the child until it is about 6 years old. During this time the jaws are growing longer, thus making room for the *permanent teeth* which are gradually developing inside the jaws. At about 6 years of age, the child's first permanent molars appear, behind the temporary molars; and during the next few years the temporary teeth loosen and fall out, being replaced by the permanent teeth, until by the 15th year only the 3rd molars (or ' wisdom teeth ') have still to appear. The complete set of permanent teeth, therefore, consists of 4 incisors, 2 canines, 4 pre-molars, and 6 molars in each jaw.

In order to summarize the number and kind of teeth, a *dental formula* is used. The dental formula for Man's permanent teeth is:—

$$2\left\{\begin{matrix}2 & 1 & 2 & 3\\ \mathrm{I} & \mathrm{C} & \mathrm{PM} & \mathrm{M}\\ 2 & 1 & 2 & 3\end{matrix}\right\} \text{ or more simply still } 2\left\{\frac{2\ 1\ 2\ 3}{2\ 1\ 2\ 3}\right\} \text{ Total 32}$$

The dental formula for the human temporary teeth is:—

$$2\left\{\begin{matrix}2 & 1 & 0 & 2\\ \mathrm{I} & \mathrm{C} & \mathrm{PM} & \mathrm{M}\\ 2 & 1 & 0 & 2\end{matrix}\right\} \quad \text{or } 2\left\{\frac{2\ 1\ 0\ 2}{2\ 1\ 0\ 2}\right\} \quad \text{Total 20}$$

Other examples are:—

$$\text{Dog} \quad .. \quad 2\left\{\frac{3\ 1\ 4\ 2}{3\ 1\ 4\ 3}\right\} \quad \text{Total 42}$$

$$\text{Cat} \quad .. \quad 2\left\{\frac{3\ 1\ 3\ 1}{3\ 1\ 2\ 1}\right\} \quad \text{Total 30}$$

$$\text{Rat} \quad .. \quad 2\left\{\frac{1\ 0\ 0\ 3}{1\ 0\ 0\ 3}\right\} \quad \text{Total 16}$$

$$\text{Elephant} \quad .. \quad 2\left\{\frac{1\ 0\ 0\ 6}{0\ 0\ 0\ 6}\right\} \quad \text{Total 26}$$

(It is interesting to note that the elephant's teeth are all molars except for 2 incisors—the *tusks*†—but these molars are so big that there is only room in the jaw for two on each side at the same time. When any of these molars is lost, the others grow to take its place.)

Examine any skulls which are available, and note the different types of teeth and their arrangement. Construct their dental formulae.

MOVEMENT IN ANIMALS (*continued*)

The chief ways in which vertebrate animals move are (*a*) by *swimming*, (*b*) by *walking*, and (*c*) by *flying*.

We have seen that fishes move by waves of muscle-contraction passing alternately along either side of the body, from head to tail. The paired fins are used for balancing and steering, while the unpaired fins serve to keep the fish upright in the water. Snakes move along the ground by a wavy motion of the backbone which is very similar to the swimming movements of a fish. In some reptiles, e.g. lizards, this movement is assisted by limbs.

In most animals, *walking* involves the use of pairs of jointed walking-limbs. Most vertebrates walk on four limbs whose jointed bones act as levers for the pull of the muscles, and two limbs support the weight of the body while the other two are lifted off the ground. In *running*, the animal raises itself on its *toes*; and in very fast running, all the feet may leave the ground for a short period between each step. Jumping animals, e.g. grasshoppers, frogs, kangaroos, have exceptionally long and powerful hind legs.

In *flying*, the wings make upward and downward movements very similar to our own movements in swimming with the breast-stroke. During the down-stroke of the wings, the feathers offer their maximum resistance to the air, and the bird lifts itself upwards. During the up-stroke, the wing-feathers offer less resistance to the air which is allowed to flow between them. Another type of flight is *gliding*, when the bird keeps its outstretched wings still and gradually descends in the air, like an aeroplane with its engine stopped.

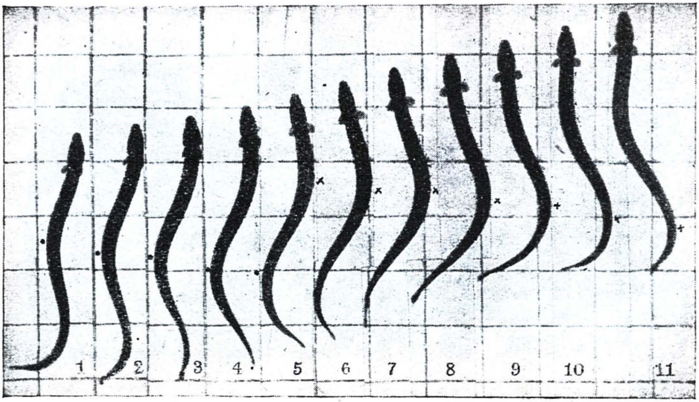


FIG. 99. MOVEMENTS OF A FISH IN SWIMMING

Many birds make use of upward currents of air to sail about for long periods without beating their wings.

MUSCLES AND THEIR ACTION

In all animals except the lowest, movement is brought about by *muscle-cells*, whose special function is *contraction*. Large groups of these muscle-cells make up muscular tissue or *muscles*.

Muscle-tissue is of three main kinds:—(*a*) *voluntary muscle*, (*b*) *involuntary muscle*, and (*c*) *heart-muscle*.

Voluntary muscles are sometimes called *skeletal muscles* because they are attached to bones and produce the movements of the skeleton.

Such muscles are usually controlled by the will, hence the name ' voluntary muscle '. Under the microscope, voluntary muscle is seen to consist of fibres which show a cross-striped appearance, hence another name for voluntary muscle is *striped muscle.* Every voluntary muscle consists of a soft, thickened, central portion (shaped like a cigar) with a tendon at either end attaching the muscle to two bones. Muscle-cells contract in length as the result of a *stimulus* received from the central nervous system through a motor-nerve. It is important to realize that when a muscle shortens its length it becomes thicker, so that the *volume of the muscle remains the same.* When we say that ' a muscle contracts ' we mean ' *contracts in length* '; it is better to say that ' *a muscle shortens and thickens* '.

The characteristic movements of voluntary muscle are *short, rapid contractions,* and this ' striped muscle ' is always present in any part of an animal where rapid movement is required, e.g. in a mosquito's wing-muscles which may have to contract and relax more than 500 times a second.

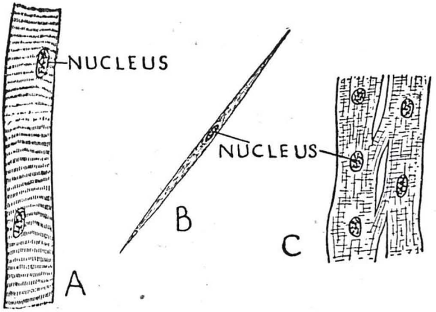


FIG. 100. MUSCLE-FIBRES (magnified) A—part of voluntary (striped) muscle-fibre (×130). B—involuntary (smooth) muscle fibre (×700). C—heart-muscle fibres (×250).

Involuntary muscle is sometimes called *smooth muscle* or *unstriped muscle* because it shows no cross-striped appearance under the microscope. Involuntary muscle is not attached to bones; but it is usually present in muscular sheets lining the internal organs, such as the food-canal and the blood-vessels, i.e. those parts of the body which are not under the control of the will. The characteristic movements of involuntary muscle are *slow, sustained contractions,* hence this ' unstriped muscle ' is useless for producing rapid movements. For example, the unstriped, involuntary muscle of a frog's stomach may take 10 seconds to contract, as compared with 0·05 seconds for the contraction of the striped, voluntary leg-muscle of the same animal.

The *heart-muscle* of vertebrates differs from voluntary and involuntary muscle in its microscopic structure and in being able to sustain fairly rapid movements *without any nervous stimulus.* A frog's heart, for example, can go on beating for a while after removal from the body.

When muscle-cells contract they *do work*, and we know that this requires *energy*. The muscle-cells, therefore, are *energy-transformers*, converting *chemical energy* into *mechanical energy*. The source of this chemical energy is the *glucose* brought to the muscle-cells by the blood-stream and the lymph. This glucose is converted into insoluble *glycogen* and is stored in the muscle-cells until required.

When a muscle contracts, glycogen is converted into *lactic acid*† (the acid which is also present in sour milk), and some chemical energy is set free as energy of movement. This change, however, needs no supply of oxygen, hence no oxygen is required for the actual contraction of muscles, and a muscle can be made to go on contracting over and over again in the absence of oxygen, *up to a point*. But the lactic acid which is produced gradually accumulates in the muscle and the contractions become weaker (or the muscle becomes *fatigued*† or tired). Finally, when sufficient lactic acid has accumulated, the muscle becomes *completely fatigued* and its contractions stop. If oxygen is then supplied to the muscle, some of the lactic acid is oxidized to yield carbon dioxide and water, and energy is set free. This energy is used to re-convert the remaining lactic acid to glycogen, which becomes available for use once more, so that the muscle is said to have *recovered** and can start contracting again.

This is a simple outline of a very complicated chemical process, but it explains why gentle exercise, e.g. walking, does not produce fatigue readily, because the lactic acid produced in the muscles is oxidized as soon as it is formed. Violent exercise, however (e.g. a 220-yards race), produces lactic acid more rapidly than it can be oxidized by the oxygen supplied by the blood, and the muscles become fatigued; until at the end of a race it is necessary to *rest* while the accumulated lactic acid is oxidized.

CHAPTER IX

HOW LIVING THINGS BEHAVE

SENSITIVITY, IRRITABILITY AND CO-ORDINATION

All living things respond to their surroundings, i.e. they are *sensitive* to changes in their surroundings and can adjust themselves to such external stimuli by suitable responses. In order that a *stimulus* shall produce the *response* which will be of most benefit to the living thing, there must be *co-ordination.*† These are the three factors of *behaviour*: (*a*) *sensitivity—the reception of a stimulus*; (*b*) *irritability—the active response to that stimulus*, and (*c*) *co-ordination—the adjustment of the response to the stimulus.*

' PLANT BEHAVIOUR '

Plants, like animals, respond to external stimuli, but as they do not have to move about to seek and capture their food or to escape from their enemies by flight, these responses can be quite slow. Such slow movements can usually be carried out merely by *growth*, and there is no need for the complicated arrangements of sense-organs, the nervous systems, the jointed skeletons and the muscles which we find in most animals. We have seen that the external stimuli which influence plants are *gravity*, *light* and *water*, and these are the main factors in ' plant behaviour '.

ANIMAL BEHAVIOUR

Unlike plants, most animals have to seek and capture their food and they have to be able to escape from their enemies. This usually requires much more sensitive arrangements for receiving external stimuli and more efficient methods to secure rapid response. We shall now trace the advance in sensitivity, response and co-ordination in the animals we have studied.

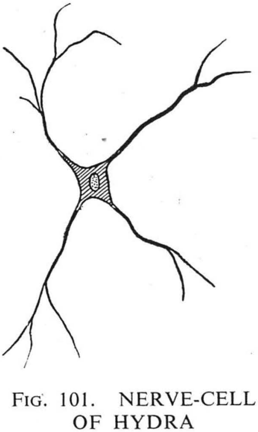


FIG. 101. NERVE-CELL OF HYDRA

In *Amoeba* and *Paramecium*, there are neither sense-organs nor a nervous system. The whole surface of the cell is sensitive to external stimuli and the whole of its protoplasm responds. In *Hydra*, some of the cells are specialized to conduct impulses, and there is a network of nerve-cells which provides for co-ordination between the various body-processes, e.g. movement, food-capture and digestion.

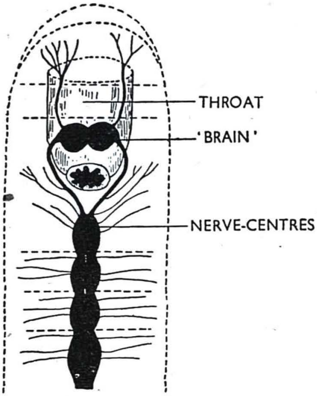


FIG. 102. NERVOUS SYSTEM OF EARTH-WORM (Anterior portion only.)

In the *earth-worm*, the nerve-cells are no longer scattered throughout the animal's body but they are grouped together to form a definite *nervous system*. There is a *ventral nerve-cord* consisting of little masses of nerve-cells (nerve-centres) joined together by nerve-fibres. These nerve-fibres are long, thin extensions of nerve-cells. Other nerve-fibres pass out from the nerve-centres to the various body-organs. The two largest nerve-centres are situated in the earth-worm's head, forming a primitive 'brain'.

The nervous system of *insects* is built on the same plan as that of the earth-worm—a ventral nerve-cord with a nerve-centre (or 'secondary brain') in each segment.

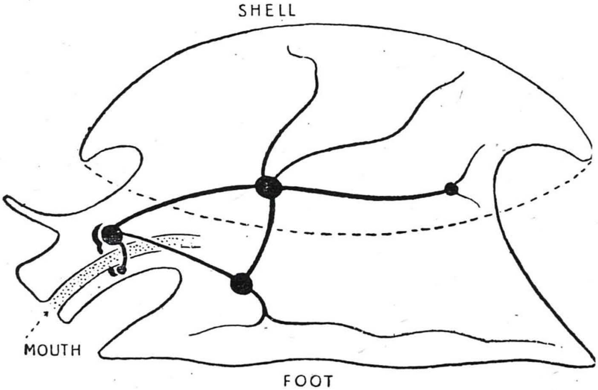


FIG. 103. NERVOUS SYSTEM OF MOLLUSC (diagrammatic)

The nervous system of *Molluscs*, e.g. the snail, is more concentrated, there being three large nerve-centres controlling the head, the digestive organs and the ' foot '.

THE NERVOUS SYSTEM OF VERTEBRATES

The nervous system of vertebrate animals is quite different from that of invertebrate animals. In all vertebrates the *central nervous system* consists of a hollow tube—the *spinal cord*—enlarged at its anterior end to form the *brain.* Unlike the ventral nerve-cord of invertebrates, this central nervous system is dorsal to the food-canal and it is protected by the backbone and the skull. We have already discussed in outline the structure and function of the vertebrate brain (Book III, p. 116).

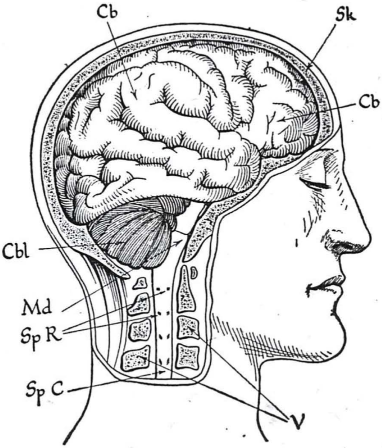


FIG. 104. SECTION THROUGH SKULL AND BACKBONE

Cb—big brain (cerebrum). Cbl—little brain (cerebellum). Md—spinal bulb (medulla). Sk—brain-case. SpC—spinal cord. SpR—roots of spinal nerves. V—backbone.

The *central canal,* which runs from end to end of the spinal cord, opens out into a number of cavities inside the brain, and both the central canal and the brain-cavities are filled with lymph. As a young vertebrate animal develops from the fertilized egg-cell, the spinal cord is the first part of the central nervous system to develop, and then the anterior end of the spinal cord begins to swell out to form the brain. Three swellings are formed in all young Vertebrates —the *fore brain,* the *mid brain* and the *hind brain.* From the primitive *fore brain* develop the two halves of the *big brain* (cerebrum). The more highly developed the big brain, the more intelligent the animal, since this part of the brain is the seat of all voluntary action and conscious sensation. The most characteristic feature of the human brain is the exceptionally large *big brain.* The *mid brain* controls the eye-muscles and it is also concerned with the control of posture,† the way in which the body is carried. The *hind brain* develops into the *spinal bulb* (medulla†) and the *little brain* (cerebellum). The *spinal bulb* controls the unconscious movements of breathing, movements of the digestive organs, heart-beat, and the activity of glands. The *little brain* appears to co-ordinate the action of all the muscles concerned with movement.

We have seen how *nerves* carry impulses between the central nervous system and the various parts of the body.

THE SENSE-ORGANS

An animal's behaviour depends mainly on what is happening in the outside world around it. For example, when an animal sees food it makes the movements necessary to capture it, or if it hears a frightening noise from one direction, it moves away in another direction. A vertebrate animal gets information about what is going on in the outside world by way of the *eyes*, the *ears*, the *nose*, and by way of sensitive cells in the *skin*. Sensory nerves carry impulses from the *sense-organs* and the sensitive cells to the central nervous system. In this Course, we shall only be able to discuss the sense-organs of Vertebrates, taking Man as our main example.

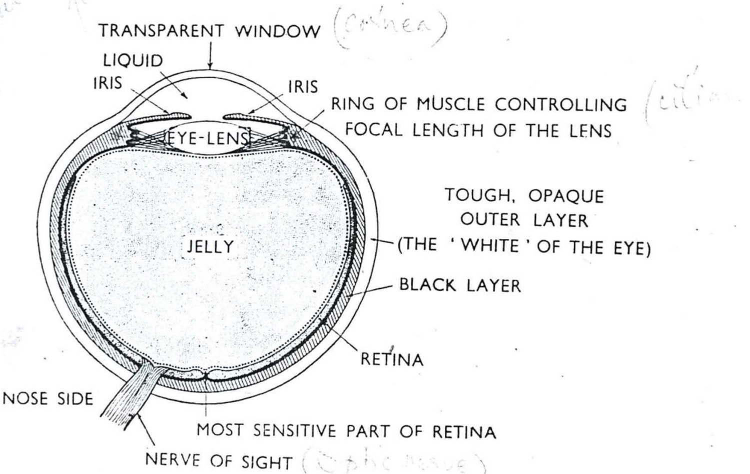


FIG. 105. THE RIGHT EYE IN HORIZONTAL SECTION (diagrammatic)

SIGHT—THE EYE

In our lessons on light, we saw that the eye resembles a photographic camera. Rays of light enter the eye through the *pupil*, and the *eye-lens* forms an image on the *retina*, the sensitive layer at the back of the eye. Branches of the nerve of sight run to all parts of the retina and these

[illegible] are distributed over its surface, and are joined by the [illegible]
the [illegible] [illegible] the image formed on the

nerve-fibres end in sensitive nerve-cells. When light falls on the sensitive nerve-cells of the retina, impulses travel along the nerve of sight to a special part of the brain, and the animal becomes conscious that it sees something.

The *eyeball* rests in a bony cavity and it can be turned up and down, and from side to side, by six flat strips of muscle. The front of the eyeball is transparent and it is kept wet by a watery liquid (' tears ') produced from a small *tear-gland* in the outer corner of the eye-socket. In this way, the transparent window in front of the eye (the cornea†) is washed each time the eyelids are closed and opened, the excess of ' tears ' draining into the nose-cavity through a tiny duct which opens at the inner corner of each eyelid. (In crying, uncontrolled laughing and violent coughing the tear-gland produces more fluid than this tiny duct can drain away into the nose, and the tears roll down the cheeks.)

The *iris*, a circular sheet of coloured muscle, corresponds to the ' *stop* ' of a camera, since it can expand or contract and thus change the size of the *pupil* (the opening in the middle of the iris). In this way the amount of light entering the eye can be controlled. If you stand in a shady place with your back to the light, and look at your eye in a hand-mirror, you will see that the pupil is large, so as to admit plenty of light. If you then turn to face the light you will see the iris contract so as to protect the delicate retina and its sensitive nerve-cells from the shock of a sudden change in the amount of light that enters.

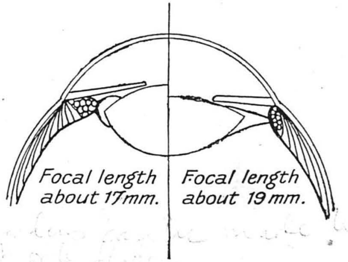


FIG. 106. HOW THE EYE-LENS CHANGES ITS FOCAL LENGTH

The part of the eye in front of the eye-lens is filled with a watery liquid. Immediately behind the iris is the double-convex *eye-lens*, a transparent, elastic body. In mammals, the curvature of this lens can be changed by the movements of rings of muscle at the circumference of the lens, so that the lens can focus light rays from both close and distant objects to produce a clear image on the retina. In other lessons, we have seen that in human beings these focussing arrangements sometimes go wrong (*as in short sight* and *long sight*) and suitable lenses (' glasses ') have to be worn in front of the eyes. The large space between the lens and the retina is filled with a transparent, jelly-like substance.

The eyeball is enclosed by a tough, fibrous outer wall, lined with a black, non-reflecting layer except in front, where the outer layer is also thin and transparent.

DISSECTION OF EYE

Examine the eye of an ox, sheep, or goat from the abattoir. Notice that, except in front, the eyeball is surrounded by fat, which forms a protective cushion in the bony eye-socket. Carefully remove as much of this fat as possible and identify (*a*) the six *eye-muscles*, (*b*) the *nerve of sight* (*optic nerve*).

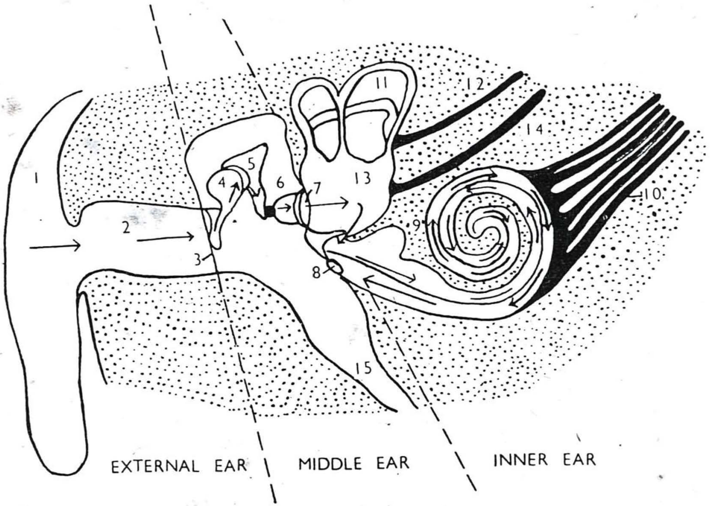


FIG. 107. VERTICAL SECTION THROUGH RIGHT EAR (diagrammatic)
1—external ear. 2—tube leading to ear-drum. 3—ear-drum. 4, 5, 6—chain of bones. 7—oval window. 8—round window. 9—organ of hearing. 10—nerves of hearing. 11—semicircular canals. 12—nerve of sense of direction of movement. 13—organ of balance. 14—nerve of sense of balance and position.

With scissors, or a sharp knife, separate the front half of the eye from the back half and identify (i) the *anterior cavity*, filled with a watery liquid, and (ii) the *posterior cavity*, filled with a jelly-like substance (you may be able to remove this, together with the eye-lens), (iii) the *eye-lens*, (iv) the *iris*, (v) the *pupil*, (vi) the tough, white, outer *layer* (*sclerotic*),† (vii) the thin *black lining* (*choroid*),† (viii) the thin grey *retina* (which can be peeled off with care).

Examine an eye which has been hardened in 40 per cent formaldehyde overnight, and then cut it in two, from front to back, in vertical section. Make a diagram of what you see. Also examine an eye-lens which has been removed undamaged.

HEARING—THE EAR

In most mammals, the visible *external ear* is a gristly structure which serves to collect the sound-waves and direct them down a tube on to the *ear-drum*. Some mammals, e.g. the horse and rabbit, can actually turn the external ear so as to collect sounds coming from different directions. (The horse, which moves its external ears a lot, has 17 muscles for moving each ear; Man has only 9 such muscles and they are all useless, except in rare cases.)

A narrow, slightly curved tube (about one inch long in Man) leads inwards from the external ear to the *ear-drum*. This tube is lined with outward-projecting hairs and wax-producing glands which help to protect the ear-drum from insects and harmful solid matter. The ear-drum consists of a very thin membrane (only a few thousandths of an inch thick in Man) stretched across the inner end of the tube leading from the external ear. Being so very thin, the ear-drum is easily damaged.

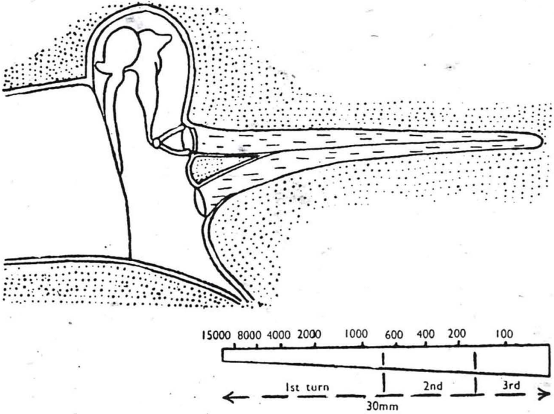


FIG. 108. DIAGRAMMATIC SECTION THROUGH EAR WITH SPIRAL TUBE UNROLLED
(after W. H. White)

The lower diagram is a plan of the elastic membrane which divides this tube into two passages marked with the frequencies of the vibrations to which the different parts of this membrane respond.

The rest of the ear—the *internal ear*—is entirely enclosed in the bone of the *ear-cases*, which form part of the skull. The *internal ear* is divided into two parts:—(*a*) the *middle ear*, and (*b*) the *inner ear*.

The *middle ear* is a cavity in the bone of the skull, filled with air, and connected to the back of the throat by a narrow passage. By means of this tube, the air on the inner surface of the ear-drum is kept at the same pressure as that of the air on its outer surface. (To see whether your own tubes are in good order, take a breath, close the mouth, pinch the nose with finger and thumb, and try to breathe out, swallowing at the same time. If your tubes are clear, you will feel and hear your ear-drums forced outwards by the increased pressure of the air in the middle ear.)

A *chain of three small bones*, hinged together, stretches across the middle ear and carries vibrations from the ear-drum to the *inner ear*, which contains the true *organ of hearing*. The external ear and the middle ear, therefore, are only concerned with receiving sound-waves and passing them on to the inner ear. (The *inner ear*, however, contains something else besides the organ of hearing: one part of it is the organ of balance and position, which we shall discuss later.) There are two small openings, closed by membranes, between the middle and the inner ear. They are called the *oval window* and the *round window*.

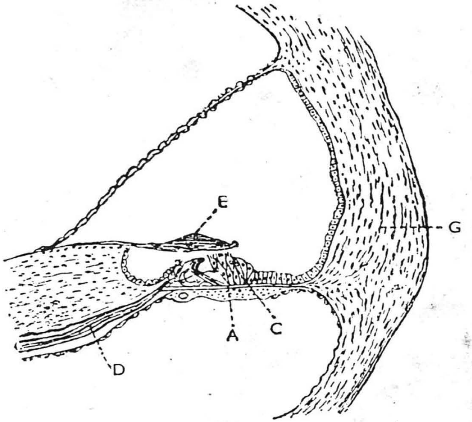


FIG. 109. SECTION THROUGH PART OF ORGAN OF HEARING

This shows an enlarged cross-section of a third, narrower passage lying between the two spiral passages in Figs. 107 and 108.
A—elastic membrane supporting rods with sensitive hair-cells (C). E—membrane forming ‘ shelf ’ overhanging the sensitive hair-cells. D—nerve-fibres. G—wall of organ of hearing.

The actual *organ of hearing* consists of a spirally-coiled passage shaped like a tiny snail-shell and filled with lymph. The vibrations of the ear-drum are passed on and magnified by the chain of bones across the middle ear to the membrane of the *oval window* and similar vibrations are set up in the lymph filling the spiral tube of the inner ear. These vibrations travel through the lymph up to the top of the spiral tube and then down again to the *round window*. (Hence, as liquids are incompressible, as the oval window moves inwards the round window moves outwards.) As the lymph vibrates, an elastic membrane (A in Fig. 109) in the spiral tube vibrates ‘ in step ’ with the vibrations of the lymph, and sensitive hair-cells (C) on this membrane rub against an overhanging shelf (E). When sounds of different frequencies are received, different hair-cells rub against the shelf and transmit impulses along the nerve-fibres (D) to the part of the brain which is concerned with hearing. The brain then recognizes these impulses as notes of varying pitch according to the particular nerve-fibres which carried them.

EXPERIMENTS ON HEARING

(i) Listen to the note of a whistle whose pitch can be gradually raised. Notice that when the pitch of the note gets very high it becomes more difficult to detect and finally it cannot be heard at all. Notice that some individuals can hear ‘ higher ’ notes than others.

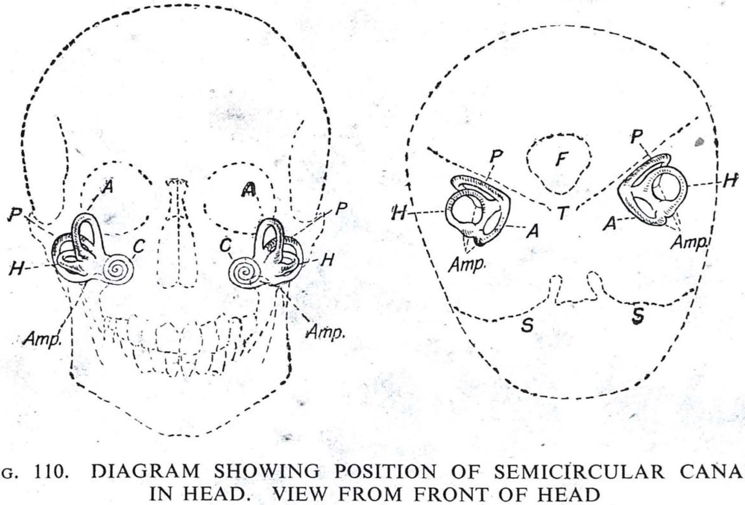


FIG. 110. DIAGRAM SHOWING POSITION OF SEMICIRCULAR CANALS IN HEAD. VIEW FROM FRONT OF HEAD

A, H, P—semicircular canals. C—organ of hearing. Amp.—swelling containing sensitive hair-cells.

(ii) Place your flat hands, with their palms facing backwards, on either side of the head, just in front of the ears. Stand in front of a loudly-ticking clock and close your eyes.

Where does the sound appear to come from ?..................

....

Your hands ' shade ' the ear-passages from sound coming from the front, hence the sound-waves can only travel round the back and sides.

Use your hands as extensions of your external ears and notice the increased loudness of sounds coming from in front of you.

SENSE OF BALANCE AND POSITION

The inner ear also contains the three *semicircular canals*, which are concerned with the sense of position and balance of the body. Each of these three semicircular canals lies in a plane at right angles to the other two, and at one end of each canal there is a swelling which contains sensitive nerve-cells provided with hairs. The semicircular canals are filled with lymph, and when the head is moved in any direction, the lymph in one of these tubes will drag behind a little and bend the hairs of these sensitive nerve-cells, which then send an impulse through their nerve-fibres to the brain. The brain recognizes this impulse as a sensation of *movement* in a particular direction.

It is clear that sensations of movement depend upon the movement of the lymph in the semicircular canals. But we can also sense the posture of the head even when no movement is taking place. This is brought about by another part of the organ of balance, lying below the semicircular canals. The cavity of this organ, too, is lined by sensitive hair-cells from which nerve-fibres lead to the brain, and among these hair-cells are tiny particles of a chalky substance which press on the surrounding hairs. These bits of chalk press on different hair-cells according to the posture of the head, and the brain recognizes the different nervous impulses as sensations of *position*. The organ of balance is helped in its work by other sense-organs. Thus our eyes tell us a great deal about our movements or position in relation to surrounding objects.

Examine carefully a model of the human ear, if one is available.

INTERNAL SENSE-ORGANS

There are other sense-organs scattered throughout the body which convey to the brain what the various voluntary muscles, tendons and joints are doing. For example, it is not necessary to have one's eyes open to know the movements and position of one's limbs.

SENSES OF TASTE AND SMELL

The *organ of taste* is the *tongue* and the *organ of smell* is in the upper part of the *nose*. These two organs nearly always work together.

In fact, it is usually difficult to decide whether a particular sensation is a taste or a smell. When one has a ' cold in the nose ' one cannot smell, and at the same time the food has very little taste. This is because most of our ideas of the ' taste ' of foods also include their characteristic smell.

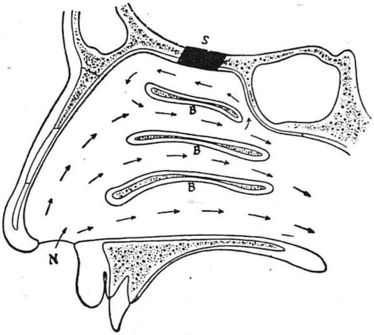


FIG. 111. SECTION THROUGH NOSE-CAVITY
(Arrows show path of air when breathing in.) N—nostril. B—curved bones. S—organ of smell.

We taste with special groups of sensitive nerve-cells called *taste-buds*† situated in tiny projections on the upper surface of the tongue, and connected with the brain by nerve-fibres.

These *taste-buds,* which are only sensitive to substances which have *dissolved* in the saliva, are of four kinds, each corresponding to one of the four pure tastes—*sweet, sour, bitter* and *salty.* Most everyday tastes are combinations of these fundamental tastes. The four kinds of taste-buds are not distributed evenly over the tongue. Thus we taste sweet things best with the tip of the tongue, where the ' sweet taste-buds ' are most numerous, while bitter tastes affect the back of the tongue. The brain recognizes the stimulus it receives from the taste-buds as a particular kind of taste.

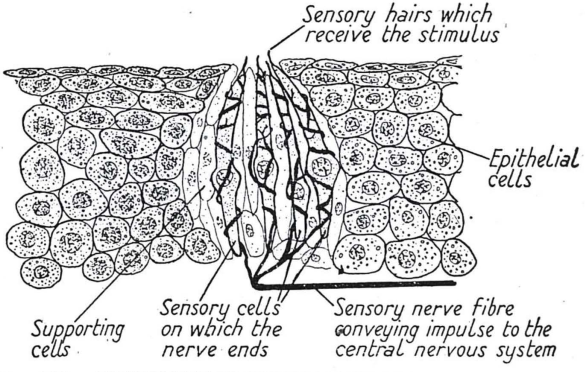


FIG. 112. SECTION THROUGH TASTE-BUD (much magnified)

We *smell* by means of special sensitive hair-cells situated on a square *smelling-patch* in the upper part of the nose-cavity. These cells are connected to the brain by nerve-fibres. As air passes through the nose, these sensitive cells are stimulated by any impurities which have a definite smell, and a stimulus is sent to the brain along the nerve-fibres. The brain recognizes this stimulus as a particular smell. (The Teachers' Handbook includes experiments in taste and smell.)

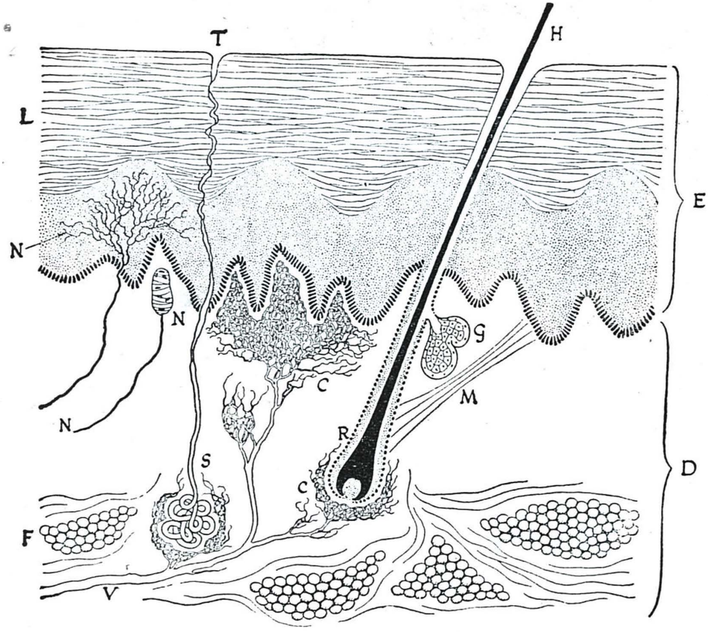


FIG. 113. SECTION THROUGH HUMAN SKIN (magnified)
C—blood capillaries. D—inner layer. E—outer layer. F—fat. G—oil-gland. H—hair. L—layer of flat, horny cells. M—hair-muscle. N—nerve-fibres and sensitive nerve-endings. R—hair-pit. S—sweat-gland. T—sweat-duct. V—large blood-vessel.

FEELING—TOUCH, TEMPERATURE AND PAIN

We get much information about our surroundings through the sense-organs of *touch* situated just beneath the outer *skin* over most parts of the body-surface. When we touch anything, even with our eyes closed, we can say whether it is hot or cold, hard or soft, and whether it hurts or not.

This is because the skin has four kinds of sense-organs, each sensitive to one only of the four types of sensation:—(i) *warmth*, (ii) *cold*, (iii) *pain*, and (iv) *touch*. Nerve-fibres carry stimuli from these sense-organs to the brain and to the spinal cord. (See Teachers' Handbook for experiments on the sense-organs of the skin.)

CHEMICAL CONTROL

HORMONES

In animals, the internal co-ordination of different parts of the body is not always carried out by the nervous system. Co-ordination is also carried out by the circulation of chemical substances with the blood. These ' *chemical messengers* ' are called *hormones* and they are passed into the blood-stream by special *ductless glands*. For example, the pancreas does not pour out pancreatic juice continuously, but only when the small intestine contains food to be digested. Whenever food enters the small intestine from the stomach, cells in the wall of the small intestine produce a hormone which is passed into the blood-stream and carried round the body. When blood containing this hormone reaches the pancreas, it supplies a stimulus to that gland, which pours out pancreatic juice. The same hormone may also stimulate the flow of bile from the liver.

In female mammals, when a fertilized egg-cell begins to develop in the womb, another hormone is produced which stimulates the growth of the milk-glands ready to supply milk for the young one after birth.

Another very important hormone is *adrenalin*,† produced by the *adrenal bodies*, a pair of ductless glands lying just anterior to the kidneys. Under great excitement, e.g. anger or fear, the adrenal glands pass adrenalin into the blood-stream, and as the hormone is distributed to every part of the body it prepares

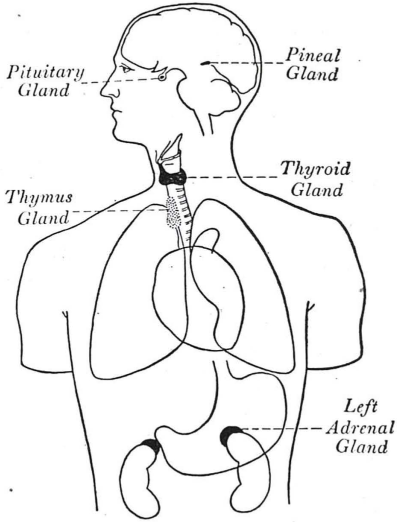


FIG. 114. POSITION OF THE MAIN DUCTLESS GLANDS

almost every organ for special efforts. For example, the heart-beat becomes more rapid, the capillaries become wider, the amount of glucose in the blood increases, the sweat-glands become more active and the activities of the food-canal are slowed down.

The *thyroid*† *gland*, in the neck, produces a hormone which regulates growth and development. For example, a tadpole which has had its thyroid gland removed continues to grow, but does not change into a frog. On the other hand, tadpoles fed with thyroid gland change into frogs very quickly, but these adult frogs are very small in size, sometimes no bigger than flies.

The *pituitary*† *gland*, situated beneath the brain, produces another hormone which also controls growth. Most human 'giants' have unusually active pituitary glands, while a lack of this hormone produces 'dwarfs'.*

Hormones are like vitamins in that exceedingly small amounts produce very striking results. In fact, it has been suggested that 'vitamins may be regarded as hormones which the body cannot make for itself'.

Similar *plant hormones*, produced by young cells, control plant growth.

A different type of chemical control is carried out in animals by the waste products carbon dioxide and urea, but these substances are not called hormones. For example, the diameter of the blood-capillaries varies according to the amount of carbon dioxide in the blood. When the body-muscles are resting, there is little carbon dioxide in the blood and the capillaries are very narrow. When the muscles are working, the blood receives more carbon dioxide and the capillaries widen, thus increasing the blood supply. At the same time, when this blood—containing more carbon dioxide—reaches the brain, nervous impulses are sent to the muscles between the ribs and to the diaphragm, and the breathing movements become deeper and more rapid.

Similarly, an increased amount of urea in the blood-stream stimulates the kidneys to greater activity in order to remove this excess of urea from the blood.

INSTINCT AND INTELLIGENCE

The behaviour of all plants and of the lower animals is *automatic*, i.e. the same stimulus always produces the same unreasoning response. For example, a root always grows towards water and a stem towards the light; a moth flies straight towards a light at night, because it turns automatically so as to allow both eyes to receive the same amount of light. These *automatic responses* to stimuli (in animals) are called *reflex actions* or *instincts*, and they are inherited* from the animal's parents in the same way as the characteristic structure of its body. Every animal is born with a set of these fixed impulses or instincts, and the lower animals go through life provided only with this inborn *instinctive behaviour*. Such animals are unable to *learn*.

Instinctive behaviour reaches its highest level in the insects and particularly in the 'social insects'—ants, bees and wasps.†

The higher animals, however, are able to modify their inherited instincts, for they can *learn by experience*. In this way, a simple reflex becomes a *conditioned reflex*. For example, when a dog smells food, its mouth 'waters' because the flow of saliva is a simple reflex response to the smell-stimulus. In a historical experiment, whenever food was offered to a dog a bell was rung, and after repeating this several times it was found that the sound of the bell alone (without any food) started the flow of saliva. That is, the bell-stimulus had replaced the food-stimulus, a conditioned reflex had been formed and the dog had *learnt*

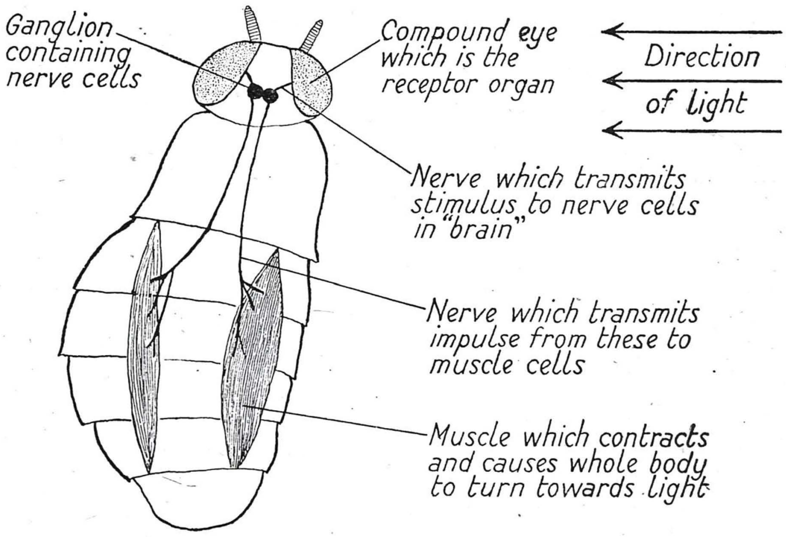


FIG. 115. REFLEX ACTION IN AN INSECT WHICH MOVES TOWARDS LIGHT

a new habit. Most mammals do not get beyond this stage of *learning new habits.*

A few mammals—some of the higher apes and particularly Man—can modify their behaviour by ' thinking things out ', i.e. by the use of *reason.* Such *intelligent behaviour* reaches its highest level in Man, in fact one of the most important characteristics which distinguishes Man from all other animals is that he has *reason* as well as *intelligence.* When faced with a new situation, he can remember his previous experiences and decide on the most suitable course of conduct, by *reasoning.* One of the factors which has enabled Man to have this capacity for reasoned behaviour is that he has *language.* Many other animals can send messages to each other by sounds, but Man is the only animal with a system of sounds which will convey complex ideas—he can speak and think *in words.* The use of language also makes it possible for experiences to be shared and also for the handing down of experiences from one generation to another. This is one of the most important factors in *education.*

Man, too, is the only creature capable of distinguishing between ' right ' and ' wrong ', i.e. he is capable of *moral* conduct,* controlling his activities with reference to an ideal, in fact men often face death itself rather than sacrifice their ideals. No other creature does this.

