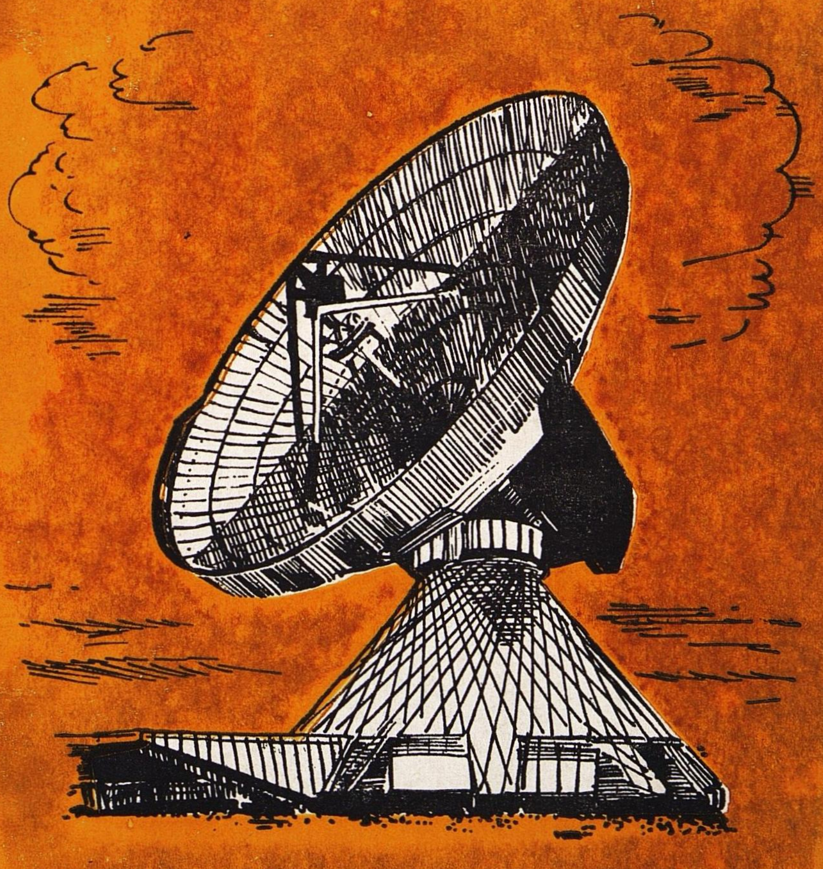
PHYSICS

HIGHER SECONDARY - FIRST YEAR

VOL. I & II



TAMILNADU TEXTBOOK SOCIETY

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PHYSICS

Vol. I & II

Higher Secondary — First Year

Untouchability is a sin
Untouchability is a crime
Untouchability is inhuman

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MADRAS

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1. BASIC PHYSICS

1. 1. UNITS AND DIMENSIONS

1. 1. 1. INTRODUCTION

Physics is a science dealing with Nature. A study of the properties of material bodies in Nature under different physical conditions forms an integral part in the study of Physics. Measurements of physical quantities such as length, mass and time are involved in the understanding of the properties of material bodies. The accurate measurements of such physical quantities form the uniqueness and the basis of Physics.

1. 1. 2. UNITS

Any physical quantity is measured accurately only in terms of a 'standard' of its own kind. For example distance is measured in Metres, msases in Kilograms and time in Seconds. These standards which are defined and accepted by all are called the units of length, mass and time respectively.

1. 1. 3. FUNDAMENTAL AND DERIVED UNITS

These three units are independent of one another and are therefore called *fundamental units*. The units of all other physical quantities are based upon them and therefore derived from them. Such units are called *derived units*.

The unit for area is the area of a square whose side is of unit length. The unit for volume is the volume of a cube whose side is of unit length. Thus the unit of area or volume is derived from the fundamental unit of length.

When we say that the speed of an object is 20 metres per second, we accept the unit of speed as $\frac{\text{metre}}{\text{second}}$ or metre sec^{-1}.

1. 1. 4. SYSTEMS OF UNITS : S. I. UNITS

As a result of the experience of scientists, an attempt has been made to simplify science by the adoption of a system of units, the *Systeme Internationale d' Unities.* This system of units, called the *SI* system, is the outcome of a resolution of the General Conference of Weights and Measures in 1948.

In the *SI* system there are six fundamental units, one each for length, mass and time, another one for electric current, yet another one for temperature and one more to measure luminous intensity, two supplimentary units for the plane angle and the solid angle, and in addition a large number of derived units. The *SI* system has certain features which make it logically superior to all other systems and more conevnient in practice. This *SI* system of units is the combination of the *MKS* units in dynamics and the Rationalised *MKS* units in electricity and magnetism with the addition of Kelvin from the field of thermal physics and of Candela from the field of optics.

1. 1. 5. THE S.I. SYSTEM OF UNITS

(i) **Length:** The Metre (m) is the length equal to 1,650,763·73 vacuum wavelengths of the orange light. $\lambda = 605{\cdot}8nm$ of the Krypton-86 discharge lamp.

(ii) **Mass:** The Kilogramme (kg) is the mass of one international prototype preserved at the International Bureau of Weights and Measures at Paris.

(iii) **Time:** The Second (s) is the duration of 9192631770 periods of the radiation corresponding to the transition between two specified energy levels of the Caesium - 133 atom.

1. 1. 6 MKS SYSTEM OF UNITS

The Metre, the Kilogramme and the Second are respectively the units used to measure length, mass and time in the *MKS* system also.

The *Metre* is the distance between the centres of two transverse lines engraved upon the polished surface of platinum-iridium bar at the temperature of melting ice (O°C). The bar is preserved at the International Bureau of Weights and Measures.

The *Kilogramme* is the mass of the cylinder made of Platinum - Iridium kept at the International Bureau of Weights and measures.

The *Second* is the time corresponding to 1/86400 th part of the mean solar day. It may be added that a solar day is the time interval between two successive noons which varies from day to day in a year. The average of all the solar days in a year is the mean solar day.

1. 1. 7. DIMENSIONS

The fundamental units of length, mass and time may be denoted by letters, L, M and T. Any other physical quantity can be expressed in terms of the fundamental units in the form $Q = KM^x L^y T^z$ where K is a measure and x, y, z are constant numbers. The physical quantity is said to have the dimensions $M^x L^y T^z$.

Examples: (i) Area $=$ Length $\times$ Length $= [L^2]$

(ii) Speed $= \dfrac{\text{Distance}}{\text{Time}} = \dfrac{L}{T} = [LT^{-1}]$

(iii) Density $= \dfrac{\text{Mass}}{\text{Volume}} = \dfrac{M}{L^3} = [ML^{-3}]$

The dimensional formulae of some important physical quantities are given in the following table.

S. No.	*Physical quantity*	*Dimensional Formula*
1.	Velocity	LT^{-1}
2.	Acceleration	LT^{-2}
3.	Momentum	MLT^{-1}
4.	Force	MLT^{-2}
5.	Impulse	MLT^{-1}
6.	Work, Energy	$ML^2 T^{-2}$
7	Power	$ML^2 T^{-3}$
8.	Pressure	$ML^{-1} T^{-2}$
9.	Angle	L^0
10.	Angular velocity	T^{-1}
11.	Angular acceleration	T^{-2}
12.	Surface tension	MT^{-2}

It is reasonable that the dimensions of all the physical quantities involved in an equation should be homogeneous. In other words the dimensions of all the terms relating to physical quantities on either side of an equation should be identical.

Examples : (i) *The displacement, the velocity and time are governed by*

$$\text{Displacement} = \text{Velocity} \times \text{Time}$$
$$s = vt$$
$$[L] = [LT^{-1}] \times [T]$$
$$= [LT^{\circ}]$$
$$= [L]$$

Hence the expression is dimensionally correct.

(ii) *The equation of motion relating to velocities* v *and* u *and acceleration* a *is*

$$v = u + at$$
$$[LT^{-1}] = [LT^{-1}] + [LT^{-2}\ T]$$
$$[LT^{-1}] = [LT^{-1}] + [LT^{-1}]$$

Hence the equation also is dimensionally correct.

1. 1. 8. USES OF DIMENSIONS

1 The method of dimensions may be used to check up the correctness of an equation.

2 The method of dimensions may be adopted to convert a physical quantity expressed in one system of units into another system of units.

Example: Let us find the number of dynes in one Newton The physical quantity involved here is the force. The dimensional formula of force is $[MLT^{-2}]$ Let x dynes = 1 Newton.

$$\text{Now, } x\,(1\text{ gm} \times 1\text{ cm} \times 1\text{s}^{-2})$$
$$= 1\,(\text{kg} \times \text{m} \times \text{s}^{-2})$$
$$= 1\,(1000\text{ gm} \times 100\text{ cm} \times 1\text{s}^{-2})$$
$$= 1 \times 1000 \times 100 \times 1\,(\text{gm} \times \text{cm} \times \text{s}^{-2})$$
$$= 10^5\,(\text{gm} \times \text{cm} \times \text{s}^{-2})$$
$$\text{or } x\,[M^1]\,[L^1]\,[T^{-2}] = 10^5\,[M^1\,L^1\,T^{-2}]$$
$$\therefore x = 10^5$$

Hence 1 Newton = 10^5 dynes.

3. The method of dimensions may be used to derive an equation relating to the physical quantities involved in a problem.

Example : Let us derive the formula for the period of a simple pendulum.