CHAPTER X

REPRODUCTION IN PLANTS AND ANIMALS

We have seen that one of the characteristics which distinguishes living things from non-living matter is that *living things can reproduce themselves* while non-living things cannot. Except for some of the simplest one-celled organisms like Amoeba and Bacteria, all living things have a limited length of life and are finally overtaken by *death.* Many plants and animals meet their death ' by accident ' before they grow old, e.g. they are often eaten by animals, they often die from lack of food, and they are often killed by parasites. But even those living things that escape ' death by accident ' finally die ' of old age ', e.g. the tissues ' wear out ' or some essential part of the organism ' stops working '.

With the exception of the simplest, one-celled organisms, every *individual* living thing must die, but there is a continuous production of new individuals to replace the dead ones and thus maintain the *race.* These new individuals develop from small pieces of living matter which have separated from other individuals in the process of *reproduction.*

REPRODUCTION IN PLANTS

The simplest type of reproduction is by *division*: one individual splits into two similar halves which grow up into two new individuals. This primitive type of reproduction is only possible in the simplest plants and animals and is commonest in one-celled organisms. For example, free-swimming, one-celled plants reproduce in this way.

In fully-grown cells of Spirogyra, the nucleus divides into two equal halves and a cell-wall grows across the cell, forming two ' daughter-cells ', each of which is a separate living unit although they remain attached end to end. This primitive type of reproduction, where the new individuals develop from one single parent and where there is no coupling or *fertilization*, is called *vegetative reproduction* or *asexual*† *reproduction.*

Spirogyra also shows a different type of reproduction in which *two cells* take part. When two threads of Spirogyra lie side by side, two opposite cells may send out projections of the cell-wall which meet and

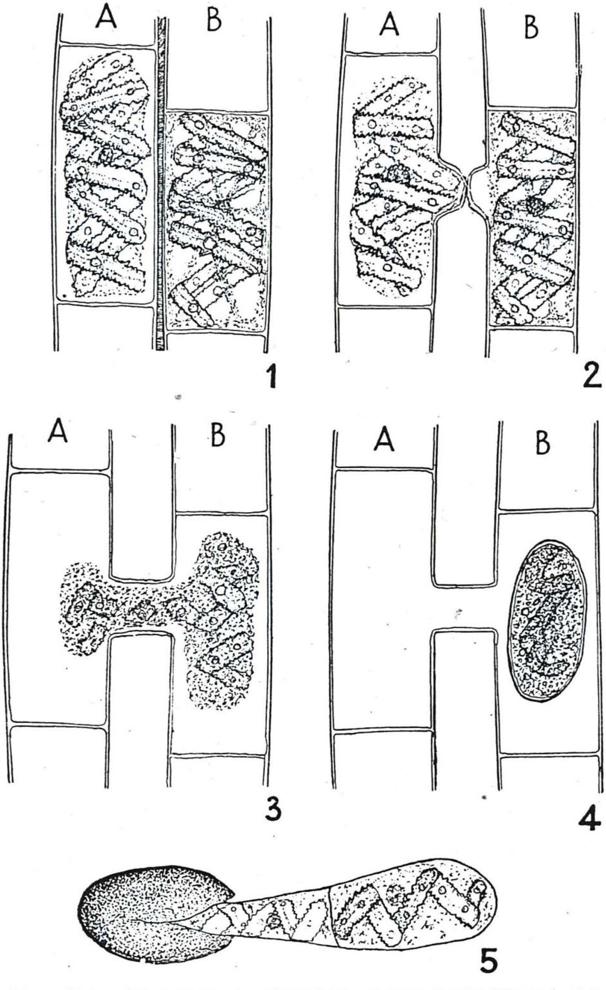


FIG. 116. PRIMITIVE SEXUAL REPRODUCTION IN SPIROGYRA

1. Threads sticking together. 2. Cells beginning to unite. 3. Cell-contents fusing together. 4. Resting-stage. 5. New plant germinating.

finally form a passage uniting the two cells. The contents of one cell pass through this passage and fuse with the contents of the other cell. This single reproductive cell, formed by the fusion of two Spirogyra cells, grows a thick, protective layer of cellulose and rests in this condition until external conditions become favourable for growth, when it grows into a new thread of Spirogyra. This type of reproduction, in which two cells take part, is called *sexual*† *reproduction* (although in the case of Spirogyra we cannot say which of the two cells taking part is the male cell and which the female cell—so that Spirogyra shows only primitive sexual reproduction).

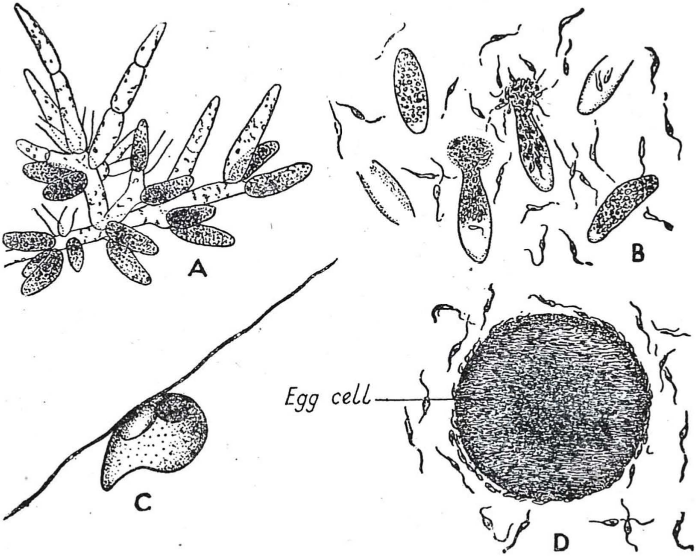


FIG. 117. FERTILIZATION IN A SEA-WEED

A—branching hair from male reproductive organ. B—sperms escaping. C—a single, free-swimming sperm (more highly magnified). D—an egg-cell surrounded by sperms.

Seaweeds show a more advanced type of sexual reproduction. There are two quite different kinds of reproductive cells: (*a*) very tiny, active, free-swimming *sperm-cells*, and (*b*) passive* *egg-cells*, several thousand times bigger than the sperm-cells. The male sperm-cells swim through the surrounding water to the female egg-cells and when one sperm-cell enters an egg-cell and their nuclei fuse together, the *fertilized egg-cell* can develop into a new Seaweed plant.

Reproduction in Ferns and Mosses is rather complicated as there are two different methods of reproduction which take place alternately. On the under-sides of fern leaves can be seen small brown or black patches. These cover groups of *spore-cases* which contain numerous microscopic reproductive cells called *spores*. The ordinary fern-plant, therefore, is the *asexual generation*, producing spores without any coupling or fertilization. These spores, however, *do not grow directly*

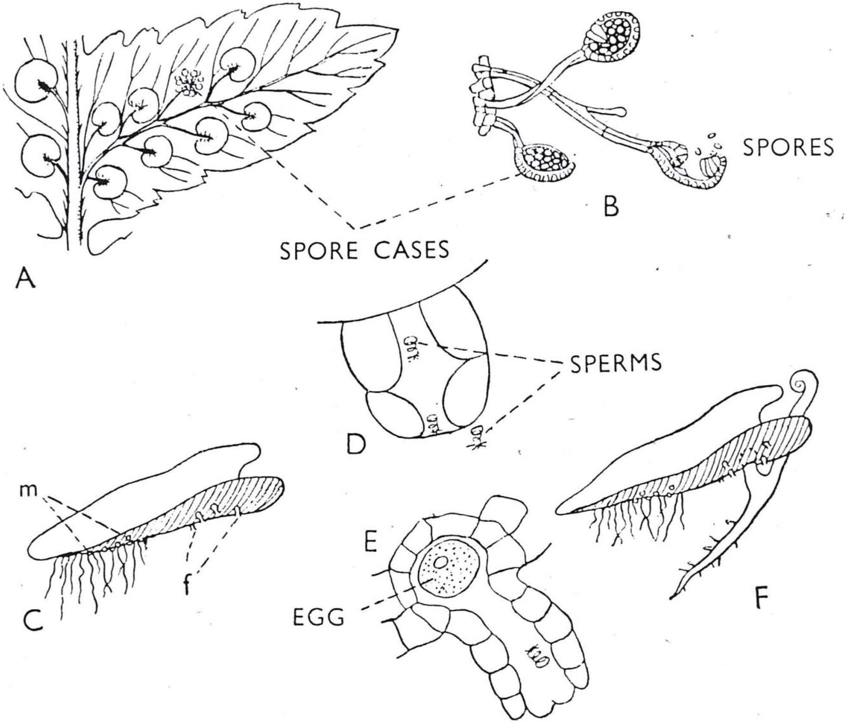


FIG. 118. THE LIFE-HISTORY OF A FERN

A—underside of fern leaflet showing groups of spore-cases (one without its cover). B—three spore-cases (one opening and setting free its spores). C—a *spore-plant* (with male and female reproductive organs on its underside). D—a male reproductive organ with sperms escaping. E—a female reproductive organ, containing an egg-cell (with one sperm entering). F—a fertilized egg-cell developing into a new *fern-plant*.

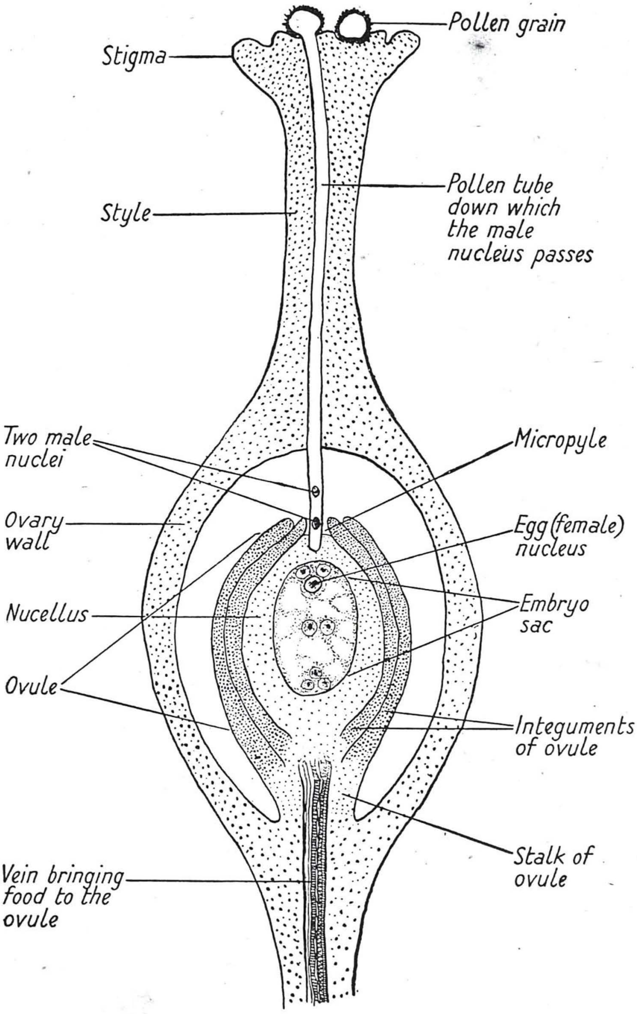


FIG. 119. FERTILIZATION IN A FLOWERING PLANT (diagrammatic)

into new fern-plants, but they produce a small, flat, green *spore-plant* (often found growing on the damp earth under the parent fern-plant). This tiny spore-plant is the *sexual generation* since it produces egg-cells and sperm-cells. The sperm-cells swim through a drop of water to the ovaries, and when a sperm-cell fuses with an egg-cell, the fertilized egg-cell develops into a new *fern-plant*. Hence there is a regular *alternation of generations*—the asexual fern-plant producing spores and the sexual spore-plant producing female egg-cells and male sperm-cells.

In *flowering plants*, the male sperm-cell no longer swims through water to reach the female egg-cell, but it is enclosed in a *pollen-grain*.

This pollen-grain is usually carried by insects or by the wind to the stigma of the flower, where it grows a pollen-tube down through the style to the ovary. The ovule contains the female egg-cell, and when the pollen-tube enters the ovule, the nucleus of the sperm-cell fuses with the nucleus of the egg-cell, and from this fertilized egg-cell the new seed-plant develops. Reproduction in flowering plants is specially adapted to life on dry land. All the lower plants depend on the presence of water in order that their sperm-cells may swim to the egg-cells. In flowering plants, both the sperm-cells and the egg-cells are well protected throughout the whole process of fertilization; and, after fertilization, the new seed-plant is protected by the ovary while it develops and receives the reserve food-material it will need later when it germinates. The dry, ripe seed is usually able to withstand unfavourable conditions for some time, resting in an inactive condition until the external conditions become suitable for germination. In most flowering plants, the parent plant

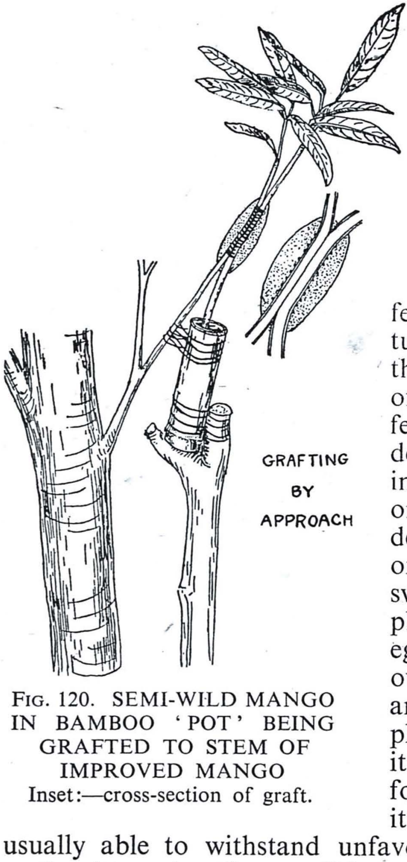


FIG. 120. SEMI-WILD MANGO IN BAMBOO 'POT' BEING GRAFTED TO STEM OF IMPROVED MANGO
Inset:—cross-section of graft.

scatters the ripe seeds so as to give them a better chance of finding a suitable spot to grow. In the higher, flowering plants, therefore, we see signs of *parental care* during every stage of reproduction.

VEGETATIVE REPRODUCTION IN FLOWERING PLANTS

We have seen that the simplest, one-celled plants can reproduce *by division.* The more complicated and specialized higher plants cannot split into two similar halves in this way, but in many flowering plants a piece cut from one plant can grow into a new and complete plant.

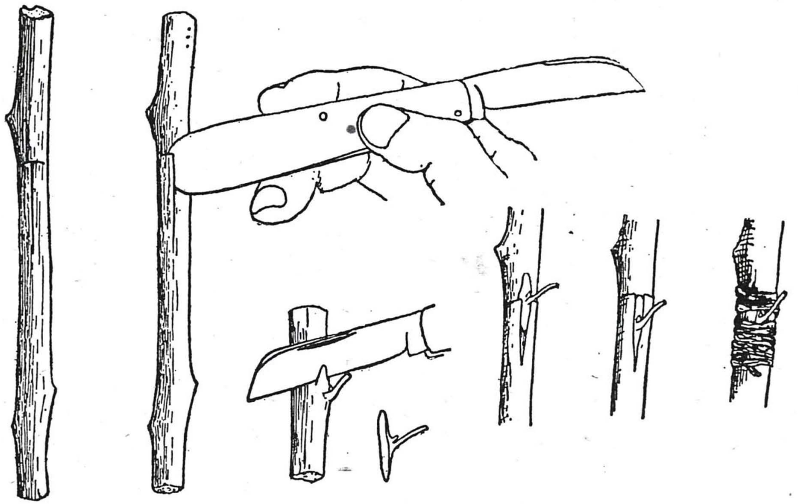


FIG. 121. BUD-GRAFTING

This method of reproduction from some portion of the plant other than the seed is called *vegetative reproduction* because only *vegetative* parts of the plant are used (as distinct from *seeds* which are derived from the male and female—or *sexual*—parts of the plant).

Vegetative reproduction is widely used by gardeners and agriculturists, particularly in tropical countries. We have already seen how *underground stems* are often used as the starting-point of a new plant. For example, Onions, Lilies and Crinums can be grown from *bulbs*; Yams, English Potatoes and Sweet Potatoes can be grown from *tubers*;

Caladiums can be grown from *corms*; Ginger and Arrowroot can be grown from *rhizomes*; Bananas can be grown from *root-suckers*; Pineapples can be grown from *stem-suckers* (as well as from root-suckers).

Another method of vegetative reproduction is by *grafting*.† In grafting, a shoot of the required plant, e.g. a good variety of Orange or Mango, is placed with its cut surface in contact with the cut surface of a rooted plant (called the ' stock '†) so that the two *cambiums* grow together. The ' stock ' supplies a strong and hardy root-system while the grafted shoot will produce fruits of the same superior quality as the parent tree from which the graft was cut. In tropical countries, however, the commonest method of grafting is *budding* (or *bud-grafting*), a single bud of the desired plant being grafted on to a suitable rootstock. This method is used for producing new rubber trees from high-yielding parents. In this way a common red Hibiscus can be made to bear several different colours and varieties of Hibiscus flowers by bud-grafting. Similarly, a Shaddock† (Pumelo) tree can be bud-grafted with any other Citrus fruit, e.g. Orange, Lime, Lemon, or Grapefruit, so that it is possible to have one tree bearing all these different fruits.

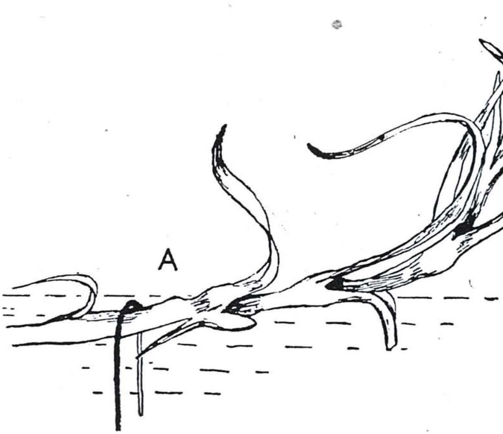


FIG. 122. LAYERING

Another method of vegetative reproduction is *layering*. In this method, a portion of a branch is covered with soil while it is still attached to the parent plant.

The buried portion grows roots and the branch can then be separated from the parent plant. This is a good method for producing new Bougainvillea plants. If the branch cannot be bent down to the ground it is passed through a bamboo ' flower-pot ' filled with damp soil.

The great practical advantage of all these methods of *vegetative* (or *asexual*) *reproduction* is that they reproduce exactly the characteristics of one parent only, just as though the cutting or other vegetative part were still growing on the original parent plant. When plants are grown from seeds there is always a possibility that new varieties of plants will be produced. In fact, very few of the seedlings are exactly like the parent plant. We shall explain the reason for this later.

REPRODUCTION IN ANIMALS

The simplest, one-celled animals, like the one-celled plants, divide into two equal parts when they reach their full size, and each half develops as a new individual. Amoeba and the Bacteria reproduce thus, by simple division into two, and we have explained that this type of reproduction is called *asexual reproduction*, because there is no fusion of two cells or any exchange of cell-contents between two cells.

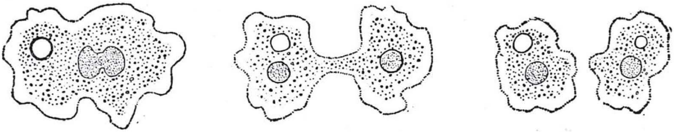


FIG. 123. AN AMOEBA REPRODUCING BY DIVISION

Paramecium, like Amoeba, usually reproduces asexually, one individual dividing into two. Under favourable conditions, this type of reproduction may take place at intervals of less than one day. Occasionally, however, a primitive type of sexual reproduction takes place in Paramecium. Two individuals couple together, ' mouth ' to ' mouth ', so that the protoplasm of both cells becomes continuous, and there is an exchange of parts of their nuclei. The two individuals then separate and again reproduce asexually, by division. Although Paramecium shows this primitive type of reproduction involving an exchange of part of the cell-contents of two individuals, it is not a means of multiplying but rather a method of renewing the vigour and activity of the animals concerned.

Hydra shows both types of reproduction—asexual and sexual. Although Hydra does not reproduce naturally by simple division, if it is cut in half each half can grow the missing parts, thus producing two new animals. The natural method of *asexual reproduction* in Hydra is by *budding*. A small swelling, or *bud*, arises on the side of the animal and grows out into a new Hydra. When the bud has developed a ‘ mouth ’ and can get its own food, it separates from its parent and lives independently.

Hydra can also reproduce *sexually*, by the fusion of special reproductive cells—egg-cells and sperm-cells. Tiny swellings are formed on the outside of the body, near the mouth end. These become the *male reproductive organs*, or *testes*, producing *sperms*, which are set free into the surrounding water when the ripe testis bursts open. The *female reproductive organs*, or *ovaries*, are larger swellings near the ‘ foot ’ of the animal, and each ovary produces a single, large *egg-cell* which is exposed to the surrounding water when the ovary becomes ripe. The sperms, which are shaped like tiny tadpoles, swim through the surrounding water by moving their ‘ tails ’, and when a sperm reaches a ripe egg-cell, it enters. The nucleus of the sperm fuses with the nucleus of the egg, producing a fertilized egg which can develop into a new Hydra (usually after a resting-stage inside a horny, protective covering).

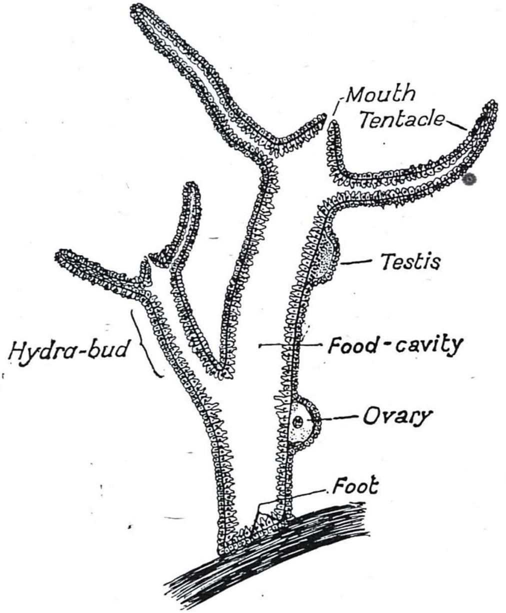


Fig. 124. LONGITUDINAL SECTION OF A HYDRA (diagrammatic)

In the higher animals, e.g. birds and mammals, each individual animal is either *male* or *female*, i.e. it has only one kind of reproductive organ, or it is *uni-sexual*.†

In *Hydra* and some other lower animals, e.g. earth-worms and snails, one animal is both male and female, both testes and ovaries being present in the same animal. Such animals are called *hermaphrodite* animals. In *uni-sexual* animals, there can be only *cross-fertilization* since the pairing egg and sperm always come from different individuals. In *hermaphrodite* animals, however, either *cross-fertilization* or *self-fertilization* may take place, according to whether the fertilizing sperm comes to the egg from the same or a different animal. Such self-fertilization, however, is rare, and we shall see that the earth-worm and the snail, for example, although hermaphrodite, are always cross-fertilized. (It is interesting to note that while the highest animals are always uni-sexual, the highest plants are nearly all hermaphrodite, having stamens and pistil present in the same flower.)

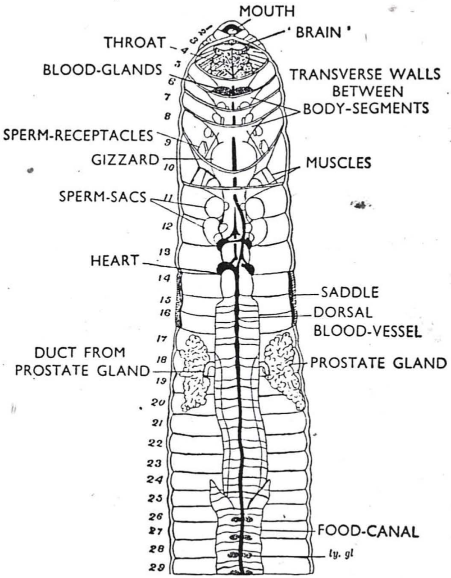


Fig. 125. EARTH-WORM OPENED TO SHOW CONTENTS OF BODY-CAVITY

In *earth-worms*, both sexes are combined in one hermaphrodite animal, yet only cross-fertilization can take place. This makes the reproductive system very complicated. In the Indian earth-worm (*Pheretima*) there are two pairs of *testes*, tiny white bodies enclosed in two pairs of *testis-sacs* lying ventrally below the food-canal in the 10th and 11th segments. The *sperm-cells* produced by the testes are passed into two pairs of *sperm-sacs* lying in segments 11 and 12 where they complete their development. Two pairs of very fine *sperm-ducts* convey the sperms to a pair of male reproductive openings in segment 18. The female reproductive organs are a pair of tiny *ovaries* lying ventrally below the food-canal in segments 13 and 14. The *egg-cells*, when ripe, are conveyed through a pair of very fine *oviducts* to the female reproductive opening in segment 14.

Besides these male and female organs, the earth-worm has several pairs of *sperm-receptacles* in segments 6-8, in which are stored *sperms* *received from another worm.* In order to exchange sperms, two coupling worms place themselves together in a head-to-tail position so that the male reproductive openings of one worm are opposite the external openings of the *sperm receptacles* of the other worm. Sperms are exchanged and the two worms then separate. Fertilization takes place later. A band of elastic skin is formed from the ‘ saddle ’ and the worm withdraws backwards out of this band. As the band of skin passes over the female reproductive opening, egg-cells are passed out into it, and as it passes over the external openings to the sperm-receptacles it receives sperms (which came originally from another worm). Fertilization then takes place, and as soon as the band of skin is clear of the anterior end of the worm it closes up to form a horny *egg-case.* In each egg-case, only one fertilized egg usually develops into a young worm. It is remarkable to find such complicated arrangements for reproduction in an animal whose other body-systems are comparatively simple, but all these complications make self-fertilization impossible and ensure cross-fertilization.

Snails, like earth-worms, are hermaphrodite but not self-fertilizing. Both egg-cells and sperm-cells are produced by the same *hermaphrodite gland* (or *ovo-testis*†) which is embedded in the liver near the apex of the shell. The eggs and sperms then pass down the wavy, white *hermaphrodite duct*, leading to the large *albumen-gland*, where the eggs receive a layer of albumen (protein) which later serves as food for the developing embryo. After leaving the albumen-gland, the eggs and sperms travel along separate ducts—an oviduct and a sperm-duct—and finally the male and female ducts join again to open to the exterior at a *common reproductive opening* on the right side of the snail's head.

Before the oviduct and the sperm-duct finally join together, the last part of the oviduct (the *vagina*) becomes wide and thick-walled, and at its posterior end arises a long tube ending in a sac—the *sperm-receptacle*—in which are stored sperms received from another snail. The lower end of the sperm-duct leads into a muscular tube—the *penis*—which can be pushed out through the reproductive opening and into the reproductive opening of another snail. Coupling snails exchange sperms by placing their reproductive openings in contact, and sperms from one animal are transferred to the sperm-receptacle of the other, where they are stored. The two animals then separate. As eggs pass down the lower end of the oviduct they are cross-fertilized by sperms from the sperm-receptacle and the fertilized eggs are enclosed in shells and are laid in a hole in the soil. As in the earth-worm, the snail's reproductive arrangements are very complicated owing to the difficulty of ensuring cross-fertilization in a bisexual,† hermaphrodite animal.

Insects are uni-sexual, and sperms are passed from the male into the body of the female, i.e. there is *internal fertilization* This type of fertilization is characteristic of all land animals, e.g. snails, insects,

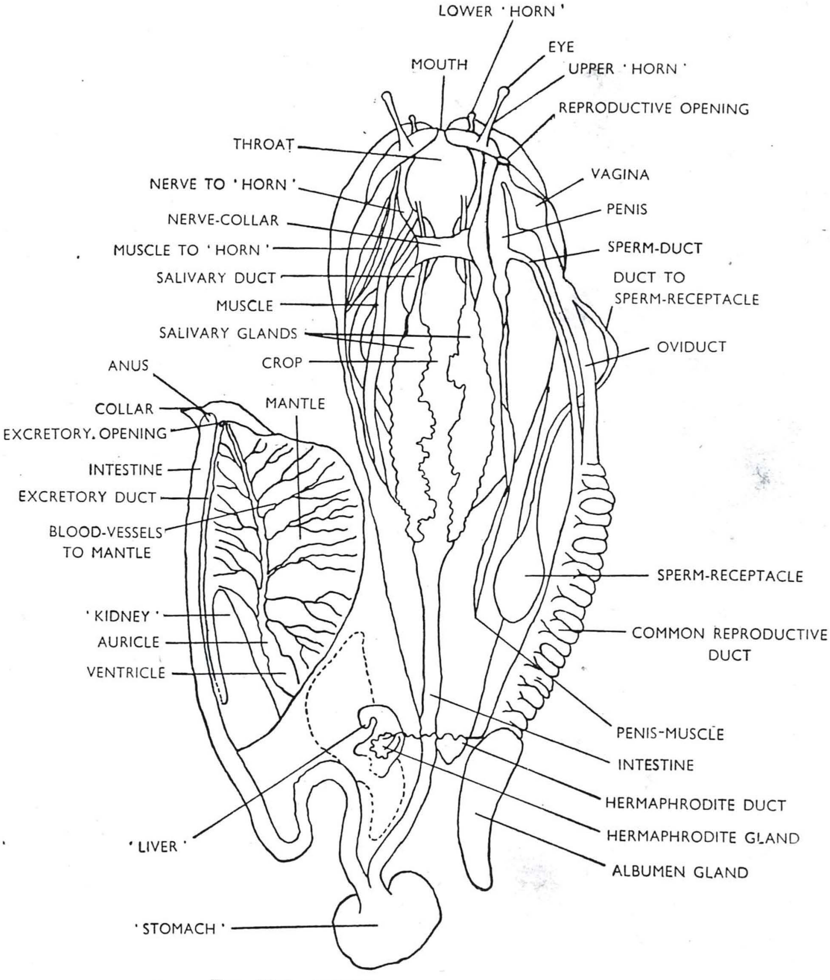


FIG. 126. DISSECTION OF SNAIL

reptiles, birds and mammals, since external fertilization is only possible in animals which breed in water, e.g. most fishes and amphibians. The reproductive system of Insects is rather complicated and difficult to make out except in the larger insects. In the male cockroach for example, the testes are a pair of tiny oval bodies lying in the 5th and 6th segments of the hind body. They can only be seen in young males since they waste away before the animal is fully grown. Sperms are passed from the testes down a pair of sperm-ducts to the sperm-sacs, which consist of a number of small, blind-ended tubes, bunched together in the shape of a mushroom. (This ‘mushroom-shaped gland’ can be seen in adult male cockroaches and was formerly mistaken for the testes.) A single duct leads from the sperm-sacs to the male reproductive opening, which is just below the anus. In the female cockroach, there is a pair of ovaries in the posterior part of the hind body. Each ovary, right and left, consists of a bunch of eight egg-tubes, and, owing to the eggs inside it, each egg-tube looks like a string of beads. The eggs pass from the egg-tubes into the oviducts, which join together before opening at the female reproductive opening into the boat-shaped structure at the posterior end of the hind body. Just above this female reproductive opening is the opening to the small, tubular sperm-receptacle, in which are stored sperms received from a male cockroach during coupling. This store of sperms is used to fertilize the eggs as they are laid. As the eggs pass out of the female reproductive opening they are fertilized by sperms from the sperm-receptacle, and a number of fertilized eggs (14-24) is then packed in a horny egg-case. The female cockroach carries this egg-case, projecting from the boat-shaped structure, until she finds a suitable place to leave it.

Fishes are uni-sexual, i.e. each individual is either a female with ovaries, producing eggs, or a male with testes, producing sperms. In the bony fishes, e.g. carp, there is *external fertilization.* The female fish discharges a very large number of eggs into the surrounding water and the male fish swims over the eggs and discharges sperms. This uncertain method of fertilization is very wasteful, and enormous numbers of eggs and sperms must be produced to compensate for the wastage. A female cod-fish, for example, may produce over four million eggs each breeding season, of which only a very small proportion will ever be fertilized and develop.

In gristly fishes, e.g. sharks, this great wastage is avoided by *internal fertilization.* The sperms are introduced directly into the female oviduct where they meet the eggs. In some gristly fishes, the fertilized egg is then enclosed in a horny egg-case, together with a large supply of food, before being passed out into the surrounding water to 'look after itself'. In some other gristly fishes, e.g. the Indian shark (*Scoliodon*), after internal fertilization, the fertilized eggs attach themselves to the lining of the lower part of the oviduct and the developing embryos obtain their food and oxygen from the parent fish. When the young fish are well developed and several inches in length, they are born alive. This type of reproduction is clearly much less wasteful than the method of passing out enormous numbers of eggs and sperms into the surrounding water and leaving them without any further attention—without any *parental care*.

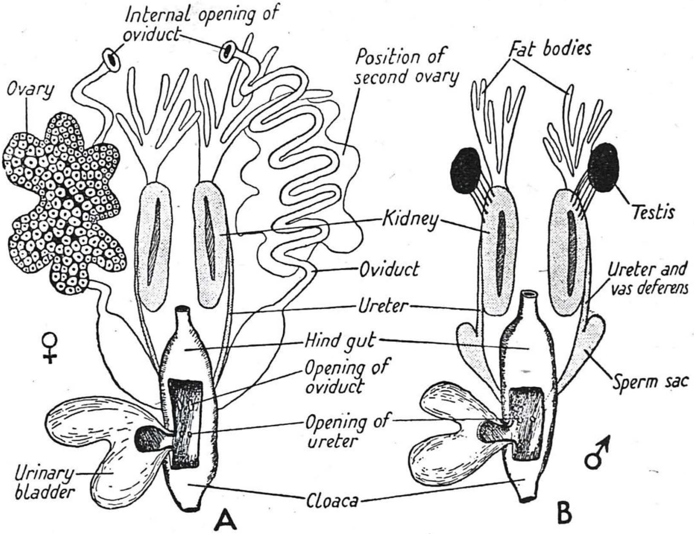


FIG. 127. REPRODUCTIVE (AND EXCRETORY) ORGANS OF (*a*) FEMALE AND (*b*) MALE FROG

(The wall of the cloaca has been partly cut away.)

In *frogs*, fertilization is *external*. The male frog pours sperms over the eggs as they are laid by the female, in water. We have already discussed the development of the frog from the fertilized egg. The reproductive organs of frogs are very similar to those of other Vertebrates. The male frog has a pair of testes: pale yellow, oval bodies attached to the anterior end of the kidneys. A number of fine ducts convey sperms from each testis through the corresponding kidney, whence they pass down the kidney-duct (which carries both sperms and urine), and are then stored in a sperm-sac ready to be poured out through the cloaca over the eggs laid by the female frog. The female frog has a pair of ovaries lying in the same position as the male testes, and consisting of irregular masses of black and white egg-cells. The egg-cells pass down through the right and left oviducts, long, coiled, white tubes. During their passage down the oviducts the egg-cells receive a thin covering of jelly which swells up as the eggs are passed out through the cloaca into water.

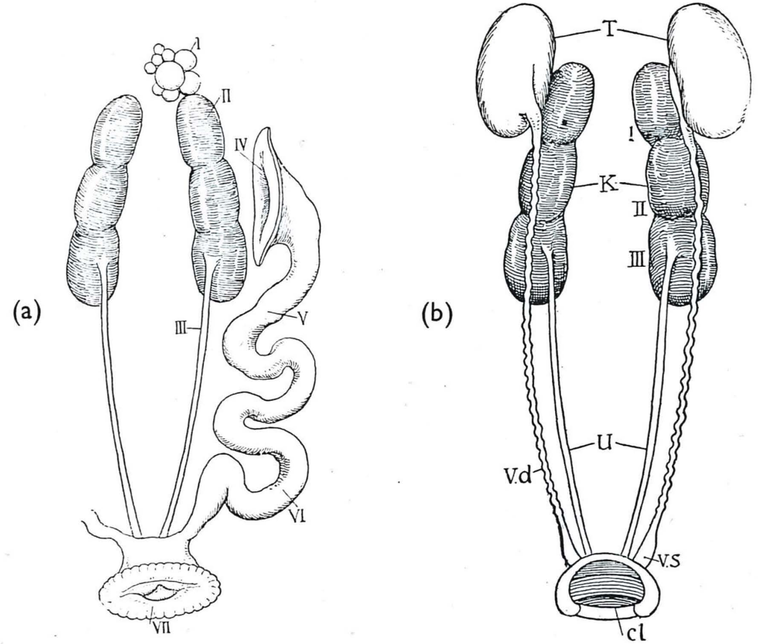


FIG. 128. REPRODUCTIVE (AND EXCRETORY) ORGANS OF (*a*) FEMALE AND (*b*) MALE BIRD

(*a*) I—ovary. II—kidney. III—kidney-duct. IV—internal opening of oviduct. V and VI—oviduct. VII—cloaca.

(*b*) T—testes. K—kidneys. U—kidney-ducts. Vd—sperm-duct. Vs—sperm-sacs. Cl—cloaca.

Birds, like the other higher animals, are uni-sexual, and in many cases there are striking external differences in the colour and arrangement of the feathers and also in the size of the body in the two sexes. The reproductive organs of birds follow the usual vertebrate plan except that there is only one ovary and oviduct (the left) in the adult female, possibly owing to the fact that the egg-cells of birds are exceptionally large and there might be difficulty in passing two large eggs (with shells) down the right and left oviducts at the same time. In an adult female domestic fowl, during the breeding season, the ovary consists of a bunch of small, rounded egg-cells, situated on the dorsal wall of the body-cavity, just anterior to the kidneys. A fully-developed egg-cell, about an inch in diameter, enters the oviduct, and, if the female bird has received a supply of sperms from a male bird, fertilization takes place in the upper part of the oviduct, the nucleus of a sperm-cell fusing with the nucleus of the egg-cell. As the egg passes down the oviduct, it first receives a thick coating of *albumen* (egg-white = about 13 per cent protein and 86 per cent water), then the transparent *shell-membrane* and finally the *shell* itself, consisting of a porous layer of calcium carbonate. About 24 hours after leaving the ovary, the egg is laid, passing out through the cloaca, the common reproductive and excretory opening. In the adult male bird, during the breeding season, the testes are two oval, light-coloured bodies, situated on the dorsal wall of the body-cavity just anterior to the kidneys, and ripe sperms are passed down the right and left sperm-ducts to a pair of sperm-sacs near the cloaca.

During sexual union, or ‘ *treading* ’,† sperms are passed from the male to the female cloaca, along with some liquid, and the sperms travel up the oviduct and fertilize the egg-cells. Fertilization, therefore, is *internal*, as in other land-animals.

The development of the chick. The new-laid egg, as shown in section in Fig. 129, consists of (*a*) a porous *shell* through which exchange of gases can take place between the developing embryo and the atmosphere, (*b*) a double *shell-membrane* forming an *air-chamber* at the broad end of the shell, (*c*) the *egg-white*, supplying food and water to the developing embryo and (*d*) the *egg-cell*, consisting of the tiny embryo—a white disc, about 5 mm. in diameter, always on the upper surface—and the large *yolk*, enclosed in a very thin membrane, supplying food for the developing embryo. (The egg-cells of birds are the largest of all cells.) Identify all these parts in a new-laid egg which has been carefully emptied out into a glass dish. Notice also that the yolk is suspended by two twisted cords of egg-white, one at either end, so that when the egg is turned the embryo always lies on top of the yolk—next to the warm body of the ‘ sitting ’ bird. The embryo, therefore, has a supply of food, water and air, and it only needs suitable warmth in order to develop. The parent therefore ‘ sits ’ on the eggs to keep them warm. Cell-division and development have already begun by the time that the egg is laid, and if the egg is kept warm (103°-105° F. or about 40° C.) development continues, until after 21 days the fully-developed chick escapes from the egg-shell. If possible, examine the developing embryo in eggs which have been kept under a hen for 1, $1\frac{1}{2}$, 2, 3 and 4 days. Most birds show much *parental care* of their offspring, both before and after hatching. The embryo is given a good supply of food and water inside the egg and is protected by the hard shell. The eggs are usually laid in a carefully prepared nest, concealed from enemies, and they are kept warm until they hatch. The young birds are then fed by their parents until they can look after themselves. Some young birds remain with their parents for some time after hatching, thus learning by imitating their parents and profiting by their experience.

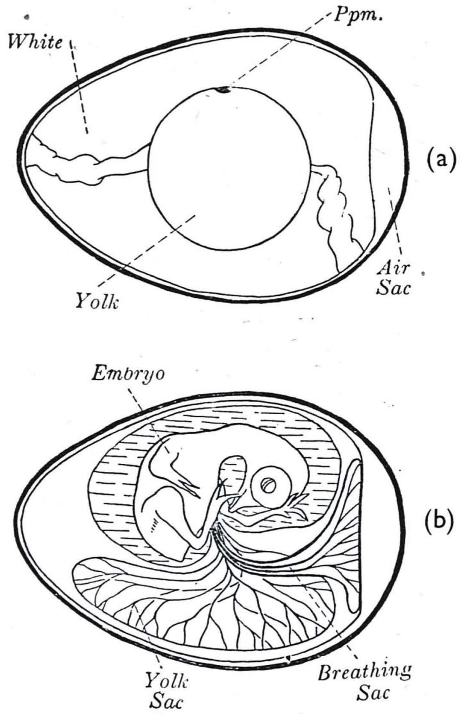


FIG. 129. DEVELOPMENT OF CHICK (diagrammatic)

(*a*)—egg when laid. Ppm.—embryo.
(*b*)—after 14 days under hen.

REPRODUCTION IN MAMMALS

Parental care reaches its highest development in *mammals*. The female parent retains the embryo inside her body throughout its development and it is fully formed at birth. After birth, the young ones are fed with milk, sheltered and defended until they can look after themselves.

In the male mammal, e.g. cavy, rabbit and Man, the testes begin their development in the body-cavity, close to the kidneys, but before birth they pass backwards into two sacs projecting from the posterior end of the body-cavity, on the ventral side (the testis-sacs). Each testis is an oval body consisting of a large number of very fine tubes from the walls of which the sperms are formed. These sperm-producing tubes lead to a long, much-coiled duct in which sperms are stored until required, when they pass along a sperm-duct leading from each testis. These two sperm-ducts join together and then open into the duct leading from the bladder to the exterior through the *penis*, or coupling organ, by means of which sperms are transferred from the body of the male to that of the female mammal. This common excretory and reproductive duct is therefore a passage for both urine and sperms. (In many mammals, e.g. cavy, there is also a pair of sperm-sacs, branching off from the sperm-ducts, in which sperms are stored.)

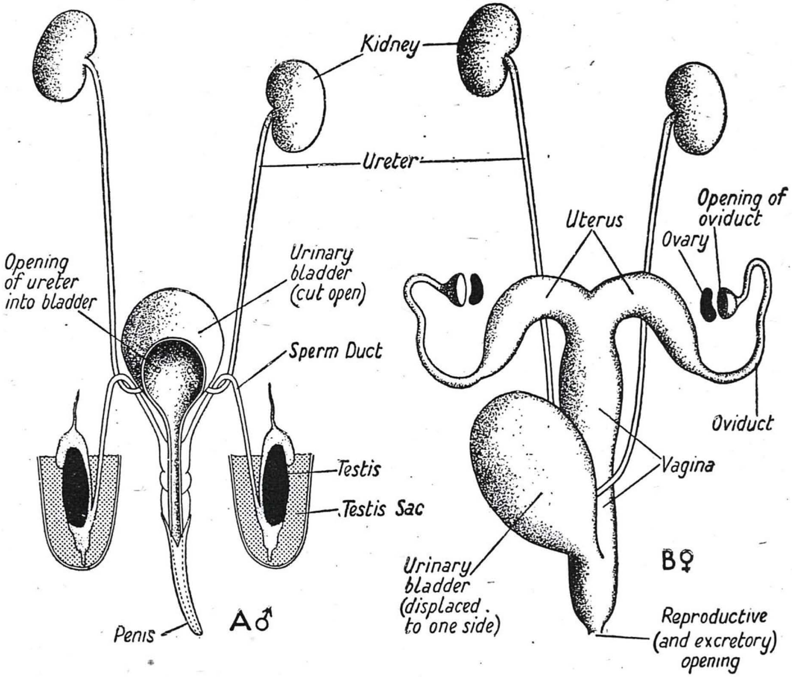


FIG. 130. REPRODUCTIVE (AND EXCRETORY) ORGANS OF (*a*) MALE AND (*b*) FEMALE RABBIT

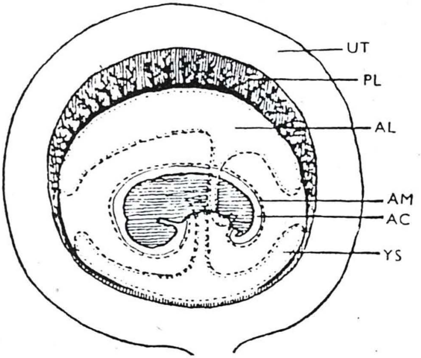


FIG. 131. EMBRYO RABBIT DEVELOPING INSIDE WOMB (diagrammatic)

ut—wall of womb (or uterus). *pl*—disc for exchange of food, oxygen and waste products between blood of parent and embryo. *al. am.*—the two 'birth-dresses' which develop round the embryo. One of these (*al*) forms the 'birth-cord' attaching the embryo to *pl.* The other (*am*) is a protective membrane enclosing the cavity (*ac*) filled with liquid in which developing embryo floats. *ys*—empty 'yolk-sac'.

In female mammals there are two ovaries on the dorsal wall of the belly-cavity. These ovaries are very small, white, oval bodies, and they produce egg-cells which are only about 0·01 inch in diameter. Ripe egg-cells are set free at intervals, and each egg-cell finds its way into the anterior end of the oviduct, where fertilization may take place if one of the sperms received from the male swims up the oviduct and fuses with the egg-cell. The fertilized egg-cell passes down an oviduct until it reaches a wider part of the duct (the *uterus*, or *womb*), when it attaches itself to the wall of this cavity and begins to develop. (In the cavy and the rabbit, there are two such cavities, and several eggs may develop at the same time, but in Man the right and left oviducts join together to form a single cavity and usually only one egg develops at a time.)

The period of development inside the female parent's body varies in different mammals; e.g. it is about 3 weeks in rats and cavies,

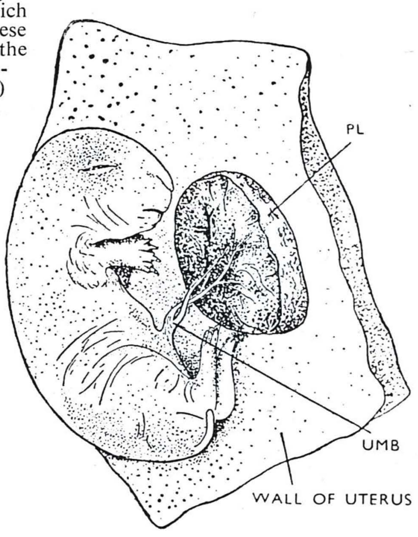


FIG. 132. RABBIT EMBRYO IN WOMB BEFORE BIRTH

Pl—disc of flesh (sometimes called the 'after-birth') covered with villi growing into wall of womb. *Umb*—'birth-cord' with blood-vessels connecting embryo and disc Pl.

9 months in Man, and 13-20 months in elephants. The developing embryo receives food and oxygen from the mother's blood-stream, and gets rid of waste material in the same way, the capillary blood-vessels of mother and embryo being in close contact (although not actually continuous—the embryo having its own independent blood circulation). When the young mammal is fully developed, it is passed out to the exterior and begins its independent existence.

Although a newly-born mammal can get its own supply of oxygen through its lungs, digest its own food and get rid of its own waste products, most newly-born mammals are very helpless and need a great deal of parental care. They are fed on milk, produced by special milk-glands of the female parent, and this milk supplies a complete diet which requires very little digestion to prepare it for absorption and assimilation. In the highest mammal—Man—there is a long period of comparative rest, during childhood, when parental care and protection reach their highest pitch and when the young one is trained and educated before being left to look after itself. The human body is not fully developed until about the age of twenty years, and under civilized conditions, the age at which human beings begin to 'earn their own living' tends to become later and later owing to the length of time required for training and education for skilled and specialized occupations. Under such conditions, the *family* becomes a very important unit of society.

CELL DIVISION

In order to understand the processes of growth and reproduction, it is necessary to know something about how cells multiply in plants and animals. During ordinary *growth*, a single cell with its nucleus divides into two similar parts and these two halves separate and form two new cells. The division of the nucleus, however, is very complicated, and we shall only concern ourselves with its main features. Under the high power of the microscope, the nucleus is seen (Fig. 132) to become a mass of fine threads (1) which shorten and thicken (2) until they become rod-like bodies called *chromosomes*† (3). These chromosomes are very important, and *every cell of the same plant or animal has the same number of chromosomes*, which appear at every cell-division. For example, the *chromosome-number* for Man is 48, Cockroach—34, Cat—36, Mouse—40, Lily—24. Each chromosome then splits lengthwise into two exactly similar halves (4), and the two halves are pulled apart to opposite ends of the cell (5), to form two exactly similar groups of half-chromosomes (6). Each group of chromosomes now forms a new nucleus (7) and the protoplasm of the cell divides into two halves, one half surrounding each new ‘ daughter-nucleus ’ (8).

After the two halves have separated there are two new ‘ daughter-cells ’, *each with a nucleus containing a half of every chromosome of the ‘ mother-cell ’*. (N.B.—The terms ‘ mother ’ and ‘ daughter ’ are not used here to mean ‘ female ’.) This whole process of cell-division usually takes an hour or two. Thus, *in ordinary cell-division, during growth and repair, every new ‘ daughter-cell ’ has exactly the same number and kind of chromosomes as the ‘ mother-cell ’. These chromosomes carry the ‘ inheritance-units ’ or genes.*† Although these *genes* are too small to be seen, even under the high power of a microscope, scientists believe in their existence because they explain the facts of inheritance (just as Dalton was forced to believe in the existence of invisible atoms because they explained the facts of chemical combination). These *genes* are the units or factors which determine the

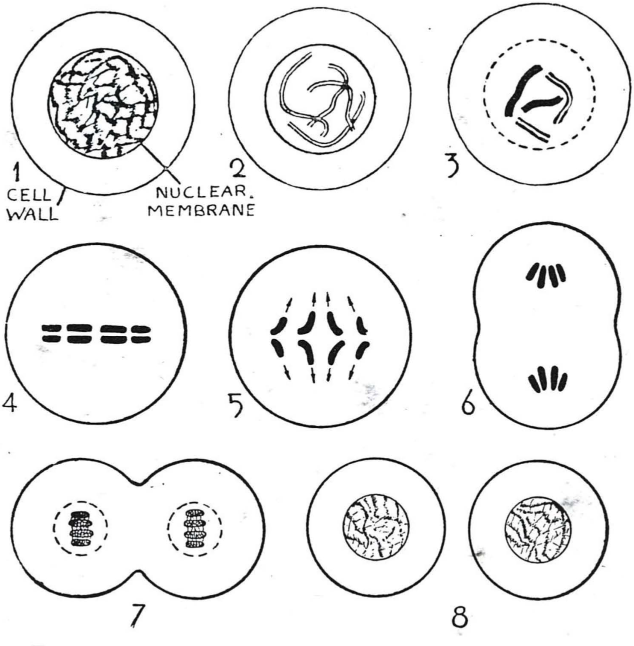


FIG. 133. ORDINARY CELL-DIVISION (diagrammatic)

inherited characteristics of each cell. The chromosomes are collections of large numbers of genes which behave as though they were a chain of tiny particles arranged in a definite order along the chromosome, like beads along a string. During ordinary cell-division, when each chromosome splits lengthwise, each 'daughter-cell' gets an exactly similar set of genes, or 'inheritance-units'. Thus, in *asexual reproduction*, which takes place by ordinary cell-division, each new individual has the same number and kind of genes as its parent and thus inherits the same characteristics.

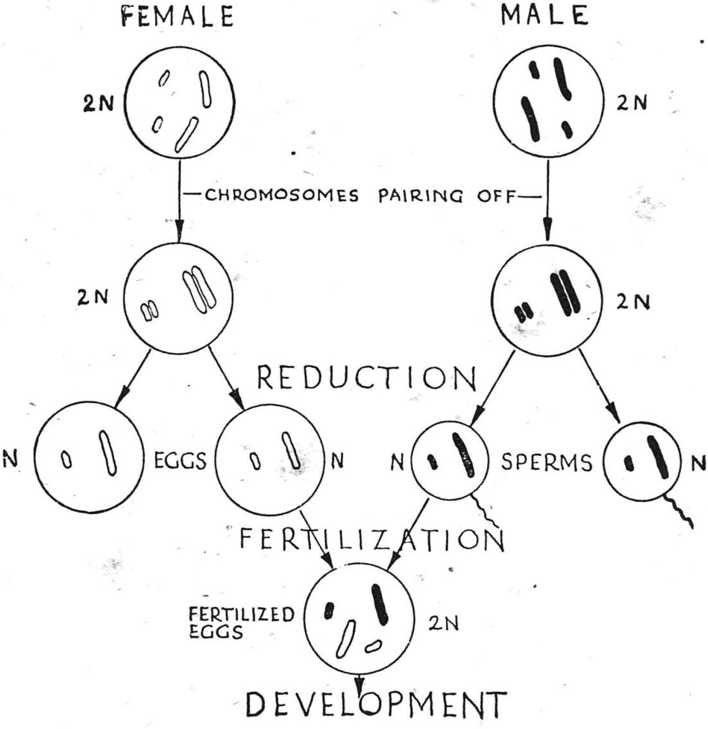


FIG. 134. FINAL STAGES OF CELL-DIVISION OF SEX-CELLS

(For simplicity, this animal is supposed to have only two pairs of chromosomes per cell. The chromosomes of the male are shown in black, and those of the female in white.

CELL-DIVISION OF SEX-CELLS

In the higher living things, special sex-cells—egg-cells and sperm-cells—are set apart for the purpose of *sexual reproduction*. Fertilization takes place when the nucleus of a male sperm-cell fuses with the nucleus of a female egg-cell. Hence the fertilized egg-cell will have *a double set of chromosomes, one set from the female parent and one set from the male parent*, so that if the sex-cells had the same number of chromosomes as the ordinary body-cells, then the body-cells of the offspring from sexual reproduction would have twice as many chromosomes as the body-cells of either parent. That is, the chromosome-number would be doubled in every new generation. *This is avoided by a special type of cell-division which halves the number of chromosomes during the formation of the sex-cells.* In Man, for example, during one of the last cell-divisions just before the formation of the ripe sperm-cells, the 24 pairs of chromosomes do not split lengthwise (as in ordinary cell-division) but *one whole chromosome from each pair* goes to either end of the cell; and when cell-division is complete, each ripe sperm-cell has only 24 chromosomes instead of the usual 48. A similar *reduction-division* yields ripe egg-cells with only 24 chromosomes. Hence, when fertilization takes place, the fertilized egg has 48 chromosomes, 24 from each parent. The fertilized egg then develops into a new individual by the ordinary process of cell-division, every body-cell containing 48 chromosomes, *of which* 24 *come from each parent.* Hence the new individual inherits some of its characteristics from the male parent and some from the female parent.

INHERITANCE OF CHARACTERISTICS

We are now able to give some explanation as to (*a*) why living things are like their parents (although not *exactly*), and (*b*) why children of the same two parents often differ from one another. This explanation is based partly on actual observations of the chromosomes in cells under the microscope and partly on the results of breeding experiments with plants and animals.

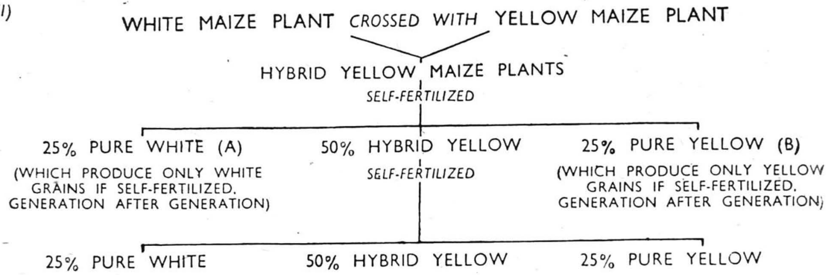
The first man to discover the essential facts of inheritance was an Austrian, *Mendel*, about 1865, long before anyone knew about chromosomes and the microscopic details of cell-division. His earliest breeding experiments were carried out with different varieties of garden Pea, but we shall discuss similar experiments with Maize, which is particularly convenient for breeding experiments since it has separate male and female flowers, and if the female flower is enclosed in a paper bag before its stigmas are ripe, pollination by wind or insects can be prevented.

If the stigmas of a ' bagged ' flower are pollinated by hand with pollen from the male flower on the same plant, self-fertilization is ensured. If a ' bagged ' flower is pollinated by hand with pollen from a different maize plant, then cross-fertilization takes place and both parents are known.

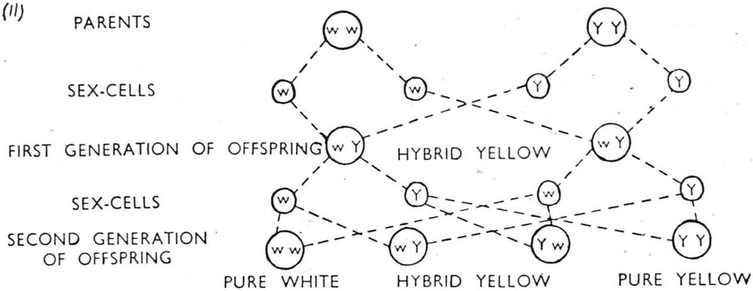
There are different pure-bred varieties of maize, some bearing yellow grains and some bearing white grains. If a pure variety of yellow maize is self-fertilized, all the seeds will give rise to maize plants also bearing yellow grains. Similarly, self-fertilized seeds from a pure variety of white maize will all give rise to maize plants bearing white grains. That is, these pure varieties, if self-fertilized, *breed true to the parental type.*

Now suppose that pollen is transferred by hand from the male flower of a yellow maize plant to a ' bagged ' female flower of a white maize plant. The white maize grains which result from this cross-fertilization, if planted, will give rise to maize plants bearing *yellow grains only.* The maize plants of this first generation of ' cross-bred ' offspring, which have arisen from parents differing in the characteristic colour of their grains, are called *hybrids.*† At first sight, it would seem that the inheritance-factor for white grains (or what we may call the ' white ' gene) has been lost in the first generation of hybrid offspring, but further experiments show that it is still present although ' hidden ' by the ' yellow ' gene. If these hybrid yellow maize grains are planted, and the resulting maize plants of the next generation are then self-fertilized, the maize-cobs which are formed bear *both yellow and white grains* in the proportion of 75 per cent yellow to 25 per cent white. Now, if all these maize-grains are planted and the resulting plants of yet another generation are again self-fertilized, *all the white maize plants bear only white cobs* (A), *one-third of the yellow maize plants bear only yellow cobs* (B), *while the remaining two-thirds of the yellow maize plants bear mixed cobs with* 25 *per cent white grains and* 75 *per cent yellow grains.* If all the seeds from this generation are planted and the flowers are again self-fertilized, the white maize plants from the unmixed white cobs (A) continue to ' breed true ', i.e. they bear only white grains: the yellow maize plants from the unmixed yellow cobs (B) also ' breed true ', bearing only yellow grains; while the self-fertilized maize plants from the mixed cobs behave as follows:—all the white maize plants bear pure white cobs; one-third of the yellow maize plants bear pure yellow cobs; while the remaining two-thirds of the yellow maize plants bear mixed cobs (25 per cent white; 75 per cent yellow) as before. A diagram will make all these results much clearer:—

(I)



(II)



The explanation of this (*i*) is simple. Every cell of a pure-bred yellow maize plant has 2 *yellow genes*, and every cell of a pure-bred white maize plant has 2 *white genes*. Diagram (*i*) can now be reconstructed in terms of genes as shown above (*ii*), representing a yellow gene by ' Y ' and a white gene by ' w '.

The character which appears in the first hybrid generation (yellow in this case) is called the *dominant gene* (Y), while the character which it ' hides ' (white in this case) is called the *recessive*† *gene* (w). It is very important to notice, however, that the recessive gene is merely ' hidden ' —it is still present in every cell of every hybrid offspring and can always reappear as a ' pure recessive ' in a later generation if another recessive gene is supplied during fertilization.

It is now clear why, in *sexual reproduction*, the offspring is not exactly like either parent. It gets half its genes from one parent and half from the other, but some of these genes are dominant and some are recessive, so that *sexual reproduction provides an opportunity for getting new varieties of plants and animals.* A new variety which is well provided for the struggle for existence will have a better chance of surviving and reproducing itself than a less well provided new variety, so that *evolution* can take place through *natural selection.*

These principles underlying the inheritance of characteristics are applied by Man to produce new and more useful varieties of crop plants and domestic animals. Improved varieties of wheat, sugar-cane, potatoes, etc., have been obtained by crossing two widely different varieties of the same plant and then raising self-fertilized generations from this first hybrid generation. In this way thousands of different re-combinations of genes are obtained from which new varieties with desirable new characteristics are produced by *artificial selection.*

SECTION III

MECHANICS

CHAPTER XI

FORCE, WORK, AND ENERGY

The scientific meanings of the terms *force*, *work*, and *energy* are different from their common everyday meanings and we must now learn how the terms are used in science.

FORCE

We have already dealt with several different kinds of *force*:—(i) The *force of gravity*. We have found the *weight* of bodies by measuring the *pull* of the Earth upon them. (ii) *Magnetic force*. We have seen that the isolated N-pole of a bar-magnet exerts a force of *attraction* on the S-poles of other bar-magnets and a force of *repulsion* on their N-poles. (iii) The force of *thrust* (or the pressure of a fluid) on an area. (iv) The force of *tension* in a stretched spring, e.g. in a spring balance.

It is difficult to explain exactly what a force *is*, and up to the present we have regarded a force as being either a *push* or a *pull*. We know, however, what a force *does*, and it is the *effects* produced by forces which concern us in everyday life.

A force may cause a body[1] *to move from rest.*

A force may cause a moving body to move faster or slower (*or stop its moving*).

A force may change the direction of a moving body.

A force may set up a strain[2] *in an elastic body which is unable to move.*

We shall consider these effects in later lessons.

THE MEASUREMENT OF FORCE

The units of force which we shall use are the *pound weight* and the *gram weight*, i.e. the pull of the Earth on a pound weight or a gram

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[1] In science, a ' body ' means ' a piece of matter '.

[2] In this case, cause a slight change in its shape and size.

weight. The simplest method of measuring forces is by means of a spring-balance.

TO SHOW THE RELATION BETWEEN THE FORCE APPLIED TO A SPRING AND THE EXTENSION PRODUCED—HOOKE'S LAW

Set up the given spring and metre rule as shown in Fig. 135. A piece of bent sheet-metal, attached to the lower end of the spring by wax, serves as a pointer for measuring changes in length. Take the reading of the pointer when there is *no load* on the scale-pan and record it below. Then place a 10 gm. weight on the scale-pan and take the new reading. Repeat this, increasing the load by 10 gm. each time until the total load is .... gm.[3] Then take off the load, 10 gm. at a time, and notice whether the extension is the same as before.

Draw a graph by plotting *load* against *extension.* The graph is a straight line, showing that *the extension produced is directly proportional to the load* (i.e. the force applied to the spring). This relationship is known as *Hooke's Law,* after its discoverer, Robert Hooke (1635-1703).

The *spring-balance* makes use of this principle that, when loaded, a spring stretches uniformly (provided that it is not over-stretched). In the making of a 100 gm. spring-balance, therefore, the weight-scale is marked zero opposite the lower end of the spring when there is no load on the pan or hook. A 100 gm. weight is added and the new position of the lower end of the spring is marked 100 gm. The distance between the two marks is then divided into a hundred equal parts, each corresponding to a load of one gram weight.

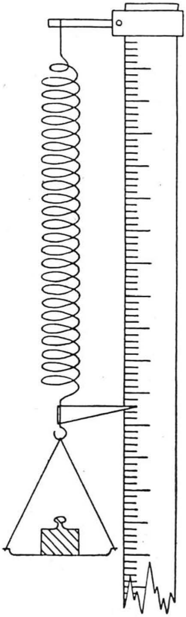

FIG. 135.
MEASUREMENT OF FORCE

A spring is said to be *elastic,* since it recovers its original shape and size when the stretching force is removed. Other very elastic substances are rubber and ivory; but lead and damp clay have very little *elasticity,* since they do not recover their original shape after bending or stretching.

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[3] Your teacher will tell you the maximum load. With stiffer springs it will be necessary to increase the load by more than 10 gm. at a time.

Load on Scale-pan	Reading on Metre Rule	Extension produced by last Increase	Increase in Load	Decrease in Load
Pan only	cm.	nil	nil	nil
gm.wt.	cm.	cm.	gm.wt.	gm.wt.
gm.wt.	cm.	cm.	gm.wt.	gm.wt.
gm.wt.	cm.	cm.	gm.wt.	gm.wt.
gm.wt.	cm.	cm.	gm.wt.	gm.wt.
gm.wt.	cm.	cm.	gm.wt.	gm.wt.
gm.wt.	cm.	cm.	gm.wt.	gm.wt.
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gm.wt.	cm.	cm.	gm.wt.	gm.wt.
gm.wt.	cm.	cm.	gm.wt.	gm.wt.

WORK

In common everyday language, the term *work* means anything that involves ' labour '. The scientific definition of *work*, however, is quite different:—*When a force MOVES a body, it is said to do work*, and the *quantity of work* done is measured by *the product of the force and the distance through which it moves*, i.e. **work = force × distance.** For practical and engineering purposes, the *unit of work* is the *foot-pound*, and this can be defined as *the work done in lifting a weight of one pound through a height of one foot*. Thus, if you lift a 10 lb. weight from the floor and put it on a table 3 ft. high, you have done 30 foot-pounds of work.

Notice carefully that, in the scientific sense, no work is done until the force succeeds in *moving* the body on which it acts. Thus, if you push against a heavy desk you are exerting a force on the desk and may, indeed, become tired with pushing, but you do not perform any *work* until the desk *moves*. It is important to remember this difference between the uses of the term *work* in the everyday sense and in the scientific sense.

ENERGY

When anything is able to *do work*, it is said to possess *energy*, i.e. *energy is the capacity for doing work* and it is measured in the same units as work. This idea of *energy* is one of the most important in the whole of science. We have already learnt something about several different kinds of energy, e.g. heat energy, light energy and chemical energy, and we shall soon learn more about energy.

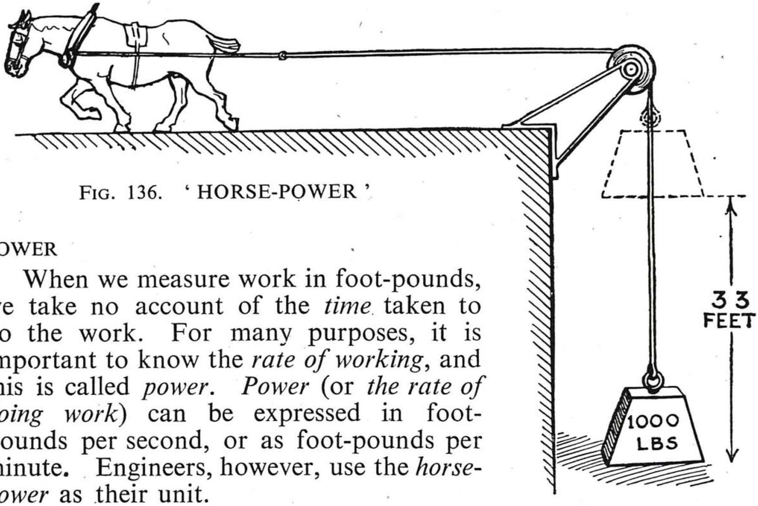


FIG. 136. 'HORSE-POWER'

POWER

When we measure work in foot-pounds, we take no account of the *time* taken to do the work. For many purposes, it is important to know the *rate of working*, and this is called *power*. *Power* (or *the rate of doing work*) can be expressed in foot-pounds per second, or as foot-pounds per minute. Engineers, however, use the *horse-power* as their unit.

This British unit of power was introduced by James Watt (1736-1819), the inventor of the modern steam-engine, who considered that an average horse could do 33,000 foot-pounds of work per minute (or 550 foot-pounds per second). *An engine is developing one horse-power when it is doing work at the rate of* 33,000 *foot-pounds per minute.*

Measure your own horse-power as follows:—Run up a long flight of steps as fast as you can while someone with a stop-watch takes the time as you reach the bottom step and as you pass the top step. In this period, you have raised your own weight through the vertical height of the steps and can now calculate your *rate of working*:—

Weight= . . . . lb. Number of steps= . . . . ft.
Height of one step = ... in.
Therefore, vertical height of . . . . steps = ... ft.
Time taken = ... sec.

$$\text{Horse-power developed} = \frac{\text{wt. in lb.} \times \text{height in ft.}}{\text{time in sec.} \times 550} = \frac{\quad\times\quad}{\quad\times\quad} = \text{. . . . h.p.}$$

This will be nearly your maximum rate of working, which you can only maintain for a few seconds. Over a longer period, your rate of working would be only about 0·1 to 0·2 h.p.

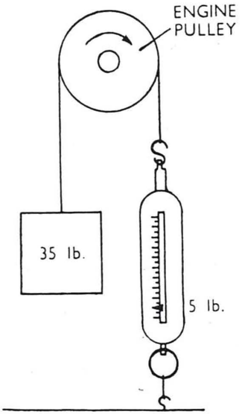


FIG. 137.
MEASUREMENT OF POWER

The *metric and electrical unit of power* is called the *watt*† (after James Watt) and the power of electric motors is usually expressed in *Kilowatts* (1 Kilowatt = 1,000 watts). 1 horse-power = 746 watts (or nearly $\frac{3}{4}$ Kilowatt), i.e. 1 Kilowatt = $1\frac{1}{3}$ horse-power (nearly). We shall discuss later the relation between the watt and other electrical units.

The rate at which an engine or motor works can be found as shown in Fig. 137. A belt or cord is placed over a pulley on the engine, one end of the belt carrying a weight of, say, 35 lb., and the other end being attached to a spring-balance which, in turn, is fastened to the ground or to the bench. When the engine is at rest, the spring-balance will read 35 lb.; but when the engine is running, the friction between the pulley and the belt will tend to* raise the weight, and the spring-balance will read less, say 5 lb. If we know the speed of the pulley, say 1,100 revolutions per minute, and its circumference, say 1 foot, we can find the rate at which the engine is doing work.

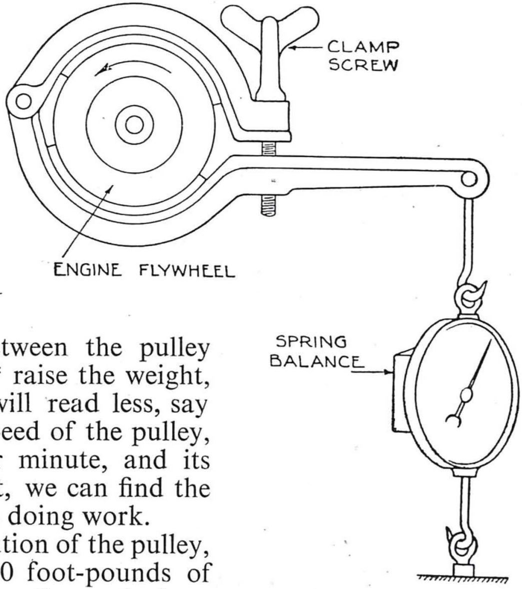


FIG. 138.
BRAKE-APPARATUS FOR MEASURING HORSE-POWER

In one complete revolution of the pulley, the engine does 35—5 = 30 foot-pounds of work against friction, hence the work done per minute is 30 × 1,100 = 33,000 foot-pounds per minute, i.e. the engine develops 1 h.p.

The power of an engine measured in this way is called its *brake horse-power* (B.H.P.). Fig. 138 shows one form of practical brake-apparatus for this purpose.

ENERGY

In the earlier parts of this Course we have often referred to different forms of *energy*; in fact, at least one-third of the whole Course is devoted to *physics*, which is *the study of the effects of different forms of energy on matter.* In biology lessons, too, we have seen that the fundamental living processes of *photo-synthesis* and *respiration* are transformations of energy.

We have defined *energy* as *the capacity for doing work*, and we have seen that it is measured in units of work, and also that work always involves movement. Hence the only form of energy which can be measured *directly* in work-units is *energy of motion* (which is sometimes called *mechanical energy* or *kinetic*† *energy*). *The kinetic energy of a moving body is the amount of work it will do before it is brought to rest.* For example, the kinetic energy of moving air (wind) can be made to do useful work in driving sailing-ships and in turning windmills; while moving water (streams and rivers) can be made to turn water-wheels and water-turbines.

There are many things, however, which can be made to do work although they are not moving.

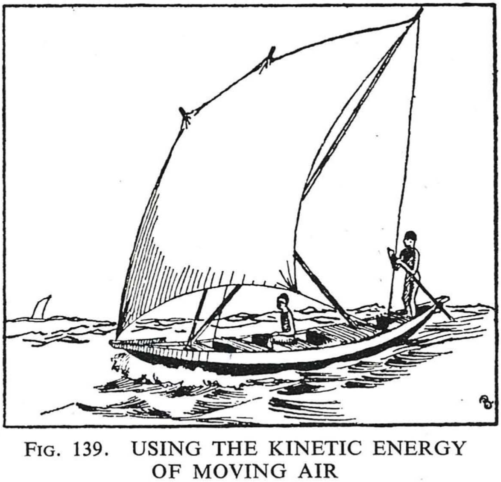


FIG. 139. USING THE KINETIC ENERGY OF MOVING AIR

For example, a wound-up clock-spring contains enough energy to keep the clock-work moving for several days; the explosive substance in a rifle cartridge contains enough energy to move the bullet at a very high speed and for a long distance through the air. This stored-up energy, which is not due to motion (but which can be converted into kinetic energy) is called *potential*[†] *energy. Kinetic energy*, therefore, is energy due to motion, while all other forms of energy are included under *potential energy*. For example, a 10 lb. weight on a table 3 ft. above the floor has potential energy because by using a suitable machine it can be made to do 30 foot-pounds of work as it falls to the floor. This potential energy of a raised weight can be called *energy of position*. Large clocks are sometimes driven by falling weights, e.g. if the clock-weight weighs 10 lb. and is raised through a height of 3 ft. when we wind up the clock, we have done 30 foot-pounds of work and the raised weight now possesses 30 foot-pounds of *potential energy* due to its position and can do 30 foot-pounds of work in driving the clock-work as the weight falls slowly to its original position.

When a small clock or watch is wound up, work is done in compressing the elastic spring, and the spring gains potential energy equal to the work done in winding it up. While the clock is working, this potential energy, stored up in the compressed spring, is gradually transformed into the kinetic energy of the moving parts of the clock.

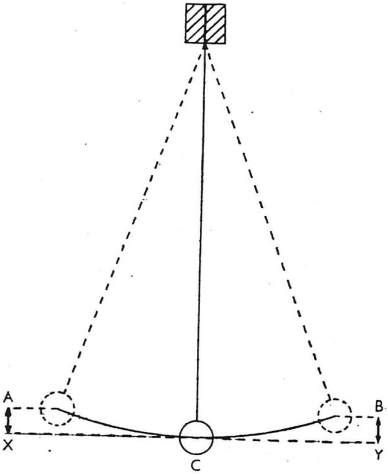


FIG. 140. POTENTIAL AND KINETIC ENERGY

The *pendulum* shows clearly this transformation from potential energy to kinetic energy and back again. As shown in Fig. 140, when the pendulum 'bob' is at rest at C, it possesses neither kinetic energy nor potential energy since it has no motion and is as low as it can be. If we pull it aside to position A, we do work in raising it a vertical height XA, and this work is stored up in the bob as potential energy. If we release the bob, it swings along the arc ACB. When the bob reaches the point C, at the bottom of its swing, the whole of its potential energy is converted into kinetic energy, which is then re-converted into potential energy once more by the time the bob reaches the point B. At A and B, therefore, the energy of the swinging pendulum bob is entirely *potential*. As it passes C, its energy is entirely *kinetic*. At points between A and C and between C and B, its energy is partly kinetic energy and partly potential energy, the sum of the two being equal to the original potential energy when at A.

WATER-WHEELS AND TURBINES

In the same way, the water at the top of a waterfall possesses *potential energy* owing to its position. If a water-wheel is placed in the path of the falling water, the water falls into the buckets at the top of the wheel and the potential energy of the water is transformed into the kinetic energy of the moving wheel before the buckets allow the water to escape at the bottom of the wheel. Practical water-wheels (which have now been largely displaced by water-turbines and Pelton wheels for large-scale purposes) are of two main types:—*undershot wheels* and *overshot wheels*, as shown in Fig. 141. The *undershot wheel*, used in flat country, is placed with its lower part in the running water, and some of the *kinetic energy* of the moving water (usually about 25 per cent) is transferred to the wheel. In the *overshot wheel*, used in hilly country, the running water falls over the top of the wheel so that use is made of both the *kinetic energy of the running water* and the *potential energy of the falling water*. This is a much more efficient arrangement than the undershot wheel, and over 80 per cent of the potential energy of the water can be converted by an overshot wheel into useful work.

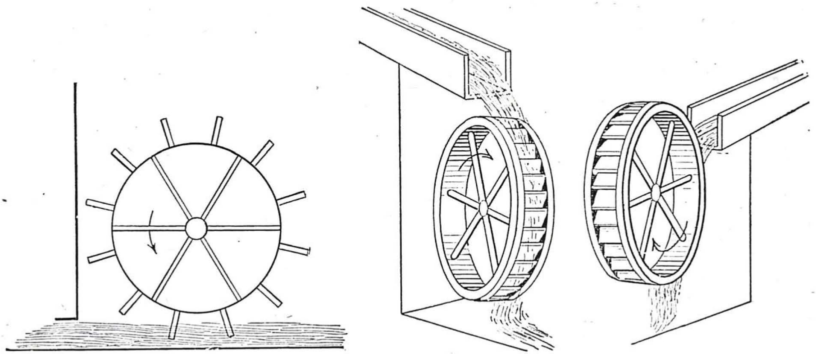


FIG. 141. WATER-WHEELS

In hydro-electric power-stations these old-fashioned water-wheels have been displaced by *water-turbines* and by *Pelton wheels*, which are more efficient still. Thus, some of the 100,000 h.p. turbines at the

Niagara Falls have an efficiency of over 90 per cent. Water-turbines are used where the ' head ' of water is only a few hundred feet, and they run at a low speed. When the ' head ' of water is over 500 ft., a Pelton wheel as shown in Fig. 142 is used. The Pelton wheel is a special type of undershot wheel in which the high-pressure water is delivered from a jet and strikes the cup-shaped buckets at a very high speed. The buckets are shaped so that they return the water in the direction opposite from that in which it strikes the wheel; thus, when the wheel is revolving, if the speed of the buckets is half the speed of the jet of water, most of its kinetic energy (over 85 per cent) is given up to the wheel (Fig. 142). The largest Pelton wheels used in hydro-electric power-stations develop about 50,000 h.p.

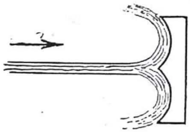


FIG. 142.
PRINCIPLE OF PELTON WHEEL

OTHER FORMS OF ENERGY

Energy can exist in several different forms, and each kind of energy can be transformed into any other kind. Some other kinds of energy are *heat energy*, *light energy*, *sound energy*, *electrical energy* and *chemical energy*, and these can all be converted into each other. For example, green plants absorb *light energy* and store it up in carbo-hydrates as *chemical energy*. When plant remains (e.g. wood or coal) are burned under a boiler, this chemical energy is converted first into *heat energy* and then into the *potential energy* of high-pressure steam. If this high-pressure steam passes into a steam-turbine, its potential energy is converted into *kinetic energy* as it passes through the turbine, and this kinetic energy is given up to the moving parts of the turbine. If a

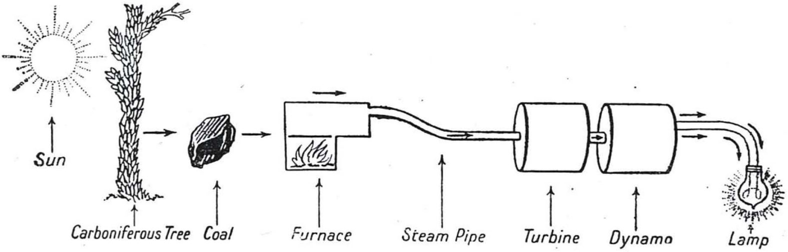


FIG. 143. TRANSFORMATIONS OF ENERGY

dynamo is coupled to the turbine, this kinetic energy (or mechanical energy) is converted into *electrical energy.* This electrical energy can be converted either into *light* and *heat* (in an electric bulb), or into *kinetic energy* (by an electric motor), or into *sound* (in an electric bell or electric gramophone), or into *chemical energy* (in a lead accumulator or during electrolysis).

It is possible to change all of these different kinds of energy into each other, although in practice some changes are easier than others. Notice particularly that electrical energy, even after travelling for long distances through wires, can be changed easily and directly into all the other forms of energy: this is one reason why electricity is so useful in everyday life.

CONSERVATION OF ENERGY

Careful experiments have shown that *in all these transformations of energy, no energy is ever lost or destroyed but is only changed into other forms of energy.* This principle is called *the Law of the Conservation of Energy*, i.e. *energy cannot be destroyed*, and the total amount of energy in the universe* always remains the same; it only changes from one form into another and finally into *heat.* This *Law of the Conservation of Energy* is the basis of quantitative *physics*, just as the *Law of the Conservation of Matter* is the basis of quantitative *chemistry.*

Although every transformation of energy yields an equivalent amount of other forms of energy, in actual practice it is seldom possible to *use* all the energy or reverse the transformation exactly, because of unavoidable losses. For example, no practical heat-engine converts *all* the heat of the working substance into mechanical energy; the used steam or exhaust gases still contain some heat energy which merely warms up the atmosphere and which cannot be recovered. This is *wasted energy* but it is *not* destroyed, although it is no longer available for doing *useful work.*

Even the most carefully-made machines always waste some energy in *friction* and *noise*, which finally yield heat energy at too low a temperature to be of practical use.

When the driver of a motor-car wishes to stop quickly, he has to change kinetic energy into some form of potential energy, and he does this by applying the brakes. The brake-drums get hot owing to the conversion of kinetic energy into heat. In theory it should be possible to start the car again by re-converting this heat into kinetic energy, but in practice this heat escapes into the atmosphere and a fresh supply of kinetic energy is obtained by burning petrol in the cylinders of the engine and transforming the heat energy into energy of motion.

When a raised weight falls freely, all its potential energy is converted into kinetic energy by the time it reaches the ground. On striking the ground, this kinetic energy is not destroyed but is changed into *heat*. In the same way, the water in a pool at the bottom of a waterfall possesses less potential energy than before it fell, but its temperature is slightly higher because this potential energy has been changed into heat.

CHAPTER XII

MACHINES

Man, in his primitive state, had to depend upon his own unaided muscles like all the other animals. It has taken about 20,000 years for Man to reach his present state of civilization and his dominant position in the world. He owes much of this advance to learning gradually how to use *tools* and *machines*. By these he can produce and control forces which, if he used his hands alone, would be far beyond his powers. The term *machine* is applied to anything by means of which a man can apply a force more conveniently and more usefully than with his unaided muscles. A bicycle, for example, is a machine which enables your leg muscles to move your body farther and faster and more easily than by walking.

PULLEYS AND LIFTING-TACKLE

If you want to raise a heavy object to the top of a building there are several ways of doing it. One way is to carry the object up the stairs or up a ladder, but this involves raising your own body as well as the object, thus 'wasting work'. Thus, if a man weighing 160 lb. carries a load of 40 lb. up a ladder to a height of 50 ft. he does 10,000 foot-pounds of work, 2,000 foot-pounds of which is 'useful work' and the remaining 8,000 foot-pounds is 'wasted work'.

Another way is to stand at the top of the building and pull up the object with a rope. Nearly all this work is 'useful work', but it is usually more *convenient* to pass the rope over a *pulley*† fixed to the top of the building and pull *down* on

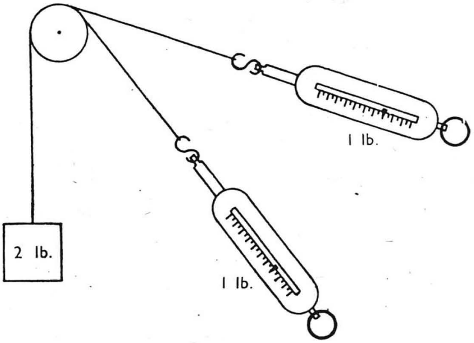


FIG. 144. SINGLE FIXED PULLEY

the rope while standing on the ground. To lift the 40 lb. load 1 ft. you would still have to do 40 foot-pounds of work by pulling the rope downward with a force of 40 lb., i.e. *a single fixed pulley merely changes the direction of the pull*, making it more *convenient*, but it gives *no mechanical advantage*.