The period of a simple pendulum t may depend on (i) the mass m of the bob (ii) the length l of the pendulum and (iii) the acceleration due to gravity g at the place

The dimensional expressions of the quantities are

Quantity	*Dimensional Expression*
m	M
l	L
g	LT^{-2}

$$t \propto m^x\, l^y\, g^z$$

Writing in terms of dimensions

$$[T^1] \propto [M^x]\,[L^y]\,(\,[LT^{-2}]^{\,z})$$

$$[M^0 L^0\, T^1] \propto [M^x\, L^{y+z}\, T^{-2z}]$$

$$[M^x\, L^{y+z}\, T^{-2z}] \propto [M^0\, L^0\, T\,]$$

Equating the dimensions of each of M, L and T on both sides,

$$x = 0$$

$$y + z = 0$$

$$-2z = 1$$

From the third equation, $z = -\frac{1}{2}$. Substituting in the second equation $y = -z = +\frac{1}{2}$. Thus $x = o$, $y = +\frac{1}{2}$ and $z = -\frac{1}{2}$

$$\therefore\ t \propto m^0\, l^{\frac{1}{2}}\, g^{-\frac{1}{2}}$$

$$\propto l^{\frac{1}{2}}\, g^{-\frac{1}{2}}$$

$$\propto \sqrt{\frac{l}{g}}$$

$$t = K\sqrt{\frac{l}{g}}$$

where K is a constant of proportionality. However, the value of the constant K cannot be determined by the method of dimensions. But it can be determined by experiment and is found to be 2π.

$$\therefore\ t = 2\pi\sqrt{\frac{l}{g}}$$

Exercise 1.1

1 *Explain fundamental units and derived units.*

2. *Define the SI units for (a) length (b) mass and (c) time*

3 *What do you understand by 'Dimensions' of a physical quantity?*

4. *Assuming that the frequency n of a vibrating string depends on (i) length l (ii) tension T and (iii) linear density m, show*

that $n \propto \frac{1}{l}\sqrt{\frac{T}{m}}$

5 *Wha are the uses of 'dimensions'?*

6. *Give the dimensions of the following physical quantities: (a) volume (b) pressure (c) specific gravity (d) density.*

7 *Verify whether the following equations are dimensionally correct:*

(i) $s = ut + \frac{1}{2} at^2$

(ii) $F = ma$

(iii) $s = \frac{1}{2} gt^2$

(s- displacement, u- velocity, a - acceleration, g - acceleration due to gravity, t-time, m- mass and F- force)

1. 2. DYNAMICS

1. 2. 1 INTRODUCTION

In dynamics, we study the motion of bodies under the action of forces.

Matter may be defined as that which occupies space and affects our senses continually. A limited portion of matter is called a material body. The mass of a body is the quantity of matter contained in it.

A particle may be defined as a small body in which th distance between its neighbouring parts is negligibly small. particle is considered to have a definite mass though small. T position of a particle at any instant may be denoted by a poi A material body may be considered to be made up of a v large number of particles.

A body is said to be in motion if it changes its position continuously with respect to its surroundings.

The path of a body is a curve drawn through the successive positions occupied by the body.

The speed of a body is the rate at which it describes its path. In measuring the speed of a body we do not take into account the direction of its motion. Speed has only magnitude. A physical quantity possessing magnitude alone and not involving direction is called a *scalar* quantity.

On the other hand a physical quantity possessing both magnitude and direction is known as a *vector* quantity.

1. 2. 2. SPEED AND VELOCITY

The speed of a moving body is said to be uniform if it covers equal distances in equal intervals of time, however small these intervals may be. When the speed is uniform, it is measured by the distance traversed in unit time.

Displacement of a moving body is its changes of position measured by the straight line joining the initial and final positions of the body (Fig. 1-1.)

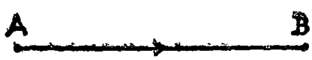


FIG - 1. 1.
Displacement

If the the initial and final positions of a moving body are represented by points A and B, the distance AB gives the displacement of the moving body.

Displacement has both magnitude and direction. The length of the line AB represents the magnitude cf the displacement and the direction of the line represents the direction of the displacement.

Velocity: *The velocity of a moving body is the rate of change of its displacement. Hence, velocity is a vector quantity.*

Uniform velocity: *The velocity of a body is said to be uniform if it undergoes equal displacements in equal intervals of time, however small these intervals may be.*

1. 2. 3. ACCELERATION

When the velocity of a body changes in magnitude or in direction, or in both, the body is said to have an acceleration. *Acceleration of a moving body is defined as the rate of change of its velocity.* Acceleration is measured by the change in velocity per unit time. If the velocity of a body changes from u to v in a time t, its acceleration is given by

$$a = \frac{v-u}{t}$$

The unit of acceleration in both the *MKS* system and the *SI* system is metre/sec^2 ($m\ s^{-2}$)

UNIFORM ACCELERATION

The acceleration of a body is said to be uniform if it moves in the same direction and equal changes in velocity take place in equal intervals of time, however small the time intervals may be.

If the velocity of the body increases, its acceleration is positive. If the velocity of the body decreases, the acceleration is negative and is called retardation.

1. 2. 4. EQUATIONS OF MOTION

Consider a body starting with an initial velocity 'u' moving under a uniform acceleration 'a'. Let its velocity change to 'v' in an interval of time t and let the distance traversed during this time be s.

From definition, the acceleration

$$a = \frac{v-u}{t} \qquad \ldots \qquad (1)$$

$$\text{or} \quad v-u = at$$

$$\therefore \quad v = u + at \qquad \ldots \qquad (2)$$

As the acceleration is uniform the velocity of the body changes uniformly.

$$\text{Average velocity of the body} = \frac{u+v}{2} \qquad (3)$$

$$\text{Distance traversed by the body} = \text{Average velocity} \times \text{time}$$

$$\text{i.e.} \quad s = \frac{(u+v)}{2}t \qquad (4)$$

Substituting $v = u + at$ from (2)

$$s = \frac{(u + u + at)}{2} \times t$$

$$= \left(\frac{2u}{2} + \frac{at}{2}\right) t$$

$$s = ut + \tfrac{1}{2} at^2 \qquad \ldots (5)$$

Now, $$s = \frac{v + u}{2}\, t \qquad \text{from (4)}$$

or $$v + u = \frac{2s}{t} \qquad \ldots (6)$$

$$v - u = at \qquad \text{from (1)}$$

$$(v + u)(v - u) = \frac{2s}{t}\, at$$

$$v^2 - u^2 = 2as \qquad \ldots (7)$$

$$v^2 = u^2 + 2as \qquad \ldots (8)$$

The equations (2), (5) and (8) *viz*,

$$v = u + at$$

$$s = ut + \tfrac{1}{2} at^2$$

$$v^2 = u^2 + 2as$$

are called equations of motion of a body moving with uniform acceleration.

1. 2. 5. ACCELERATION DUE TO GRAVITY

The most familiar example of motion with uniform acceleration is that of a body which falls towards the earth. In the absence of the resistance of air, it is found that all objects, regardless of their size or weight, are subjected to the same acceleration at a given place on earth. This acceleration exerted by the earth on the body is called the acceleration due to gravity of the earth and is denoted by the letter 'g'. The value of 'g' varies from place to place, being maximum at the poles and minimum at the equator, the maximum value being 9·83 $m\,s^{-2}$ and the minimum value being 9·78 $m\,s^{-2}$. For practical calculations, we use the approximate value of 9·8 $m\,s^{-2}$ for g.

(i) Bodies thrown vertically upwards

Consider a body thrown vertically upwards from a point on the ground with a velocity u. Here its initial velocity is in the upward direction while the acceleration due to gravity acts in the downward direction. So the body is subjected to a negative acceleration,

i. e. $a = -g$

The equations of motion become

$$v = u - gt \quad \text{... (9)}$$
$$s = ut - \tfrac{1}{2} gt^2 \quad \text{... (10)}$$
and $$v^2 = u^2 - 2gs \quad \text{... (11)}$$

(ii) Bodies thrown vertically downwards

For a body thrown vertically downwards, from a point above the ground, with a velocity u, the direction of u is downwards and the acceleration due to gravity also is in the same direction. The body is therefore subjected to the positive acceleration g. Hence the equations of motion are

$$v = u + gt \quad \text{... (12)}$$
$$s = ut + \tfrac{1}{2} gt^2 \quad \text{... (13)}$$
$$v^2 = u^2 + 2gs \quad \text{(14)}$$

(iii) Bodies dropped from a point

If a body is dropped from a point, it is known as a freely falling body. In this case $u = 0$ and the equations of motion become

$$v = gt \quad \text{... (15)}$$
$$s = \tfrac{1}{2} gt^2 \quad \text{... (16)}$$
and $$v^2 = 2gs \quad \text{... (17)}$$

Example 1

A body moving in a straight line with an initial velocity of $4\ m\ s^{-1}$ travels a distance of 140 metres in 10 seconds. Calculate (i) the acceleration (ii) the velocity at the end of 5 seconds and (iii) the velocity when it has travelled 32 metres.

(i) $u = 4\ m\ s^{-1}$, $s = 140\ m$, $t = 10\ s$ $a = ?$

We know that, $s = ut + \frac{1}{2}at^2$

or $ut + \frac{1}{2}at^2 = s$

$\frac{1}{2}at^2 = s - ut$

$$a = \frac{2}{t^2}(s - ut)$$

$$= \frac{2}{100}(140 - 4 \times 10)$$

$$= \frac{2}{100}(140 - 40)$$

$$= \frac{2}{100} \times 100$$

i. e. acceleration $= 2\ ms^{-2}$

(ii) $u = 4\ m\ s^{-1}$, $a = 2\ ms^{-2}$, $t = 5\ s$, $v = ?$

We know that $v = u + at$

$= 4 + (2 \times 5)$

$= 4 + 10$

$= 14\ m\ s^{-1}$

i. e. the velocity at the end of 5 seconds is $14\ ms^{-1}$

(iii) $u = 4\ m\ s^{-1}$, $s = 32\ m$, $a = 2\ m\ s^{-2}$, $v = ?$

$v^2 = u^2 + 2\ as$

$= (4 \times 4) + (2 \times 2 \times 32)$

$= 16 + 128$

$= 144$

$v = 12\ m\ s^{-1}$

The velocity when it has travelled $32m$ is $12\ m\ s^{-1}$

Example 2

A body is moving in a straight line with an initial velocity $20\ m\ s^{-1}$ and a uniform acceleration of $6\ m\ s^{-2}$ Find the distance travelled in the 10th second of its motion.

Distance travelled in the 10th second

= Distance travelled in 10 seconds — Distance travelled in 9 seconds.

$= (20 \times 10 + \frac{1}{2} \times 6 \times 10^2) - (20 \times 9) + (\frac{1}{2} \times 6 \times 9^2)$

$= (200 + 300) - (180 + 243)$

$= 500 - 423$

= 77 metres.