Pass a string over a single fixed pulley as shown in Fig. 144. Hang a 2 lb. weight on one end and fasten a spring-balance to the other. Hold the spring-balance in different positions so that the string attached to it makes different angles with the vertical. What is the reading of the spring-balance in each case ? ....lb.

In everyday life, it is often necessary to lift a big weight (or *load*) by applying a small force (or *effort*) so as to get a *mechanical advantage*, which is defined as :—

$$\text{Mechanical advantage} = \frac{\text{load}}{\text{effort}}$$

The *mechanical advantage* (or *force ratio*) of any machine, therefore, is *the ratio of the load to the effort.*

If we turn the last arrangement (Fig. 144) upside down, as shown in Fig. 145, we can increase the effect of our pull. *Half* the 2 lb. weight

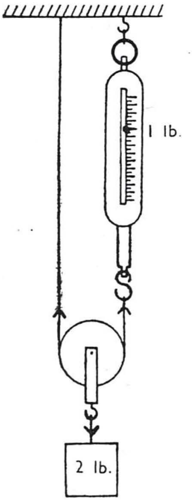


FIG. 145. SINGLE MOVABLE PULLEY

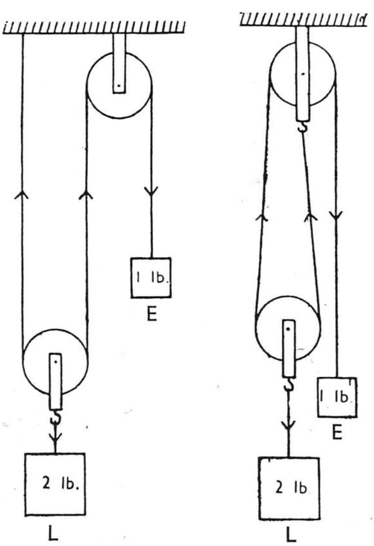


FIG. 146. SIMPLE TACKLE

is now supported by one string and *half* by the other string. The *effort* *E* required to hold the *load L* in place is therefore only *half* the load = 1 lb. Verify this by experiment, measuring the force *E* with a spring-balance as shown in Fig. 145.

Such an arrangement of pulleys is called a *tackle*,† and in this case we can lift a load twice as great as the effort applied to the string, i.e. this tackle gives us a *mechanical advantage* of 2. In practice, the arrangement shown in Fig. 145 is inconvenient because it requires an upward pull. We have seen, however, that a single fixed pulley will change the direction of the pull, hence the arrangement shown in Fig. 146 is more convenient.

Fit up the arrangement shown in Fig. 146.

(i) First balance the movable pulley with hanging weights. Then hang a 2 lb. weight from the movable pulley and find what weight is required at *E* to balance this load. . . . . .lb.

What is the mechanical advantage? . . . .

(ii) Measure the distance the load *L* rises when the effort *E* descends one foot. . . . . .ft. The ratio between these two distances is called the *velocity* ratio* of the machine.

$$\text{Velocity ratio} = \frac{\text{distance moved by effort}}{\text{distance moved by load}} = \ldots\ldots = \ldots\ldots \text{in this case.}$$

Notice that in this case (when the movable pulley is counter*-balanced) the *mechanical advantage is the same as the velocity ratio*, but this is only true of a ' perfect ' lifting tackle in which the movable pulleys have no weight and in which all the pulleys turn without friction.

(iii) Now remove the hanging weights you used to counter-balance the movable pulley, and find the effort *E* required to raise a 2 lb. load, using a spring-balance. Effort = ...lb.

$$\text{Therefore, actual mechanical advantage} = \frac{\text{load}}{\text{effort}} = \ldots\ldots = \ldots\ldots$$

The velocity ratio, however, is unchanged. In this case, therefore, and *in all practical machines, the mechanical advantage is less than the velocity ratio* because some work is always wasted (as heat) in overcoming *friction* and in lifting moving parts.

THE PRINCIPLE OF WORK

In Experiment (ii) above (where friction was too small to affect our rough measurements) we see that the product *load × distance moved by load* is the same as the product *effort × distance moved by effort*, i.e. *the WORK done on the load BY the machine = the WORK done by the effort ON the machine.*

We found the same relation in the hydraulic press (Book II, p. 90) where the force acting on the small piston × the distance through which it moved = the force acting on the large piston × the distance through which it moved, i.e. **the work put into the machine is equal to the work got out of the machine** (if there is no friction). This important *Principle of Work* was first stated by Sir Isaac Newton in 1687, and it applies to all machines. In other words, *no machine can do more work than is done ON it*, or (in a ' perfect ' machine) output = input. Hence, although a machine enables us to increase the *force* available, it does *not* enable us to do more *work*.

EFFICIENCY OF MACHINES

In all actual, practical machines, some energy is wasted (as heat) in overcoming friction, i.e. *work put into machine = work got out of machine + work done in overcoming friction*: so that *less useful work* is got out of the machine than is put in. This fraction (usually expressed as a percentage) is called the *efficiency*† of the machine.

$$\text{Efficiency} = \frac{\text{useful work got out of machine}}{\text{total work put into machine}} = \frac{\text{output}}{\text{input}}$$

$$= \frac{\text{load} \times \text{distance load moves}}{\text{effort} \times \text{distance effort moves}} = \frac{\text{mechanical advantage}}{\text{velocity ratio}}$$

Find the efficiency of the arrangement used in Experiment (iii) above.

COMBINATIONS OF PULLEYS

If the number of pulleys is increased we can increase the mechanical advantage of the lifting-tackle. In the arrangement shown in Fig. 147, the load is supported by 6 strings, hence (neglecting friction and the weight of the movable pulleys) the pull in any one string is one-sixth of the load; hence the effort *E* required to hold the load in place is one-sixth of the load, and the theoretical mechanical advantage is 6. In order to raise the load through 1 ft. we have to pull in 6 ft. of rope at *E*, hence the velocity ratio is 6. (N.B.—The velocity ratio and the *theoretical* mechanical advantage of any such system of pulleys can be found by counting the number of strings passing through the *lower* pulley-block.)

Set up the arrangement shown in Fig. 147.

(iv) First balance the hanging pulley-block with weights at E. Then hang 6 lb. from the lower pulley-block and find what weight is required at E to balance this load. $E = \ldots\ldots\ldots\ldots$ lb. Therefore, mechanical advantage $= \ldots\ldots\ldots\ldots$

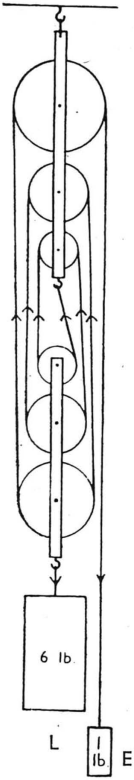


FIG. 147. LIFTING-TACKLE

Verify the velocity ratio by measurement as in Experiment (ii) above $= \ldots\ldots\ldots\ldots$ When the effort E descends through 6 ft. *the work put into the machine* = effort × distance moved by effort $= \ldots\ldots\ldots$ foot-pounds, and *the work got out of the machine* = load × distance moved by load $= \ldots\ldots\ldots\ldots$ foot-pounds.

(v) Now remove the weights used to counter-balance the hanging pulley-block and find what weight is required at E to balance the system. $E = \ldots\ldots\ldots\ldots$ lb. The *total work put into the machine* when the effort E descends through 6 ft. $= \ldots\ldots\ldots \times 6 = \ldots\ldots\ldots$ foot-pounds. The *useful work got out of the machine* during the same time $= 6 \times 1 = 6$ foot-pounds.

$$\text{Efficiency} = \frac{\text{useful output}}{\text{total input}} = \frac{6}{\quad\quad} \times 100 = \ldots\ldots\ldots\%$$

THE SIMPLE WHEEL AND AXLE

This consists of a wheel of large diameter attached to an axle of smaller diameter. A similar machine, with a long handle in place of the wheel (as shown in Fig. 148) is often used for raising buckets of water from wells.

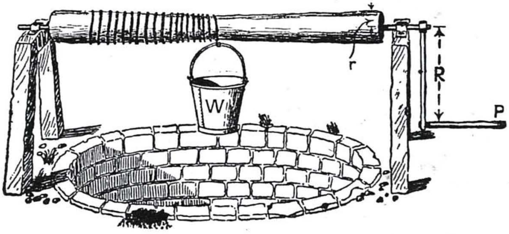


FIG. 148. ‘WHEEL AND AXLE’

The principle of the wheel and axle is shown in Fig. 149. Ropes are wound in opposite directions round the wheel and the axle, the rope round the axle B carrying the load W, while the effort P is applied to the rope wound round the wheel A. In one complete revolution of the wheel and axle, *the work PUT INTO the machine = effort × circumference* *of wheel*, and *the work GOT OUT of the machine* = *load* × *circumference of axle.*

Therefore, *by the Principle of Work,*

load (W) × *circumference of axle* = *effort* (P) × *circumference of wheel.*

Therefore,

$$\text{Mechanical advantage}=\frac{\text{load}}{\text{effort}}=\frac{W}{P}=\frac{\text{circumference of wheel}}{\text{circumference of axle}}$$

$$=\frac{\text{radius of wheel}}{\text{radius of axle}}=\frac{R}{r}$$

During one complete revolution of the wheel and axle, the effort P descends a distance equal to the circumference of the wheel and the load W rises through a distance equal to the circumference of the axle.

Therefore,

$$\text{Velocity ratio}=\frac{\text{circumference of wheel}}{\text{circumference of axle}}=\frac{\text{radius of wheel}}{\text{radius of axle}}=\frac{R}{r}$$

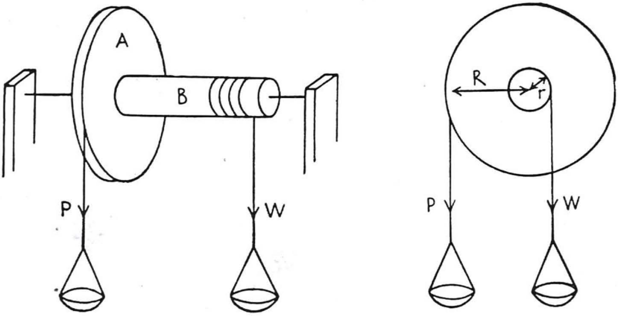


FIG. 149. PRINCIPLE OF WHEEL AND AXLE

Set up the arrangement shown in Fig. 149.

(*a*) Hang a 4 lb. weight from the string wound round the axle and hang weights from the string wound round the wheel until it descends slowly but steadily. Effort P = ...lb.

(*b*) Pull down the string wound round the wheel until the load W rises 4 in. exactly and measure the distance through which the effort P has descended. Load rises . . . . .in. Effort descends . . . . .in.

$$\text{Efficiency}=\frac{\text{useful output}}{\text{total input}}=\frac{\text{load}\times\text{distance load moves}}{\text{effort}\times\text{distance effort moves}}=\frac{\quad\times\quad}{\quad\times\quad}100$$

$$=\ldots\ldots\ldots\ldots\%$$

Since the mechanical advantage $=\frac{\text{radius of wheel}}{\text{radius of axle}}$, by having a very large wheel (or a very long handle) and a very thin axle, a very small effort would lift an enormous load *if the thin axle would stand it.* The same result is obtained with the *differential*[†] *wheel and axle* shown in Fig. 150, where the axle has two halves of different thicknesses and the rope *winds up* round the thicker part A (of radius r_1) as it *unwinds* from the thinner part B (of radius r_2). Hence, in one complete revolution of the machine, the effort P moves through $2\pi R$ (where R is the radius of the handle) and the rope shortens by $2\pi r_1 - 2\pi r_2$, i.e. the difference between the circumferences of A and B. But as the load W hangs from a movable pulley, it rises only *half* the length of rope wound up.

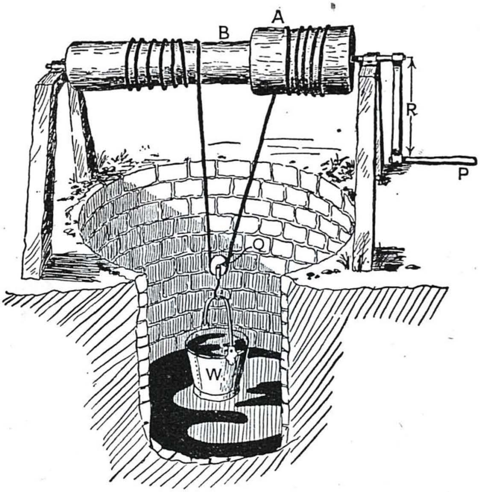


FIG. 150. DIFFERENTIAL WHEEL AND AXLE

Hence, *work done on load* $= W \times \frac{2\pi(r_1 - r_2)}{2}$

$= W\,\pi\,(r_1 - r_2)$

And *work done by effort* $= P \times 2\pi R$

Hence, *by the Principle of Work*, $W \times \pi(r_1 - r_2) = P \times 2\pi R$

Therefore,

velocity ratio (and *theoretical* mechanical advantage W/P) =

$$\frac{2\pi R}{\pi(r_1 - r_2)} = \frac{2R}{r_1 - r_2}$$

Set up a differential wheel and axle and find its actual mechanical advantage and its efficiency.

THE DIFFERENTIAL PULLEY-BLOCK

The same principle is used in the *differential pulley-block* shown in Fig. 151. This consists of two pulleys of slightly different radius r_1 and r_2, *joined together* on the same axle, with an endless chain passing first over one fixed pulley, then round the lower movable pulley (which carries the load) and then up again over the other fixed pulley. All the pulleys have projections to stop the chain from slipping. As in the differential wheel and axle, during one complete revolution of the fixed pulleys the chain between the upper and lower pulleys is shortened by the difference between the circumferences of the two fixed pulleys, i.e. by $2\pi r_2 — 2\pi r_2$, and the load is raised by *half* this amount. Hence, *the work done on the load W* during one complete revolution of the fixed pulleys is $W \times \frac{2\pi r_1 — 2\pi r_2}{2} = W \times \pi(r_1 — r_2)$, and *the work done by the effort* $= P \times 2\pi r_1$ (since the effort P moves a distance equal to the circumference of the larger fixed pulley). Hence, *by the Principle of Work,* $W \times \pi(r_1 — r_2) = P \times 2\pi r_1$

Therefore,

Velocity ratio (and *theoretical* mechanical advantage $W/P) = \frac{2\pi r_1}{\pi(r_1 — r_2} = \frac{2r_2}{r_1 — r_2}$

Hence, the smaller the difference in diameter of the two fixed pulleys, the greater the mechanical advantage of the tackle.

If a differential pulley-block is available, find its actual mechanical advantage and its efficiency.

The efficiency of this type of machine is always less than 50 per cent, but this low efficiency is actually an advantage, because the load does not fall when the effort is removed as in ordinary lifting tackle whose efficiency is over 50 per cent.

FIG. 151.
DIFFERENTIAL
PULLEY-BLOCK

THE INCLINED PLANE

Suppose that two men want to put a heavy drum of oil, weighing 250 lb., into a lorry 4 ft.

from the ground. It will be necessary to lift a weight of 250 lb. to a height of 4 ft., and this will require 1,000 foot-pounds of work. An easier way of putting the drum on the lorry will be to roll it up sloping pieces of wood placed against the side of the lorry to form an *inclined plane*, as shown in Fig. 152.

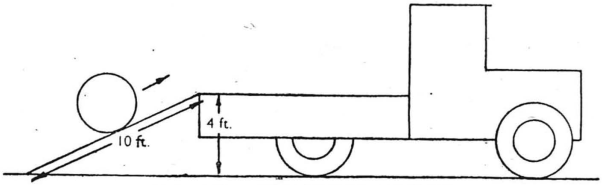


FIG. 152. INCLINED PLANE

Suppose the length of this inclined plane is 10 ft. By the *Principle of Work*, it is necessary to do 1,000 foot-pounds of work, no matter what method is used, hence by using an inclined plane 10 ft. long, the men can get the drum into the lorry by exerting a force of 100 lb. over a distance of 10 ft., which will be much easier than exerting a force of 250 lb. over a distance of 4 ft. by a direct lift.

This method of getting *mechanical advantage* is also used in *wedges* for splitting wood or stone, in *axes* and in *knives*. Suppose the wedge *E* shown in Fig. 153 is 10 in. long and 2 in. wide at its flat end. For every 10 in. it is driven forward into a piece of wood, the wood is forced outwards 2 in. By the *Principle of Work*, this outward thrust *L* will be five times as great as the forward thrust (neglecting friction). The efficiency of such a wedge, however, is usually less than 50 per cent, fortunately for the user. (Why ?)

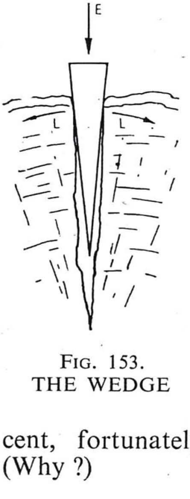


FIG. 153.
THE WEDGE

THE SCREW AS A MACHINE

Screws have many practical applications for getting a mechanical advantage. A screw combines the principle of the

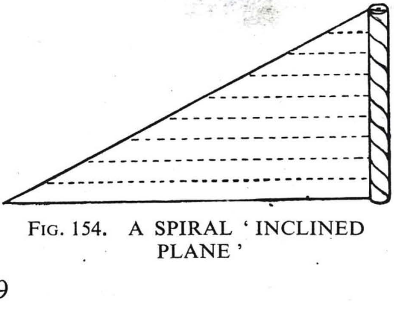
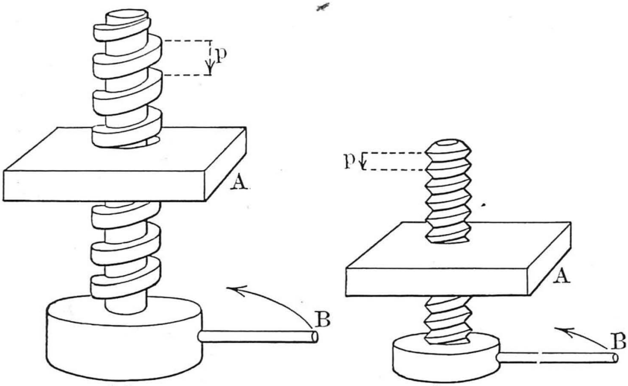


FIG. 154. A SPIRAL 'INCLINED PLANE'

inclined plane and that of the lever. To show that a screw is a spiral inclined plane, cut a rectangular sheet of paper along a diagonal and wrap one of your right-angled triangles round a pencil as shown in Fig. 154. You will notice that the hypoteneuse of the triangle forms a spiral '*screw-thread*' round the pencil.

In any screw, the distance from one thread to the next, measured along the length of the screw, is called the *pitch* of the screw, and this is the distance the head of the screw moves forward into a fixed *nut* when the screw-head is turned through one complete revolution.

In Fig. 155, the effort E acts at the end of the handle B (of length r) and raises a load L (not shown). During one complete turn of the screw through the fixed block A, the work done by the effort $= E \times 2\pi r$, and the work done on the load $= L \times p$ (where p is the pitch of the screw).



(*a*) (*b*)

FIG. 154. SCREWS

Hence, by the *Principle of Work*, $E2\pi r = Lp$

Therefore,

Velocity ratio (and *theoretical* mechanical advantage L/E) $= \dfrac{2\pi r}{p}$

Hence, if the pitch is small (*b*) and the handle is long an enormous mechanical advantage can be obtained. For example, if a load of 2,112 lb. rests on the head of a *screw-jack*† having 4 threads per inch (i.e. of $\frac{1}{4}$ in. pitch) turned by a handle 12 in. long, during one complete turn of the screw the load of 2,112 lb. is raised through a height of $\frac{1}{4}$ in. ($= \frac{1}{48}$ ft.) so that 44 foot-pounds of work is needed (if there is no friction). By the *Principle of Work*, this is equal to the work done on the handle of the screw-jack, i.e. to a force of F lb. acting through a distance of $2\pi \times 12$ in. $=6\cdot28$ ft. Hence, $6\cdot28\ F=44$, i.e. $F=7$. Hence, with a ' perfect ', frictionless screw, a force of 7 lb. acting at the end of the 12 in. handle could raise a weight of 2,112 lb.

In actual practice, however, friction is very great in screws and a much greater force than 7 lb. would be necessary. Although the mechanical advantage of a screw-jack, for example, may be very great, its efficiency is often as low as 25 per cent owing to the work wasted by friction. A screw-jack, therefore, holds its load, however great, when the effort is removed.

If a motor-car screw-jack is available, find its actual mechanical advantage and its efficiency. A suitable ' load ' will be a friend of known weight standing or sitting on a board fitted to the head of the jack. Use a spring-balance to measure the effort which must be applied to the ends of the handle to raise the load. Then remove the ' load ' and find how far the head of the jack moves forward during 10 complete turns of the handle.

What is the velocity ratio of the jack ?..........................

What is its actual mechanical advantage ?....................

What is its efficiency ?..................................

Many practical screw-jacks make use of the differential principle by having a *compound screw*, one part of which has a *right-hand thread* and screws *upward* while the other part has a *left-hand thread* and screws *downward*. The right-hand screw is of slightly larger pitch than the left-hand screw, hence the arrangement is mechanically equivalent to a single right-hand screw *of very small pitch* (which normally has a very fine, weak thread) but the threads of the two parts of the compound screw can be made sufficiently strong to bear a heavy load.

CHAPTER XIII

FORCES IN EQUILIBRIUM

INERTIA

In all our examples of *machines*, the force applied to the machine makes the load *move*. This is one of the most important effects of forces: *producing motion*. In fact, one of the best definitions of force is: *Force is that which changes, or tends to change, the state of rest of a body or its uniform motion in a straight line*. This definition is obtained from Newton's First Law of Motion, which states:—*A body continues in its state of rest or of uniform motion in a straight line, unless it is acted on by a force*. Hence no non-living body at rest can move *of itself*, and no non-living moving body can change its speed or direction *of itself*.

This fundamental property of matter, of resisting any attempt to move it when at rest, or to stop it when in motion, is called *the principle of the inertia*† (or *inertness*) *of matter*.

It is this property of *inertia* that keeps the planets from falling on to the Sun; that makes mud and water fly off at a tangent from a revolving bicycle-wheel; and that makes the Earth's diameter greater at the Equator than at the Poles.

This outwards force, due to inertia, which shows itself when a weight is swung round and round at the end of a string, is usually called '*centrifugal force*'. But for the gravitational attraction of the Sun, the Earth would fly off at a tangent owing to this '*centrifugal force*' instead of continuing to rotate round the Sun.

If you are cycling at a high speed and your bicycle is stopped very suddenly (by a fallen tree, for example), you will be thrown forward over the handle-bars because *a body which is moving tends to continue moving in the same straight line*.

FORCES WHICH PRODUCE NO MOTION

Forces do not always produce movement. Thus, when you sit down on a chair the force of gravity is pulling you downwards (as you notice when someone removes the chair before you sit down !) but your body remains at rest because the upward pressure of the chair exactly balances your weight. These two *equal and opposite forces*, which produce no motion, are said to be *in equilibrium*.†

In the same way, when an object is hung from a spring-balance it descends until the upward pull of the spring is equal to the downward pull of gravity, when it comes to rest, the two forces being balanced or in equilibrium. These examples also illustrate Newton's Third Law of Motion (or the Law of Reactions):—*To every action there is an equal and opposite reaction.* Consider a book resting on a table. It is acted upon by gravity, and if this were the only force acting on the book it would produce motion and the book would fall vertically downwards through the table. The book remains at rest because the table presses vertically upwards on the book with an equal and opposite force.

The forces we are concerned with when dealing with bodies at rest are (*a*) *tensions*† (or pulls), e.g. in strings and muscles; (*b*) *thrusts* (or pushes), e.g. in rigid rods and bones; (*c*) *attractions* and *repulsions*, e.g. the force of gravity and magnetic forces; (*d*) *pressures*, e.g. in fluids; and (*e*) *friction*, i.e. the force which opposes motion when one body slides over another. All these forces have *size* (which we measure in units of weight) and *direction*. They can therefore be represented graphically by straight lines; the *length* of the line representing the *size* of the force, and the *direction* of the line representing the *direction* in which the force acts. (An arrow-head on the line shows which way the force acts along the line.) To represent a force completely, we must also know at what point of the body it acts, because our force-line must start from a definite point. For example, in Fig. 155, the line *AB* represents a force of 15 lb. acting at the point *A* in a north-easterly direction, the scale being 1 in. = 10 lb.

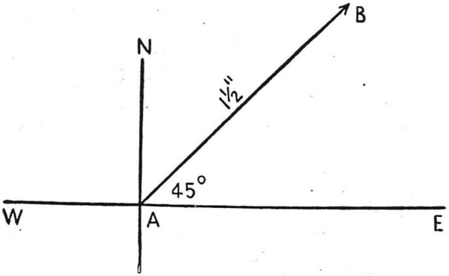


FIG. 156. A FORCE REPRESENTED GRAPHICALLY

CENTRE OF GRAVITY

The Earth's attraction for a body is made up of a large number of little forces applied to the individual molecules of the body. For the purpose of *force-diagrams*, however, it is much more convenient to consider a single downward force applied at a point called the *centre* *of gravity* at which the Earth's pull seems to act. When a body is hung from a thread, it always comes to rest with its centre of gravity vertically below the point of support.

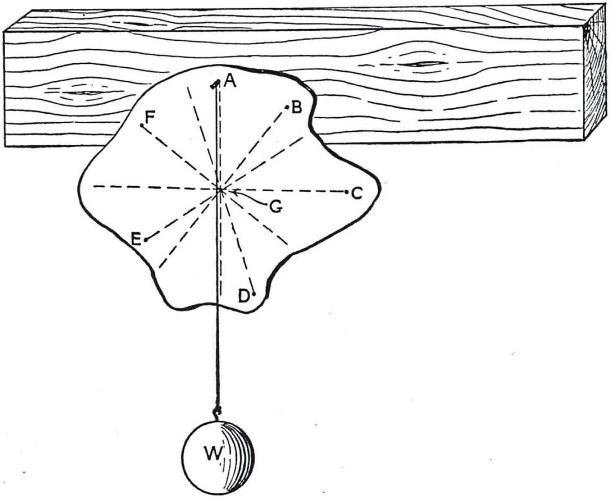


FIG. 157. CENTRE OF GRAVITY

(i) Find the centre of gravity of an irregularly shaped piece of cardboard as follows:—Hang the sheet by a pin, e.g. at *A*. The sheet will hang so that its centre of gravity is vertically below the point of support. If a plumb-line is hung from the pin, the centre of gravity

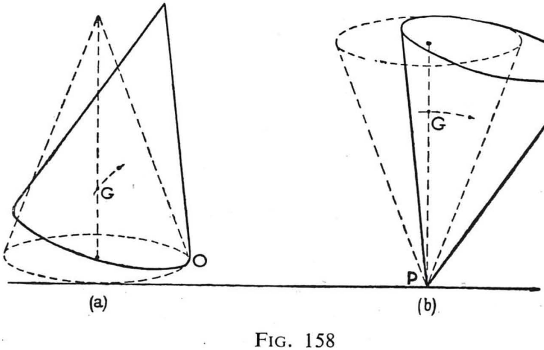


FIG. 158

(*a*) STABLE EQUILIBRIUM (*b*) UNSTABLE EQUILIBRIUM will lie somewhere along this line. Now hang up the sheet by a pin through *B* and draw the vertical line through *B* in the same way. Then the centre of gravity lies somewhere along these two lines. It must therefore be at their point of intersection *G*.

(ii) Trace a map of your country on a sheet of cardboard and then cut round the outline with scissors. Find the best place for a ‘ central ’ broadcasting station (from the point of view of distance only).

(iii) Find the centre of gravity of (*a*) a tripod stand, (*b*) a retort stand.

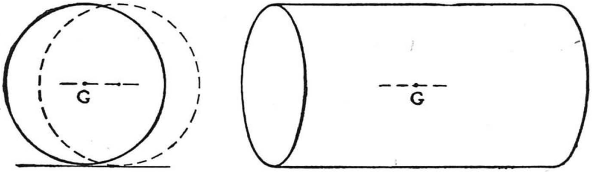


FIG. 159. NEUTRAL EQUILIBRIUM

STABILITY

The *stability*† of a body (i.e. whether it is easy to upset or not) depends upon the position of its centre of gravity in relation to its point of support.

A body is said to be in *stable equilibrium* if it returns to its original position after slight displacement, e.g. a pendulum, a brick resting on its larger face, a cone or a filter-funnel standing on its base. For a body to be in *stable equilibrium, its centre of gravity must be as low as it can get*. See Fig. 158 (*a*).

A body is said to be in *neutral equilibrium* when, after a slight displacement, it remains in whatever position it has been put, e.g. a sphere or cylinder lying on a horizontal surface, a cone or a filter-funnel

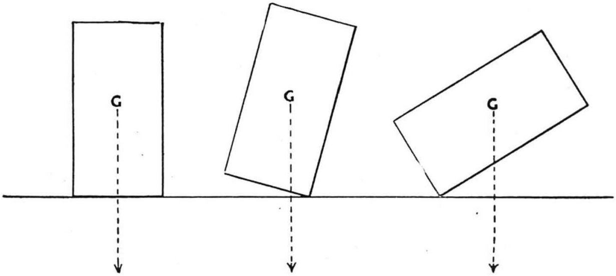


FIG. 160. STABILITY

lying on its side. A body is in neutral equilibrium when its centre of gravity is neither raised nor lowered when it is displaced slightly. See Fig. 159.

A body is said to be in *unstable equilibrium* when, after a slight displacement, it moves still farther from its original position, e.g. a cone balanced on its point. A body is in unstable equilibrium when a slight displacement *lowers its centre of gravity*, i.e. its centre of

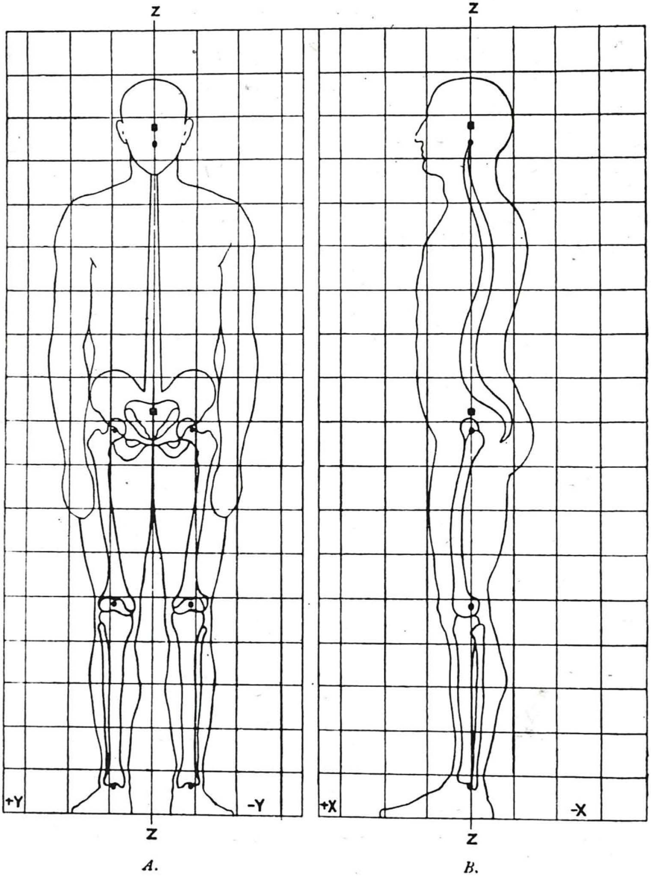


FIG. 161. CENTRE OF GRAVITY OF HUMAN BODY

gravity was originally vertically above the point of support. See Fig. 158 (*b*).

For any object resting on a support to be in *stable* equilibrium, *the vertical line drawn through its centre of gravity must fall within its base.* Stand with your heels against a wall and try to pick up a book from the floor in front of you without bending your knees and without moving your feet. This will enable you to verify the truth of the above principle.

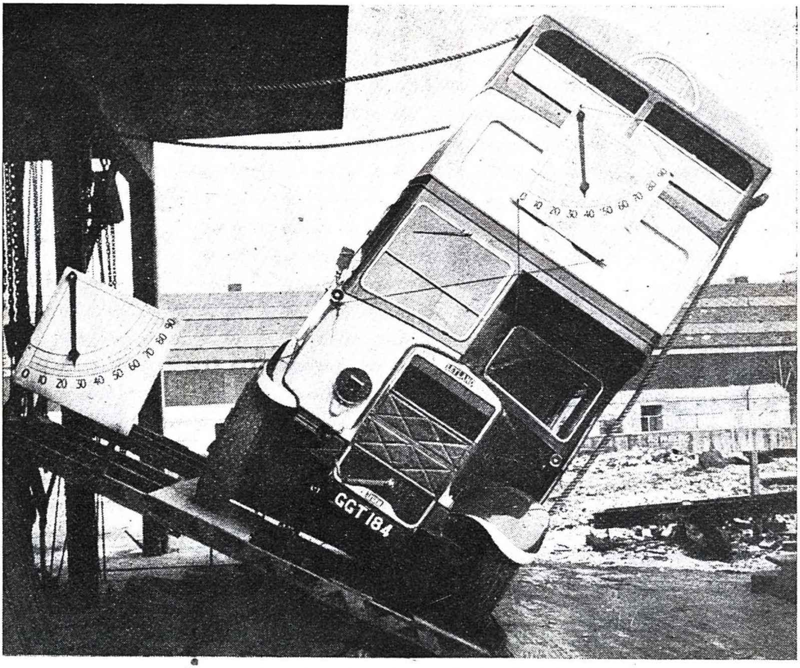


Fig. 162. BUS BEING TESTED FOR STABILITY

Any article which must not upset easily, e.g. a motor-car, is therefore built with the engine and all the heavier parts *as low as possible*, so as to produce a low centre of gravity. Similarly, a car with wide wheel-tracks is less likely to overturn than one with narrower wheel-tracks. Motor-buses are usually tested for stability by putting them on a hinged floor and gradually raising one side until the bus begins to fall over, i.e. until the vertical line drawn through the centre of gravity falls outside the wheel-track. A modern passenger-bus when fully loaded will stand on a sideways slope of 30° without falling over.

COMPOSITION AND RESOLUTION OF FORCES

The resultant† *of two forces is that single force which will produce the same effect as the joint action of the two forces.* This substitution of a single force to produce the same effect as a number of forces acting together is called *the composition of forces.* The substitution of two forces which act together to produce the same effect as a single force is called *the resolution*† *of forces.*

When two forces acting *in the same direction* are applied to the same point, the *resultant force* is equal to the *sum* of the two forces. Similarly, when two forces acting in opposite directions are applied to the same point, the resultant force is equal to the *difference* between the two forces and acts in the direction of the greater force. In other words, **when a number of forces act on a body in the same straight line, their resultant force is the ALGEBRAIC SUM of the constituent forces.** For example, a telephone post supports two wires, one having a tension of 100 lb. and the other a tension of 80 lb. It is clear that if one wire runs north and the other wire south from the post, then the resultant force is a pull of 20 lb. in the direction of the greater pull, i.e. north. If both wires run south from the post, then the resultant force is a pull of 180 lb. in the same direction.

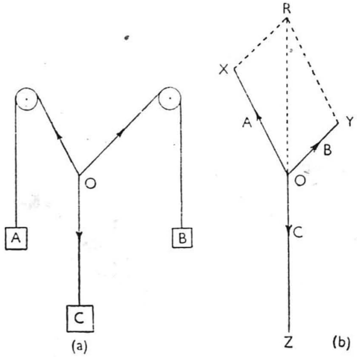


FIG. 163. FORCES IN EQUILIBRIUM

We must now consider forces which do not act in the same straight line, but at an angle to each other.

INCLINED FORCES ACTING AT A POINT

(i) Tie three different weights, *A* (=......lb.), *B* (=......lb.), *C* (=......lb.), to three pieces of string knotted together at *O*, and hang the weights *A* and *B* over two pulleys as shown in Fig. 163. When the system comes to rest we have three forces acting at the point *O*, which are in equilibrium. It is clear that the downward force *C* is balancing the inclined forces *A* and *B*, and the force *C* is called the *equilibrant*† of *A* and *B*.

This force C must be equal and opposite to the *resultant* of A and B.

(ii) Place a sheet of paper behind the strings and mark the direction of all three strings on the paper. Remove the paper and mark off lengths OX, OY, and OZ, proportional to the weights A (......), B (......) and C (......) respectively. Produce ZO to R so that $ZO = OR$. Then OR represents the *resultant* of A and B in size and direction. Join XR and YR and examine the figure $OYRX$. If your pulleys are easy-running, if your strings are light, and if you have done the experiment carefully, $OYRX$ is a *parallelogram*.

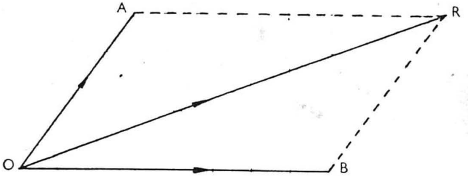


FIG. 164. THE PARALLELOGRAM OF FORCES

We have now a simple graphical method for finding the resultant of two inclined forces acting at a point, using the *Principle of the Parallelogram of Forces*:—**If two forces are represented in size and direction by two neighbouring sides of a parallelogram, then their resultant is in size and direction represented by the diagonal of the parallelogram.** For example, in Fig. 164 if OA and OB represent two forces drawn to scale, then OR represents their resultant drawn to the same scale.

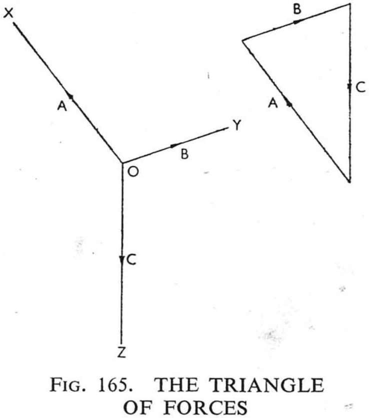


FIG. 165. THE TRIANGLE OF FORCES

THE TRIANGLE OF FORCES

(iii) On the same sheet of paper that you used in Experiment (ii), draw three lines parallel to the forces OX, OY and OZ, in such a way as to form a triangle. Measure the sides of the triangle. They are proportional to the forces OX, OY and OZ. This result illustrates *the Principle of the Triangle of Forces*:—**If three forces acting at a point are in equilibrium, they can be represented in size and direction, though NOT in position, by the three sides of a triangle taken in order.** Conversely, **if three forces acting at a point can be represented in size and direction by the three sides of a closed triangle taken in order, then they are in equilibrium.** Hence, when you try to draw a triangle of forces and it does not close, the three forces are *not* in equilibrium.

Notice that it saves time to draw the triangle *OYR* instead of completing the whole parallelogram *OYRX*.

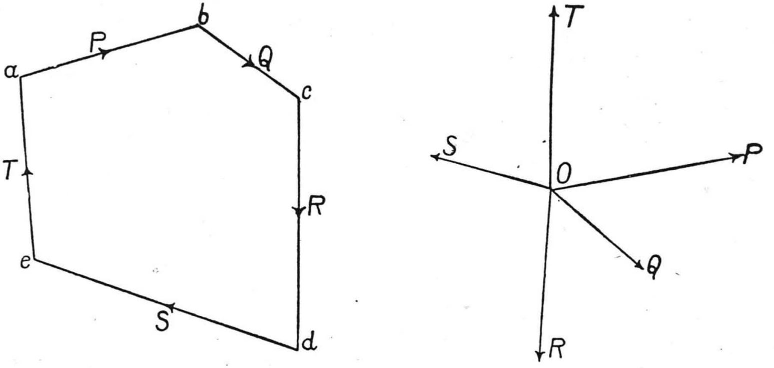


FIG. 166. POLYGON OF FORCES

THE POLYGON OF FORCES

The *parallelogram of forces* and the *triangle of forces* apply to systems of three forces only. A similar principle states the conditions for equilibrium between more than three forces:—**If any number of forces *P, Q, R, S, T,* etc., acting at a point are in equilibrium, they can be represented in size and direction, but not in position, by the sides of a closed polygon**[†] ***abcde,*** **etc., taken in order.** (See Fig. 166.) Conversely, **if a system of forces acting at a point can be represented in size and direction, but not in position, by the sides of a closed polygon taken in order, then the forces are in equilibrium.** Hence, when you try to draw a polygon of forces and it does not close, the extra line required to close the figure represents the *equilibrant* of the unbalanced forces. (See Fig. 167.)

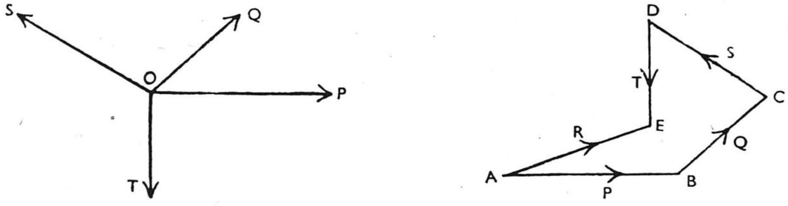


FIG. 167. FINDING THE RESULTANT OF FOUR FORCES

The application of these principles is of the greatest practical importance to engineers who design structures and who have to work out the forces acting on the various parts, e.g. in the framework of a steel bridge. If you are to understand these principles (as distinct from merely learning them by heart) you must practise applying them by solving as many simple number problems as possible, by graphical methods.

THE RESOLUTION OF FORCES

We have seen how two or more forces can be replaced by a single force called their *resultant.*

It is sometimes necessary to replace a single force by two or more forces which, acting together, will produce the same effect. This is called the *resolution of forces* since one force is *resolved* or converted into two or more forces acting in different directions. For example, a small railway coach is being pulled along the rails by a man *X* walking outside the rails, so that the rope *OX* makes an angle of 30° with the rails. If the man's pull is 100 lb., find how much of this pull is doing useful work in moving the coach along the rails, and how much is being wasted in pulling the wheels at right-angles to the rails. We have to *resolve* the force of 100 lb. into two other forces along *OY* and *OX* (see Fig. 168). We therefore mark off a length *OX* proportional to 100 lb. (say 2 in.) and then draw *XY* perpendicular to *OY*. Then *OY* represents the useful forward pull (= . . . . in. = ... lb.) and *XY* represents the side-thrust wasted against the rails (= . . . . in. = ... lb.).