Example 3

A stone is dropped from the top of a tower of height 122·5 m. After what time will it reach the ground? What will be the velocity with which it will strike the ground. (Assume $g = 9{\cdot}8\ m\ s^{-2}$)

Since the stone is dropped, its initial velocity

$$u = 0$$
$$a = g = 9{\cdot}8\ m\ s^{-2}$$

Now $s = ut + \frac{1}{2} at^2$

$$= 0 + \frac{1}{2} gt^2$$
$$t^2 = \frac{2s}{g}$$

Substituting $t^2 = \dfrac{2 \times 122{\cdot}5}{9{\cdot}8} = 25$

$$\therefore\ t = 5 \text{ seconds.}$$

It will reach the ground after 5 seconds.

Next $v = u + at,$

Hence $v = gt$

$$= 9{\cdot}8 \times 5 = 49\ m\ s^{-1}$$

It reaches the ground with a velocity 49 $m\ s^{-1}$

Example 4

A body is thrown vertically upwards from the ground with a velocity of 39·2 $m\ s^{-1}$ Calculate (i) the maximum height reached by the body (ii) the time elapsed before it reaches the ground. (Assume $g = 9{\cdot}8\ m\ s^{-2}$)

(i) When the body reaches the maximum height, its final velocity $v = 0$

$$u = 39{\cdot}2\ m\ s^{-1}$$
$$a = -g = -9{\cdot}8\ m\ s^{-2}$$
$$s = ?$$

We know $v^2 = u^2 + 2as$

$$\therefore\ s = -\frac{u^2}{2g}$$

$$= \frac{39 \cdot 2^2}{2 \times 9 \cdot 8}$$

$$= \frac{39 \cdot 2 \times 39 \cdot 2}{2 \times 9 \cdot 8}$$

$$= 78 \cdot 4\ m$$

The maximum height reached is 78·4 metres.

(ii) When the body falls back to the ground the total

$$\text{displacement} \quad + s - s = 0$$

$$\text{or} \quad ut + \tfrac{1}{2} at^2 = 0$$

$$ut - \tfrac{1}{2} gt^2 = 0$$

$$u - \tfrac{1}{2} gt = 0 \qquad (\because t \neq 0)$$

$$\therefore \quad t = \frac{2u}{g} = \frac{2 \times 39 \cdot 2}{9 \cdot 8}$$

$$= 8 \text{ seconds.}$$

It reaches the ground after 8 seconds.

1. 2. 6. LAWS OF MOTION

A force has to be applied on a body to bring about its motion or change of motion.

A train moves because it is pulled by the engine. The train is stopped when the brakes are applied and now, a force is applied in a direction opposite to the direction of motion of the train. Similarly, to catch a moving cricket ball or to stop the ball, the fields-man has to exert a force in a direction opposite to the direction of motion of the ball. From these examples, it may be seen that a force has to be applied whenever a change of its state of rest or of motion of a body has to brought about.

Momentum: The momentum of a body is measured by the product of its mass and velocity. If m is the mass of the body and v its velocity, the momentum of the body is mv. Momentum has both magnitude and direction and it is therefore a vector quantity.

Newton's Laws of Motion

Law I: *Every body continues in its state of rest or of uniform motion in a straight line unless it is compelled by an external force acting on it.*

There are two parts in this law. The first part says that a body at rest will continue to remain at rest unless a force acts on it. The second part of the law says that a body which is in uniform motion will continue to do so unless a force acts on it. This at first sight appears contrary to our experience, because, we find that a ball set rolling on a floor, stops after some time.

A more careful observation shows that the ball stops because external forces like the force of friction and the resistance of air act on it. If the ball and the floor are both made smooth, the ball will travel a much longer distance before it comes to rest. If the forces due to friction and air resistance are completely eliminated and floor is perfectly smooth, then the ball will move continuously on the floor establishing the truth of the first law.

The first law gives us the principle of the *inertia* of matter. The inability of any object to change, by itself, its state of rest or of uniform motion in a straight line is called the *inertia* of the body. Any material substance has intertia depending on its mass.

Example: Suppose a person jumps out of a moving train on to the platform. His feet are brought to rest suddenly, whereas his head and the upper part of his body are trying to move in the direction of motion of the train. Hence normally, he falls forward. If he wants to avoid falling down, he will have to move his feet also in the direction of motion of the train. Thus he can avoid the accident of falling down on the platform, if he runs forward upto some distance and then stops.

The first law also provides the definition of force. *Force is action exercised on a body so that it changes or tends to change the state of rest or of uniform motion of the body.*

Law II : *The rate of change of momentum of a body is directly proportinal to the impressed force acting on it and takes place in the direction of the force.*

This law gives a method of measuring force. From the first law, we know that when a force acts on a body, it produces a change in its velocity and therefore a change in momentum. Suppose a constant force F acts on a body of mass m moving with a velocity u and changes its velocity to v in a time t.

Initial momentum of the body $= mu$

Final momentum of the body $= mv$

Change of momentum $= mv - mu$

$$\therefore \text{Rate of change of momentum} = \frac{mv - mu}{t} = \frac{m(v-u)}{t}$$

$$\text{But the acceleration } a = \frac{v-u}{t}$$

$\therefore$ Rate of change of momentum $= ma$.

The second law states that the rate of change of momentum is ditectly proportional to the force F i.e. $ma \propto F$ or $F \propto ma$.

$$\therefore F = k\, ma.$$

where k is a constant of proportionality. Its value depends on the choice of the unit in which the force is to be measured.

Let us define the unit of force as that force which acting on unit mass produces unit acceleration. When the unit of force is so chosen,

$$F = 1 \quad \text{when } m = 1 \text{ and } a = 1.$$

$$F = k\, ma \text{ becomes } 1 = k \times 1 \times 1 \text{ or } k = 1$$

$$F = ma. \qquad \dots \quad (18)$$

Hence the force acting on a body is measured by the product of the mass of the body and the acceleration produced by the force on the body.

Unit of force: The unit of force in the M. K. S. system is the *newton*. It is the force acting on a body of mass 1 kilogram producing an acceleration of 1 $m\ s^{-2}$, In the *SI* sys em also the unit of force is the newton. It is represented as N ($N = kg\ m\ s^{-2}$). Newton is the absolute unit of force. The unit of force depending on the gravitational attraction of the earth is called the gravitational unit.

In the MKS system, the gravtiational unit of force is the kilo gram weight. It is the forcc of attraction exerted by the earth on a body of mass 1 kilogram.

$\therefore$ 1 kilogram weight $= g$ newtons

where g is the acceleration due to gravity. If $g = 9{\cdot}8\ m\ s^{-2}$ at a given place, 1 kilogram weight $= 9{\cdot}8\ N$. The gravitational unit of force varies from place to place since it depends on the acceleration due to gravity. However, the value of the acceleration due to gravity at a place is always a constant.

Mass and weight: The mass of a body is the quantity of matter contained in it. But its weight is the force of attraction exerted by the earth on it. If m is the mass of a body and g the acceleration due to gravity at the given place, the weight of the body $=$ mg newtons. The mass of a body is a scalar quantity, whereas the weight is a vector quantity.

Law III. *Every action has an equal and opposite reaction.*

This law implies the presence of two bodies. Action is the force exerted by one body on another. Reaction is the force exerted. at the same moment, by the second body on the first. These forces, according to this law are equal in magnitude but opposite in direction.

The following illustrates the third law of motion. When you strike against a stationary wall, you are exerting a force on the wall. This is action. The wall exerts an equal force on your hand at the same instant; This is reaction. These two forces are equal, but their directions are opposite.

The third law gives us the *law of conservation of momentum* The law of conservation of momentum states *that the total momentum of a closed system of bodies in a given direction remains unaltered as a result of collisions among themselves.*

We can prove this law in the case of a direct impact between two bodies.

Suppose a body A of mass m_1 moving with a velocity u_1 collides with another body B of mass m_2 moving with a velocity u_2. After collision let their velocities be changed to v_1 and v_2 respectively. Let us for the sake of simplicity take the direction of u_1, u_2, v_1 and v_2 to be the same. During the collision, let F_1 be the force exerted by A on B. Let F_2 be force exerted by B on A in the opposite direction. Let t be the time of contact during collision. Then by the third law of motion,

$$F_1 = -F_2$$

Now, F_1 acting on m_2 for a time t changes the velocity of m_2 from u_2 to v_2.

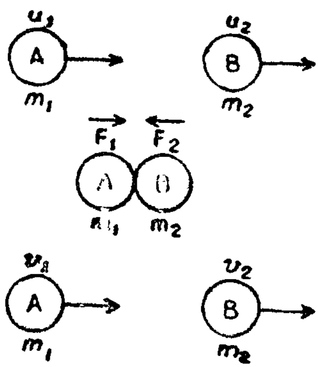


Fig. 1-2
Collision between two bodies

$\therefore$ F_1 = mass $\times$ acceleration of the second body

$$= m_2 \left(\frac{v_2 - u_2}{t}\right)$$

Similarly F_2 acting on m_1 for the same time t changes its velocity of m_1 from u_1 to v_1.

F_2 = mass $\times$ acceleration of the first body

$$= m_1 \left(\frac{v_1 - u_1}{t}\right)$$

Since $F_1 = - F_2$

$$m_2 \left(\frac{v_2 - u_2}{t}\right) = - m_1 \left(\frac{v_1 - u_1}{t}\right)$$

$$\therefore m_2 (v_2 - u_2) = - m_1 (v_1 - u_1)$$

i.e. $m_2 v_2 - m_2 u_2 = - m_1 v_1 + m_1 u_1$

or $m_2 v_2 + m_1 v_1 = m_2 u_2 + m_1 u_1$ ... (19)

The total momentum after impact

= the total momentum before impact.

The collision has not produced any change in total momentum of the closed system of two bodies. This is the law of conservation of momentum. The law can be applied to any number of impacts among any number of bodies in a closed system.

Example 1: A train of mass 240 tonnes starts from rest and attains a speed of 45 *kmph* in one minute. Calculate the net force exerted by the engine on the train assuming it to be uniform.

$$\text{Speed} = 45\ km/h = \frac{45 \times 1000}{60 \times 60}\ m\,s^{-1} = 12{\cdot}5\ m\,s^{-1}$$

Initial velocity $u = o$. After 60 seconds the final velocity

$$v = 12{\cdot}5\ m\,s^{-1}$$

$$a = \frac{v - u}{t} = \frac{12{\cdot}5}{60}\ m\,s^{-1}$$

Mass m = 240 tonnes

= 240 × 1000 kg

The net uniform force exerted by the train.

$$F = ma$$

$$= \frac{240 \times 1000 \times 12{\cdot}5}{60}$$

= 50,000 newtons

Example 2: A body A of mass 10 kg moving with a velocity of 5 ms^{-1} impinges directly on another body B of mass 20 kg at rest. If after impact A comes to rest, calculate the velocity of B.

$m_1 = 10\ kg \qquad u_1 = 5\ m\ s^{-1} \qquad v_1 = 0$

$m_2 = 20\ kg \qquad u_2 = 0 \qquad v_2 = ?$

From the law of conservation of momentum,

total momentum before impact

= total momentum after impact

$$m_1 u_1 + m_2 u_2 = m_1 v_1 + m_2 v_2$$

$$\therefore m_1 u_1 = m_2 v_2$$

$$v_2 = \frac{m_1}{m_2} u_1 = \frac{10}{20} \times 5 = 2{\cdot}5\ m\ s^{-1}$$

After collision, *B* moves with a velocity of 2·5 metres per second.

1. 2. 7. WORK, POWER AND ENERGY

If a force acts on a body and the point of application of the force moves, then work is said to be done by the force.

If the force is F and it moves the body through a distance s in its direction, then the work done by the force is given by

$$W = F \times s \qquad \dots (20)$$

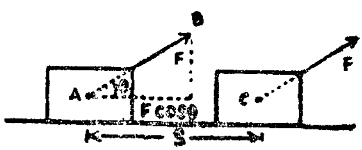


Fig. 1-3
Work done by a force

Suppose a force F acts along the direction AB and the body moves along AC. Then the work done in this case is given by the product of the component of the force in the direction of motion and the distances through which the body is moved. If θ is the angle between the direction of the force and direction of motion of the body, work done, $W = F \cos \theta \times s$. *Hence work done is equal to the product of the component of the force in the direction of motion and the distance travelled.*

If $\theta = 0°$; $W = Fs$

If $\theta = 90°$ i.e. F is in a direction perpendicular to the direction of motion, $\cos 90 = 0$ and $W = 0$.

Unit of work : *In the MKS system the unit of work is joule. It is the amount of work done when the point of application of a force of one newton is displaced through a distance of one metre in the direction of the force.*

In the *SI* units, it is represented as J

$$J = N\, m$$

Power : *The rate of doing work is called power. It is measured by the amount of work done in unit time.*

$$\text{Power} = \frac{W}{t} = \frac{F \times s}{t}$$

Unit of Power: *In the MKS system the unit of power is called watt. A watt is the power when work is done at the rate of one joule per second.*

In the *SI* units, it is represented as W.

$$(W = J\, s^{-1})$$

A larger unit of power is called the *kilowatt*. A kilowatt is equal to 1000 watts. Power consumed at the rate of 1 kilowatt for 1 hour is called 1 kilowatt-hour and it is called the Unit in electrical energy consumption.

Energy: *The energy of a body is its capacity to do work and is measured by the amount of the work that it can perform.*

Since the energy of a body is measured by the amount of work it can perform the unit of energy is the same as that of work.

There are many forms of energy such as mechanical energy, heat energy, light energy, electrical energy, chemical energy and atomic or nuclear energy.

There are two forms of mechanical energy viz. (i) **potential energy** and (ii) **kinetic energy.**

The potential energy of a body is the energy it possesses by virtue of its position or state of strain.

Examples : Water stored up in a reservoir; a wound up spring; a stretched rubber cord; compressed air.

For a body of mass m remaining at rest at a height h above the ground, the potential energy $= mgh$.

Let a body of mass m be initially at rest at a height h above the ground. The work done in raising the body from the ground to that height is stored as potential energy in the body. When the body falls, the same amont of energy can be obtained from it. To lift the body vertically up, a force mg equal to the weight of the body should be applied upwards When this force moves the body vertically through a distance h,

$$\text{Work done} = \text{Force} \times \text{distance} = mgh$$

$$\therefore \text{Potential energy} = mgh$$

If m is in Kg, h is in metres, and g is in $m\ s^{-2}$,

the potential energy $= mgh$ joules

The kinetic energy of a body is the energy possessed by it on account of its motion.

Consider a body of mass m moving with a velocity v Suppose a force F acts on the body in the oppsite direction and brings it to rest in a distance s. Now, the work done by the body against the opposing force will be equal to the initial kinetic energy of the body.

$$\text{Kinetic energy} = F \times s$$

$$= ma \times s$$

Since the final velocity is zero, from the equation of motion

$$(v^2 = u^2 + 2as)$$

we get $\quad 0 = v^2 + 2\ as$

$$\therefore \quad a = -\frac{v^2}{2s}$$

The opposing force $= - ma = m \dfrac{v^2}{2s}$

$\therefore$ K. E. $= \left(m \dfrac{v^2}{2s}\right) \times s = \frac{1}{2} mv^2$

If m is in kilograms and v is in ms^{-1}, the kinetic energy of the moving body is in joules.

Law of Conservation of Energy

Energy can neither be created nor destroyed, but can be transformed from one form to another.

This law holds good for all forms of energy. Consider the case of a freely falling body. When it falls, it experiences a loss of potential energy but there is a gain of kinetic energy. At any position during its fall, the loss of potential energy is equal to the gain of kinetic energy. Similarly when a current is passed through a coil of wire, the amount of heat energy produced in it is equal to the amount of electrical energy consumed. Thus during any transformation of energy, there is no loss or gain of energy but there is only a change of energy from one form to another.

Verification of the law of conservation of energy for a freely falling body

(i) Let a body of mass m be at rest at a point A at a height h above the ground.

Potential energy at $A = mgh$

Kinenic energy at $A = 0$

$\therefore$ At A, $P.E. + K.E. = mgh + 0 = mgh$ joules.

(ii) Let the body strike the ground with a velocity v. Then

$$v^2 = 2gh$$

$$\therefore \text{ Kinetic energy at } B = \tfrac{1}{2} mv^2$$
$$= \tfrac{1}{2}\, m\, 2gh$$
$$= mgh \text{ joules.}$$

Potential energy at $B = mgh = mg \times 0 = 0$.

At B, $P.E. + K.E. = 0 + mgh = mgh$ joules

(iii) Consider any intermediate point C at a depth x below A. C is at a height $(h-x)$ above the ground.

$\therefore$ Potential energy at $C = mg\ (h-x)$

If after falling through a distance x the velocity is v, then $v^2 = 2gx$

$\therefore$ The kinetic energy at $C = \frac{1}{2}\ mv^2$

$= \frac{1}{2}\ m.\ 2\ gx = mgx$

At C, $P.E. + K.E. = mg(h-x) + mgx$

$= mgh$ joules.

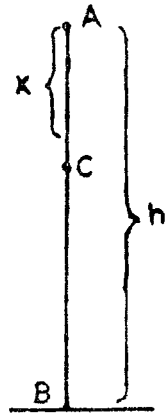


Fig. 1-4
Freely falling body

Hence the total energy of the body remains a constant throughout the motion.

Example 1: Calculate the amount of work done in drawing a bucket of water weighing 5 kg from a well of depth 10 m.

Work done in lifting 5 kg of water through a height of 10 m

$= mgh$

$= 5 \times 9{\cdot}8 \times 10$

$= 490$ joules.

Example 2 : A man weighing 60 kg runs up a flight of stairs 3 m high in 4 seconds. Calculate the power developed by him.

Work done $= mgh$

$= 60 \times 9.8 \times 3$ joules

Power develoqnd $= \dfrac{W}{t}$

$= \dfrac{60 \times 9{\cdot}8 \times 3}{4}$

$= 441$ watts.

Exercise 1. 2.

1. *Explain what is meant by (a) a particle and (b) a body*
2. *Explain what is meant by (a) displacement and (b) velocity of a particle.*
3. *Define 'uniform velocity; What is meant by 'average velocity'?*
4. *Define uniform acceleration.*
5. *Derive the equations of motion for a particle subjected to a uniform accleration.*
6. *What is 'acceleration due to gravity'. Derive the equations of motion for a body subjected to acceleration due to gravity.*
7. *State Newton's laws of motion.*
8. *Define Force. How does the second law of motion help us to measure force ?*
9. *Explain 'Inertia' of a body with examples.*
10. *Explain the 'law of conservation of momentum'.*
11. *Establish the law of conservation of momentum in the case of a direct collision betwe n two bodies.*
12. *What is the unit of force?*
13. *Define the terms: work, power and energy. What are the units in which they are measured?*
14. *State the law of conservation of energy and explain.*
15. *Define kinetic energy and potential energy. Derive the expressions for potential energy and kinetic energy.*
16. *Establish the law of conservation of momentum in the case of a freely falling body.*

1. 3. STATICS

1. 3. 1. INTRODUCTION

Statics is the branch of mechanics dealing with the equilibrium of a body acted upon by a system of forces.

1. 3. 2. SIMPLE MACHINES

A simple machine is a device by which a force applied at one point in one direction is made available at some other point, altered in magnitude or dirction or both. The lever, the pulley, the wheel and axle and the inclined plane are some examples of simple machines.

The force applied to the machine is called the *power* or *effort*. The force which is overcome is called the *weight* or *load*.

MECHANICAL ADVANTAGE

In a simple machine, when the power P just balances the load W, the ratio of the load to the power is called the *mechanical advantage* of the machine. The mechanical advantage of the simple machine = $\frac{W}{P}$.

VELOCITY RATIO

Let the point of application of the power applied to the machine be moved through a distance 'x' in the direction of the power. As a result of this if the point of application of the load moves through a distance 'y' along the line of action of the load, the ratio x/y is called the "*velocity ratio*" of the machine.

Velocity ratio of the simple machine is defined as the ratio of the displacement of the point of application of power to the displacement of the point of application of the load in the same time.