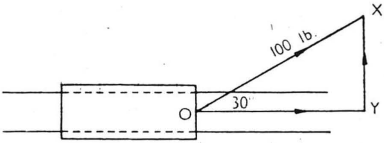


FIG. 168. RESOLUTION OF FORCES

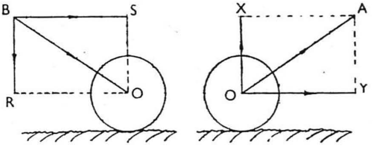


FIG. 169. RESOLUTION OF FORCES

To take another example, why is it easier to *pull* a garden-roller on a soft piece of ground than to *push* it ? See Fig. 169. Resolving the *push BO* into a horizontal force and a vertical force, the horizontal force *BS* tends to move the roller along the ground while the vertical force *BR* acts *downwards,* pushing the roller deeper into the ground. Resolving the *pull OA* into a horizontal force and a vertical force, the horizontal force *OY* tends to move the roller along the ground while the vertical force *OX* acts *upwards* and reduces the pressure of the roller on the ground; so that it moves more easily when it is pulled than when it is pushed.

Solve some simple number problems graphically for yourself.

PARALLEL FORCES

We have now considered (*a*) forces acting in the same straight line and (*b*) forces inclined at an angle to each other. The remaining possibility is (*c*) forces acting parallel to each other.

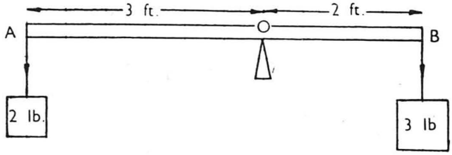


FIG. 170. PARALLEL FORCES

If two forces act in parallel directions we cannot find their resultant by using the parallelogram or triangle of forces. For example, if weights of 2 lb. and 3 lb. are hung from each end of a bar *AB*, 5 ft. long, as shown in Fig. 170, there will be two parallel forces acting on the bar. If we put a knife-edge *O* under the bar, the force at *B* tends to move in a clockwise direction and the force at *A* tends to move in an anti-clockwise direction; but we can find one point where the bar will balance at *O* and the resultant of the two downward forces must act in a vertical line passing through this point *O*. The point *O* is called *the centre of the two parallel forces*, and it is in such a position that *each downward force multiplied by its distance from O gives the same product*. This product is called the *leverage* (or the *turning-effect*, or the '*moment*'†) of the force about the point *O*, and *when the bar is in equilibrium, the clockwise leverage at B is equal to the anti-clockwise leverage at A*. This is *the Principle of Leverage* (or '*moments*') and it also applies to any number of parallel forces acting on a bar.

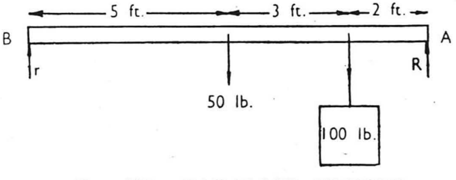


FIG. 171. PARALLEL FORCES

In Fig. 170, the *leverage* of the force at *A* is 2 (lb.)×3 (ft.) =6, while the *leverage* of the force at *B* is 3 (lb.)×2 (ft.) =6. In this case, the *resultant* is a force of 5 lb. acting vertically downwards at *O*, which is 3 ft. from *A* and 2 ft. from *B*.

The Principle of Leverage (or ' moments ') enables us to calculate the forces acting on a loaded beam or bridge. For example, in Fig. 171, AB is a uniform beam, 10 ft. long and weighing 50 lb., resting on two supports at A and B, and a weight of 100 lb. is hung 2 ft. from A. The downward forces acting on the two supports A and B are the same as the upward reactions R and r at A and B.

As the beam is in equilibrium, the upward forces R and r must balance the downward forces, i.e. $R+r=50+100=150$ lb., and the clockwise leverage about either A or B must be equal to the anti-clockwise leverage. Consider the leverage about A:—

$$\begin{aligned}10r&=(100\times 2)+(50\times 5)\\&=200+250\\&=450\end{aligned}$$

Therefore $r=45$ lb. But $R+r=150$ lb.
Therefore $R=105$ lb.

We can check this by considering the leverage about B:—

$$\begin{aligned}10R&=(50\times 5)+(100\times 8)\\&=250+800\\&=1050\end{aligned}$$

Therefore $R=105$ lb.

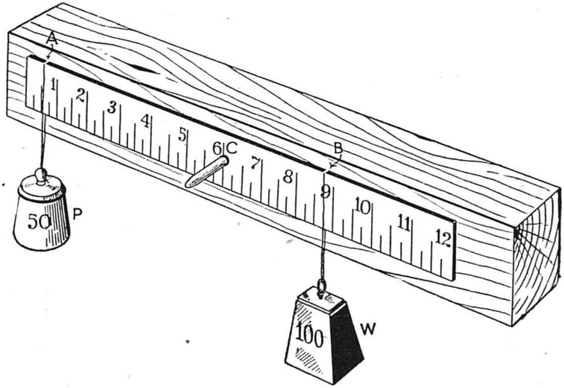


FIG. 172. 'SEE-SAW LEVER' (CLASS I)

LEVERS

The apparates shown in Fig. 172 is a simple *lever*, i.e. *a rod free to turn about some fixed point called the pivot*. It is clear that such a lever can be used as a *machine*, the force (or *effort*) acting at A and

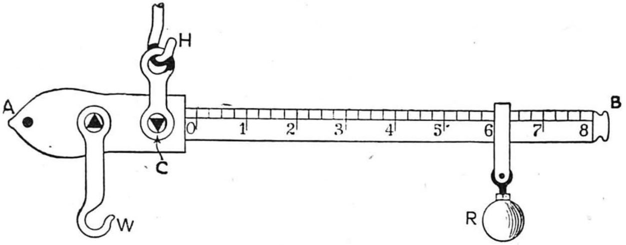


FIG. 173. STEEL-YARD

moving another force (or *load*) acting at *B*. The pivot need not always be placed between the effort and the load: in fact, there are three

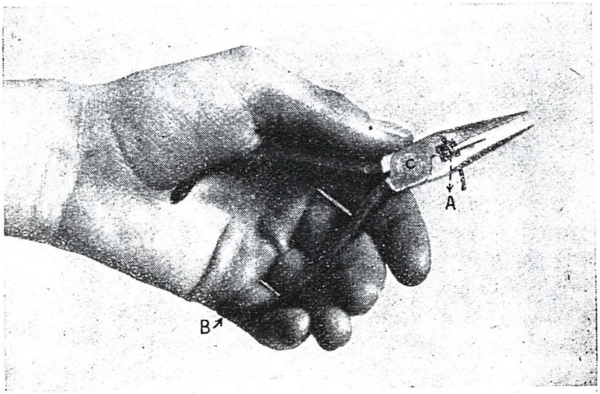


FIG. 174. WIRE-CUTTERS

possible arrangements of the *pivot*, the *effort* and the *load*, giving rise to three classes of levers.

Class I.—' See-saw '† levers. In levers of the *first class*, *the pivot is between the effort and the load*, as shown in Fig. 172, where the *effort P* acts downwards at *A* and the *load W* acts downwards at *B*, on the opposite side of the *pivot C*. Examples of such levers are (i) a see-saw, (ii) the beam of a balance, (iii) a steel-yard† (Fig. 173),

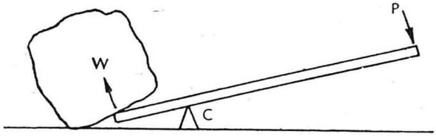


FIG. 175. ONE WAY OF USING A CROW-BAR

(iv) a pair of scissors or wire-cutters (Fig. 174), (v) a crow-bar† when used in the ordinary way for raising a heavy stone (Fig. 175).

Class II.—'Wheel-barrow'† levers. In levers of the *second class, the load is between the effort and the pivot*, as shown in Fig. 176, where the *effort P* acts *upwards* at *A* and the *load W* acts downwards at *B*, between the effort and the *pivot C*. Examples of the second class of levers are (i) a wheel-barrow (Fig. 177), (ii) the human foot and ankle (Fig. 95 (*b*)), (iii) a bottle-opener for 'crown corks', (iv) a crow-bar when used as in Fig. 178.

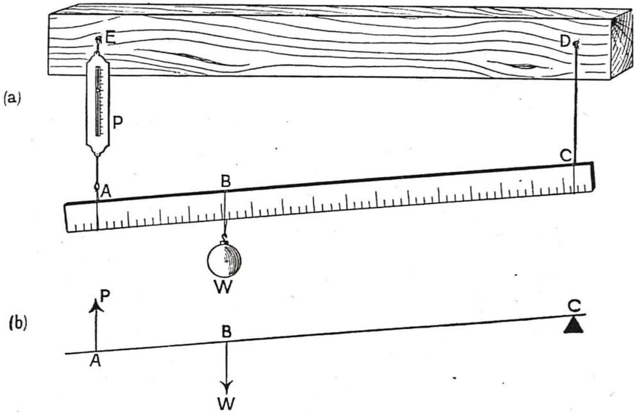


FIG. 176. 'WHEEL-BARROW LEVER' (CLASS II)

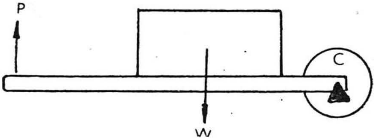


FIG. 177. PRINCIPLE OF WHEEL-BARROW

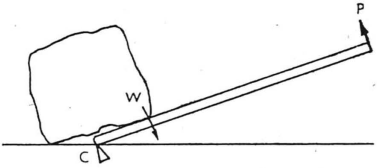


FIG. 178. ANOTHER WAY OF USING A CROW-BAR

Class III.—'Forearm levers'. In levers of the *third class, the effort acts between the load and the pivot*, as shown in Fig. 179, where the *load W* acts downwards at *B* and the *effort P* acts upwards at *A*, between the load and the *pivot C*. Examples of the third class of levers are (i) the human forearm when lifting a weight with the hand (Fig. 95 (*c*)), (ii) a pair of forceps, (iii) a steam-engine safety valve (Fig. 180).

Set up levers of each class as shown in Figs. 172, 176, and 179. Use

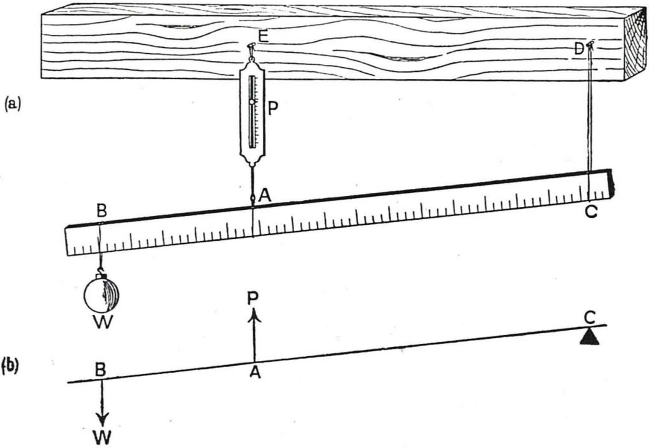


FIG. 179. 'FOREARM LEVER' (CLASS III)

large weights so that the weight of the lever-beam itself is small in proportion to the effort and load. Enter your results below:—

Class of Lever	I	II	III
Effort (*P*)			
Distance of effort from pivot (*x*)			
Load (*W*)			
Distance of load from pivot (*y*)			
Leverage of effort (*Px*) ('*Effort-arm*') ..			
Leverage of load (*Wy*) ('*Load-arm*') ..			

If you have worked carefully, you should find that the last two quantities Px and Wy are equal, i.e. *effort-arm* = *load-arm*.

Notice that levers of the first and second classes can give a mechanical advantage, i.e. a small force applied as the *effort* can move a larger *load*. Levers of the third class, however, never give a mechanical advantage since the *effort* is always greater than the *load*, but they give *convenience* (e.g. in forceps) and *speed* (e.g. any movement of the biceps muscle is greatly magnified in the corresponding movement of the hand).

Use *the Principle of Levers* to solve some number problems.

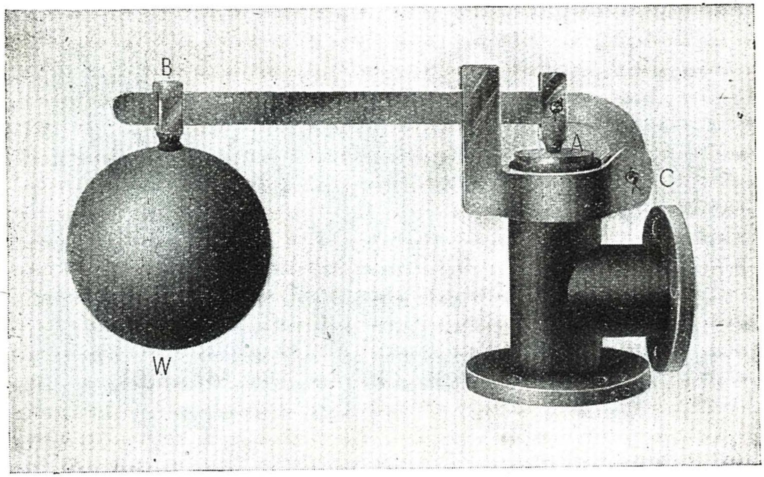


FIG. 180. STEAM-ENGINE SAFETY-VALVE

CHAPTER XIV

MOTION

The word '*motion*' gives the idea of *change of position*, and we can classify most objects around us as either *moving* or *at rest*. But the terms *motion* and *rest* are only *relative*. This building, for example, is at rest in relation to the ground on which it stands, but the Earth's surface is not at rest since the Earth is revolving round its own axis and is also moving round the Sun. For all ordinary purposes, however, we describe position in relation to the Earth, and if an object is not changing its position relative to the Earth, we say that it is *at rest*.

SPEED AND VELOCITY

In order to describe the motion of a body completely, it is necessary to state the *direction* in which it moves and also the *speed* with which it moves. Another name for *speed* is *velocity*,[1] and this can be defined as *the distance travelled in unit time*.

Speed may be either *uniform* or *variable*, e.g. if a train travels exactly 88 ft. in every second, it is moving with a *uniform speed* of 88 ft. per second, i.e. 60 miles an hour. Such perfectly regular speeds are most unusual in everyday life, however; for example, the speed of fast trains is sometimes greater and sometimes less according to the slopes and curves of the railway-track, and the train is then said to move with *variable speed*. If a train, in spite of its variable speed, still travels 60 miles in one hour (although it has not travelled 88 ft. in each individual second) we say that the *average speed* has been 60 miles an hour.

$$\text{Velocity} = \frac{\text{distance travelled}}{\text{time taken}} \text{ or } v = \frac{s}{t}$$

$$\text{or distance travelled} = \text{velocity} \times \text{time taken, or } s = vt$$

$$\text{or time taken} = \frac{\text{distance travelled}}{\text{velocity}} \text{ or } t = \frac{s}{v}$$

---

[1] In more advanced work you will find that there is a distinction between *speed* and *velocity*. Strictly speaking, *speed* refers to the distance travelled in a given time *in any direction*, while *velocity* refers to the distance travelled *in a straight line*.

Fig. 181 shows the *velocity-time graph* of a train moving with a *uniform velocity* of 60 m.p.h. Notice that the shaded *area* gives *the distance travelled.*

THE COMPOSITION OF VELOCITIES

It is possible for a body to be moving in two directions at the same time, e.g. when a man walks from one side of a moving ship to the other side. If the ship is travelling north with a velocity of 20 ft. per second and the man walks east at 5 ft. per second, it is clear, that his actual path, relative to the Earth, is somewhere between north and east. *The Parallelogram of Velocities* is a similar principle to the Parallelogram of Forces:—

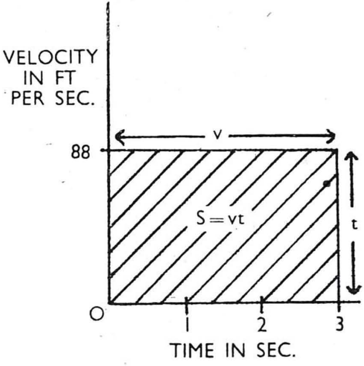


FIG. 181. UNIFORM VELOCITY

If a body has two velocities represented by two neighbouring sides of a parallelogram, then the actual (or resultant) velocity is represented by the diagonal† of the parallelogram.

Similarly, **if a body has two velocities represented by two sides of a triangle taken in order, then its actual (or resultant) velocity is represented by the third side.**

A velocity is completely described if we can give its *rate* (e.g. in ft. per sec., or m.p.h.) and its *direction*. Hence we can represent a velocity completely by a straight line, drawn to scale, bearing an arrowhead. In Fig. 182, *OA* represents the velocity of the ship and *OB* the man's velocity. The *resultant velocity* is *OC*, the diagonal of the parallelogram (which in this case is a rectangle).

Work out some number problems graphically.

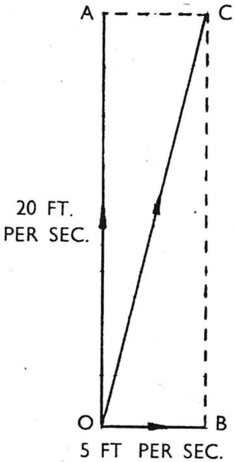


FIG. 182. PARALLELOGRAM OF VELOCITIES

ACCELERATION

In everyday life, steady speeds or uniform velocities are very rare, and the velocity of most moving bodies changes from time to time.

When its velocity changes, a body is said to have an *acceleration.*[†] This acceleration is positive if the speed is increasing and negative if it is decreasing. **Acceleration is rate of change of velocity.**

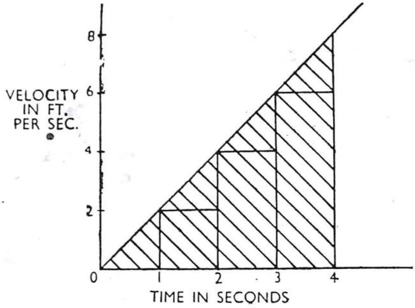


FIG. 183.
UNIFORM ACCELERATION

For example, if a train starting out of a station has a velocity of 2 ft. per second after 1 second, a velocity of 4 ft. per second after 2 seconds, a velocity of 6 ft. per second after 3 seconds, and a velocity of 8 ft. per second after 4 seconds, then its velocity is increasing by 2 ft. per second in every second; i.e. it has *an acceleration of two feet per second per second.*

Fig. 183 shows the *velocity-time graph* of the train. The shaded *area* gives the *distance travelled.*

Expressed in symbols: if a body starts from rest and acquires a velocity of v feet per second in t seconds, then its acceleration

$f=\frac{v}{t}$ *ft. per sec. per sec.*, i.e. *acceleration = change of velocity ÷ time.*

THE EQUATIONS OF MOTION FOR A BODY STARTING FROM REST

Suppose a body starts from rest and acquires a velocity of v feet per second in t seconds, after travelling a distance of s feet.

The *acceleration* $f=\frac{v}{t}$ feet per second per second (1*a*)

or $v=ft$ (velocity = acceleration × time).

The *average velocity* is $\frac{s}{t}$ feet per second, and since the body started from rest, the *final velocity* v is *twice the average velocity.*

That is, $v=\frac{2s}{t}$ or $s=\frac{1}{2}tv$ (2*a*)

Hence, for a body starting from rest, we have

$$f=\frac{v}{t} \text{ or } v=ft, \text{ and } v=\frac{2s}{t}$$

Substituting $v=ft$ in $v=\frac{2s}{t}$ we have $ft=\frac{2s}{t}$ or $s=\frac{1}{2}ft^2$ (3*a*)

That is, *distance travelled = half the acceleration × the square of the time.*

It is sometimes useful to express the velocity v in terms of the acceleration f and the distance travelled s.

From (1a), $v = ft$, $\therefore\ v^2 = f^2t^2$. From (3a), $s = \frac{1}{2}ft^2$, $\therefore\ t^2 = \frac{2s}{f}$

Substituting $\frac{2s}{f}$ for t^2 in f^2t^2, we have $v^2 = \frac{f^2 2s}{f}$; $\therefore\ v^2 = 2fs$ (4a)

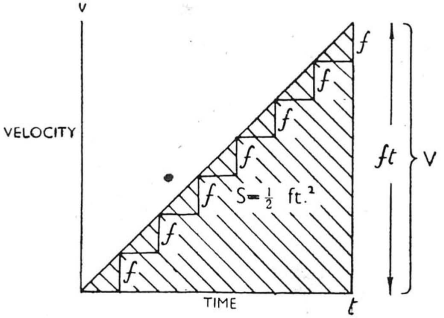


FIG. 184. UNIFORM ACCELERATION FROM REST

Fig. 184 shows the *velocity-time graph* for a body starting from rest and moving with a uniform acceleration of f feet per second per second.

Use the four equations of motion

(1a) $v = ft$
(2a) $s = \frac{1}{2}tv$
(3a) $s = \frac{1}{2}ft^2$
(4a) $v^2 = 2fs$

to work out numerical problems on the motion of a body which starts from rest and travels with uniform acceleration.

TO SHOW THAT THE DISTANCE TRAVELLED IS PROPORTIONAL TO THE SQUARE OF THE TIME

Mark off a long, grooved piece of wood into four equal sections and raise one end on a wooden block as shown in Fig. 185. Start a steel ball rolling from the top of the slope and, as it passes the first mark, start another ball from the third mark. It is found that the two balls both arrive at the end of the groove together. Since the time taken by the first ball to travel through the top section equals the time taken by the second ball to travel through the bottom section, the first ball will have travelled through four sections in twice the time it took to travel through one section. *Thus the distance it travelled is proportional to the square of the time it took.*

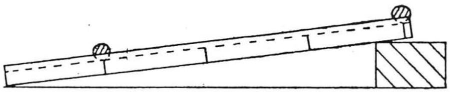


FIG. 185. GALILEO'S EXPERIMENT

THE EQUATIONS OF MOTION FOR A BODY WHICH DOES NOT START FROM REST

Now let us consider the case of a body starting with a velocity of u feet per second, moving for a time t seconds, travelling for a distance of s feet, and acquiring a final velocity of v feet per second.

The *change of velocity* is $(v-u)$ feet per second, and this takes place in t seconds.

Therefore, the acceleration $f=\frac{v-u}{t}$ or $v-u=ft$, or $v=u+ft$ (1*b*)

Now the *average velocity* is *half the sum of the first and the final velocities*, i.e. $\frac{s}{t}=\frac{1}{2}(v+u)$, or $s=\frac{1}{2}t(v+u)$, or $v+u=\frac{2s}{t}$ (2*b*)

Substituting $v=u+ft$ from (1*b*), we have $s=\frac{1}{2}t(u+ft+u)$

Therefore $s=ut+\frac{1}{2}ft^2$ (3*b*)

It is sometimes useful to express the change of velocity in terms of the acceleration f and of the distance travelled s.

Taking $v+u=\frac{2s}{t}$, and multiplying by $v-u=ft$

we have $v^2-u^2=2fs$ (4*b*)

When the body starts from rest, $u=0$, and the above equations become:—

$$(1a)\ f=\frac{v}{t} \text{ or } v=ft \quad (2a)\ s=\tfrac{1}{2}tv \quad (3a)\ s=\tfrac{1}{2}ft^2 \quad (4a)\ v^2=2fs$$

Fig. 185 is the *velocity-time graph* of a body starting with a velocity of u feet per second and travelling for t seconds with a uniform acceleration of f feet per second per second, and acquiring a final velocity of v feet per second.

Use equations (1*b*), (2*b*), (3*b*) and (4*b*) to solve some numerical examples.

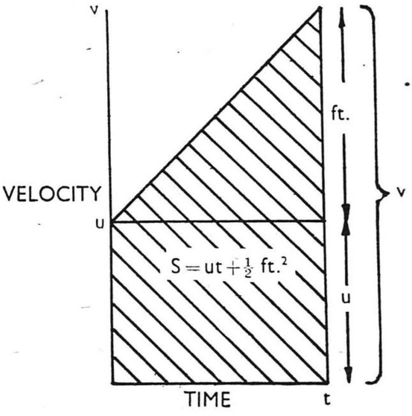


FIG. 186.
UNIFORM ACCELERATION OF A BODY WHICH DOES NOT START FROM REST

FALLING BODIES

Falling bodies provide a good example of *uniform acceleration*. About 1590, Galileo (an Italian who invented the telescope and was the leading scientist of his time) carried out experiments on dropping bodies of different sizes and materials from the top of a high tower. He concluded that *all bodies would fall at exactly the same rate if it were not for the resistance of the air.* After the invention of the *air-pump,* Sir Isaac Newton (who was born in 1642, the year in which Galileo died) tested Galileo's conclusion by his *coin and feather experiment.* Newton enclosed a gold coin and a small feather in a long glass tube and pumped out the air.

On inverting the tube, he found that the heavy coin and the light feather fell side by side, showing that, **in a vacuum, all freely falling bodies acquire exactly the same velocity.**

A simple way of illustrating this principle is to place a circle of thin paper on a large coin so that the paper does not project beyond the edge of the coin. If the coin is held in a horizontal position and then released, the paper falls with it, because the coin overcomes the resistance of the air in front of the paper. (See Fig. 187.)

FIG. 187. FALLING BODIES

THE ACCELERATION DUE TO GRAVITY

If freely falling bodies, starting from rest, all acquire the same velocity after travelling for the same time, their acceleration must be the same. *This constant acceleration of all falling bodies is called the acceleration due to gravity,* and it is usually represented by the symbol 'g'.

The value of g varies very slightly in different parts of the world. Thus in London it is 32·19 feet per second per second; in New York it is 32·16 feet per second per second; in Calcutta it is 32·11 feet per second per second; and in Singapore $g = 32\cdot 09$ feet per second per second. We shall take $g = 32$ *feet per second per second* as being accurate enough for practical purposes.

We have seen that the *weight* of a body is not exactly the same everywhere on the Earth's surface, being greater at the Poles than at the Equator. It is clear that the variation in the value of g and the variation in the *weight* of a body are due to the same cause.

Galileo, who laid the foundations of physics with his experiments on moving bodies, was also responsible for the earliest observations of the properties of *pendulums*. In 1583, it is said, he watched the swinging of a number of lamps hanging from the roof of a building. All the lamps were of the same size and they all hung by chains of equal length. As they swung backwards and forwards, Galileo noticed that the time of swing was the same for every lamp, although some were swinging through a much wider arc than others: that is, a complete swing always took the same time, whether the angle of swing was large or small.

THE SIMPLE PENDULUM

A ' perfect ' simple pendulum is defined as *a heavy particle suspended by a weightless thread.* The nearest we can get to this in practice is to hang a small metal ball or ' *bob* ' on a very thin thread, and we shall now use this arrangement to carry out experiments to illustrate the most important properties of pendulums.

TO SHOW THAT THE TIME OF SWING IS INDEPENDENT OF THE ANGLE OF SWING

Thread a piece of cotton thread, about 2 metres long, through a small hole in a cork, and hold the cork in a retort-clamp attached to a stand. Hang a metal ' bob ' to the end of the thread and adjust the length so that the *centre* of the bob is exactly 1 metre below the *bottom* of the cork. This distance is *the length of the pendulum.*

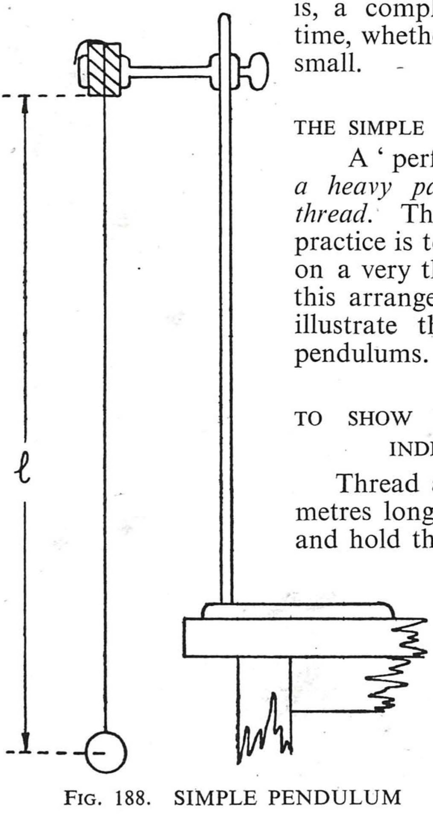


FIG. 188. SIMPLE PENDULUM

Set the pendulum swinging by pulling the bob about 1 in. to one side. Then, with a stop-watch, take the time of 50 *complete swings* (i.e. ' there-and-back '), being careful to count 'nought' as you begin timing and 'one' at the *end* of the first complete swing.

Time of 50 swings = ... seconds.

Repeat this, pulling the bob about 2 in. to one side, and again take the time for 50 swings.

Time of 50 swings = ... seconds.

Repeat this once more, pulling the bob about 3 in. to the side when you start the pendulum swinging, and again take the time of 50 swings.

Time of 50 swings = ... seconds.

If your experiment is done carefully, you will find that *the time of swing is independent of the angle of swing (provided this is small).*

TO SHOW THAT THE TIME OF SWING IS INDEPENDENT OF THE WEIGHT OR THE MATERIAL OF THE BOB

Fit up two simple pendulums of exactly the same length (about 1 metre) but with bobs of different material and different weight, e.g. use lead, brass, steel, wood or glass bobs. Pull both bobs about 3 in. to one side and let them go at the same moment. Notice that their times of swing are exactly the same. Hence, *the time of swing is independent of the weight or the material of the bob.*

TO SHOW THAT THE TIME OF SWING DEPENDS ON THE LENGTH OF THE PENDULUM

Measure the time of 50 swings of a pendulum 1 metre long. Shorten the pendulum to 81 cm. and again find the time of 50 swings. Repeat this with pendulums 64 cm., 49 cm., 36 cm., and 25 cm. long, recording your figures below. Plot graphs of (*i*) *length* (*l*) *against time* (*t*), (*ii*) $\sqrt{l}$ *against t.*

Length of Pendulum (l)	Time of one complete swing (t)	$\sqrt{\text{Length}}$ ($\sqrt{l}$)	$\sqrt{l} \div \sqrt{t}$
cm.	$\div 50 =$ sec.	cm.	

If your experiments have been carried out carefully all the figures in the last column will be the same, and your graph of $\sqrt{l}$ against t will be a straight line, showing that, for a simple pendulum, *the time of swing is proportional to the square root of the length of the pendulum.*

You have now examined the most important properties of swinging bodies and pendulums, and these universal properties have many everyday applications. The property of swinging in exactly equal intervals of time makes the pendulum the basis of accurate *time measurement*, which we shall discuss shortly.

It can be shown mathematically that, for a simple pendulum where t is the *time of swing*, l is the *length* of the pendulum, and g is the *acceleration due to gravity*:—

$$t = 2\pi \sqrt{\frac{l}{g}}$$

This gives us an accurate method for finding the value of g since we can easily find t and l by experiment.

Squaring both sides of the above formula, we have

$$t^2 = \frac{4\pi^2 l}{g} \quad \therefore \quad g = \frac{4\pi^2 l}{t^2}$$

Use the figures from your last experiment to find g, substituting your values of l and t in the last formula. Your result will be the value of g in centimetres per second per second. Since 1 ft. = 30·5 cm., divide your result by 30·5 to get the value of g in feet per second per second.

Now use your value of g to find the length of a simple pendulum which will ‘beat’ seconds accurately. (N.B.—The time of each complete swing, ‘there-and-back’, will be 2 seconds.)

Also find the length of the ‘seconds pendulum’ from your graph of l and t.

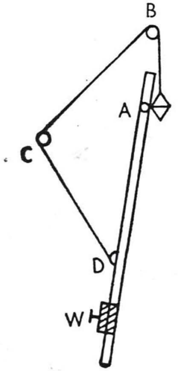


FIG. 189.
FINDING ‘G’

TO FIND ‘g’ BY USING A PENDULUM AND A FALLING WEIGHT

The acceleration due to gravity can also be found by measuring the time taken by a freely falling body to travel a known distance from rest. This can be done experimentally with the apparatus shown in Fig. 189. A long wooden rod swings on a pivot A at its upper end, and its time of swing can be adjusted by moving the sliding weight W. A strip of white paper is gummed to one side of the pendulum-rod and a thread is attached at D about half way down the opposite face of the rod. This thread passes round a pin C and then over another pin B fixed above the top of the pendulum, and a weight (blackened with soot) is hung from the other end of the thread. The pendulum is first allowed to hang freely and the pin B is adjusted so that the weight *just touches* the side of the pendulum as it hangs vertically. The pendulum is now drawn aside through a small angle by drawing back the pin C, and a mark is made near the top end of the pendulum by swinging the weight against it.

The thread is then burnt between B and C, setting free the pendulum and the weight at the same instant. The weight falls vertically and strikes the paper when the pendulum reaches the vertical, i.e. after it has made *one-quarter of a complete swing*. The length between the two marks on the side of the pendulum is the distance the weight has fallen in one-quarter of a complete swing.

The pendulum is now set to swing freely and the time of 20 complete swings is taken with a stop-watch.

Experiment Number	I	II	III
Fall in feet (s)	ft.	ft.	ft.
Quarter-swing in seconds (t)	sec.	sec.	sec.

Take the average values for s and t from three experiments (without changing the length of the pendulum).

In our equation of motion 3*a* (p. 250), $s=\frac{1}{2}ft^2$. Substituting g for f we have $s=\frac{1}{2}gt^2$ since g is the acceleration due to gravity, therefore

$$g=\frac{2s}{t^2}=\frac{2\ \times\qquad}{\qquad\times\qquad}=\ldots\ldots\ldots\text{feet per second per second.}$$

Chapter XV

THE MEASUREMENT OF TIME

The measurement of time is of great practical importance both in everyday life and in scientific work. In fact, it is possible to state all scientific measurements in the three fundamental units of *length*, *weight*, and *time*. Physical measurements are usually based on the *C.G.S. system* (*centimetre-gram-second system*), and engineering measurements on the *F.P.S. system* (*foot-pound-second system*).

The scientific unit of time is the *second*. This is got from the time taken for the earth to travel round the sun—365·25 days. The average length of all these days is a *clock-day* (or *mean solar*[†]*-day*) and this is divided into 24 hours. Each hour is divided into 60 minutes and each minute into 60 seconds. All this is worked out by the careful observations and calculations of astronomers who control the master-clocks from which we get our *standard time*. For example, the wireless time-signal sent out from Greenwich Observatory each day marks a mean solar-day of 86,400 seconds, and its accuracy never varies by more than 0·01 seconds per day. Nowadays, accurate time can be kept by clocks and watches adjusted to these wireless time-signals, but centuries before clocks and watches were invented there were other methods of measuring time.

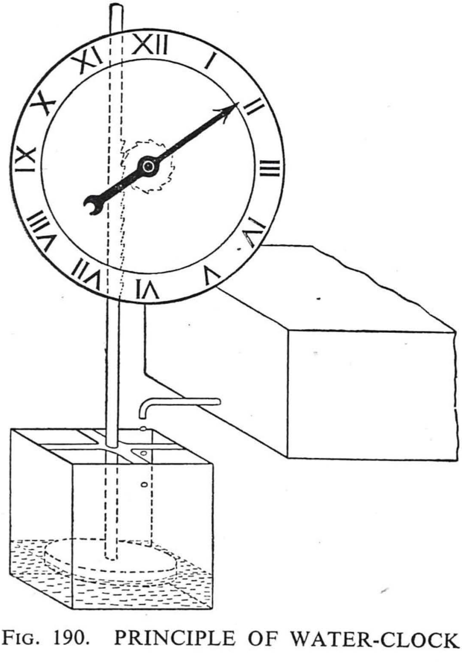


FIG. 190. PRINCIPLE OF WATER-CLOCK

WATER-CLOCKS

One of the earliest methods of measuring time was by means of water-clocks. In the simplest type of water-clock, water was allowed to fall slowly but regularly from one vessel to another. The rise of water in the lower vessel could then be used to measure time, e.g. if the level rose 24 in. in one day, then a rise of one inch represented one hour. By adding a float arranged to move a pointer over an hour-scale, the water-clock was made more convenient for everyday use. (See Fig. 190.)

HOUR-CANDLES

Another early method for measuring time was by using candles which burned at a steady rate. These candles were marked off in equal units of length, the usual unit being the length of candle which burned away in an hour.

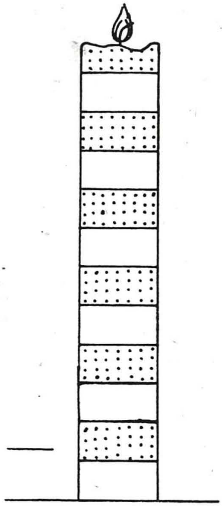


FIG. 191.
AN HOUR-CANDLE

SAND-GLASSES

Before the invention of clocks and watches, sand-glasses were used for measuring shorter periods of time. In the hour-glass shown in Fig. 192, the fine sand runs through the narrow opening between the two bulbs in one hour. Smaller sand-glasses were used for measuring smaller intervals of time down to one minute.

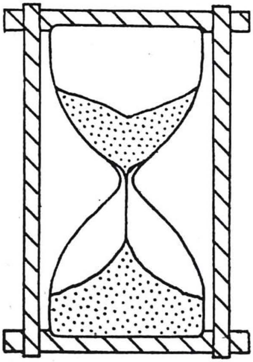


FIG. 192. SAND-GLASS

SUN-DIALS

One of the earliest instruments for measuring time was the *sun-dial.*† If a stick is fixed upright in the ground, when the sun rises in the east the stick casts a long shadow westwards, and this shadow moves round the stick until, when the sun sets, the shadow points eastwards. If a mark is made on the ground in the path of this moving shadow, the shadow will fall on the mark once each day and the time-interval between two successive shadows passing this mark is a *solar-day.* The length of this time-interval is slightly different from day to day, varying with the distance of the earth from the sun as it follows its path round the sun along an *ellipse*† and not in a circle. Only four times each year—on 15 April, 15 June, 1 September, and 24 December—are the *solar-day* and the *clock-day* of the same length. The greatest difference between them is about 16 minutes at the beginning of November.

In a practical sun-dial, like the one shown in Fig. 193, which is specially suitable for places near the Equator, the rod which casts the shadow forms the axis of a cylinder, the bottom half of which catches the shadow. The *shadow-rod* points due north and south and it is inclined to the horizontal at an angle equal to the *latitude*† of the place where the sun-dial is to be used. That is, the axis of the cylinder (the shadow-rod) is *parallel to the earth's axis.* Noon is marked vertically below the shadow-rod, the horizontal edges of the half-cylinder are marked 6 a.m. and 6 p.m., and 12 equal intervals (hours) are marked off round the half-cylinder.

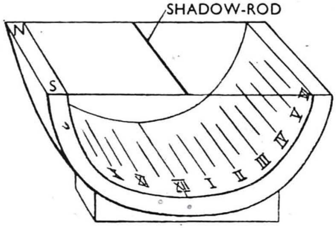


FIG. 193. CYLINDRICAL SUN-DIAL FOR USE NEAR EQUATOR

Such a sun-dial gives *sun-time* (or *local apparent time*) and this agrees with *clock-time* only on four days each year. Hence a table of corrections is needed to get clock-time from the sun- or solar-time shown by the sun-dial.[1] (In countries with ' daylight-saving ' laws,

TABLE OF CORRECTIONS FOR CONVERTING SOLAR-TIME (' SUN-DIAL-TIME ') TO LOCAL STANDARD TIME (' CLOCK-TIME ')

MONTH	Date 1-10	Date 11-20	Date 21-31	MONTH	Date 1-10	Date 11-20	Date 21-31
January ..	min.	min.	min.	July ..	min.	min.	min.
February ..	,,	,,	,,	August ..	,,	,,	,,
March ..	,,	,,	,,	September	,,	,,	,,
April ..	,,	,,	,,	October ..	,,	,,	,,
May ..	,,	,,	,,	November ..	,,	,,	,,
June ..	,,	,,	,,	December ..	,,	,,	,,

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[1] TEACHER'S NOTE:—This can be worked out from the ' Nautical Almanack '.

a further correction is necessary, e.g. *plus* 33 minutes in Malaya.) A sun-dial, however, is useless at night or in cloudy weather.[1]

None of these early methods for measuring time was of any use for the accurate measurement of time for scientific purposes, and accurate time-measurement had to wait until the discovery of the properties of the *pendulum* in the sixteenth century.

PENDULUMS AND BALANCE-WHEELS

We have seen that the time of swing of a pendulum is always the same provided that its length is not altered and if the angle of swing is small. In other words:—

(*a*) *The time of swing is independent of the angle of swing* (*provided this is small*).

(*b*) *The time of swing is independent of the weight or the material of the pendulum-bob.*

(*c*) *The time of swing depends only on the length of the pendulum.*

Under suitable conditions, a spring can be made to behave like a pendulum, always taking the same time for a complete backwards and forwards movement.

Most modern clocks and watches are controlled by *pendulums* or by *balance-wheels*, making use of this property of constant time of swing.

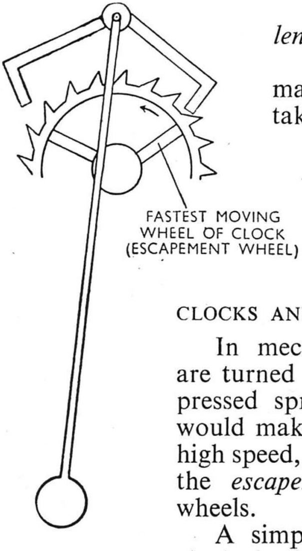

FIG. 194. PENDULUM-CLOCK ESCAPEMENT

CLOCKS AND WATCHES

In mechanical clocks and watches, the 'works' are turned either by a hanging weight or by a compressed spring. Left to itself, the weight or spring would make the hands of the clock spin round at a high speed, but every clock or watch has a device called the *escapement*† for controlling the speed of the wheels.

A simple type of escapement used in pendulum-clocks is shown in Fig. 194. This allows the fastest-moving wheel of the 'clock-work' to move forward from one tooth to the next as the pendulum reaches

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[1] A concrete half-cylinder, 3 ft. in diameter, as used for large concrete drains, makes an excellent permanent sun-dial. If the table of corrections is worked out for ten-day intervals throughout the year, it will be sufficiently accurate for school purposes.

either end of its swing. At the same time, the escaping tooth of the escapement-wheel gives the escapement a slight push which keeps the pendulum swinging. In a tall, ‘ grandfather ’ clock, the pendulum is a seconds-pendulum, about 39 in. long, and the ‘ seconds-wheel ’ and ‘ seconds-hand ’ move forward one division per second. A clock with a pendulum one-quarter this length will beat half-seconds.

In smaller clocks and in watches, the *balance-wheel* serves the same purpose as the pendulum, carrying the escapement which controls the movement of the fastest-moving wheel in the ‘ clock-work ’. The balance-wheel swings backwards and forwards under the action of a spiral *hair-spring* in much the same way as a pendulum swings under the pull of gravity.

A *balance-wheel*, in fact, is a *compound pendulum*, with the same property of always taking exactly the same time for a complete swing provided that the size and weight of the balance-wheel and the strength of the hair-spring do not alter. In pendulum-clocks, the time of swing is adjusted by altering the length of the pendulum. In small clocks and watches the time of swing is adjusted by altering the length of the hair-spring, shortening it to make the wheel swing faster, and *vice versa*.