EFFICIENCY OF A SIMPLE MACHINE

Suppose a power P applied to a machine moves its point of application through a distance x in its direction. As a result of it, let the point of application of the load move through a distance y in the same interval of time.

Work done by the power $= P \times x$. This is the work supplied to the machine.

Work done by the load $= W \times y$. This is the useful work turned out by the machine.

Efficiency of a simple machine is the ratio of the useful work done by the machine to the total work supplied to the machine.

$$\therefore \text{Efficiency} = \frac{\text{Useful work done by the machine}}{\text{Total work supplied to the machine}}$$

$$= \frac{W \times y}{P \times x}$$

$$= \frac{W}{P} \div \frac{x}{y}$$

$$= \frac{\text{Mechanical advantage}}{\text{Velocity ratio}}$$

Efficiency is usually expressed as a percentage.

In an ideal machine no friction is present. Therefore all the work supplied to the machine is spent in overcoming the load and the efficiency is equal to unity. In practice, a certain amount of friction is always present in the machine. So the work turned out by the machine is less than the work supplied to the machine and the efficiency is less than unity.

As the efficiency is always less than unity, the machanical advantage is always less than the velocity ratio.

1. 3. 3. PULLEY

A pulley consists of a small wheel or disc made up wood or metal. It can rotate about an axis though its centre and perpendicular to its plane. It has a groove along its circumference in which a string can be passed. It is held in a frame called the block. If the block is attached to a fixed support it is called a fixed pulley. If the block is free, it is called a movable pulley.

SINGLE FIXED PULLEY

Here the block of the pulley is fixed to the rigid support, for example a roof or a beam. This is used to raise loads. The load is attached to one end of the string passing round the groove of the pulley and the power is applied to the other end of the string. (Fig. 1-5). A small fixed pulley has a mechanical advantage unity if friction could be neglected. It is chiefly used to facilitate application of the effort in a convenient direction.

SINGLE MOVABLE PULLEY

Here one end of the string is passed around a movable pulley and is attached to a fixed point. The effort P is applied to the other end passing round the fixed pulley B (Fig. 1-6).

The load W is attached to the block of the movable pulley A. If we assume that there is no friction, the tension of the string is the same at all points. When the segments are vertical the total upward force in the movable pulley $= P + P = 2P$. If we neglect the weight of the movale pulley, the total downward force $= W$. If the pulley is in equilibrium,

$$W = 2\,P.$$

$$\therefore \text{ Mechanical advantage } \frac{W}{P} = 2$$

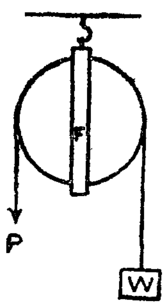


Fig. 1-5
Single fixed pulley

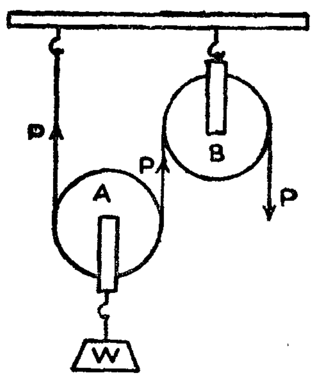


Fig. 1-6
Single movable pulley

SYSTEMS OF PULLEYS

With a single movable pulley, the mechanical advantage that can be obtained is only 2. To obtain a larger mechanical advantage, a number of movable pulleys may be used. The pulleys can be arranged in three different systems.

FIRST SYSTEM OF PULLEYS

Here a separate string passes through each movable pulley. One end of each string is attached to a fixed support, the other end being attached to the block of the next movable pulley. The load is suspended from the block of the lowermost movable pulley. The power is applied at the free end of the string passing over the

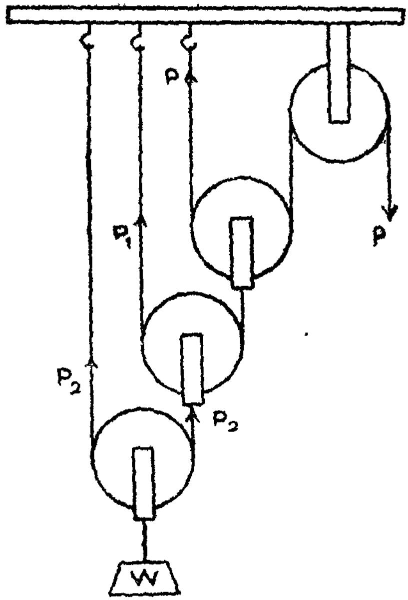


Fig. 1·7
First system of pulleys

top-most movable pulley. For convenience of direction, the power is applied after passing the string over a fixed pulley.

Let the tension in the first string be equal to P the power applied. Let P_1 and P_2 be the tension in the other two strings. Considering the equlibrium of the lower-most pulley $W = 2P_2$ neglecting the weight of the pulley.

Considering the equilibrium of the next movable pulley.

$$P_2 = 2P_1; \text{ Similarly } P_1 = 2P;$$

$$\therefore \quad W = 2P_2 = 2\times2\,P_1 = 2\times2\times2P = 8P$$

$\therefore$ Mechanical advantage $\frac{W}{P} = 8 = 2^3$

When there are three movable pulleys, mechanical advantage is = 2^3.

If there are n movable pulleys, the mechanical advantage

$$\frac{W}{P} = 2n. \qquad \dots \qquad (22)$$

Here the weights of the pulleys have been neglected.

SECOND SYSTEM OF PULLEYS

Here there are two separate blocks, each containing usually the same number of pulleys, the upper one being fixed and the lower movable. One end of the string is attached to the bottom of the upper block, and the same string passes round all pulleys in succession. The load is attched to the lower block and the power is applied to the free end of the string.

Let P be the power applied to the free end of the string which just balances a load W. Since the same string is passing round all the pulleys. the tension is same at all points of the string and is equal to P (The weight of the lower block is neglected).

Considering the equilibrium of the load

$$W = 6P$$

$$\therefore \frac{W}{P} = 6$$

$\therefore$ Mechanical advantage $\dfrac{W}{P} = 6.$

This is equal to the total number of segments of the string supporting the load i.e. total number of pulleys in both the blocks.

$$\therefore \frac{W}{P} = n \qquad \dots \qquad (23)$$

where n is the total number of pulleys in both the blocks.

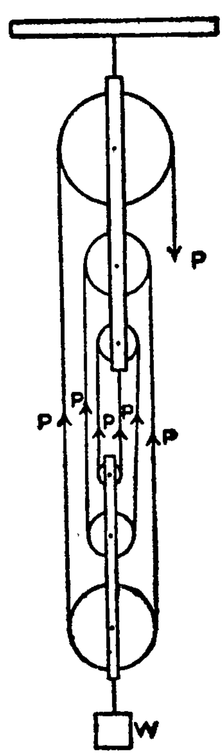


Fig. 1-8
Second system of pulleys

THIRD SYSTEM OF PULLEYS

Here a separate string passes over each pulley. One end of each string is attached to the load. The power is applied to the free end of the last string.

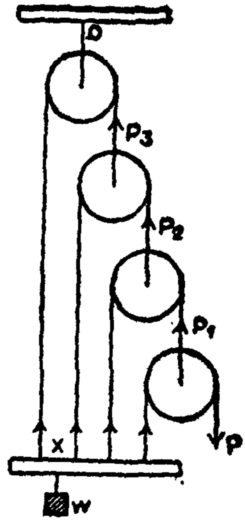


Fig. 1-9
Third system of pulleys

Let us suppose the power P applied to the last string just balances the load W. Then P is the tension in the last string. Let P_1, P_2 and P_3 be the tensions in the other strings:

Considering the equilibrium of the lowest pulley, $P_1 = 2P$

Similarly $P_2 = 2P_1 = 2 \times 2P = 4P$
$P_3 = 2P_2 = 2 \times 2 \times 2P = 8P$

Considering the equilibrium of the weight W,

$$\begin{aligned} W &= P + P_1 + P_2 + P_3 \\ &= P + 2P + 4P + 8P \\ &= 15\,P \\ &= (2^4 - 1)\,P \end{aligned}$$

$\therefore \quad W/P = (2^4 - 1)$

$\therefore$ The mechanical advantage $= (2^4 - 1)$

If the total number of pulleys is n, the mechanical advantage

$$\frac{W}{P} = 2^n - 1 \qquad \ldots \quad (24)$$

Here also the weights of the pulleys are neglected.

1. 3. 4. REQUISITES OF A GOOD BALANCE

A good balance is one which can measure a mass correctly accurately aud quickly. In order to achieve this, a balance must have the following requisities: *(a) truth (b) sensitiveness* and *(c) stability*.

(a) TRUTH

A balance is said to be true, if the beam remains horizontal when both the pans are empty or equally loaded. It can be shown that for a balance to be true,

(i) the arms of the balance should be of equal length and

(ii) the scale pans should be of equal mass.

(b) SENSITIVENESS

A balance is said to be sensitive if the beam deflects through a large angle even for a small difference between the weights placed in the two scale pans. The sensitiveness of a balance can be shown to be high if,

(i) the arms are long
(ii) the beam is light and
(iii) the centre of gravity of the beam is close to the central knife edge.

(c) STABILITY

A balance is said to be stable if the beam returns quickly to the equilibrium position after being disturbed. The stability of a balance will be high, if

(i) the beam is heavy and
(ii) the centre of gravity of the beam is far away from the central knife edge.

Thus it is seen that the conditions for sensitiveness are opposed to the conditions for stability. In practice this is not a serious handicap because quick weighings need not be done accurately and accurate weighings can take some time and there are separate balances designed for the different purposes.

1. 3. 5. GAUSS METHOD OF WEIGHING WITH A FAULTY BALANCE

Suppose a balance has unequal arms a and b. Let a body weigh W_1 and W_2 when placed successively in the two pans. Let W be the true weight.

Taking moments about the central knife edge,

$$W \times a = W_1 \times b \quad ...(25)$$
$$W \times b = W_2 \times a \quad ...(26)$$

(1) × (2) gives

$$W^2 a b = W_1 W_2 a b$$
$$W^2 = W_1 W_2$$
$$\therefore W = \sqrt{W_1 W_2} \quad ...(27)$$

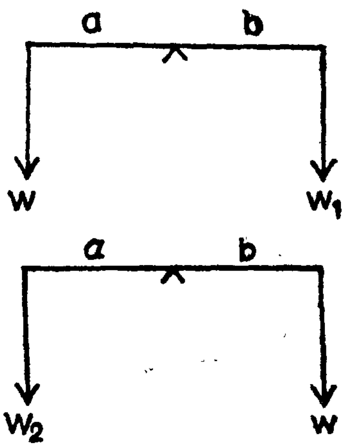


Fig. 1-10
Gauss method of weighing

Also $\frac{(1)}{(2)}$ gives $\frac{a}{b} = \frac{W_1}{W_2} \times \frac{b}{a}$

$$\frac{a^2}{b^2} = \frac{W_1}{W_2}$$

$$\therefore \quad \frac{a}{b} = \sqrt{\frac{W_1}{W_2}} \qquad \ldots (28)$$

1. 3. 6. FRICTION CLUTCH

A clutch is a mechanism by which the rotatory motion of oue shaft can be transmitted to another shaft the two shafts being mounted co-axially. In one type of clutch, called gradual engagement clutch, one of the shafts rotates rapidly while the other is either stationary or moving with a low speed. When the engagement or the clutch proceeds. rapidly moving shaft is retarded while the slowly moving shaft is accelerated. The process goes on until the two shafts rotate as one, with the same speed. The clutch is now said to be fully engaged. The clutch used in a motor car between the engine and the gear box is based on the action of the frictional force which is called into play between two rotating bodies when they are pressed together. This type of clutch is known as friction clutch.

To understand the action of the friction clutch, consider two shafts, C and D supported on bearings A and B, so that they are free to rotate about a common axis PQ. E and F are two circular discs which face each other and which are keyed to the ends of the shafts. Suppose the shaft C with the disc E is rotating rapidly while the shaft D with the disc F is stationary. The two shafts are pressed together endways. When the faces of the two discs come into contact, the force of friction between them tends to retard the speed of E. As the forces with which the dics press on each other gradually increase, the frictional force between them also increases gradually. At a certain stage it is sufficiently great to overcome the resistance between the discs. There-

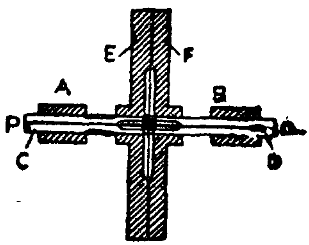


Fig. 1-11
Friction clutch

fore the disc F begins to rotate with its speed gradually increasing. This goes on until the two discs move with the same speed. At this stage, there is no slip between the discs, and the clutch is fully engaged

In the motor car clutches, the discs are kept pressed against each other by means of a spring. The elasticity of the spring always keeps the clutch in engagement. To disengage the clutch, one of the discs is pulled back against the pressure of the spring.

Exercise 1.3

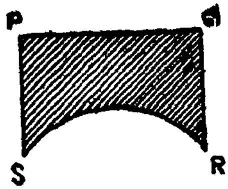
1. *What are simple machines? What is called 'power' or 'effort' and what is called 'load' or 'weight'?*
5. *Define the following: (i) mechanical advantage (ii) velocity ratio and (iii) efficiency of a simple machine. Obtain the relation between them.*
3. *What is a pulley? What are the mechanical advantages of (a) a single fixed pulley and (b) a single movable pulley?*
4. *Explain any one of the three systems of pulleys and obtain an expression for its mechanical advantage.*
5. *What are the requisites of a good balance? When do you say a balance possesses these requisites?*
6. *How can you find the correct weight of an object using a false balance?*
7. *What is meant by a 'clutch'? Describe the action of a friction clutch.*

1. 4 FLUIDS

1.4.1. SURFACE TENSION

The surface of a liquid behaves like a stretched elastic membrane, so that it has always a natural tendency to contract, thereby reducing the surface area to a minimum. A number of experiments can be performed to demonstrate such behaviour in the case of the surface of any liquid.

(i) Consider a copper wire $SPQR$ bent twice at right angles as shown in Fig 1-12 Let a thread be tied loosely to the ends S and R. If the frame is dipped in soap solution and taken out,



a soap film is formed between the frame and the thread. The thread is drawn inwards in the form of an arc of a circle so that the area of the film is a minimum.

Fig. 1-12
Thread SR pulled inwards

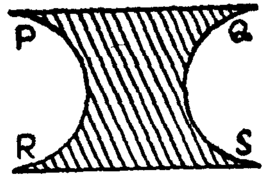


Fig. 1-13
Threads PR, QS pulled inwards

(ii) Consider two glass rods *PQ*, *RS* suspended parallel to each other by means of two vertical silk threads. If the whole arrangement is dipped in soap solution and taken out, a soap film is enclosed between the glass rods, and the threads are drawn inwards in the form of an arc of a circle thereby reducing the area of the film to a minimum (Fig. 1-13).

(iii) Consider a ring of copper wire dipped in soap solution and taken out. There is a film, enclosed in the ring. If a small loop of silk thread moistened with soap solution is placed on the film and the film inside the loop pricked, the string will immediately assume the form of a circle. For a given perimeter, the circle has maximum area and therefore the remaining surface of the film has minimum area.

These experiments clearly illustrate that the surface of a liquid behaves as if it is in a state of tension with a tendency to contract its surface area to a minimum value. This tendency is due to the **surface tension** of the liquid. *It is defined as the force acting on unit length of a line drawn on the surface of the liquid, the force acting normal to the line and parallel to the surface of the liquid.* It is expressed in newton per metre *(N/m)*

FREE SURFACE ENERGY

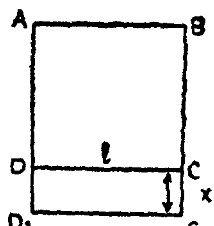


Fig. 1-14
Work done in stretching a film

Consider a liquid film held in a frame *ABCD* (Fig.1-14) in which *CD* can slide freely. Let the side *CD* move through a small distance x. The increase in area of the film $= 2\,l.x$, taking into account both surfaces of the film.

Work done against the surface tension T in increasing the area of the film by $2lx = T \times 2\,l \times x$, Besides this, a certain amount of heat energy is absorbed by the extended film from the atmosphere to make up for the cooling produced when it is extended. If H is the heat energy absorbed per unit increase of area of the surface, expressed in joules, the energy absorbed for the increase of area by $2lx = H \times 2lx$. Let the total energy of the newly formed surface be $E(2lx)$, where E is the total energy per unit area.

$\therefore$ From the law of conservation of energy

$$E \times 2lx = (T \times 2lx) + (H \times 2lx)$$

$$\therefore \quad E = T + H$$

T represents the mechanical part of the surface energy of the film. It is referred to as the *free surface energy.* Hence surface tension of a liquid is numerically equal to the free surface energy per unit area.

1.4.2 ANGLE OF CONTACT

When a liquid meets a solid, its surface near its line of contact with the solid is generally curved. The angle between the tangent to the liquid surface at the point of contact and the solid surface inside the liquid is called the angle of contact of the liquid with the solid.

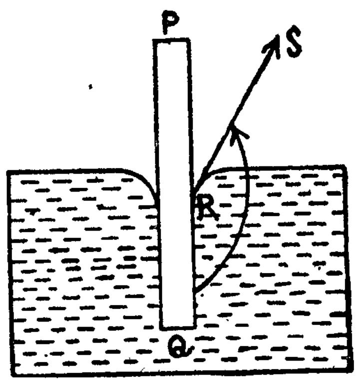


Fig. 1-15
Angle of contact of mercury

Let PQ (Fig 1-15) be a glass plate dipped vertically in a trough containing mercury. It is found that the mercury surface near the glass plate goes down to meet it at R.

The liquid surface at R is called the meniscus of the liquid. RS is the tangent drawn to the liquid surface at R. The angle $Q\hat{R}S$, which is the angle between the tangent to the liquid surface at R and the portion of the glass plate QR inside the liquid is called the angle of contact of mercury with glass. The angle of contact of mercury with glass is obtuse, whereas the angle of contact of any other liquid with glass is acute.

DETERMINATION OF THE ANGLE OF CONTACT OF MERCURY WITH GLASS

AB is a glass plate immersed vertically in a vessel containing mercury (Fig. 1-16). The mercury bends and meets the plate at C. C is a point at which the liquid, glass and air meet. CD is the tangent to the liquid surface at C. $\angle DCB$ is the angle of contact of mercury with glass.

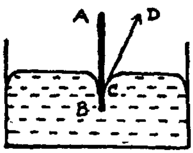


Fig. 1-16
Glass plate vertically held

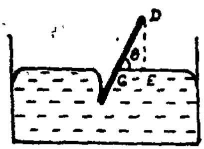


Fig. 1-17
Glass plate inclined to make meniscus disappear

If the plate is slowly tilted from its vertical position, so that it ultimately coincides with the tangent CD, the meniscus dissappears on one side and appears exaggerated on the other side (Fig 1-17) The disappearance of the meniscus when the plate coincides with CD is used to determine the angle of contact of mercury with glass. In that position a vertical line is drawn from D to meet the surface of the liquid at E. Then tan $\theta = DE\ CE$

$$\text{or } \theta = \tan^{-1}\left(\frac{DE}{CE}\right)$$

The supplement of this angle i.e. $(180-\theta)$ is the angle of contact of mercury with glass.

1.4.3 CAPILLARY RISE OF A LIQUID

When a capillary tube is dipped vertically in a liquid like water, the liquid rises inside the tube to a certain height. If mercury is used instead of water, the level of mercury inside the capillary tube goes below the level outside. The phenomenon of rise or fall of liquid level inside a capillary tube is an important property of liquids.

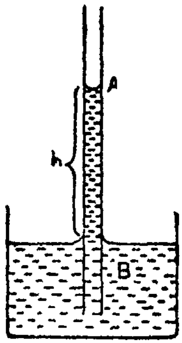


Fig. 1-18
Capillary rise

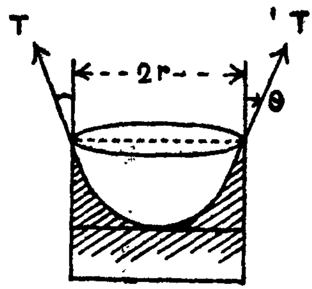


Fig. 1-19
Liquid meniscus

By measuring the rise of liquid in capillary tube, the surface tension of the liquid can be determined.