SECTION IV

ELECTRICITY

CHAPTER XVI

THE PRODUCTION OF ELECTRIC CURRENTS

Until the end of the eighteenth century, continuous electric *currents* were unknown, although electric *charges* and momentary electric *shocks* had been produced by rubbing different substances together. For example, when dry glass is rubbed with dry silk, one kind of electric charge appears on the glass and another kind of charge on the silk, and two bodies can be charged in opposite ways until a momentary electric *spark* passes between the two charged bodies when they are brought near together. This '*frictional electricity*', however, is of little use for everyday purposes, although it produces important effects in Nature, e.g. lightning, and it is '*current electricity*' which is used in all our modern electrical devices, e.g. batteries, dynamos and electric motors.

The first step towards the production of a continuous electric current was

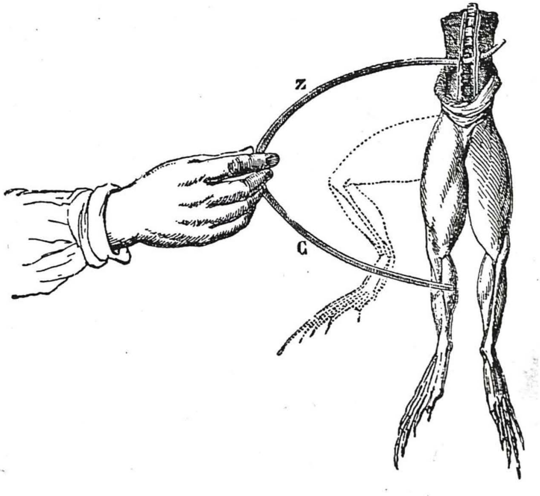

FIG. 195. GALVANI'S EXPERIMENT
Z—zinc. C—copper.

made accidentally by Galvani, an Italian scientist, in 1786, while he was carrying out experiments on the nerves and muscles of frogs, using a small electric *shock* as a stimulus to produce muscular contraction. One day, he got a similar result by holding a piece of copper and a piece of zinc in contact between his fingers and allowing the other end of the copper to touch a nerve and the other end of the zinc to touch a leg muscle, as shown in Fig. 195. Galvani explained this by saying that the frog's muscles contained electricity, but Volta, another Italian scientist, thought that the electricity which made the frog's legs kick came from the contact between the two different metals, copper and zinc. Volta experimented with pairs of different metals and, in 1800, he showed that **when two different metals are separated by a solution which reacts chemically with one of them, an electric current flows along a wire joining the two metals.** He made the first *voltaic*† *cell* from discs of zinc and copper separated from each other by discs of porous paper soaked in salt solution.

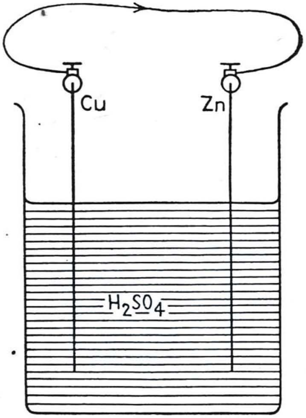

FIG. 196.
SIMPLE VOLTAIC CELL

THE SIMPLE VOLTAIC CELL

The simplest voltaic cell consists of a sheet of *zinc* and a sheet of *copper* dipping into dilute sulphuric acid.

When a sheet of commercial zinc is dipped into dilute sulphuric acid, bubbles of hydrogen appear, showing that the zinc is being dissolved by the acid. This is called ' *local action* ' and is caused by the presence of impurities, e.g. iron and carbon, in the commercial zinc. *Pure zinc* does not dissolve in dilute sulphuric acid. This ' local action ' can be prevented by covering the zinc plate with a thin layer of mercury which dissolves some of the zinc but not the impurities, forming a bright coating which consists of a solution of pure zinc in mercury. If sheets of mercury-coated zinc and copper are put into a beaker of dilute sulphuric acid, neither metal dissolves *until the copper touches the zinc*, when the zinc begins to dissolve. (This should remind you of your preparation of hydrogen, when you added some copper sulphate solution to deposit a little copper on the zinc and make it dissolve.)

When the lower halves of the zinc and copper plates are separated from each other by dilute sulphuric acid, as shown in Fig. 196, and their upper ends are joined by a wire, *a current flows through the wire.* The direction of the current *in the wire* is said to be *from the copper towards the zinc*; and it flows *from the zinc towards the copper through the acid inside the cell.* The *zinc* is called the *negative pole* of the cell and the *copper* is called the *positive pole.*

The current from a simple copper-zinc cell soon gets weaker and weaker, because a layer of hydrogen collects on the copper plate and prevents the acid from reaching the copper, while the hydrogen tends to set up a current in the opposite direction. That is, the layer of hydrogen forms a *new pole* between the copper and the zinc, and the cell is said to be *polarized.*† Polarization in the simple voltaic cell is a very serious disadvantage, because any cell which is to be used for practical purposes must give a steady current. Hence all practical voltaic cells have special arrangements for reducing polarization, i.e. for *de-polarizing* the cell. In most practical voltaic cells an oxidizing agent is used (e.g. manganese dioxide or potassium bichromate†—$K_2Cr_2O_7$) to oxidize the hydrogen on the positive plate to water and thus de-polarize the cell.

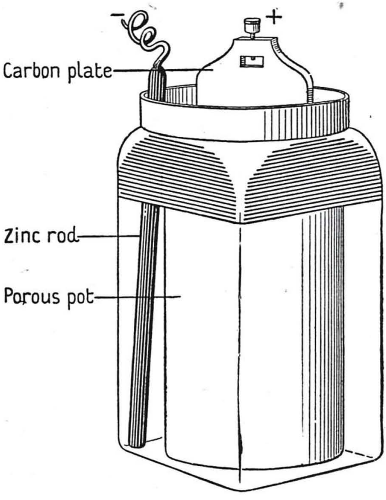

FIG. 197. LECLANCHÉ CELL

THE LECLANCHÉ CELL

The disadvantages of (*a*) *polarization*, and (*b*) '*local action*' are overcome in the voltaic cell invented by Leclanché in 1868, and the *Leclanché cell* is the only voltaic cell in everyday use to-day. The positive pole is made of *carbon* and the negative pole of *zinc*. The carbon rod or plate is surrounded by a coarse mixture of *manganese dioxide* (an oxidizing agent) and *carbon* (to increase the surface area of the positive pole and to absorb some of the hydrogen) contained in a porous pot or bag. The cell is filled with a saturated solution of *ammonium chloride*. When the poles are connected by a wire, a current flows and the zinc begins to dissolve:—

$$Zn + 2NH_4Cl \longrightarrow ZnCl_2 + 2NH_3 + H_2$$

The grains of carbon absorb some of the hydrogen which the manganese dioxide begins to oxidize to water:—

$$H_2 + MnO_2 \longrightarrow H_2O + MnO \text{ (manganous oxide)}$$

If only a small current is being taken from the cell (e.g. when used as part of the ‘ high tension battery ’ of a wireless set) the manganese dioxide can oxidize the hydrogen as fast as it is formed, de-polarizing the cell continuously and maintaining a steady current. If a bigger current is taken from the cell (e.g. when used for lighting a pocket-lamp bulb) the manganese dioxide cannot oxidize the hydrogen as fast as it is formed and the current soon falls off owing to polarization. After a rest, however, the manganese dioxide de-polarizes the cell. Hence a Leclanché cell is only suitable for supplying small currents for a short time, so that the cell has time to de-polarize before it is used again. It is a very useful cell, however, for working electric bells.

THE DRY CELL

The Leclanché cell is most commonly met with in its ‘dry’ form. The *zinc* container forms the negative pole. The central *carbon* rod forms the positive pole and this is surrounded by a *damp* de-polarizing mixture of *ammonium chloride*, *manganese dioxide* and *carbon*, contained in a porous bag. The space between this bag and the zinc container is filled with a *damp* paste of ammonium chloride and flour. These *dry cells* are made in millions each year for use with pocket-lamps, bells, telephones and wireless sets, where only small currents are required at intervals.

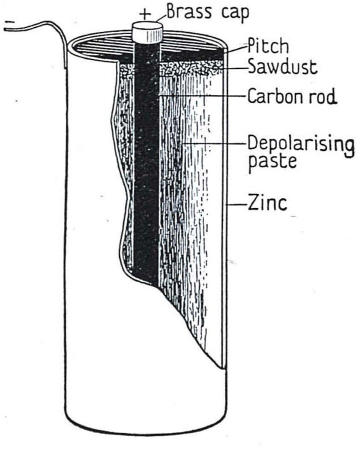

FIG. 198. DRY CELL

THE SOURCE OF ENERGY IN VOLTAIC CELLS

By *the Principle of the Conservation of Energy*, electrical energy can only be obtained by changing an equivalent amount of some other form of energy into electricity. In *voltaic cells*, one of the metals (usually zinc) dissolves to produce electricity and a salt of the metal. This salt possesses less potential (chemical) energy than the metal from which it was formed, for example when 1 gm. of zinc dissolves to form zinc sulphate solution, about 1,300 calories of heat are set free.

In a simple copper-zinc cell, this energy appears as electricity and not as heat. Hence the electricity is obtained at the expense of the zinc which dissolves.

This is a very expensive way of producing electricity, but as a ‘ dry ’ cell is self-contained and easy to carry about, it provides a very *convenient* way of producing small electric currents. It would be far too expensive, however, to light a house or drive an electric fan with electricity obtained from voltaic cells. It is much cheaper to produce large quantities of electricity from a *dynamo* driven by a steam-engine whose energy is obtained from burning coal.

ACCUMULATORS

Voltaic cells, producing electrical energy directly from the chemical energy of a metal, are called *primary batteries*. When larger quantities of electricity are required out of reach of a wired supply from a dynamo (e.g. in a motor-car), an *accumulator* (or *storage cell*) is used. This type of cell is called a *secondary battery* because it does not produce electrical energy directly, but must first be charged with electricity, electrical energy being converted into chemical energy and stored in that form until an electric current is required. The stored-up chemical energy is then converted into electrical energy once more as the cell is discharged.

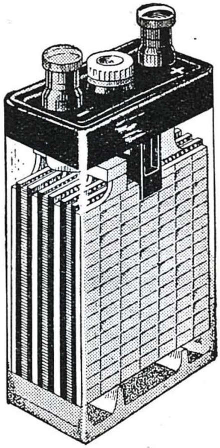

FIG. 199. LEAD ACCUMULATOR

Most accumulators consist of two *lead* plates surrounded by dilute sulphuric acid (1 part of acid in 4 parts of water). When *charged*, one plate becomes covered with grey, spongy, metallic *lead* ($-ve$), and the other with brown *lead peroxide* ($+ve$). As the cell is *discharged* (i.e. when current is taken from it), both the metallic lead on the negative plate and the lead peroxide on the positive plate are converted into insoluble *lead sulphate*† which covers both plates. Some sulphuric acid is removed from the battery acid during discharge, hence the specific gravity of the acid becomes less, in fact the best way to tell whether a lead accumulator is fully charged or not is to measure the specific gravity of the battery acid with a hydrometer. When the cell is re-charged by passing an electric current through it (from a dynamo) the lead sulphate is changed back again to metallic lead on the negative plate and to lead peroxide on the positive plate, sulphuric acid being set free at the same time, so that the specific gravity of the battery acid increases. When the battery is fully charged, it ' gasses ' freely, bubbles of hydrogen being set free at the negative plate and bubbles of oxygen at the positive plate.

The chief disadvantage of lead accumulators is their great weight. They are also damaged by too rapid charging or discharging.

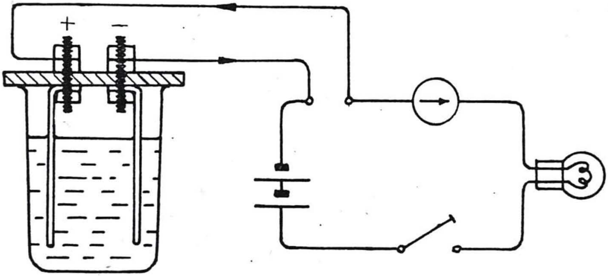

FIG. 200. CHARGING A SIMPLE LEAD ACCUMULATOR

Set up a simple lead accumulator by fixing two lead plates to a strip of wood by two terminals and dipping the plates in a jar containing dilute sulphuric acid (1 in 4). (*Care.* Battery acid is many times stronger than the dilute sulphuric acid in your bench reagent bottles.) Set up the circuit shown in Fig. 200, with a pocket-lamp battery (2 or 3 cells), a pocket-lamp bulb, a simple current-detector,[1] a tapping-key and your simple lead accumulator in series. Hold down the tapping-key and notice that the bulb lights faintly and the needle of your current-detector is deflected* to one side, showing that a small current is flowing through your ' accumulator '. Notice that bubbles of gas form on both plates and that after a few minutes one plate turns brown while the other remains grey. After ' charging ' for a few minutes, replace the pocket-lamp battery with a piece of copper wire and ' discharge ' your ' accumulator ' by pressing the tapping-key.

Notice that the bulb lights brightly for a short time, showing that your ' accumulator ' is discharging much more rapidly than you charged it. Notice also that the needle of your current-detector is

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[1] e.g. a pocket-compass lying over a wire or strip of metal as in Book III, p. 246, Expt. (i).

deflected to the opposite side, showing that the current is flowing in the reverse direction.

THE SCREW RULE

The last experiment shows that we can use a compass-needle to show which way a current is flowing in a neighbouring wire. Learn 'the *Screw Rule*': *Imagine a right-handed screw being screwed along the wire in the direction in which the current is flowing* (copper to zinc with a voltaic cell). *Then the direction in which the screw turns gives the direction of the magnetic lines of force* (i.e. the direction in which the N-pole of the compass-needle will be deflected).

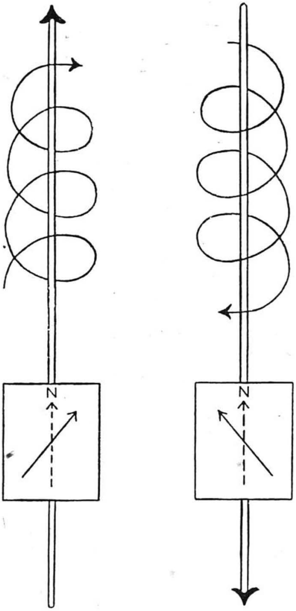

FIG. 201. THE SCREW RULE

Verify this rule with your 'charging-circuit'.

CHAPTER XVII

THE EFFECTS PRODUCED BY ELECTRIC CURRENTS

Electrical energy is easily converted into heat and into chemical energy. We have also seen that there is a magnetic field round a wire carrying an electric current. In everyday life we are mainly concerned with the *effects* produced by electric currents. These can be grouped under three headings: (i) *the Heating Effect*, (ii) *the Chemical Effect*, and (iii) *the Magnetic Effect*. We use the *heating effect* in electric heating and lighting; we use the *chemical effect* in electrolysis, electroplating and secondary batteries; we use *magnetic effect* in dynamos, motors, telephones and most electrical measuring-instruments.

THE HEATING EFFECT OF AN ELECTRIC CURRENT

We have seen that one of the great practical advantages of electricity over other forms of energy is that it can travel long distances through *conductors*, usually copper wires. Whenever an electric current passes through a wire, however, some electrical energy is converted into heat-energy. When small currents are passed through thick copper wires, the rise of temperature is only small and the heat produced soon escapes into the surrounding air; but when large currents are passed through thin wires of high ‘ *resistance* ’, much heat is produced and the wire gets very hot (as we have seen by the expansion produced in a nichrome resistance wire, Book I, p. 25). By passing measured electric currents through coils of fine resistance wire surrounded by water in a calorimeter, it can be shown that *the heat produced is proportional to* (*a*) *the square of the current,*[1] (*b*) *the resistance*[1] *of the wire*, and (*c*) *the time for which the current flows.*

ELECTRIC HEATERS

The heating effect of an electric current is widely used in everyday life to provide a clean and convenient source of heat. Electric heaters, cookers, irons, etc., get their heat from the heating effect of an electric current passing through coils of resistance wire. The wire used for

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[1] We shall learn in later lessons how *currents* and *resistances* are measured.

this purpose is usually *nichrome*, an alloy of nickel, chromium and iron, which has a resistance 50 times greater than that of copper and does not oxidize and burn away at a red heat. It is often flattened out into ribbon. This resistance wire or ribbon is wound on supports made of fire-clay, mica, or some other non-conductor which will stand high temperatures. In many electrical heating devices, the heat can be regulated by a ' heat-control switch ' which puts extra resistances into the circuit if greater heat is required.

Power-stations usually supply the same electrical energy at a cheaper rate for heating than for lighting purposes, because electric heaters and cookers are used mainly in the daytime when very little current is being used for lighting. By reducing the rate for electric power used for heating purposes, a power-station gets a steady load for its dynamos during the day as well as a heavy load in the evening when everyone switches on the electric light. A steady output results in cheaper running expenses for the power-station per unit of current produced.

FUSES

A *fuse*† is a short length of thin wire of fairly high resistance and of low melting-point, introduced into an electric light or ' power ' circuit to protect the rest of the wiring from a dangerous ' overload '. If the current becomes too great for safety, the fuse-wire gets hot, melts, and stops the flow of current before the low resistance wires in the rest of the circuit become dangerously hot. The fuse is usually enclosed in a fire-proof porcelain holder. In this way, the damage done by an ' overload ' or ' short-circuit ' is done in a safe place where it is easily repaired. If the new piece of fuse-wire ' blows ' when the current is switched on again, it is important to find the cause of the trouble—probably a ' short-circuit ' (i.e. a direct connection between the two opposite wires of the circuit).

ELECTRIC LIGHTING

In an electric heater, the resistance wire seldom becomes more than red-hot. At a higher temperature, the metal would combine with the oxygen of the air and burn away. If a wire is to be heated white-hot so that it gives out light (as well as heat), it must be heated in the absence of oxygen. The earliest electric bulbs used a thin thread (or *filament*†) of *carbon*, enclosed in a glass bulb from which all the air had been removed. These *carbon filament bulbs* gave a yellowish light because the filament could not be heated above about 1,750° C., and they converted about 0·5 per cent of the input of electrical energy into light-energy, i.e. over 99 per cent was converted into heat; in fact, carbon filament bulbs are used to-day for heating instead of lighting.

The filaments of modern electric bulbs are made of *tungsten*,† a metal which melts above 3,500° C. The first tungsten filament lamps had vacuum bulbs, and the filament could be heated to about 2,500° C., giving an efficiency of about 1 per cent, i.e. about twice the efficiency of a carbon filament bulb. At higher temperatures, however, the tungsten gradually vaporized and condensed on the inside of the bulb, blackening it and reducing the output of light. Modern electric lamps are 'gas-filled', the bulb being filled with *argon* to slow down the vaporization of the tungsten. In such 'gas-filled' bulbs, the filament can be run at about 3,000° C., giving a more efficient conversion of electricity into light. A later improvement was to wind the tungsten filament into a spiral coil so that the neighbouring coils 'keep each other warm', and a still later improvement was to first make a spiral of very tiny coils and then wind this coiled filament into a larger spiral, producing the 'coiled-coil' filament.

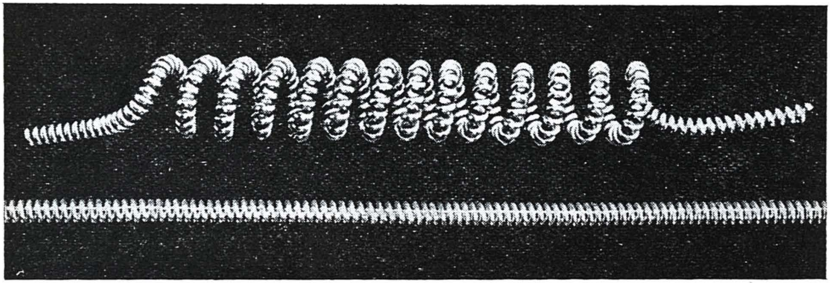

FIG. 202. COILED-COIL FILAMENT LAMP

Pieces of the Mazda coiled-coil filament (above) and the filament coil in an ordinary gas-filled lamp (below).

An electric bulb is usually marked with two numbers, e.g. 230V, 60W. The first number shows the *voltage*† (or 'electrical pressure') of the circuit on which the lamp is designed to be used and on which it will give 'the best value for money'. If a 230V lamp is used on a 110 volt circuit, it will not give out much light but will last for a very long time. On the other hand, if a 110V lamp is used on a 230 volt circuit, it will give out a very bright light but only for a very short time, because the filament gets too hot on the higher voltage.

The second number on the bulb shows the number of *watts* used, i.e. the *electrical power* required to light the lamp at the given voltage.

Modern electric bulbs use from 0·5 to 1·0 *watts per ' candle-power '*, which means that only about 2 per cent of the electrical energy is converted into light-energy. The remaining 98 per cent is radiated as heat and wasted. It is interesting to compare the lighting efficiency of our best electric bulbs (about 2 per cent) with that of the fire-fly† (nearly 100 per cent).

Electric discharge lamps, now used in many towns for street-lighting, work on an entirely different principle, an electric discharge being passed through mercury vapour or sodium vapour. Mercury vapour lamps give a greenish-blue light while sodium lamps give a golden-yellow light. These discharge lamps are several times more efficient than any filament lamps, but as they do not produce white light they are unsuitable for house-lighting. Another type of discharge lamp produces light by passing an electric discharge through a gas under a low pressure. *Neon* discharge tubes, for example, give a bright red light and are often bent into complicated shapes for use in advertising signs.

THE ARC LAMP

If two *carbon rods* are allowed to touch each other and a current of at least 50 volts is passed, the ends of the rods become red-hot. If the carbon rods are now drawn 2-3 mm. apart, a blinding white *electric arc* is formed at the gap. This arc consists of white-hot carbon vapour at a temperature of nearly 4,000° C., the highest temperature that Man can produce. An arc lamp converts into light energy about 10 per cent of the electrical energy supplied to it and provides our most efficient source of artificial light, but it has the disadvantage that the light is of an unpleasant colour. Arc lamps are used mainly for cinema-projectors and for search-lights, where it is important that the source of light should be concentrated at one point.

The high temperature of the electric arc is made use of in *arc welding*. An electric arc is produced at the point where two pieces of metal are to be joined, and the high temperature soon melts a little of the metal and welds the two pieces together.

THE CHEMICAL EFFECTS OF ELECTRIC CURRENTS

In earlier lessons we have seen that water containing a little sulphuric acid will conduct an electric current. This kind of conduction, however, is different from the conduction of a current by a copper wire, where the conductor remains unchanged after the passage of the current. The acidified water is split up into hydrogen and oxygen when a current passes through it, and this decomposition of a substance by electricity is called *electrolysis*. Substances which conduct electric currents and at the same time undergo decomposition are called *electrolytes*.† Water solutions of practically all *acids*, *bases* and *salts* are electrolytes. *Pure* water and most organic liquids, e.g. alcohol and oils, are *non-electrolytes*. During *electrolysis*, the current enters and leaves the electrolyte through conductors (usually metal plates) called *electrodes*. The electrode at which the current enters is called the *anode* (+*ve*); and the other, at which the current leaves, is called the *cathode* (−*ve*). During electrolysis, *hydrogen and the metallic elements are set free at the cathode, and all other elements at the anode.*

ELECTROLYSIS OF COPPER SULPHATE

Set up the circuit shown in Fig. 203. *X* is a pocket-lamp battery and *Y* a pocket-lamp bulb.

The anode *A* is a strip of thin copper sheet and the cathode *C* is a carbon rod[1] dipping into a 10 per cent solution of copper sulphate in water acidified with dilute sulphuric acid. After the current has been flowing for 5 minutes, take out the electrodes and notice that the part of the carbon cathode which was beneath the surface is now covered with copper (or *copper-plated*). Reverse the battery connections (i.e. make the copper-plated carbon the anode and the copper strip the cathode) and again pass the current for 5 minutes. Notice that the coating of copper disappears from the carbon. It is clear that *the electric current removes copper from the anode and deposits it on the cathode.*

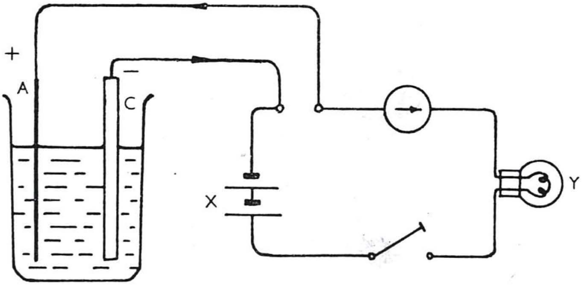

FIG. 203. ELECTRO-PLATING

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[1] e.g. the positive pole from an old pocket-lamp battery.

ELECTRO-PLATING

In the last experiment we copper-plated the carbon cathode. If we use a nickel anode instead of copper and a solution of a nickel salt as the electrolyte, we can *nickel-plate* the cathode. This process of covering a surface with a thin layer of metal, by electrolysis, is called *electro-plating* and it has many everyday applications. Gold, silver and chromium can be deposited on other metals in the same way, but in actual practice it is usual to apply a layer of copper before plating with these other metals.

ELECTROTYPES

Copper-plating is used in the printing industry for *electrotyping*,† i.e. producing a number of printing-plates from a single page of type. When a page of type has been set up, it is pressed into a sheet of wax or papier-mâché,† a paper-like material. This wax mould is covered with a thin layer of graphite to form a conducting surface, and is then made the cathode in a solution of copper sulphate with a copper anode. When a current is passed, a layer of copper is deposited on the wax mould, giving an exact copy of the original type. The wax is then melted away from the copper and the thin sheet is strengthened by pouring melted type-metal over its back. The copper *electrotype* is then used for printing, and the type distributed for use in another book. ' Electros ' are much lighter than lead type and take less room to store. By making several electrotypes of each page, a large number of copies can be printed much more rapidly than by using only the original type for the book.

The same principle is used in making gramophone-records. The original record is traced by the vibrating needle on a disc of wax. This wax record is then dusted with graphite and electro-plated with copper. The thin sheet of copper, with ridges corresponding to the wavy grooves in the original wax record, is backed with metal and pressed on round discs of a black, plastic material which has been softened by warming, thus ' printing ' a copy of the original sound-track on the record we play on a gramophone.

ELECTRO-METALLURGY

Several common metals are purified by electrolysis, e.g. the copper which is deposited on the cathode of an electrolytic cell is chemically pure. Such pure copper is in great demand for electrical purposes owing to its low resistance. Hence, copper used for electric wiring is refined by electrolysis. Thick sheets of crude copper are hung in a solution of copper sulphate and form the anodes. The cathodes are thin sheets of pure copper, and, when the current flows, copper is removed from the crude anodes and pure copper is deposited on the cathodes. Impurities fall to the bottom of the liquid.

Gold and silver can be refined by electrolysis in the same way.

Aluminium is now produced in enormous quantities by electrolytic methods. Although aluminium is one of the commonest elements in the Earth's crust (7 per cent), only in recent years has it become possible to extract the metal from its ores, because of the difficulty of reducing its oxide.

Carbon, for example, will not reduce aluminium oxide to metallic aluminium. *Bauxite*† (Aluminium oxide with some loosely-combined water) is the ore used as a source of aluminium. It is first purified and then heated to convert it into the anhydrous oxide, Al_2O_3, which is dissolved in melted *cryolite*† (a fluoride of aluminium and sodium) which melts at about 1,000° C. This liquid mixture is electrolyzed in an iron vessel lined with carbon (which serves as the cathode) with large carbon anodes dipping into the mixture. The liquid aluminium collects on the negative carbon floor and oxygen is set free at the carbon anodes, which gradually burn away. The cryolite is not used up during electrolysis and it is only necessary to add more aluminium oxide from time to time.

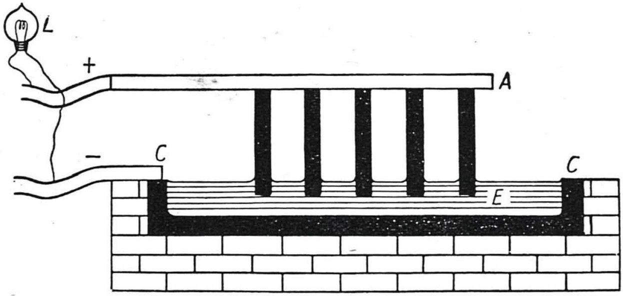

FIG. 204. ELECTROLYTIC PREPARATION OF ALUMINIUM

A—iron vessel lined with carbon which acts as the cathode. B—copper rod with carbon anodes attached. C—fused electrolyte. D—liquid aluminium. E—tapping-hole.

In chemistry lessons (p. 34) we have referred to the commercial importance of the electrolysis of brine (saturated sodium chloride solution) in the manufacture of sodium hydroxide. Metallic sodium and potassium are also prepared by electrolysis.

QUANTITY OF ELECTRICITY

As Faraday showed by experiment in 1832, when solutions of copper sulphate or silver nitrate are electrolyzed, the weight of copper or silver deposited depends only on *the quantity of electricity* which passes. This gives us a direct method of measuring *quantity of electricity*, and the practical unit of quantity of electricity is called the *coulomb*† (after the French scientist of that name). *A coulomb is the quantity of electricity which will deposit* 0·001118[1] *grams of silver.*

The ' strength ' of an electric current is *the rate of flow of electricity* through the circuit, i.e. the quantity of electricity passing per second. The practical *unit of current* is the *ampere*† (so called after the French scientist of that name). *An ampere is a flow of one coulomb per second*, or, *an ampere is a current which will deposit* 0·001118[1] *grams of silver in one second.*

We can test the accuracy of an *ammeter*† (or ampere-meter), therefore, by comparing the reading shown by the instrument with the weight of silver deposited during a known time from a solution of silver nitrate in an electrolytic cell placed in series with the ammeter.

The following everyday examples will give you a better idea of this unit of current: (i) An ordinary 60-watt bulb takes a current of about one-quarter of an ampere from the 230-volt house supply; (ii) a motor-car battery is charged at a rate of about 8-10 amperes when the car is running in the daytime; (iii) a 36-watt motor-car head-lamp bulb takes 3 amperes from a 12-volt accumulator; (iv) the ' self-starter ' motor of a car may take as much as 100 amperes from the accumulator (hence it is very bad for the car-battery if the starter-switch is kept closed for more than a second); (v) a 'valve' in a wireless receiver takes only a few milli-amperes (i.e. thousandths of an ampere) from the 'high tension' supply; (vi) the currents flowing in long-distance telephone-wires are only small fractions of a milli-ampere; (vii) fuses in house-lighting circuits usually melt when the current reaches 5 amperes.

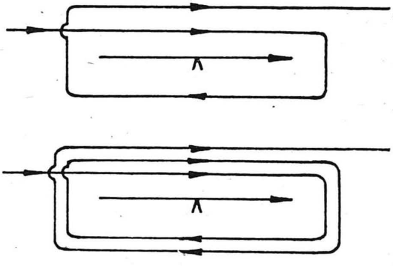

FIG. 205. OERSTEDT'S EXPERIMENT

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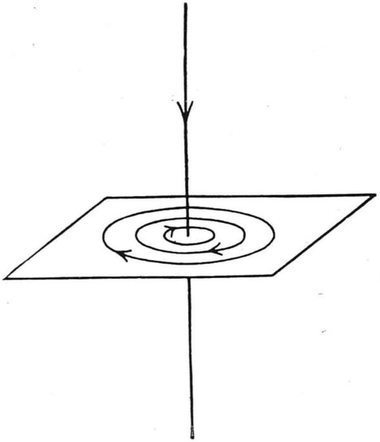
[1] You need not remember this figure.



THE MAGNETIC EFFECTS OF ELECTRIC CURRENTS

When a wire is held parallel to a compass-needle and an electric current is passed through it, the needle is deflected to one side. The Earth's magnetic field tends to turn the compass-needle into a north and south direction, while the action of the current tends to turn the needle into an east and west direction.

Under the combined influence of these two turning forces, at right angles to each other, the needle sets itself in such a direction that the two forces balance.

This magnetic effect of an electric current was first noticed in 1820 by Oersted, a Danish scientist, who thus discovered the close relation between *magnetism* and *electricity*: *that there is always a magnetic field round a conductor carrying an electric current.*

FIG. 206. MAGNETIC FIELD ROUND STRAIGHT CONDUCTOR

If the wire carrying the current is held first *over* and then *under* the compass-needle, the needle is deflected in opposite directions, but if the wire is carried first above the needle and then bent back again below the needle, as shown in Fig. 205, we get twice the turning effect. If we take the wire several times round the needle, over and under in turn, we can multiply the turning effect as much as we like. Such an arrangement gives us a very delicate method for detecting electric currents and for comparing their strengths, since the stronger current produces the greater deflection of the needle. This principle is applied in one type of current-measuring instrument (see p. 282).

THE MAGNETIC FIELD ROUND A STRAIGHT CONDUCTOR

If a thick copper wire is fixed vertically through a small hole in a horizontal sheet of cardboard covered with iron filings and a strong current (of at least 10 amperes) is then passed through the vertical wire, the filings arrange themselves in circles round the wire. With the current flowing downwards, pocket-compasses placed on the cardboard show that the direction of the lines of force is as shown in Fig. 206 (as you would expect from the Screw Rule). When the current is reversed, the direction of the lines of force is also reversed.

THE MAGNETIC FIELD ROUND A CIRCULAR CONDUCTOR

If a vertical *ring-coil* of wire is used instead of a single vertical wire, and iron filings are spread on the horizontal sheet of cardboard (*AB* in Fig. 207), when the current is passed the lines of force are found to be arranged as in Fig. 207. This magnetic field is of the same type as that surrounding a very short bar-magnet or a magnetic disc. If the number of

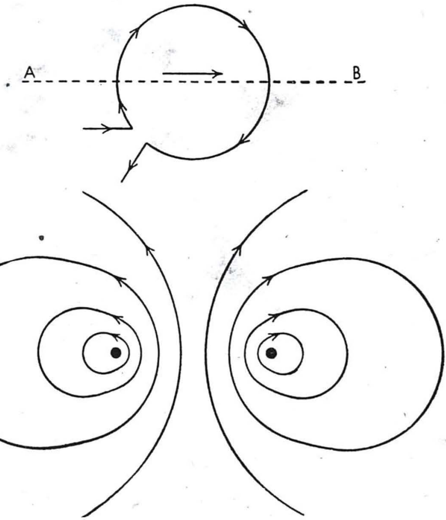


FIG. 207. MAGNETIC FIELD ROUND RING-COIL

circular turns of wire in the ring-coil is increased, the magnetic field will be strengthened in the same proportion.

If the wire is wound into a close spiral to form a cylindrical coil, the magnetic field produced on passing a current has its lines of force arranged as shown in Fig. 208. Notice that these lines of force are just like those round a long bar-magnet. If a soft-iron rod is placed inside such a cylindrical coil it becomes a strong magnet *while the current is flowing*. If a steel rod is placed inside the coil it becomes a permanent magnet; and this is how permanent magnets are usually made.

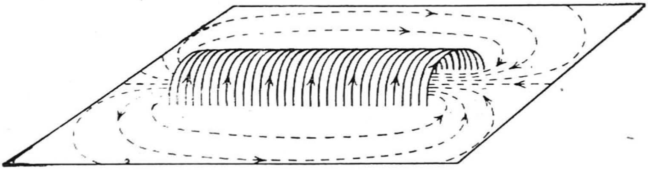


FIG. 208. MAGNETIC FIELD ROUND CYLINDRICAL COIL

It is sometimes necessary to find (*a*) which end of a coil corresponds to the N-pole of a magnet, or (*b*) which way the current is flowing in the coil. The easiest way of doing this is to remember the diagrams in

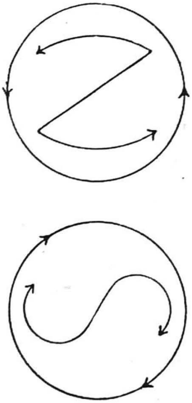


FIG. 209. RELATION BETWEEN POLARITY AND DIRECTION OF CURRENT

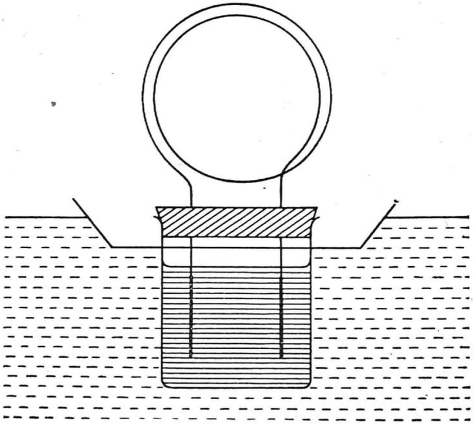


FIG. 210. FLOATING RING-COIL

Fig. 209. If the current flows as in (*a*), that end of the coil is a N-pole (or conversely, if that end of the coil is found to be a N-pole then the current is flowing *anti-clockwise* as we face that end of the coil). If the current flows as in (*b*), that end of the coil is a S-pole (or conversely, if that end of the coil is found to be a S-pole, then the current is flowing *clockwise* as we face that end of the coil).

This rule can be verified with the floating ring-coil shown in Fig. 210. The ring-coil, of many turns of covered copper wire, is mounted on a cork float and its ends are connected to copper and zinc plates below the cork. The plates dip into some dilute sulphuric acid in a beaker, forming a simple cell which sends a current round the coil from copper to zinc. When floating in a trough of water, the beaker turns until the faces of the coil look north and south. When the N-pole of a strong bar-magnet is brought near the coil, the face representing its S-pole is attracted and the other face is repelled.

CHAPTER XVIII

THE MEASUREMENT OF ELECTRICITY

CURRENT-MEASURING INSTRUMENTS

For everyday practical purposes, electric currents are nearly always measured by their magnetic effect, and there are three types of magnetic current-measuring instruments: (1) *moving-iron instruments*, (ii) *moving-magnet instruments*, and (iii) *moving-coil instruments*.

MOVING-IRON INSTRUMENTS

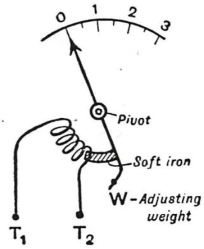


FIG. 211. MOVING-IRON INSTRUMENT (attraction type)

The simplest and cheapest kind of magnetic current-measuring instrument is the *moving-iron* instrument, which has the further advantage that it will measure both *direct currents* (D.C.) and *alternating currents* (A.C.). In the type shown in Fig. 211, the current to be measured passes through a coil between the two terminals† T_1 and T_2, drawing a piece of soft iron into the coil and thus moving the pointer across the scale. The attraction of the soft iron and the movement of the pointer depend upon the strength of the current passing through the coil.

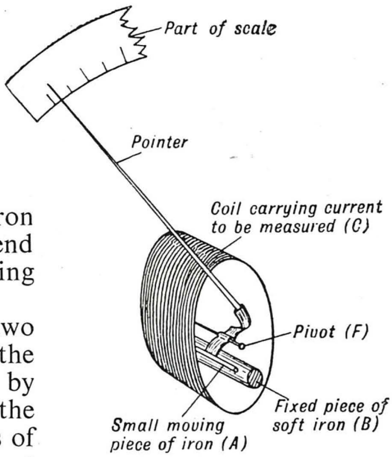


FIG. 212. MOVING-IRON INSTRUMENT (repulsion type)

In the type shown in Fig. 212, two pieces of soft iron (one fixed and the other pivoted to a pointer) lie side by side surrounded by the coil carrying the current to be measured. Both pieces of iron are magnetized when a current passes through the coil, and as their like poles are together they repel each other, moving the pointer across the scale. The repulsion between the two pieces of iron and the movement of the pointer depend on the strength of the current passing through the coil.

The pointer-scale of a moving-iron instrument is graduated by passing currents of different strengths through the instrument in series with an accurate standard instrument.

MOVING-MAGNET INSTRUMENTS

The earliest current-measuring instruments were of the *moving-magnet* type, but although they are cheap and easy to make they are never used for everyday electrical measurements outside an elementary laboratory.

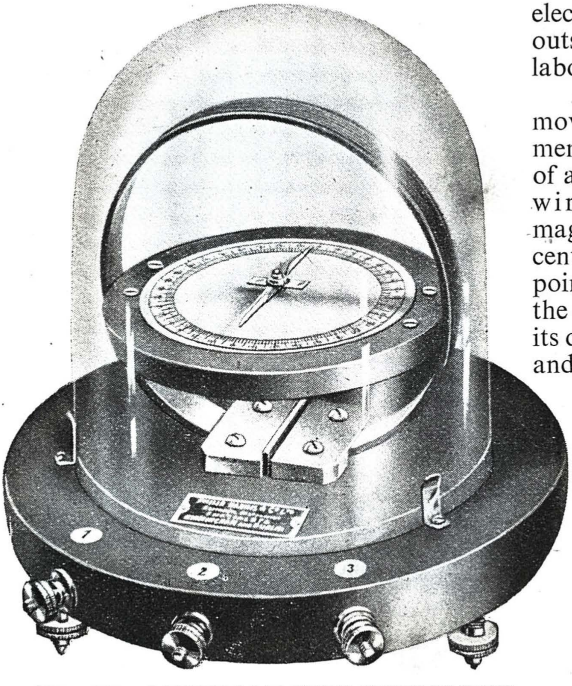


FIG. 213. MOVING-MAGNET INSTRUMENT

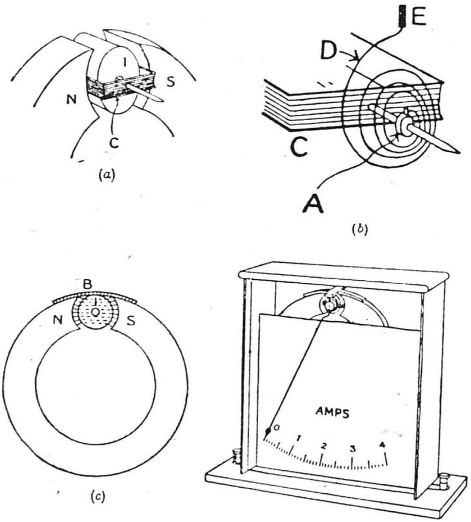
A simple type of moving-magnet instrument (Fig. 213) consists of a vertical ring-coil of wire with a small magnet pivoted at its centre. A long, light pointer is attached to the small magnet so that its deflection can be seen and measured more easily. The instrument is first placed with its coil and magnet in the magnetic meridian when no current is flowing. When a current is passed, the magnetic field due to the current will tend to deflect the magnet into an east and west direction while the Earth's magnetic field will tend to keep the magnet pointing in the plane of the coil. The magnet will come to rest in such a direction that these two forces just balance. The amount of deflection depends upon (*a*) *the strength of the current*, (*b*) *the strength of the*

Earth's magnetic field, (*c*) *the number of turns of wire in the coil*, and (*d*) *the radius of the coil*. The factors (*b*), (*c*) and (*d*) will be constant if the instrument is always used in the same place, hence the strength of the current can be measured.

Moving-magnet instruments have two serious disadvantages: (*a*) they must be used with the coil lying in the magnetic meridian, and (*b*) the pivoted magnet is affected by the presence of iron, or other magnets, near by. For these reasons, moving-magnet instruments are only of theoretical and historical importance.

MOVING-COIL INSTRUMENTS

Practically all accurate current-measuring instruments are of the *moving-coil* type, which is free from the disadvantages of the moving-magnet and moving-iron types. In this kind of instrument, there is a large, strong, *fixed magnet*, and a small, light, *moving coil* through which the current flows.