Let θ be the angle of contact of the liquid with the tube Now the vertical upward force $T \cos \theta$ per unit length acts on the circumference $2\pi r$ of the meniscus. (Fig. 1-19)

$\therefore$ the total upward force $= 2\pi r \times T \cos \theta$

This balances the weight of the column of the liquid of height h. Neglecting the weight of small amount of liquid in the curved position of the meniscus,

$2\pi r\, T \cos \theta = \pi\, r^2 h \rho g$ ($\rho =$ the density of the liquid).

$\therefore$ The surface tension $T = \dfrac{rh\rho g}{2 \cos \theta}$

DETERMINATION OF THE SURFACE TENSION OF WATER BY THE CAPILLARY RISE METHOD

A capillary tube AB of uniform cross-sectional area is first cleaned with dilute acid and rinsed with water. It is then clamped vertically with one end inside water taken in a beaker (Fig. 1-20). Water rises to a definite height in the tube due to surface tension and the meniscus is at a certain height. A pointer *P* is mounted vertically so that its lower end just touches the surface of water in the beaker.

A travelling microscope *M* is adjusted such that the horizontal cross-wire is tangential to the lower meniscus of the liquid in the tube. The reading on the vertical scale is taken. The beaker of water is removed and the microscope is brought down and focussed on the tip of the pointer and again the reading is taken. The difference between the two readings gives the rise *h* of the liquid in the capillary tube.

The diameter $2r$ of the capillary tube is measured using the travelling microscope.

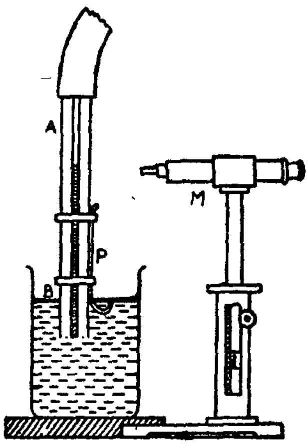


Fig. 1-20
S.T. by Capillary rise method

$$\text{The surface tension } T = \frac{rh\rho g}{2 \cos \theta}$$

$$\text{As } \theta = 0 \text{ in th case of water, } T = \frac{rh\rho g}{2}$$

If h is measured in metres and ρ is in kgm^{-3} then T will be in netwons/meter ($N\,m^{-1}$)

1.4.4. SURFACE TENSION BY SEARLE'S TORSION BALANCE

The arrangement consists of a tripod with a screw *S*, provided on one of its legs (Fig. 1-21). The tripod carries a metallic frame which can be raised or lowered.

A steel wire *AB* is attached to the metallic frame. To the middle of the steel wire, a long pointer is attached. The longer arm of the pointer moves in front of a vertical graduated scale. The shorter arm carries a small sliding weight *W*.

P is a scale pan suspended from the pointer The pan has a hook at the bottom. From the hook a glass plate is suspended.

The glass plate is first cleaned with dilute acid and then with water. It is supended from the hook at the bottom of the pan. The pointer is adjusted to be horizontal by raising or lowering the metal frame and also by adjusting the sliding weight. A trough of water is brought below the glass plate. The lower edge of the glass plate is made horizontal and parallel to the water surface in the trough is adjusted so that the glass plate is just in contact with water surface. The screw S is adjusted so that the glass plate just gets detached from the water surface. The reading R of the pointer on the vertical scale is noted just when the glass plate gets detached.

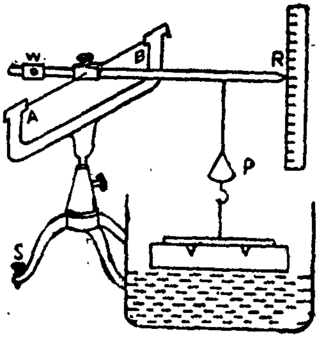


Fig. 1-21
S. T. by Torsion balance method

The trough is removed, the glass plate is wiped dry and weights are added to the pan P till the pointer indicate the same reading R. Let the mass placed on the pan now be m. The force due to surface tension acting on the lower edge of the glass plate $= 2T\,(l+t)\cos\theta$

where l and t are the length and thickness respectively of the glass plate. T is the surface tension of the liquid and θ is the angle of contact of the liquid with glass. (For water $\theta=0$)

This force is equal to the weight placed on the scale pan.

$$\therefore mg = 2T\,(l+t)$$

$$T = \frac{mg}{2\,(l+t)}\ N\,m^{-1} \qquad (31)$$

Exercise 1.4

1. Explain surface tension with examples.
5. What is meant by surface energy? How is it related to surface tension?
3. What is the phenomenon of capillary rise?
4. Define 'angle of contact'. Illustrate with examples.
5. Explain how the surface tension of a liquid can be determined by measuring the capillary rise.
6. Describe an experiment to determine the surface tension of a liquid in the laboratory.

1.5 OSCILLATIONS AND WAVES

1.5.1 SIMPLE HARMONIC MOTION

A simple type of motion often encountered is the vibratory motion of a particle about an equilibrium position under the influence of a force that is proportional to the distance of the particle from the equilibrium position. The force always acts such that it directs the particle back to its equilbrium position and hence is called the restoring force. This kind of motion is known as *simple harmonic motion.* The oscillations of a simple pendulum, vertical oscillations of a mass suspended from a spring, vibrations of a freely suspended magnet in a magnetic field, motion of particles of air through which sound waves are passing are some examples of such motion.

DEFINITION OF SIMPLE HARMONIC MOTION

A particle is said to execute simple harmonic motion. if its acceleration is directly proportional to the displacement from a fixed point and is always towards that fixed point.

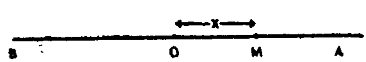


Fig. 1-22
S.H.M. of a particle

Let M be a particle executing simple harmonic motion along a straight line AOB with O as a fixed point. (Fig. 1-22). By definition, its acceleration and the restoring force on it is directed towards the fixed point, O, and is directly proportional to its displacement, X, from O. ie.,

$$\text{acceleration } (\alpha) \propto x$$

ar $$\alpha = -\omega^2 x \qquad (32)$$

where ω is a constant for a given simple harmonic motion and is the **angular frequency** of the simple harmonic motion; the negative sign indicates that the acceleration is opposite to the direction of increase of x, (i.e.,) always directed towards O. O is also the equilibrium position of the particle. If m is the mass of the particle, the restoring force.

$$F = m\,\alpha$$

or $$F = -m\omega^2 x \qquad (33)$$

The simple harmonic motion is closely associated with circular motion. This can be illustrated as follows: Let us consider a particle P, moving along a circle with uniform speed. Let O be the centre of the circle, a, the radius, AOB, any diametre and M, the projection of P on AOB. (Fig. 1-23). As the particle, P, moves along the circle, the point M will move to and fro along AOB. It can be shown that the acceleration of M is always directed towards O and is proportional to the displacement, x, of M from O. Hence, M executes simple harmonic motion along AOB. If AOB is along X - direction the simple harmonic motion is said to be along X - direction Similarly, if YOY_1, is another diameter perpendicular to AOB and N, the projection of P on YOY_1, N also will execute simple harmonic motion along the Y direction. On this basis the simple harmonic motion can also be defined as follows: *The simple harmonic motion is the projection of uniform circular motion on any diameter.* The circle is known as the generating circle.

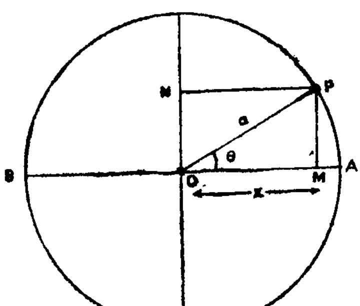


Fig. 1-23
Simple harmonic and circular motions

If ω is the angular velocity of the particle, P, about O, it can be shown that at any instant when the displacement of M from O is x, the acceleration of M

$$\alpha = -\omega^2 x$$

Hence the angular frequency of the simple harmonic motion as defined in Eqn. 32 is nothing but the angular velocity of P about O

Characteristics of Simple Harmonic Motion

The characteristics of the simple harmonic motion can be well understood with the help of the above definition based on circular motion. The charateristics of the simple harmonic motion are (i) amplitude, (ii) frequency, and (iii) phase.

Amplitude:- *The maximum displacement from the equilibrium position of the particle exerting simple harmonic motion is called the amplitude of the simple harmonic motion.*

In Fig. 1-23, $OA = a$ is the amplitude of the simple harmonic motion along the X- direction and OY is that along Y- direction.

Frequency:- Let us now consider a simple harmonic motion executed by N along YOY_1 as shown in Fig. 1-24. Let the particle, P start its motion along the circle from A (when N is at O); i.e let $\phi = 0$ when $t = 0$. Let at any instant t, $D\hat{O}A = \phi$ and $ON = y$ is the displacement of the simple harmonic motion. Then from the figure it is evident that

$$y = a \sin \phi \tag{34}$$

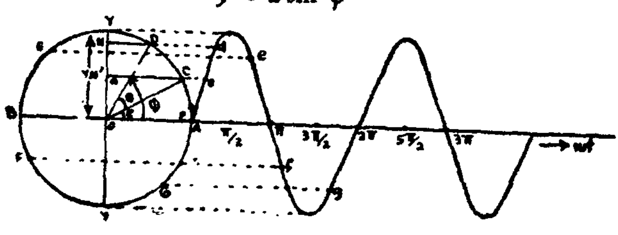


Fig. 1-24

Simple hormonic motion along Y direction

If ω is the angular velocity of the particle P, about the centre O of the circle, then, $\phi = \omega t$ and

$$y = a \sin \omega t$$

Since ω is a constant, the displacement of the particle executing imple harmonic motion (N) veries sinusoidally with time Hence, a graph drawn between the displacement, y and the time t will be a sine curve as shown in Fig 1-24.