(*a*) (*b*) (*c*) (*d*)

FIG. 214. MOVING-COIL INSTRUMENT

We have seen (p. 279) that when a current flows through a ring-coil of wire, the coil behaves like a very short bar-magnet or a magnetic disc. If such a coil is hung between the poles of a circular magnet, it will tend to turn so that the N-pole side of the coil faces the S-pole of the permanent magnet.

As shown in Fig. 214, a cylindrical soft-iron core *I* is usually placed inside the coil so as to concentrate the magnetic lines of force between the poles *NS* of the permanent magnet. A rectangular coil *C*, of many turns of thin, covered wire, is free to turn round this soft-iron core, and its two ends are connected by two spiral springs to the terminals of the instrument (*D* is one of these springs).

When no current is passing, the springs hold the pointer at the zero mark on the scale, i.e. with the plane of the coil parallel to the magnetic lines of force crossing the gap between the poles of the permanent magnet. But for the springs, when a current flows round the coil it would turn until its N-pole side faced the S-pole of the magnet, but the spring stops the coil before it reaches this position. As the coil turns, it twists the spring, and the larger the current, the more it can twist the spring.

Such an instrument is used for measuring very small currents (e.g. thousandths of an ampere). For larger currents, a similar instrument called an *ammeter* is used, with its scale marked in *amperes*. A moving-coil instrument is made so that the turning effect on the coil is directly proportional to the current, hence the pointer-scale can be marked in equal divisions.

Moving-coil instruments have the great practical advantage that they are not affected by the Earth's magnetic field or any neighbouring magnets or pieces of iron, and can therefore be used anywhere.

HOT-WIRE INSTRUMENTS

A very cheap and simple current-measuring instrument can be made by using the *heating effect* of the current to be measured. As shown in Fig. 215, a fine resistance-wire *W* is stretched between the terminals of the instrument; and when a current passes through it, its temperature rises and it expands in length. The spring *S* pulls the thread *C* which is wrapped round the axle *F* of the pointer, thus moving the pointer across the scale. A hot-wire instrument will measure both D.C. and A.C.

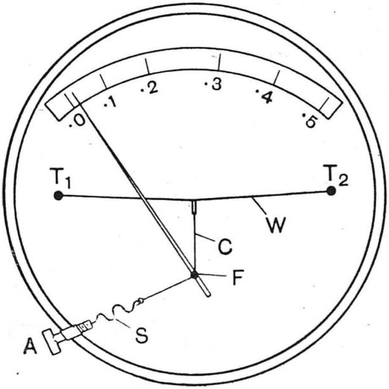


FIG. 215. HOT-WIRE INSTRUMENT

ELECTRICAL UNITS

In our lessons on heat we found that for heat to pass from one body to another, a difference of *temperature* is necessary, i.e. heat flows from the body at the higher temperature to the body at the lower temperature.

In the same way, in order that water may flow along a pipe, there must be a difference between the *pressures* at the two ends of the pipe. In much the same way, in order that an electric current may flow along a wire, there must be a difference of '*electrical pressure*' at the two ends of the wire.

ELECTROMOTIVE FORCE

The scientific name for this 'electrical pressure' is *electric potential*, and if one point in a circuit is at *a higher potential* than another, then electricity will flow from the first point to the second. The *potential difference* between two points is measured in *volts*, i.e. *the volt is the unit of electric potential* (just as the ampere is the unit of electric current).

If we wanted to describe the flow of water through a pipe we could state the *quantity* of water flowing per second (i.e. the rate of flow) and also the '*head*' of water (i.e. its pressure or 'driving-power'). In the same way, we measure the *rate of flow* of an electric current in *amperes* and the '*electric pressure*' in *volts* (using a voltmeter). For example, a single-cell lead accumulator has a *potential difference* between its terminals of about 2 volts, or it is said to have an *electromotive*† *force* of 2 volts for driving a current along a wire connecting the two terminals. In a Leclanché cell, or a 'dry' cell, the chemical action inside the cell produces an electromotive force (E.M.F.) of 1·5 volts. A simple copper-zinc cell has an E.M.F. of just over 1 volt.

ELECTRIC POWER

The main advantage of electricity over other forms of energy is that electric currents can be made to travel for long distances through wires, thus *carrying energy* to a distant point where it can be made to *do work*. We have learnt that *power* is the *rate of doing work*, and that mechanical energy is measured in *horse-power* units (33,000 foot-pounds per minute). Electrical energy is measured in *watts*. There is a simple relation between the units of *electric power*, *current* and *potential*: *If a current of one ampere flows at an E.M.F. of one volt, then the power is one watt*, or

Current (*in amperes*) × E.M.F. (*in volts*) = power (*in watts*), or
Amperes × volts = watts.

One Kilowatt = 1,000 watts. One H.P. = 746 watts. Therefore one Kilowatt = about $1\frac{1}{3}$ H.P.

Electrical energy is measured (and sold) in '*Board of Trade Units*', usually called simply '*units*', and this practical unit is the '*Kilowatt-hour*'. Thus, for example, when a current of 10 amp. at an E.M.F. of 10 volts flows for 10 hours it uses one Kilowatt-hour of electrical energy.

To understand the relation between amperes, volts and watts, you should work out various examples.

ELECTRICAL RESISTANCE

We now know something about two of the quantities involved in the flow of a current along a conductor: the *current* (measured in amperes) and the *electromotive force* (measured in volts). The third quantity is the *resistance of the conductor.*

Materials which allow a current to pass through them are called *conductors* (e.g. all metals and carbon), and their electrical *resistance* is said to be *low*. Materials which do not allow a current to pass are called *non-conductors* (e.g. glass, porcelain, rubber, paper, cotton, silk) and their electrical *resistance* is said to be very *high*. Some conductors allow a current to flow more readily than others, hence it is often necessary to *measure* the *resistance* of materials.

The *unit of resistance* is called an *ohm*,† after the German scientist of that name who, in 1827, carried out experiments on the flow of electric currents through metallic conductors, and found that *the current* (C) *flowing in a conductor is exactly proportional to the electromotive force* (E) *driving the current*, or $\frac{E}{C} =$ *a constant* (R). This is known as *Ohm's Law of Electrical Conduction.*

Hence every conductor has its own value for the ratio $\frac{E}{C}$, and this electrical constant R is called its *resistance*. When the current C is measured in *amperes* and the E.M.F. E in *volts*, then the resistance R is in *ohms*; e.g. if *one volt* applied to the ends of a wire causes *one ampere* to flow through it, then the wire has a resistance of *one ohm.* For practical purposes, the *ohm* is defined as *the resistance of a column of pure mercury at* 0° *C.*, 106·3 *cm. long and* 1 *sq. mm. in cross-section.*[1]

We have now defined two of the international electrical units: the *ampere* (p. 277) and the *ohm*, so we can now define the *volt* (by Ohm's Law) as *that electromotive force which drives a current of one ampere through a conductor whose resistance is one ohm.*

Ohm's Law is of the greatest importance in nearly all electrical measurements and calculations. To take a very simple example: If a motor-car head-lamp bulb is marked 12V. 36W. we know that it takes a current of 3 amps. and we can find the resistance of its filament when hot from $R = \frac{E}{C} = \frac{12}{3} = 4$ ohms.

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[1] There is no need to remember these figures.

You should apply Ohm's Law in calculating

$$\text{Current: } C \textit{ in amperes} = \frac{E \textit{ in volts}}{R \textit{ in ohms}}$$

$$\text{E.M.F.: } E \textit{ in volts} = C \textit{ in amperes} \times R \textit{ in ohms}$$

$$\text{Resistance: } R \textit{ in ohms} = \frac{E \textit{ in volts}}{C \textit{ in amperes}}$$

AMMETERS AND VOLTMETERS

A moving-coil current-measuring instrument like that described on page 284 can be used either as an *ammeter* for measuring *current* in *amperes* (Fig. 217), or as a *voltmeter*† for measuring *E.M.F.* in *volts* (Fig. 216) by connecting known resistances with the instrument.

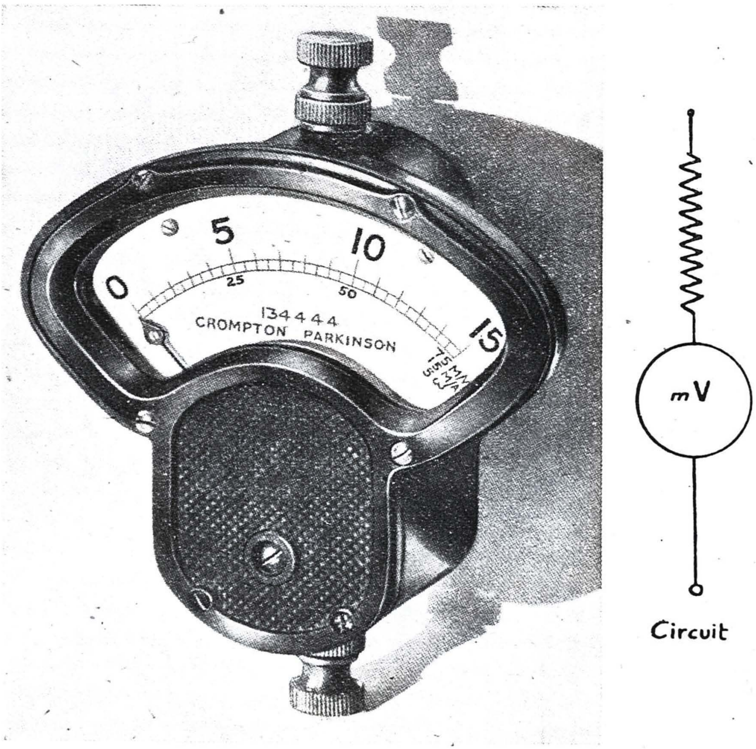


FIG. 216. VOLTMETER

Suppose that the moving coil has a resistance of 5 ohms and that a current of 0·015 amp. moves the pointer right across the pointer-scale. The scale could be divided into 15 divisions, each of which would be 0·001 amp., i.e. one milli-ampere, and the instrument would be a *milli-ammeter*. By Ohm's Law (E = CR), when 0·015 amp. flows through a resistance of 5 ohms, the E.M.F. is 0·015 × 5 = 0·075 volts, so that each division of the scale would represent 0·005 volts, i.e. 5 *milli-volts*, and the same instrument can also be used as a *milli-voltmeter*.

Now suppose we want to measure volts instead of milli-volts. If a resistance of 495 ohms is connected *in series* with the instrument, as shown in Fig. 216, the current will have to pass through a total resistance of 495 + 5 = 500 ohms and it will need an E.M.F. of 0·015 × 500 = 7·5 volts to drive the same *current* through the instrument and produce a full-scale deflection. Hence the instrument becomes a *voltmeter*, reading 0-7·5 volts, instead of a milli-voltmeter reading 0-75 milli-volts. Similarly, by putting a 4995 ohm resistance in series with the 5 ohm moving coil, the same instrument can be used to measure any E.M.F. between 0 and 75 volts. Notice that the same *maximum current* of 0·015 amp. passes through the moving coil in every case.

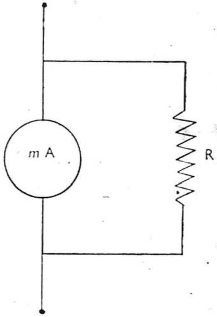


FIG. 217. AMMETER

If the same instrument is to be used for measuring 0-15 *amperes* instead of 0-15 milli-amperes, it is still necessary to limit the maximum current passing through the moving coil to 0·015 amp., and a lower resistance (999 times smaller) must be connected *in parallel*, as shown in Fig. 217, so that *exactly* 0·1 *per cent* of the current to be measured passes through the moving coil and the remaining 99·9 per cent is 'by-passed' or 'shunted' through the added resistance. Such a resistance, arranged *in parallel* across the terminals of the instrument, is called a *shunt*.† If the resistance of the moving coil is 999 times greater than that of the shunt, a 15 amp. current will split up so that 0·015 amp. goes through the moving coil and the remaining 14·985 amps. takes the easier way through the shunt, and the instrument gives a full-scale deflection which represents 15 amps.

In this way, a single moving-coil instrument, together with a set of *high* resistances for connection *in series*, can be used as a *voltmeter* for measuring a wide range of E.M.F.s, and the same instrument, together with a set of *low* resistances for connection *in parallel*, can be used as an *ammeter* for measuring currents of any strength.

CONDUCTORS IN SERIES AND IN PARALLEL

When cells, lamps, resistances and measuring instruments are connected up in an electrical circuit, they can be arranged in two different ways. If *all the current* passes through each conductor *in turn*, they are said to be '*in series*'. (In Fig. 218 (i), for example, the two resistances r_1 and r_2 are arranged in series.) If they are arranged so that *only part of the current* passes through each of them, they are said to be '*in parallel*'. (In Fig. 218 (ii), for example, the resistances r_1, r_2 and r_3 are arranged in parallel.)

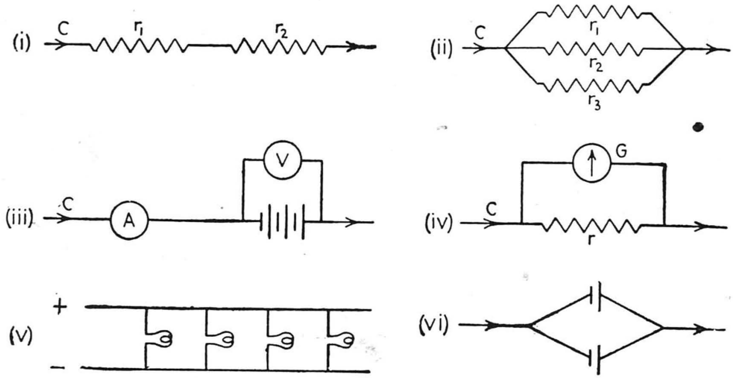


FIG. 218. 'SERIES' AND 'PARALLEL' CIRCUITS

When resistances are arranged *in series*, the total resistance of the series is the *sum* of the individual resistances, i.e. $R = r_1 + r_2$. In the same way, if three dry cells (E.M.F. $= 1 \cdot 5$ volts) are connected in series, the total E.M.F. of the combination is $4 \cdot 5$ volts. (Fig. 218 (iii) shows three cells connected in series.)

When resistances are connected *in parallel*, the current is offered a choice of conductors and splits up so that the largest part of the current flows through the smallest resistance, i.e. the current flowing through any one conductor is *inversely proportional to its resistance*. In Fig. 218 (ii), for example, if the equivalent single resistance is R, the current flowing through r_1 is $\frac{R}{r_1}$ of the total current, r_2 carries $\frac{R}{r_2}$ of the total current, and r_3 carries $\frac{R}{r_3}$ of the total current. The relation between the total resistance R and the individual resistances r_1, r_2 and r_3 in parallel is $\frac{1}{R}=\frac{1}{r_1}+\frac{1}{r_2}+\frac{1}{r_3}$. Notice that the sum of a number of resistances in parallel is always *less* than any single resistance, because each additional resistance provides a path for more current to flow.

When several similar *cells* are connected *in parallel* (Fig. 218 (vi)), the total E.M.F. is the same as that of each single cell.

Since an *ammeter* measures the *quantity of electricity* flowing in a circuit, it must be connected *in series* so that *all the current flows through it* (see Fig. 218 (iii)). A *voltmeter*, however, measures only the *potential difference* between two points in the circuit and is therefore connected *in parallel* across these two points (see Fig. 218 (iii)). Perhaps you will find this easier to understand if you think of the air in a motor-car tyre. If we want to measure the *quantity* of air in the tyre we have to let out *all* the air and measure its volume, just as we have to make *all* the current pass through the *ammeter*. If we only want to measure the *pressure* of the air in the tyre, it is not necessary to make all the air pass through the pressure-gauge: only a very little air is used, just as only a very small fraction of the current flows through the *voltmeter* when measuring potential difference.

CHAPTER XIX

MOTORS, DYNAMOS AND ELECTRIC POWER SUPPLIES

ELECTRIC MOTORS

If you understand the principle of the moving-coil instrument (p. 284) it is easy to explain how a simple electric motor works. If a current is passed through the moving coil of an ammeter from which the spiral springs have been removed (Fig. 219), the coil *ABCD* will turn so that its N-pole side faces the S-pole of the permanent magnet, and it will stop in that position as long as the current flows. If the current is then reversed, however, what was the N-pole side of the coil becomes the S-pole, and the coil will turn through 180° so that this S-pole faces the N-pole of the permanent magnet. Hence, *if we can reverse the direction of the current at exactly the right moment and if we supply current to the coil in such a way that the coil is free to turn, then the coil will go on revolving as long as a current is passed.* Such an arrangement (shown in Fig. 220) would be a simple *electric motor.*

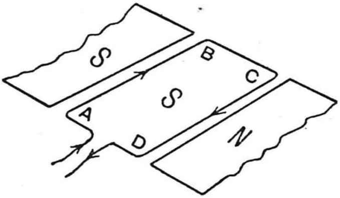


FIG. 219. MOVING COIL

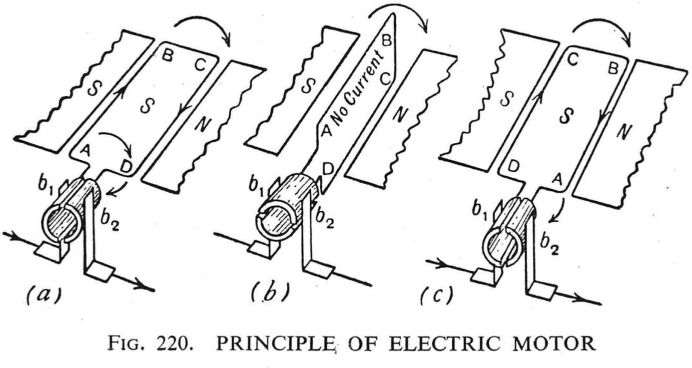


FIG. 220. PRINCIPLE OF ELECTRIC MOTOR

In most practical electric motors, *electro-magnets*† are used instead of permanent magnets (Fig. 221). These electro-magnets (or *field-magnets*) are made of soft iron and in such a shape as to form a cylindrical gap between the poles. The moving part of the motor is called the *armature*† and it consists of an axle carrying a soft-iron core wound with covered wire. The ends of each coil of wire in the armature are connected to a *commutator*† (or *current-reverser*). In its simplest form, i.e. with a two-pole armature, this consists of the two halves of a brass ring which revolve with the armature as shown in Fig. 227. Two *brushes*, strips of copper or carbon, press one against each half of the commutator so that as the armature revolves, each brush makes contact with each end of the coil-winding in turn. In this way, the current is reversed as

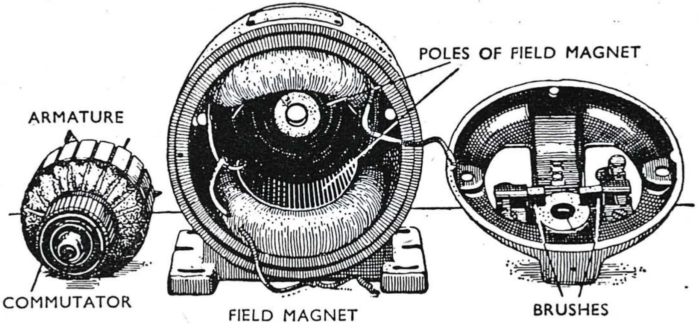


FIG. 221. PRACTICAL ELECTRIC MOTOR

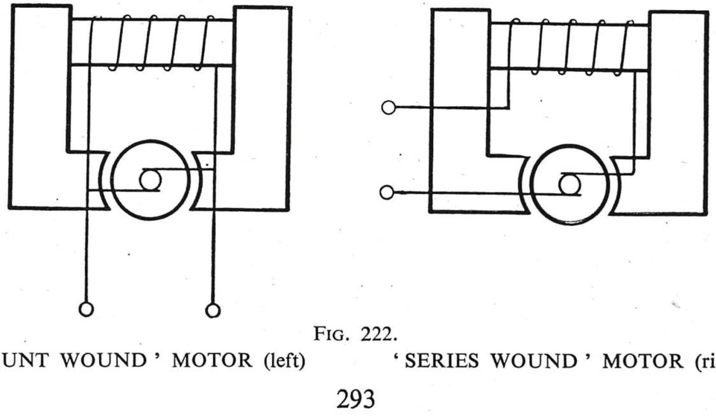


FIG. 222.
'SHUNT WOUND' MOTOR (left) 'SERIES WOUND' MOTOR (right)

the brushes slip from one half of the commutator to the other, the current entering through one brush and leaving through the other.

Practical electric motors (see Fig. 221) always have more than one coil in the armature, and the greater the number of coils the smoother the running of the motor.

The current for the field-magnets is taken from the same source as that supplying the armature, but the connections can be made in two different ways, as shown in Fig. 222, giving either a *series-wound motor* or a *shunt-wound motor*. In a series-motor, the field-magnet coils are in series with the armature-coils, so that the whole current passes through both sets of coils. Series-motors ‘ pull ’ hard at slow speeds and are therefore used for electric trains and tram-cars, giving rapid acceleration when starting from rest. Shunt-motors have their field-magnet coils connected in parallel across the armature-coils, so that part only of the current passes through the field-magnet coils and part through the armature. Shunt-motors will maintain nearly the same speed under varying loads and are therefore used for driving machinery, e.g. lathes,† where a fairly steady speed is required under varying loads.

DYNAMOS

When we supply *electrical energy* to an *electric motor*, the armature revolves and supplies *kinetic energy* which can be made to *do work*. Conversely, if we supply the same motor with *kinetic energy* by turning the armature, it supplies *electricity*. To understand how a dynamo works we must first learn something about *electro-magnetic induction*.

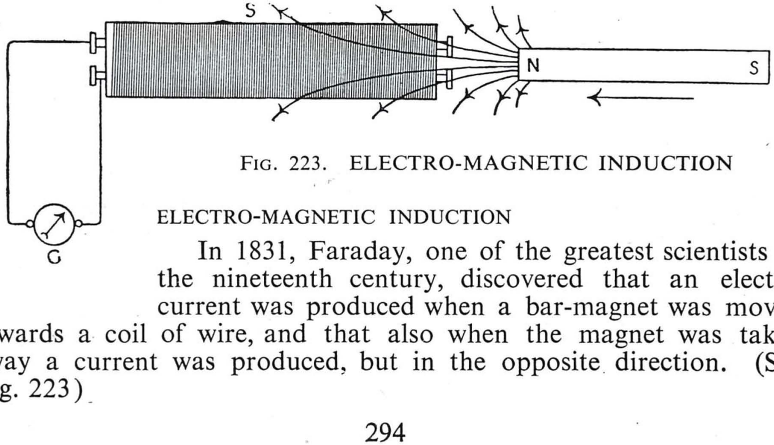


FIG. 223. ELECTRO-MAGNETIC INDUCTION

ELECTRO-MAGNETIC INDUCTION

In 1831, Faraday, one of the greatest scientists of the nineteenth century, discovered that an electric current was produced when a bar-magnet was moved towards a coil of wire, and that also when the magnet was taken away a current was produced, but in the opposite direction. (See Fig. 223)

This effect is called *electro-magnetic induction* and the electric current is said to have been *induced* in the coil.

Faraday also found that another coil carrying a current could be used instead of a bar-magnet. When the smaller coil S_2 in Fig. 224, carrying the current, is either pushed into or withdrawn from the larger coil S_1, a current is induced in the larger coil.

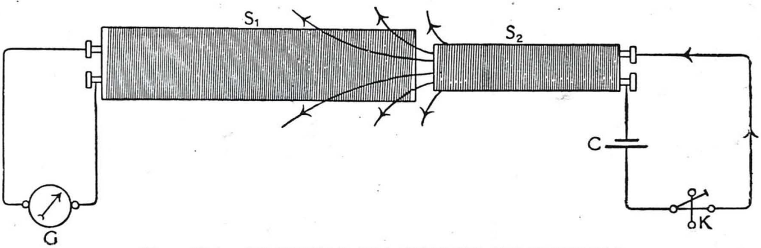


FIG. 224. ELECTRO-MAGNETIC INDUCTION

The same effect can be produced without moving the coils if a switch *K* is included in series with the smaller coil S_2 which carries the current. Faraday put a small coil (the *primary coil*) connected to a battery and switch, inside a larger coil (the *secondary coil*) connected to a current-detector *G*. On closing the switch in the primary circuit, the compass-needle in the secondary circuit was deflected, and when the switch was opened again, there was another deflection of the needle, but in the opposite direction. Similarly, when the current was reversed, the deflections of the needle were reversed.

This production of an *induced current* in a circuit when a current flows in a neighbouring circuit is the basis of all dynamos, transformers and induction coils, and it is used in telephones and in wireless transmitters and receivers. You should remember, therefore, that *the movement of a magnet, or the change in an electric current, always induces currents in neighbouring circuits.*

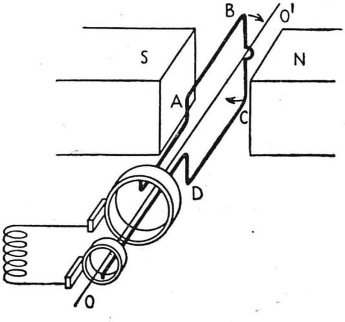


FIG. 225.
PRINCIPLE OF DYNAMO

DYNAMOS

We have seen that the *dynamo* is the converse of the *electric motor*, and practical dynamos are made in the same way as motors with armatures consisting of many turns of covered wire wound on a soft-iron core, revolving between the poles of an electro-magnet. When the armature is turned, a current is induced in its coils and passes through the coils of the field-magnets. After every half-turn of a simple, 2-pole, single-coil armature, the induced current is reversed; and if the ends of the armature-coil are connected to an external circuit, as in Fig. 225, we shall get an *alternating current* (A.C.), flowing first one way and then the other. Fig. 226 shows how the current starts from zero with the coil in the position shown in Fig. 225, rises to a maximum when the armature has turned through 90°, and falls to zero again after 180°. For the next half-turn, the current flows in the opposite direction, increasing to a maximum (−ve) after another 90° and falling to zero again at the end of the complete turn. The whole cycle of current changes is then repeated during the next revolution.

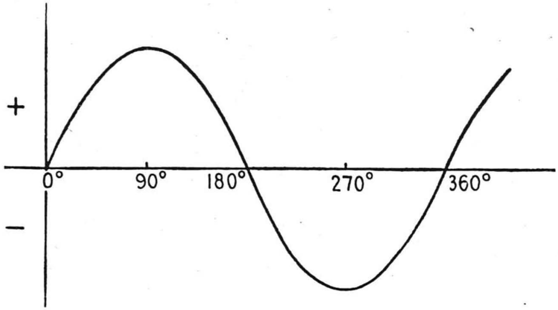


FIG. 226. GRAPH OF ALTERNATING CURRENT FROM TWO-POLE ARMATURE

If a *direct current* (D.C.) is required, a commutator or current-reverser (Fig. 227) must be used to reverse the current every half-revolution of the 2-pole armature. Fig. 228 shows the effect of the commutator in producing a current flowing in one direction only (although not a *steady* current).

Practical dynamos are never made with a single-coil, 2-pole armature, because such an arrangement does not give a steady current. The larger the number of coils the steadier the current produced (just as a modern four-cylinder motor-engine gives smoother running than the early single-cylinder engines). Practical dynamos, therefore, have the armature wound with a large number of coils, and have the same number of segments in the

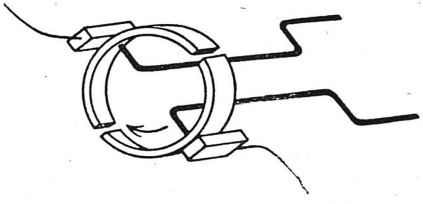


FIG. 227. PRINCIPLE OF COMMUTATOR OR CURRENT-REVERSER

commutator. Fig. 229 shows the current produced by a 4-pole armature (i.e. two coils fixed at right-angles to each other). Notice that when the current from one coil is at a maximum, the current from the other coil is at a minimum (as shown by the dotted lines) so that the combined current never falls to zero. The dark line is a graph of the current produced by the double-coil, 4-pole armature, showing that the current is steadier (or 'smoother') than with a single-coil, 2-pole armature.

It is clear that the larger the number of coils, the smoother will be the current.

A *magneto*† is a dynamo with permanent steel magnets producing an alternating current of high voltage. In the small permanent-magnet dynamo used for lighting bicycle-lamps, the coils are usually stationary and the magnets revolve.

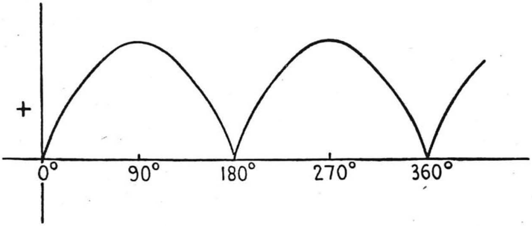


Fig. 228. GRAPH OF DIRECT CURRENT FROM TWO-POLE ARMATURE

THE INDUCTION COIL

For some purposes, small currents at a very high voltage are necessary. Thus 'X'-ray tubes and the spark-plugs† of motor-cars require small currents of only a few milli-amperes with a large E.M.F. of thousands of volts. Such currents are most conveniently produced by an *induction coil.* The *primary coil,* consisting of a few hundred turns of thick, covered copper wire, is wound on a soft-iron core (usually a bundle of soft-iron wires). The *secondary coil,* consisting of many thousands of turns of very thin, covered copper wire, is wound over the primary coil.

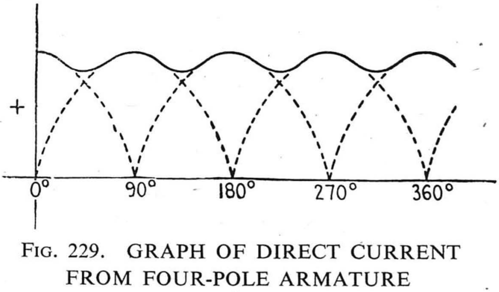


Fig. 229. GRAPH OF DIRECT CURRENT FROM FOUR-POLE ARMATURE

The *primary coil* is connected to a battery in series with a ' *make-and-break* ' (or current-interrupter) like that used in electric bells (see Book III, p. 250). The ends of the *secondary coil* are connected to terminals or two adjustable arms between which sparks can pass.

At each ' make-and-break ' (i.e. every time the primary current starts or stops), a current is induced in the secondary coil, but since the ' break ' is much more sudden than the ' make ', its effect is much greater, and the induction coil gives a rapid succession of rushes of high-voltage direct current instead of the alternating current which would be produced if the ' make ' and the ' break ' had equal effects.

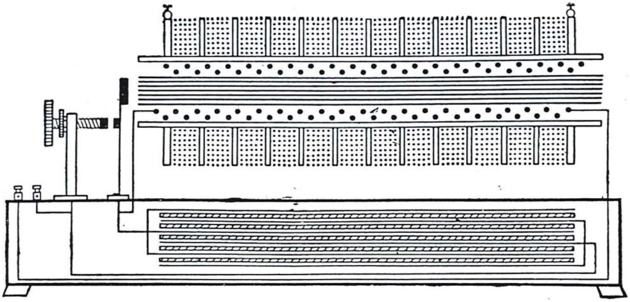


FIG. 230. INDUCTION COIL

If the potential difference between the terminals of the secondary coil is big enough, a spark passes between them. In order that a spark may pass between two sharp points one inch apart (in air), a potential difference of about 20,000 volts is necessary. To produce a spark across the normal gap of the spark-plug of a motor-engine requires a potential difference of about 5,000 volts at the pressure inside the cylinder at the top of the compression-stroke. In the open air, about 2,000 volts is sufficient.

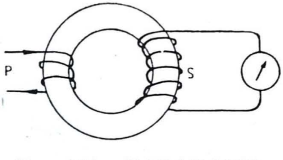
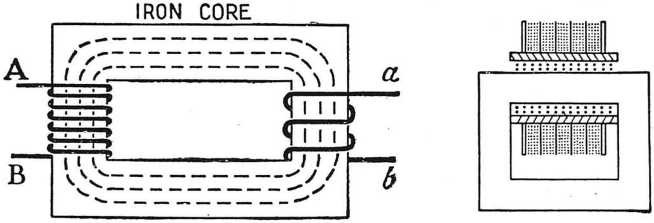


FIG. 231. FARADAY'S FIRST TRANSFORMER

THE A.C. TRANSFORMER

One of the most important everyday applications of electro-magnetic induction is in *transformers*. Faraday, in 1831, made a soft-iron ring and wound two separate coils of covered copper wire round it, as shown in Fig. 231. The ends of one coil (the *primary*) were connected to a battery through a switch, while the ends of the other coil (the *secondary*) were connected to a current-detector. When the switch in the primary circuit was closed, the compass-needle in the secondary circuit was deflected for a moment, but returned to its original zero position until the switch was opened again. The needle was then deflected in the opposite direction for a moment.

It is important to notice that there is no induced current in the secondary coil as long as the primary current is steady, *but only when it changes.* An *alternating current,* however, is rapidly changing in strength and direction (see Fig. 226), hence if an alternating current is passed through the primary coil, an alternating current of the same frequency will be induced in the secondary coil. This is the principle of the *transformer* which is used in the distribution of electrical power. Practical transformers consist of a primary and a secondary coil of covered wire wound on a soft-iron core as shown in Fig. 232, and by varying the number of turns in the primary and secondary windings, an alternating current can be '*stepped up*' or '*stepped down*' to any voltage.



AB—secondary. ab—primary.

FIG. 232. TRANSFORMER

If the primary and secondary coils have the same number of turns, the induced E.M.F. in the secondary coil will be the same as the E.M.F. applied to the primary coil. If the secondary coil has twice as many turns as the primary, then the E.M.F. induced in the secondary will be twice that in the primary coil. In general:—

$$\frac{\text{Secondary voltage}}{\text{Primary voltage}} = \frac{\text{number of turns in secondary}}{\text{number of turns in primary}}$$

Hence, by varying the number of turns in the primary and secondary windings of a transformer, we can change the voltage of an alternating current supply to suit any requirements. A 'step-up' transformer has fewer turns in the primary than in the secondary, and it 'steps up' the supply from a lower to a higher voltage. A ‘ step-down ’ transformer has more turns in its primary coil than in its secondary, and ‘ steps down ’ a high-voltage alternating current to a lower voltage.

We cannot, however, get more *electrical energy* out of the secondary coil than we put into the primary coil; i.e. the *output* (in watts) is equal to the *input* (in watts) if we neglect small heat losses, and what we gain in *voltage* we lose in *current* (and *vice-versa*). For example, suppose a transformer has 1,000 turns in the primary to every 100 turns in the secondary (i.e. it is a 10-to-1 step-down transformer), and the input to the primary is a current of 1 amp. at 100 volts (=100 watts). Then the output from the secondary will be a current of 10 amp. at 10 volts (=100 watts), and we have transformed a small current at a high voltage into a large current at a low voltage.

TRANSMISSION OF ELECTRICAL POWER

In most of our elementary lessons on electricity we have dealt with the effects of direct currents (i.e. currents flowing steadily in one direction) but all modern power supplies are *alternating current* (i.e. a current flowing backwards and forwards many times a second—50 cycles per second in most public lighting supplies). Although A.C. is as good as D.C. for driving motors, heating, and lighting, it is useless for electrolysis or for charging accumulators, and gives much more dangerous shocks than D.C. at the same voltage. The fundamental reason why most power-stations supply A.C. is because transformers can be used with A.C. but not with D.C. An alternating current can be transformed to any required voltage, either up or down, with very little loss of energy. With D.C. it is much more difficult and wasteful operation to change the voltage.

The necessity for changing the voltage of power supplies arises in connection with the *long-distance transmission* of electric currents. We have seen (p. 270) that some heat is produced whenever a current flows through a conductor. The rate at which energy is wasted in useless heating of electric transmission wires depends only on two factors: (*a*) the *resistance* of the wires, and (*b*) the *current* flowing through them (the heat produced being proportional to the *square* of the current). The resistance of the wires is kept low by making them of copper or aluminium. The current is kept low by raising the voltage, and since the quantity of electrical power (in watts) transmitted is the product of the current (in amperes) and the voltage (in volts), the same amount of electrical *energy* is carried by a small current at a high voltage as by a large current at a low voltage, and the loss of energy will be much less with the smaller current.

To take a practical example: Suppose it is required to supply 230 Kilowatts from a power-station to a factory 3 miles away. Let us compare the relative cost and efficiency of transmitting this energy (*a*) at 230 volts and (*b*) at 6,000 volts.

	(*a*)	(*b*)
Distance for transmission	3 miles	3 miles
Electrical energy to be supplied	230,000 watts	230,000 watts
Voltage of supply at factory end	230 volts	6,000 volts
Current required (watts ÷ volts)	1,000 amps.	38·3 amps.
Cross-section of wire required to carry this current without overheating	1 sq. in.	0·04 sq. in.
Approximate weight of wire required	29 tons	1·1 tons
Approximate cost of this copper wire	£2,320	£90
Resistance of wire	0·125 ohms	3·265 ohms
'Voltage Drop' (Ohm's Law: E=C×R)	0·125×1,000= 125 volts	38·3×3·265= 125 volts
Voltage of supply at power-station end	230+125= 355 volts	6,000+125= 6,125 volts
Energy wasted in transmission by useless heating of wires. (Voltage Drop × Current)	125×1,000= 125 Kw.	125×38·3= 4·8 Kw.
Relative Efficiency of Transmission	45·6%	97·9%

Notice that by increasing the voltage 26 times, the transmission losses are reduced 26 times. Besides this, the cost of copper is 26 times less when the higher voltage is used.

Transformers are used for stepping up the voltage of power supplies for long-distance transmission. Such high voltages, however, are very dangerous and have to be stepped down by other transformers to give a lower voltage which will be safe for household use. To take an actual example: The output of a power-station in a large town is at 6,600 volts. For transmission to another town 20 miles away, this is stepped up to 66,000 volts. On arrival at the second town it is stepped down to 11,000 volts for transmission to the 'sub-stations' in different parts of the town, from whence it is distributed to houses at 400 and 230 volts, the latter being the highest voltage which is at all safe for household use.

To change a direct current from one voltage to another involves feeding an electric motor with the first current and making this motor drive a dynamo producing current at the required voltage, a much more complicated, expensive and wasteful process than changing the voltage of an alternating current with a transformer.

Another advantage of transformers is that when the primary is connected to a supply of A.C., only a small current passes when the secondary is ‘ open ’ and no current is being taken from it. The primary current is said to be ‘ choked ’. When the secondary circuit is ‘ closed ’ and current is taken from it, the primary is no longer ‘ choked ’ and a larger current flows through it. For example, a *bell-transformer*, used instead of dry cells for ringing electric bells, steps down 230 volts A.C. to 6 or 8 volts, and can be left permanently connected to the electric light mains because the primary takes a very small current indeed until the secondary is closed by pressing a bell-push. Then, as soon as a current flows in the secondary, the primary takes its full current from the mains.

CONVERSION OF A.C. TO D.C.

There are several methods of converting A.C. to D.C. One method is to feed the A.C. to an A.C. motor driving a D.C. dynamo. This arrangement is called a *rotary* converter* and it gives a *smooth* D.C. output.

Another method is to use an *electrical valve*, which allows current to flow in one direction but not in the other.

Such a *valve-rectifier* converts (or *rectifies*†) an A.C. into pulses of D.C. This pulsating D.C. can be ‘ smoothed ’ by passing it through a ‘ *choke* ’ or ‘ choking-coil ’ (which is simply a transformer with only one coil).

Small D.C.s at a low voltage are very often required in everyday life, e.g. for charging accumulators and for wireless sets, and these are most easily obtained by using a *copper oxide rectifier*, which consists of discs of copper covered with a layer of copper oxide on one side only. Currents flow through such a copper-copper oxide contact much more readily in one direction than in the other, so that when an A.C. is passed, if the positive half of the wave flows across the copper-copper oxide boundary, the negative half is stopped. In this way, an A.C. is *rectified* to produce a pulsating D.C. (which can be ‘ smoothed ’ by passing through a ‘ choke ’).

CHAPTER XX

TELEGRAPHS AND TELEPHONES

We have seen that one of the most important everyday uses of electric currents is to convey electrical *energy* to a distant point where it can be converted into other forms of energy. Almost equally important is the use of electric currents to convey *messages* over long distances. Before the invention of telegraphs and telephones, news and messages had to be written and sent by land or sea, so that they travelled with no greater speed than Man himself (except, perhaps, when carrier-pigeons were used). Using telegraphs and telephones, messages arrive at a distant ' receiver ' at almost the same instant as they are sent by the ' transmitter '.

The *telegraph* was the earliest instrument to use an electric current for carrying messages. One of the simplest types uses an electro-magnet at the receiving end, worked by currents sent along a wire by the *tapping-key* at the transmitting end. When the tapping-key is pressed, the circuit is closed and a current flows along the wire and through the electro-magnet at the receiving end, attracting a piece of soft iron which either makes a sharp sound or presses an inked wheel against a moving strip of paper. By using the Morse Code,* where each letter of the alphabet and every number is represented by a combination of ' dots ' and ' dashes ', a message can be transmitted as a series of impulses of current, a short impulse being a ' dot ' and a longer impulse a ' dash '.

In modern *teleprinters*,† the operator taps out the message on a transmitting instrument which looks like a typewriter and sends electrical impulses along the wires to a similar machine at the receiving end which types out the message on a strip of paper. In post-offices, this strip of paper is cut up and gummed on a telegram form.

THE TELEPHONE

Nowadays, the telegraph is used mainly for long-distance transmission, and the *telephone* is a more familiar instrument in daily life. In the telephone *transmitter*, sound-waves (i.e. waves of compression and rarefaction, or variations of air-pressure) produce variations of electric current which travel to the *receiver* where these variations of current are converted once more into variations of air-pressure (or sound-waves).

To understand how the telephone transmitter works, you should first examine a simple *carbon microphone*,† consisting of a hard carbon rod, sharpened at both ends and held loosely between two carbon plates as shown in Fig. 233. The electrical resistance at these carbon contacts depends on the force pressing them together, and anything which changes the pressure on these contacts also changes the current flowing in the circuit. Sound-waves will do this, so that the microphone will convert variations of air-pressure into variations of electric current.

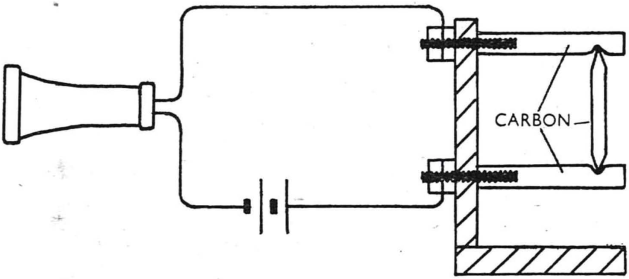


FIG. 233. SIMPLE CARBON MICROPHONE

If a watch is placed near such a carbon microphone, connected in series with a dry battery and a telephone receiver, the ticking of the watch can be heard in the receiver.

In practical telephone transmitters like the one shown in Fig. 234, a large number of grains of carbon are used instead of the two contacts of the simple microphone, and the thin plate

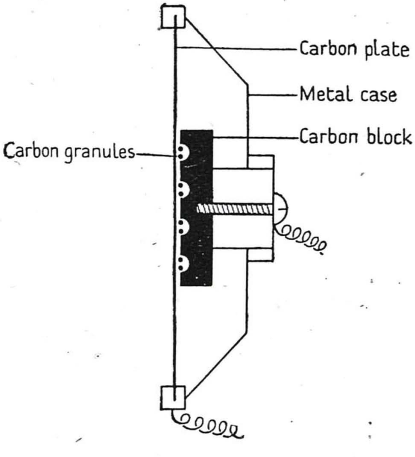


FIG. 234. TELEPHONE TRANSMITTER

(or *diaphragm*) vibrates backwards and forwards ' in step ' with any sound-waves falling on it. When the diaphragm is pushed inwards by a wave of compression, the grains of carbon are pressed more tightly together and more current flows, and when the diaphragm moves outwards during a wave of rarefaction, the current is weaker. Hence the variations in air-pressure caused by the sound-waves are converted into variations of electric current by the transmitter.

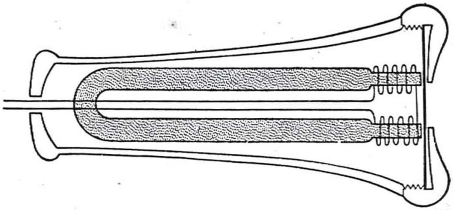


FIG. 235. TELEPHONE RECEIVER

In the telephone *receiver*, this varying current passes round coils of wire surrounding the poles of a permanent magnet which attracts a thin iron plate. Variations in this current produce corresponding variations in the strength of this magnet, so that when a wave of compression strikes the diaphragm of the *transmitter* and a stronger current flows round the coils of the *receiver*, the iron plate is pulled inwards and when a wave of rarefaction at the transmitter produces a weaker current, the iron plate in the receiver flies back, thus vibrating ' in step ' with the diaphragm of the transmitter. The vibrations of the diaphragm in the receiver set up sound-waves corresponding to those reaching the transmitter. Fig. 236 shows a common type of combined transmitter and receiver.

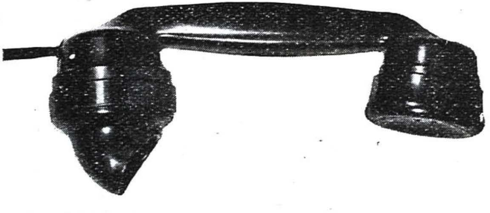


FIG. 236. COMBINED TRANSMITTER AND RECEIVER

Fig. 237 shows a practical telephone circuit. Only low voltages can be used in the microphone-circuit (not more than 3 volts), otherwise the carbon contacts burn and crackling noises are produced, hence the low-voltage current from the transmitter is stepped up for transmission through the long wires leading to the distant receiver (just as in power-transmission) to reduce losses due to line-resistance.

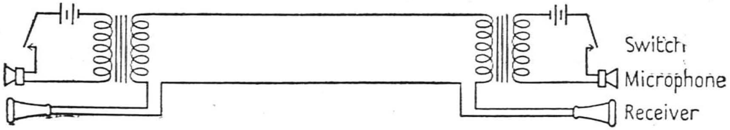


FIG. 237. TELEPHONE CIRCUIT

CHAPTER XXI

THE NATURE OF ELECTRICITY

In order to understand the nature of electricity, some knowledge of the *constitution* of matter* is necessary, since, during recent years, it has been found that *all matter is built up of electrical charges, and of nothing else.*

In our earlier lessons we have learnt that matter is made up of very small particles called *molecules*; a molecule being the smallest portion of a substance which possesses the characteristic properties of that substance. Molecules themselves are built up of chemical *atoms* held together by electrical attractions, an atom being the smallest portion of an element which can enter into chemical combination.

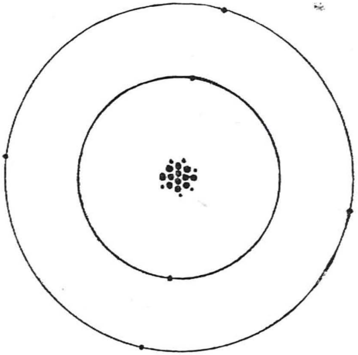


FIG. 238. DIAGRAM ILLUSTRATING THE STRUCTURE OF A CARBON ATOM, WITH SIX 'FREE' ELECTRONS REVOLVING ROUND A CENTRAL NUCLEUS (WHICH CONSISTS OF TWELVE PROTONS AND SIX ELECTRONS)

Up to the end of the nineteenth century, the *atom* was considered to be the fundamental unit of matter, but, at the beginning of the twentieth century, it was found that *atoms*, too, are compound things and not the fundamental units, since they are built up of still smaller particles. In the atom, electrified particles of *negative electricity* called *electrons* are arranged like a tiny solar system round a central *positive nucleus*, containing one or more electrified particles of *positive electricity* called *protons*, and the much smaller negative *electrons* revolve round the positive nucleus, at a very great speed.

The *electron* is the natural unit of negative electricity, since all electrons, whatever their source, are exactly alike. The weight of an electron is exceedingly small, about 1/1800th part of the weight of the hydrogen atom (which is the lightest atom).

The *proton* is the natural unit of positive electricity, and its weight is about 1,800 times greater than that of an electron; hence nearly all the weight of an atom is concentrated in its nucleus.

In chemistry lessons we have learned that the universe contains only 92 simple kinds of matter called *elements*, which combine together to form *compounds*. But these 92 elements consist of 92 different kinds of *atoms*, and each atom consists of a central *positive nucleus* surrounded by *negative electrons*. The differences between the properties of the various elements are due to differences in the number and arrangement of the electrons in the atom. Thus one arrangement gives atoms of zinc, another gives atoms of sulphur, others give atoms of silver, carbon, etc. In fact, the 92 elements can be arranged in a '*periodic table*'* according to the number of 'free' negative electrons revolving round the central positive nucleus. Hydrogen has 1 'free' electron, helium has 2, lithium† 3, beryllium† 4, boron 5, carbon 6, nitrogen 7, and so on, up to uranium† with 92 'free' electrons. This number of 'free' electrons is called the *atomic number* of the element.

We see, therefore, that everything in the universe consists of positive *protons* and negative *electrons* differently grouped, i.e. *all matter consists of different arrangements of electrical charges.*

Electrically charged objects have either an excess or else a shortage of electrons, e.g. vulcanite, when rubbed with fur, *gains* electrons which have been removed from the fur, and thus the vulcanite becomes negatively charged; glass, on the other hand, when rubbed with silk, *loses* electrons to the silk, and thus the glass becomes positively charged.

Insulators, i.e. substances which do not allow electricity to flow through them, have very stable arrangements of electrons in their atoms, so that the electrons are not readily disturbed.

Conductors, i.e. substances which allow electricity to flow through them, have electrons which are readily disturbed and separated from their atoms.

The action of a battery is simply to pile up electrons at the negative pole. Consider a single electron at the negative pole. When this pole is joined by a conductor to the positive pole, this electron rushes into the nearest atom of the conductor, drives out another electron, and takes its place. This displaced electron repeats the process with the next atom of the conductor, and so on with succeeding atoms along the conductor. This goes on at the rate of several million exchanges per second, and this movement of electrons is the *electric current*. Different atoms allow this displacement of electrons with different degrees of opposition, and this opposition to the displacement of electrons is called the *resistance* of the conductor.

A *dynamo* is simply a device for making electrons move, and, in general, electrons may be set in motion by (*a*) *heat*, (*b*) *chemical action*, and (*c*) *electrical charges*.

QUESTIONS

BIOLOGY

1. Describe the egg-cells of any named animal with which you are acquainted, the organs in which they are formed, and the ducts along which they travel. Where does fertilization take place in the example you choose ?
2. What are the normal requirements of the leaf for the manufacture of carbohydrates ? Carefully describe ONE experiment showing clearly what happens in the absence of ONE of these requirements. What control experiment would you perform ?
3. By what methods do some plants make provision for living on from one growing-season to the next ? Illustrate your answer by short accounts, with drawings, of ONE named example of each method you mention.
4. When do you call a plant a weed ? Name ONE common weed and make drawings to show its general features when fully grown. Explain how you would attempt to get rid of this weed.
5. Explain clearly what happens to the water which enters the roots of a tall leafy plant. Trace the path of the water till the surplus leaves the plant, and describe ONE experiment to illustrate your answer.
6. What do you understand by ' metamorphosis ' as applied to animals ? Illustrate your answer by *short* accounts of TWO named animals in which metamorphosis takes place.
7. To what extent do (*a*) plants depend upon animals, (*b*) animals depend upon plants, (*c*) animals depend upon animals, for their nutrition ? Give definite examples.
8. Make drawings to show the appearance and relative positions of the following organs as seen in a dissection of a frog: stomach, heart, kidneys, fat body, ovaries (or testes). What is the main function of the kidneys ?
9. Describe briefly the treatment given to any familiar area of uncultivated land in order to make it suitable for growing a

certain (named) crop. Point out the essential changes brought about by the treatment described.

10. Name ONE igneous, ONE sedimentary, and ONE metamorphic rock. In what kinds of natural situations have you seen examples of each of these rocks ? How were you able to recognize each kind of rock ?

11. From what natural sources do green plants obtain their supplies of nitrogen ? How can supplies of nitrogen in agricultural soil be increased other than by adding chemical fertilizers ?

12. Of what non-living materials does a fertile soil generally consist ? Explain how you could find out the proportions of its most important constituents.

15. Name three fertilizers, explain their general chemical nature, and state the special purposes for which they are used in agricultural practice.

16. Describe one experiment to show quite clearly that water escapes from the above-ground parts of plants and another to demonstrate the exact region where this water-loss takes place. Give practical details in connection with both these experiments.

17. What is the relation between photo-synthesis and respiration ? Describe in detail any ONE experiment you have performed or seen performed, to show that leaf-green is necessary for photo-synthesis.

18. Construct a series of large, labelled diagrams to illustrate the structure of a maize flower. Describe exactly how it is adapted for wind-pollination.

19. Describe from your own observations the way a cockroach moves. In what situations do these insects live and on what do they feed ? Give an account of their life-history.

20. How far is it possible to classify soils according to the size of their soil-particles ? How does the size of the soil-particles affect the water-retaining power of the soil ?

21. Describe a simple experiment to compare the volume of air contained within two equal blocks of dry soil. What is the importance of air in the soil ?

22. What do you understand by the term ‘ weathering ’ ? Of what importance is this in the formation of soil ? Give ONE instance where you yourself have seen the effects of weathering.

23. For what general reason do farmers not grow the same crop on the same piece of ground year after year ? What is the effect on the chemical nature of the soil of growing leguminous plants ?

24. Give an account of different ways in which loss of water from the soil may be minimized in hot dry weather. Discuss the relative value of the different methods you mention.

25. Describe the circulation of nitrogen in Nature.

26. Describe a series of THREE simple experiments you could quickly carry out which would enable you to discover the main properties of any given soil.

27. How do THREE of the following animals obtain their food: house-fly, mosquito, butterfly (or moth), any one (named) beetle ? What is the nature of the food in the case of each of the animals you choose ?

28. Explain clearly, with the aid of diagrams, giving a definite example in each case, how plants are grown from cuttings, bulbs, tubers, suckers.

29. Respiration is one of the characteristics of living organisms. Explain exactly what is meant by respiration, illustrating your answer with reference to both plants and animals.

30. Draw on a large scale a longitudinal section of a named flower you have examined carefully. Label all the parts. Describe briefly the functions of the different parts.

31. Why are mosquitoes dangerous to health ? Give an account of the life-history of a named mosquito.

32. What exactly do you understand by pollination ? How is this process connected with the formation of seeds ? From what part of the flower do seeds arise ?

33. Describe the main differences in structure and function between a root and a shoot of a flowering plant. Mention by name a plant that you have studied and refer to this plant in your answer. Illustrate your answer by diagrams.

34. What are the functions of the blood ?

35. To what extent is parental care of the offspring shown during the life-history of (*a*) fish, (*b*) birds, (*c*) mammals ?

36. In what ways do different external stimuli affect the growth of roots of plants ? Describe an experiment to show the effect

of ONE of the stimuli you mention. Make a sketch of the apparatus you would use.

37. In what ways do light and gravity influence the direction of growth of stems of green plants? Describe with practical details ONE experiment in connection with light and ONE experiment in connection with gravity, to illustrate your statements.

38. What do you understand by 'vegetative reproduction'? Illustrate your answer by means of drawings to show how vegetative reproduction is brought about in ONE named plant.

39. Describe TWO experiments which show that all the carbon contained in a flowering plant comes from the carbon dioxide of the air.

40. What exactly do you understand by the term 'cell' as used in biology? Compare typical animal cells with typical plant cells. Make accurate drawings of the appearance of any cells you have yourself examined under the microscope, stating where the cells came from, and giving their approximate size.

41. Some of the carbon contained in the body of an animal which dies may some day become part of your body. Describe different ways in which this might happen and outline the various stages in each way you mention.

42. What are the advantages and disadvantages of reproduction by means of (*a*) seeds and (*b*) vegetative structures such as suckers and tubers?

43. Describe TWO experiments you have performed, or seen performed, in the laboratory to illustrate the movements of water in the soil.

44. Some organisms are adapted for living in air, others for living in water. Illustrate this with reference to named examples of one animal and one plant living in air and one animal and one plant living in water.

45. Heat is a form of energy and energy cannot be created. Trace by careful steps as far back as you can the source of the energy that enables you to maintain your body-temperature.

46. What do you know of the structure and functions of TWO of the following: (*a*) the liver, (*b*) the stomach, (*c*) the spinal cord?

47. Why is it important for growing crops that the soil should contain plenty of air? Describe methods for ensuring a proper supply of air in the soil.

48. What are the chief elements absorbed by green plants from the soil ? Make a sketch of the absorbing region and describe the general method of absorption.

49. Give an account of the way you have studied the germination of the seeds of any ONE named plant. Make labelled drawings to illustrate the different stages during germination.

50. What are the main factors in the gradual breakdown of rocks to form soil in your country ?

51. For what processes is energy required by a (named) animal ? It is known that the energy of animals comes in the first place from the sun. Explain how this comes about.

52. Describe the digestive organs of a named animal (excluding the structures present in the mouth). Explain the parts played by the various organs in the digestion of a meal.

53. Compare the properties of sand and clay. By what means could you improve (*a*) a very sandy soil, (*b*) a heavy clay soil, in order to make them more suitable for the growth of crops ?

54. Describe how oxygen is taken in and carried to the region in which it is used in (*a*) a named plant, (*b*) a named vertebrate animal.

55. As a potato tuber gradually increases in size it accumulates a quantity of food-reserve. What is the nature of this reserve ? Explain, as fully as you can, the origin of this food, and the various steps by which this accumulation is brought about.

56. Name THREE chemical elements (apart from carbon, hydrogen and oxygen) necessary for plant growth. Describe an experiment to show the necessity for any ONE of these elements.

57. Write a brief account of the changes, favourable to plants, which are produced in soils by (*a*) earth-worms, (*b*) bacteria.

58. What are the chief reasons for crop rotation ? Illustrate your answer by describing any ONE course of crop rotation.

59. What is meant by the circulation of carbon in nature ? Make a carefully labelled diagram to illustrate this process and show the parts played in it by both plants and animals.

60. Describe simple experiments to show that both air and water are essential for the germination of a seed. Sketch the apparatus you would use.

61. Explain as fully as you can the part played by the soil in the nutrition of a *green* plant. How could you grow plants without soil ?

62. How would you do an experiment to show that pollination is necessary for the formation of fruit ?

63. Give an account of the changes which take place in the flower of a named plant from pollination to the formation of the ripe fruit.

64. Describe an experiment to show that nitrates are essential for the healthy growth of plants. What is the use of nitrates to the plant ?

65. In a named animal give a full account of the *method* of breathing. Under what conditions is breathing deeper and faster than usual, and for what reasons ?

66. Write a brief account of the following: green manuring, silt, humus, liming of soil, the harmful effects of weeds on crops.

67. Plants and animals show irritability (i.e. they respond to stimuli). Illustrate this statement by reference to ONE plant example and ONE animal example, and in each case explain how the response is made.

68. Name four important differences between animals and plants. Why are animals said to be dependent on plants for their existence ?

69. What are the uses of the liver ?

70. Explain carefully what happens to food in the small intestine.

71. Explain fully why it is necessary to manure cultivated soils. What are the advantages of adding to the soil (*a*) artificial manure, (*b*) animal manure ?

72. Growth is characteristic of living things. Some non-living things, such as crystals and stalactites, grow too. Show how the growth of a living thing differs from that of something which is not alive.

CHEMISTRY

1. How would you show by experiment that hydrogen is a reducing-agent ? Sketch and label the apparatus you would use. How would you make the hydrogen for your experiment ?

2. What reasons have you for believing: (*a*) that carbon dioxide contains carbon, (*b*) that water is a compound, (*c*) that chalk contains oxygen ?

3. Outline the method of manufacturing any gas on a large scale. Mention TWO important industrial or commercial uses of the gas.

4. Mention FOUR properties of metals, and state how metals differ in these properties from non-metals. Then answer any TWO of the following: (*a*) Give reasons why it is advisable to paint iron-work. (*b*) Discuss the uses of aluminium. (*c*) Compare the effect of heating in air a 'noble' metal, e.g. platinum or gold, with that of heating in air a 'base' metal, e.g. magnesium or copper.

5. State the Law of Constant (or Definite) Proportions. The formula of calcium carbonate is $CaCO_3$. What does this mean ? What weight of quicklime is obtained when 100 grams of chalk are heated until all the carbon dioxide is driven off ?
$$(Ca = 40,\ C = 12,\ O = 16)$$

6. Draw a fire-extinguisher and explain how it works.

7. How would you prepare a few jars of chlorine ?

8. How would you tell rain-water from distilled water ?

9. 10 gm. of one specimen of black copper oxide when reduced by hydrogen yielded 8 gm. of copper. 2 gm. of another specimen of black copper oxide (made by an entirely different method) when reduced by hydrogen yielded 1·6 gm. of copper. State the law which these facts illustrate and show how they illustrate it. Draw a labelled diagram to show how the reduction named above could be carried out. Name another compound which can be used to illustrate the law, and outline TWO different ways of preparing it.

10. Describe the preparation of the following: (*a*) quicklime from chalk, (*b*) clear lime-water from quicklime, (*c*) coke from coal, (*d*) water-gas from coke. What is the composition of water-gas ? For what purpose is it used ?

11. Why is a Bunsen flame non-luminous when the burner is properly adjusted ?

12. How would you ensure that the air entering an apparatus was free from (*a*) carbon dioxide, (*b*) water vapour ?

13. Give the chemical formulae of hydrochloric acid, sulphuric acid, caustic soda, sodium carbonate (anhydrous), sodium bicarbonate, common salt.

14. Explain the terms ‘ acid ’, ‘ alkali ’, ‘ salt ’, and give TWO examples of each. State TWO methods by which a salt can be produced and give an example in each case. For what practical purposes are alkalis used ?

15. How would you measure the proportion by volume of oxygen in air ? Air also contains carbon dioxide, water-vapour, and other gases. Give the name of one of these other gases and state any use to which it is put. What is the source of the carbon dioxide in the air and what important part does it take in supporting life ?

16. You are given five white powders which you are told are quicklime, slaked lime, chalk, common salt, soda ash. State and explain the tests by which you would identify them.

17. When clear water from a well in a chalky district is boiled, a white powder is deposited; how would you prove that this powder is chalk ? Explain how the chalk was dissolved and why it was precipitated on boiling the water.

18. What are the products obtained when steam is passed over (*a*) red-hot iron, (*b*) red-hot carbon ? How would you distinguish between the gaseous product of (*a*) and (*b*) ?

19. How is oxygen prepared on the large scale ? State one industrial purpose for which it is used.

20. You are given five jars, each containing one of the following gases: air, oxygen, carbon monoxide, carbon dioxide, nitrogen. Describe exactly what you would do to find out which gas was contained in each jar. You may assume that a second jar of each gas is available.

21. Describe the preparation of a gas which is soluble in water and state its most important properties. Give the name of the gas.

22. State exactly what is meant by H; H_2; H_2O; $2H_2 + O_2 = 2H_2O$.

23. Explain (*a*) the blue flame that often appears over a charcoal fire; (*b*) why modern oil-burners produce a blue flame.

24. How would you prove that (*a*) chalk is a compound of quicklime and carbon dioxide, (*b*) carbon dioxide is a compound of carbon and oxygen ?

25. Describe the production of hydrogen by the action of a metal on steam.

26. Describe the production of oxygen from air.

27. Give a careful diagram of the longitudinal section of a burning oil-lamp. Describe what is going on in each part of the flame and explain how both the oil and the air reach the seat of combustion. Why does the lamp smoke if turned up too high ? What are the chemical products of combustion ?

28. Describe fully how you can obtain (*a*) a metallic oxide from a metal, and (*b*) a metal from a metallic oxide. You may select any metal you please.

29. Describe, in detail, what you would do to find out what happens when the following are burnt in air: a candle, alcohol (methylated spirit), kerosene (paraffin oil), petrol (gasoline). What new substances will be formed in each case ? What change takes place in the air ?

30. Give a short, scientific description of the following substances: air, brass, wood, glass, steel, sulphuric acid, alcohol, clay, sugar.

31. Make a list of eight metals, other than the four given below in this question, which you have seen, and discuss the properties which they all have in common. (Alloys may be included, but, if you include any in your list, you should state which they are.) How would you quickly distinguish between aluminium, silver, iron, platinum ?

32. What do you understand by the terms ‘ element ’, ‘ atom ’, ‘ molecule ’ ? How do these ideas help us to understand the Law of Definite Proportions ?

33. Water is found in the following cases: (*a*) on the outside of a metal vessel containing water at room temperature, when it is first held over a flame; (*b*) on the outside of a jug of very cold water when it is brought into a warm room; (*c*) inside a test-tube, at the top, when a piece of washing-soda is heated at the bottom. Describe fully where the water comes from in each case.

34. How would you obtain the first-named substance in the following mixtures free from the second substance: (*a*) nitrogen and carbon dioxide; (*b*) calcium carbonate and sodium carbonate; (*c*) oil and water; (*d*) brass filings and sawdust ? Give a brief account of what you would do in each case.

35. How can sulphuric acid be used for the preparation of (*a*) hydrogen, (*b*) carbon dioxide, (*c*) hydrogen chloride. In each case, state whether (i) the acid is concentrated or dilute, (ii) heat is required or not. Illustrate your answer by sketches showing how each gas is collected. How would you distinguish between these three gases ?

36. You are given two black powders, one of which is copper oxide and the other charcoal. What tests would you apply to tell which is which ?

37. State clearly what you understand by the terms ‘ oxidation ’ and ‘ reduction ’, using the following examples to illustrate your answer: (*a*) a method of obtaining a metal from its oxide; (*b*) the reaction between white-hot coke and steam; (*c*) the rusting of iron.

38. Explain the following: (*a*) Calcium chloride is used for drying gases. (*b*) The flame of a Davy lamp is surrounded by copper gauze. (*c*) Water cannot be used to extinguish burning petrol. (*d*) The air-supply of a Bunsen burner should be lessened before lowering the flame.

39. How would you distinguish between jars of oxygen, nitrogen, hydrogen, water-gas ? Describe the manufacture of one of these gases on the large scale, and state the purposes for which it is used.

MECHANICS AND PROPERTIES OF MATTER

1. Describe with a diagram a simple machine which would enable a man who can exert a force of 50 lb. weight to lift an iron rail weighing 200 lb. from the ground on to a lorry.

2. Explain why a wall which is strong enough to keep in the water of a narrow canal will be strong enough to keep in the water of a very broad pool if the water comes to the same height in both cases.

3. Explain why, even when pulled with a great force, a telegraph wire hangs in a curve.

4. Explain why an ordinary suction pump will not raise water more than about 30 feet.

5. Explain why a boat floats at different levels (*a*) when full of people and when empty; (*b*) in fresh water and in sea water.

6. What do you understand by the conservation of energy ? Explain where the heat-energy comes from in the following cases: (*a*) the heating of the brakes of a motor-car when they are applied; (*b*) the heating of the air when it is pumped into a tyre; (*c*) the heating of the engine of a motor-car when it is working.

7. Describe an experiment to verify the principle of the parallelogram of forces.

8. Explain why a man with the aid of a crowbar can move a stone which he cannot move unaided.

9. How would you find the density of a brick ?

10. How would you measure the area of an irregular field on a plan ?

11. How would you find the weight of a small lump of iron by means of a metre rule, a 100 gm. weight, and a knife-edge ?

12. Describe, with a diagram, a pump suitable for pumping water to a height of 50 feet.

13. A strong wooden board stands on two bricks, *A* and *B*, which are 6 feet apart. What will be the additional weight on each brick if a man stands on the board between them at a point 2 feet from *A*, his weight being 150 lb. ?

14. Describe, with a diagram, a simple pulley system in which the effort moves four times as fast as the load. What effort will be required to lift 72 lb. if the system is 60 per cent efficient ?

15. A steel beam of uniform construction whose weight is 30 tons and length 30 feet is supported at its ends on two walls. A load of 5 tons is suspended from a point on the beam 6 feet from one end. What is the total force exerted on each of the two walls ?

16. Explain what is meant by saying that the displacement of a ship is 20,000 tons.

17. Three strings are knotted together. From one string a weight of 10 lb. is suspended; a second string passes over a pulley and supports a weight of 8 lb.; the third passes over another pulley and supports a weight of 6 lb. Find graphically the inclination of the second and third strings to each other, the three strings all being in the same vertical plane.

18. A rectangular rod of wood 2 cm. by 1 cm. cross-section is weighted at one end so that it floats vertically in a liquid. Its length is divided into cm. and when floating in water it sinks to the mark 8 cm. from the bottom. Find (*a*) the volume of water displaced by the wood, (*b*) the weight of the wood.

It is now floated in another liquid and this time sinks to the mark 10 cm. from the bottom. (*c*) Is this liquid denser or less dense than water ? Find (*d*) the volume of liquid displaced, (*e*) the weight of the liquid displaced, (*f*) the specific gravity of the liquid.

To what depth would the rod sink in a salt solution of specific gravity 1·25 ?

19. A picture is hung from a nail by means of a single piece of string attached to it at two points so as to form an equilateral triangle. If the string is just strong enough to support in this way a picture weighing 10 lb., find what weight it would support directly.

20. Describe the effect on the readings of the following instruments of the changes in the conditions or construction mentioned: (*a*) Clock: increasing (i) the length, (ii) the extent of swing, of the pendulum. (*b*) Barometer: (i) increasing the width of the tube, (ii) substituting water for mercury (S.G. 13·6), (iii) inclining the barometer-tube. (*d*) Balance: moving the knife-edge supporting the weights farther from the central knife-edge.

21. Explain how a man and a boy can carry a heavy wooden beam so that the man takes twice as much of the weight as the boy.

22. Describe any form of mercury barometer. If the height of the barometer is 30 in., calculate the pressure of the air in pounds per square inch. (S.G. of mercury = 13·6; one cubic inch of water weighs 0·037 pounds.)

23. Explain why a boy finds that although he is not strong enough to lift a large stone out of the water he can raise it from the bottom of a pool of water to the surface.

24. Show how a man weighing 12 stone and a boy weighing 6 stone can balance each other on opposite sides of a see-saw.

25. Describe a specific gravity bottle.

26. When a 2-oz. weight is placed on a 12-inch ruler at a point 2 inches from one end, the ruler (with the weight on it) balances about

a point 3 inches from the same end. What is the weight of the ruler ?

27. Calculate the atmospheric pressure in Kilograms per square centimeter when the mercury barometer stands at 76 cm. (Density of mercury = 13·6 gm. per c.cm.)

28. Draw diagrams of EITHER a common balance OR a steel-yard so as to show all the essential features.

29. What is the increased pressure in lb. per sq. in. on a diver when he descends from the surface to a depth of 90 ft. in sea-water of density 64 lb. per cu. ft. ?

30. A weight of 10 lb. is suspended by two strings, each 3 ft. long, from two points at the same level and 4 ft. apart. Find graphically the tension in the strings.

31. A balloon of volume 5,000 cu. ft. is filled with hydrogen of density 0·0056 lb. per cu. ft. If the density of the air is 0·0806 lb. per cu. ft., calculate the maximum load which the balloon can carry. Show that the lifting power would be halved if the balloon were filled with helium, which has double the density of hydrogen.

32. Explain how you would find experimentally the time of vibration of a simple pendulum. State the effect of increasing (*a*) the length of the pendulum, (*b*) the weight of the bob, (*c*) the angle through which the pendulum swings.

33. Use the following examples to illustrate the advantages obtained by the use of simple machines: the crowbar; a single fixed pulley; the human fore-arm.

34. If a given object were weighed on a spring-balance on the surface of the Moon, the weight would not be the same as it is on the Earth. Explain this.

35. Explain why it is possible to screw a nut more tightly with a spanner than with the fingers. Draw a diagram and explain the principle involved.

36. Why does a steel ball-bearing sink in water ? Why is it that, sometimes, a steel razor-blade does not sink in water ?

37. Suppose that you give some gold to a goldsmith and ask him to make a solid bracelet. If the finished article weighs the same as the original gold and looks like gold, how can you tell if it contains all your gold ?

ELECTRICITY AND MAGNETISM

1. How would you arrange a single electric circuit to show the chemical, magnetic and heating effects of an electric current? Give a diagram.
2. Describe with a diagram a simple current-measuring instrument.
3. What are the various parts of an accumulator? How can you tell when an accumulator is fully charged?
4. When an electric lamp is knocked over so that there is a short-circuit in the filament it shines very brightly for a moment before going out. Explain this.
5. Give a wiring diagram showing how you would provide from 100-volt mains one light, complete with switches and fuses, in each of three rooms. If each has a 40-watt lamp, what current is supplied by the mains?
6. Describe a Leclanché cell and explain its action. For what purposes is it suitable? Why is it less harmful to short-circuit a Leclanché cell than an accumulator?
7. Explain why a bird can perch with safety on a bare 400-volt overhead wire.
8. Give a wiring diagram for the charging of a 12-volt accumulator from direct current mains at 200 volts. A 12-volt accumulator is used to light three 12-volt 24-watt bulbs in parallel. Calculate (*a*) the current which passes through each bulb, (*b*) the resistance of each bulb, stating the units in each case.
9. Describe and explain the principle of an electric bell.
10. Explain why a 40-watt gas-filled bulb gives more light than a 40-watt vacuum bulb and why its filament is smaller.
11. A room is lighted by two lamps whose resistances are 600 and 1,200 ohms respectively. Would the lamps be arranged in series or in parallel? Give reasons. If the mains are at 200 volts, what current will pass through *each* lamp if they are arranged (*a*) in series, (*b*) in parallel? Find also the number of watts used by each lamp in (*a*) and (*b*) respectively.
12. What reasons would you give for advising a friend not to connect a large electric heater to a lighting circuit? How much would it cost to keep an electric heater going for two hours if it took a current of 6 amps. when connected to a 200-volt circuit?

(1 Kilowatt hour costs 5d.) If the heater consists of two similar heating coils, show how they should be connected so that one coil can be switched off when desired. Calculate the resistance of each coil.

13. Give a diagram of an electric lighting system consisting of 50 2-volt accumulators charged by a dynamo and working an electric heater and lamps. Show ammeters connected to read (*a*) the charging current through the battery when the dynamo is running and the heater is on, (*b*) the discharge current when a lamp is on and the dynamo is stopped. The position of the necessary switches should be shown. If the heater uses 1,000 watts, what current does it take? Find also the charging current in (*a*) if the dynamo delivers 25 amps.

14. Explain, giving a diagram, how you would find the number of watts taken EITHER by an electric iron OR by a pocket-lamp bulb. A two-Kilowatt, 230-volt electric heater is run for three hours. Find (*a*) the current taken, (*b*) the cost at 1½d. per unit.

15. The 12-volt battery of a motor-car supplies current to two 36-watt headlamp bulbs arranged in parallel. What current is taken by each lamp? What current flows through the battery? What is the resistance of each lamp?

16. Explain the principle of a pocket-lamp battery.

17. A wire, forming part of an electric circuit, runs north and south. How would you find by means of a small pocket-compass (*a*) if a current was flowing, and (*b*) its direction?

18. What changes take place when an electric current is passed (*a*) between copper electrodes in copper sulphate solution, and (*b*) through a resistance wire surrounded by oil in a beaker?

19. State and explain the advantages and disadvantages of an accumulator when compared with a dry battery. State one purpose for which you would prefer a dry battery and give the reasons for your choice. State also one purpose for which you would prefer an accumulator and give the reasons for your choice.

20. Give reasons for the use of fuses in the electric wiring system of a house. Why is a fuse-wire which is used in the lighting circuit of a small house not suitable for use in a power circuit?

GENERAL

1. Why is it not advisable to choose coloured cloth by artificial light ?
2. Give TWO examples, as different as possible, to illustrate the fact that sound takes some time to travel in air.
3. Show, by a diagram only, how the image of a tree is formed in a single-lens camera. Explain (*a*) why the inside of the camera is blackened, (*b*) the need for focussing if ' close-up ' photographs are being taken.
4. Describe experiments to show the following (the number of experiments expected is given in brackets): (*a*) sound obeys the same laws of reflection as light (one), (*b*) white light is compounded of several colours (two), (*c*) soft iron has different magnetic properties from steel (one or two), (*d*) convection takes place both in gases and in liquids (two).
5. On a day when there is a wind blowing you feel distinctly cooler when you come out of the water than you felt when bathing in the water, although the air is much warmer than the water. Explain this.
6. What is the British Thermal Unit ? What is the cost of heating the water for a 20-gallon bath from 60° F. to 100° F. with gas costing 10d. per therm ? (1 gallon of water weighs 10 lb. A therm =100,000 B.Th.U.)
7. Define the thermal capacity of a body. A lump of metal weighing 500 gm. is heated to 100° C. and then dropped into 800 gm. of water at 15° C. The temperature rises to 20° C. Find the thermal capacity of the lump of metal and also the specific heat of the metal.
8. A coloured advertisement consists of letters printed, some in green and others in red, on a white background. Describe and explain the effect of lighting it with (*a*) green light, (*b*) red light, (*c*) both lights at once.
9. What type of mirror would you consider most suitable for use (*a*) as a shaving mirror, (*b*) as a back-reflecting mirror in a motor-car. Give reasons for your choice and illustrate by diagrams.
10. Why should the compass on a modern warship be more liable to error than the compass on a warship of the time of Nelson ? Describe how such errors can be reduced.

11. Suppose that you have no sense of smell or taste: describe experiments which will enable you to distinguish between water, alcohol and petrol.

12. The distance of a ship from the shore can be found by observing the difference between (*a*) the time of receiving a wireless signal from the shore and (*b*) the time of receiving (under water) the sound of a bell struck under water at the same time that the wireless signal was transmitted. Explain this. Why is this method better than the one in which the bell is struck in air ?

13. A beaker of water is placed on a letter-balance which records 52 gm. A glass stopper is hung on a spring-balance and is found to weigh 25 gm. When the spring-balance is lowered so that the stopper is surrounded by the water in the beaker its reading falls to 15 gm. Find the volume and the specific gravity of the stopper. What is the new reading of the letter-balance ?

14. Explain why a straight stick dipping into water at an angle of 30° to the vertical appears to be bent. Show in a diagram how it seems to be bent.

15. A flower which is bluish-purple in daylight appears red in an artificial light which is lacking in blue. Explain this.

16. When a cloud passes in front of the sun one feels colder immediately although the temperature of the air is unchanged. Explain this.

17. Explain why a sound travels farther along the ground with a following wind than against a head wind.

18. Explain why curtains and other soft materials are often hung in a large room to improve it as a lecture-room.

19. Explain why a pond freezes at the top, while the water at the bottom is above freezing-point.

TABLE OF THE ELEMENTS

with atomic numbers (in brackets before the name), symbols, and atomic weights.

(89)	Actinium	Ac	(226)
(13)	**Aluminium**	**Al**	26·97
(51)	**Antimony**	**Sb**	121·76
(18)	**Argon**	**A**	39·944
(33)	**Arsenic**	**As**	74·91
(56)	**Barium**	**Ba**	137·36
(4)	Beryllium	Be	9·02
(83)	**Bismuth**	**Bi**	209·0
(5)	**Boron**	**B**	10·82
(35)	**Bromine**	**Br**	79·916
(48)	Cadmium	Cd	112·41
(55)	Caesium	Cs	132·91
(20)	**Calcium**	**Ca**	40·08
(6)	**Carbon**	**C**	12·01
(58)	Cerium	Ce	140·13
(17)	**Chlorine**	**Cl**	35·457
(24)	**Chromium**	**Cr**	52·01
(27)	**Cobalt**	**Co**	58·94
(41)	Columbium[1]	Cb	92·91
(29)	**Copper**	**Cu**	63·57
(66)	Dysprosium	Dy	162·46
(68)	Erbium	Er	167·2
(63)	Europium	Eu	152·0
(9)	**Fluorine**	**F**	19·00
(64)	Gadolinium	Gd	156·9
(31)	Gallium	Ga	69·72
(32)	Germanium	Ge	72·60
(79)	**Gold**	**Au**	197·2
(72)	Hafnium	Hf	178·6
(2)	**Helium**	**He**	4·003
(67)	Holmium	Ho	164·94
(1)	**Hydrogen**	**H**	1·008
(61)	Illinium	Il	—
(49)	Indium	In	114·76
(53)	**Iodine**	**I**	126·92
(77)	Iridium	Ir	193·1
(26)	**Iron**	**Fe**	55·85
(36)	**Krypton**	**Kr**	83·7
(57)	Lanthanum	La	138·92
(82)	**Lead**	**Pb**	207·21
(3)	**Lithium**	**Li**	6·94
(71)	Lutecium	Lu	174·99
(12)	**Magnesium**	**Mg**	24·32
(25)	**Manganese**	**Mn**	54·93
(43)	Masurium	Ma	—
(80)	**Mercury**	**Hg**	200·61
(42)	Molybdenum	Mo	95·95
(60)	Neodymium	Nd	144·27
(10)	**Neon**	**Ne**	20·183
(28)	**Nickel**	**Ni**	58·69
(86)	Niton[2]	Nt	222·0
(7)	**Nitrogen**	**N**	14·008
(76)	Osmium	Os	190·2
(8)	**Oxygen**	**O**	16·000
(46)	Palladium	Pd	106·7
(15)	**Phosphorus**	**P**	30·98
(78)	Platinum	Pt	195·23
(84)	Polonium	Po	(210)
(19)	**Potassium**	**K**	39·096
(59)	Praseodymium	Pr	140·92
(91)	Protactinium	Pa	(231)
(88)	Radium	Ra	226·05
(75)	Rhenium	Re	186·31
(45)	Rhodium	Rh	102·91
(37)	Rubidium	Rb	85·48
(44)	Ruthenium	Ru	101·7
(62)	Samarium	Sa	150·43
(21)	Scandium	Sc	45·10
(34)	Selenium	Se	78·96
(14)	**Silicon**	**Si**	28·06
(47)	**Silver**	**Ag**	107·88
(11)	**Sodium**	**Na**	22·997
(38)	**Strontium**	**Sr**	87·63
(16)	**Sulphur**	**S**	32·06
(73)	Tantalum	Ta	180·88
(52)	Tellurium	Te	127·61
(65)	Terbium	Tb	159·2
(81)	Thallium	Tl	204·39
(90)	Thorium	Th	232·12
(69)	Thulium	Tm	169·4
(50)	**Tin**	**Sn**	118·70
(22)	Titanium	Ti	47·90
(74)	Tungsten	W	183·92
(92)	Uranium	U	238·07
(23)	Vanadium	V	50·95
(54)	**Xenon**	**Xe**	131·3
(70)	Ytterbium	Yb	173·04
(39)	Yttrium	Yt	88·92
(30)	**Zinc**	**Zn**	65·38
(40)	Zirconium	Zr	91·22

[1] or Niobium.

[2] or Radon.

LIST OF NEW TECHNICAL TERMS

(Referring to their first appearance)

GLOSSARY

(Simple definitions of new general words in the text which are not included in the standard vocabulary on which the Course is based.)

abattoir, place built for the killing of animals for food, 131

blast, strong current of air in smelting, 38
bristle, short stiff hair of pig and some other animals, 29
brittle, easy to break, 27

cancel out, said of e.g. balancing quantities on each side of an equation which can be struck out, 10
cartridge, charge of explosive for guns which is contained in cylinder of paper, metal, etc., 24
cast (of metal), formed into some shape in a mould, 39
choke, stop the breathing of, 29
code, set of letters or figures or words given special meanings for shortness or secrecy, 303
constitution, way in which something is formed, 307
counter (*adv.*), in the opposite direction, 223
crude, in the natural state, not yet purified, 51

deflect, turn (something) aside, 268
disinfectant, substance which destroys germs and cleanses of infection, 29
distribute, spread about, give out shares of, 115
dwarf, person, animal or plant much below ordinary size, 180

geology, science of the earth's crust, 59
glide, change place by smooth continuous movement, 155

hearth, floor of fireplace or furnace, 39

identity, absolute sameness, 45
ignite, take fire, 44
indestructible (*n.*) *indestructibility*, that cannot be destroyed, 2
inherit, receive (characteristics, property, etc.) from parents, 181
intermediate, occurring between two others, 67
intoxicate, make drunk, excite beyond self-control, 70

jam, fruit preserved by boiling with sugar to a sticky pulp, 64

main, principal pipe, etc., for the supply of water, gas, etc., 51
moisten, make slightly wet, make damp, 90

moral, according to accepted rules of behaviour, 182

parchment, skin of sheep prepared for writing on, kind of paper similar to this, 100
passive, acted upon, not active, 185
periodic table, table giving particulars of the revolution of heavenly bodies, 308
plastic, able to be easily moulded, 27
propellent, driving force behind bullet, etc., 24

quantitative, concerned with quantity, 2

reciprocal (see Fig. 1, page 5, and explanation of it), equivalent, corresponding, 2
recover, (here) get back to normal condition, 166
refine, purify, 58
reject, not accept, put aside without using, 142
rotary, acting like a wheel turning on its axis, 302

stable (*adj.*), not easily moved or changed or destroyed, 27
sterilize, make free from living germs, 130
still (*n.*), distilling apparatus, consisting of boiler and condensing chamber, 58
strain (mech.) (*n.* and *v.*), (exert) force that tries strength, 141
symbol, (in chemistry) a letter used in equations to represent a chemical element, 7

tend to, be moving or straining (towards the direction stated), 214
tile, flat plate of glazed baked clay, 69
timber, wood regarded as building material, 140

universe, the, all things everywhere, 219
unstable, see *stable*

velocity, rate of motion, 223 (see also footnote, page 248)
volcanic, of a *volcano*, a mountain with an opening through which steam, ashes, etc., escape from beneath the earth's crust, 24

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