If however, the particle P starts its motion from a position C (when N is at N') such that $C\hat{O}A = \varepsilon$ at $t=O$, then at any instant t when P is at D such that $D\hat{O}C = \theta$, the displacement of the simple harmonic motion is given by

$$y = a \sin(\theta + \varepsilon) \tag{36}$$

or

$$y = a \sin \omega (t + t') \tag{37}$$

where $\omega t' = \varepsilon$ and t' is the time that would have elapsed if P had started from A. Eqn. (36) and (37) are the most general equations of the simple harmonic motion.

It can be seen from Eqn. 35 that the displacement, y, is the same (zero) when ϕ $(= \omega t) = O, 2\pi, 4\pi, \ldots\ldots\ldots\ldots$, i.e., when $t = O, \dfrac{2\pi}{\omega}, \dfrac{4\pi}{\omega} \ldots\ldots\ldots\ldots$ (or a when $\omega t = \pi/2, 5\pi/2 \ldots\ldots\ldots$ i.e., when $t = \pi/2\omega, 5\pi/2\omega, \ldots\ldots\ldots\ldots$). The time interval between two consecutive instants at which the displacement is the same is called the period of the simple harmonic motion

$$\text{i.e., Period } T = \frac{2\pi}{\omega} \tag{38}$$

(Though $y = O$ when $\omega t = O$ as well as π, i.e. when P is at A as well as B, there is a difference. When $\omega t = O$, N tends to move upwards from O where as when $\omega t = \pi$ i.e. P is at B, N tends to move downwards.) It is evident that the period of simple harmonic motion is also the time taken for one complete to and fro motion of N along YOY_1. One such complete to and fro motion is called one *oscillation*. The number of oscillations made in one second by the particle executing simple harmonic motion is defined as the frequency of simple hormonic motion. It is given by the reciprocal of the period of simple harmonic motion; i.e.,

$$\textit{frequency } n = \frac{1}{T} = \frac{\omega}{2\pi} \tag{39}$$

It is expreesed in the unit *cycles/sec*. The unit cycle per sec.is called *hertz* (Hz). The angular frequency

$$\omega = 2\pi n \tag{40}$$

Phase:- The displacement, at any instant t, of a particle executing simple harmonic motion is given by Eqn. 36 or Eqn. 37 In Eqn. 36, the factor $(\theta+\varepsilon)$ is known as the *phase* of the simple harmonic motion; ε is known as the initial phase or *epoch*. Here the phase is expressed in terms of angle. In Eqn. 37, $(t + t')$ is also called the phase of the simple harmonic motion. Here it is expressed in terms of time. It may be clear that $(t + t')$ is the time that has elapsed from the instant N (Fig. 1-24) crosses it equilibrium position, O in the positive direction. Hence, *the phase of the simple harmonic motion can be defined as the time that has elapsed since the particle crosses its equilibrium position in the positive direction*. It is the usual practice to express the phase as a fraction of the period $\left(\frac{T}{4}, \frac{T}{2}\right)$ of the simple harmonic motion. Since $\omega t = 2\pi$, a phase of T corresponds to a phase of 2π in angle.

If there are two simple harmonic motions represented by equations,

$$y_1 = a \sin (\theta+\varepsilon_1) \tag{41}$$

and

$$y_2 = a \sin (\theta+\varepsilon_2) \tag{42}$$

they are said to differ in phase by $\varepsilon_1 - \varepsilon_2$. If $\varepsilon_1 - \varepsilon_2 = O, 2\pi, 4\pi,$ ...., the two simple harmonic motions are said to be in phase; if $\varepsilon_1 - \varepsilon_2 = \pi, 3\pi, 5\pi,$ ...., they are said to be in opposite phases.

1.5.2 WAVE MOTION

The principal methods by which energy is transferred from place to place are (1) by mechanical motion of particles and (2) by wave motion. A stone dropped into a quiet pond gives rise to a set of circular ripples or waves spreading outward over the surface at constant speed. A piece of cork floating on the surface of water does not move forward with the waves but merely moves up and down about its mean position. It is only the disturbance that moves on the surface of water while the water particles vibrate up and down perpendicular to the direction of the wave.

1.5.3. TRANSVERSE WAVES

When the motion of the individual particles of the medium is perpendicular to the direction in which the wave advances, the wave is called a transverse wave. **Hence waves on the surface of water are transverse waves.**

In transverse waves, due to the vibrations of the particles in a direction perpendicular to that of the wave, crests and troughs are formed. Crests are points having maximum upward displacement and troughs are points having maximum downward displacement. These crests and troughs move in the direction of the wave with a velocity known as the wave velocity.

1.5.4. LONGITUDINAL WAVES

When the vibratory motion of the individual particles of the medium is parallel to the direction in which the wave travels, the wave is called a longitudinal wave. Longitudinal waves may be observed in fields when a breeze makes the top of the plants move to and fro. In longitudinal waves due to the vibrations of the particles parallel to the direction of the wave, compressions and rarefactions are formed. Compressions are points where the density of the vibrating particles is maximum; rarefactions are points where the density is minimum. These compressions and rarefactions move in the direction of the wave with a velocity known as the wave velocity.

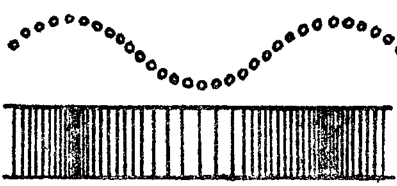


Fig. 1-25
Transverse (above) and Longitudinal (below) waveforms

1.5.5. DEFINITIONS

The wavelength of a wave is the distance measured in the direction of the wave between two successive points on it which are in the same state of vibration.

It is usual to measure the distance between two successive crests or troughs in the case of transverse waves as the wavelength. Similarly in the case of longitudinal waves the distance between two successive compressions or rarefactinos will give the wavelength.

The frequency of the wave is the number of wavelengths that cross any point in one second. It is also equal to the number of vibrations made by any single particle in the medium about its equilibrium position in one second.

The velocity of a wave is the distance moved in one second by a crest or a trough in the case of transverse wave or the distance moved in one second by a compression or rarefaction in the case of a longitudinal wave.

1.5.6. FORMULA

If n is the frequency of a wave and λ its wavelength, the distance travelled by a crest or trough in one second is $n\lambda$. Similarly the distance travelled in one second by a compression or rarefaction also is $n\lambda$. Since the distance is travelled in one second, it is equal to the velocity v of the wave.

$$\therefore v = n\lambda. \tag{43}$$

1.5.7. PROGRESSIVE AND STATIONARY WAVES

When waves are produced in an extensive medium, they travel) forward without being disturbed. These are known as progressive waves. But if a progressive wave meets an obstacle, it is reflected by the obstacle. The reflected wave is superposed on the incident wave. Each particle of the medium is disturbed both by the in-

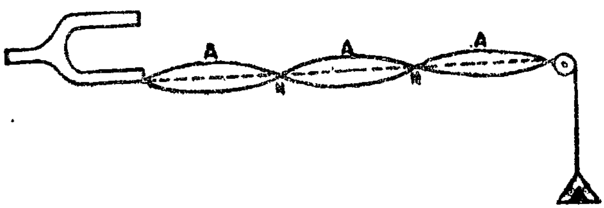


Fig. 1-26
Stationary waves

cident and the reflected waves at the same time. At certain points of the medium, the two sets of waves cancel each other's effect. So the particle of the medium will be at rest at these points. These points are called **nodes**. At points midway between two successive nodes, the particle experiences maximum displacement.

These points are called **antinodes.** Thus if two identical waves travelling in opposite directions are superposed, stationary waves are produced. Nodes and antinodes are the characteristics of stationary waves.

Transverse stationary waves can be produced in strings. A string is attached to one prong of a fork, and passed over a smooth pulley and kept stretched by placing weights in the scale-pan. The fork is set into vibration electrically. The length of the string, and the weights in the scale pan are adjusted so that the string appears to be in the form of loops (Fig. 1-26). Points marked as N, N are at rest and they are the nodes Points marked A, A having maximum displacements are the antinodes. The distance between two successive nodes is equal to half of the wavelength. The distance between a node and the next antinode is equal to one-fourth of the wavelength.

Longitudinal stationary waves can be produced through the air column in a tube. Consider a tube closed at one end. Let a progressive longtiudinal wave advance through the air column in the tube. When it reaches the closed end, it is reflected. The reflected wave is superposed on the incident wave producing stationary longitudinal waves in the air column.

1.5.8. VIBRATIONS OF AIR COLUMNS AND PIPES

Sounding bodies are capable of vibrating in a variety of ways giving rise to a complex note consisting of tones of different frequencies. The lowest frequency in this mixture is usually the the strongest and is called the fundamental. Overtones whose frequencies are in the ratio 1 : 2 : 3 etc., form a series called *Harmonic series*. Organ pipes are capable of producing a number of harmonics.

OVERTONES IN ORGAN PIPES

(i) Closed Organ Pipe

A closed organ pipe has one end closed and the other end open. When it is emits a note, a node is formed at the closed end and an antinode is formed at the open end. The fundamental note having the lowest frequency and the longest wavelength will be the loudest note.

The pipe will accommodate one node and one antinode (Fig. 1-27).

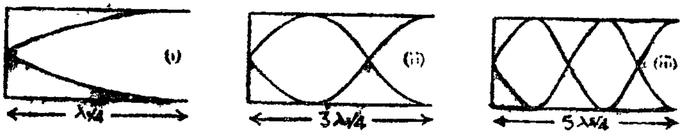


Fig. 1-27
Closed organ pipe

The length of the pipe corresponding to half a loop or a quarter of the wavelength is given by

$$l = \frac{\lambda_1}{4} = \text{or } \lambda_1 = 4l$$

If v is the velocity of sound

$$v = n_1 \lambda_1$$
$$v = n_1 \times 4l$$
$$\text{or } n_1 = \frac{v}{4l}$$

The first overtone will correspond to the next possible higher mode of vibration of the air column. Here the length of the pipe corresponds to three quarters of a wavelength (Fig. 1.27 ii).

$$l = \frac{3}{4}\lambda_2 \qquad \therefore \lambda_2 = \frac{4}{3} l$$

$$v = n_2 \lambda_2 = n_2 \times \frac{4}{3} l$$

$$n_2 = \frac{3v}{4l} = 3 n_1.$$

Thus the first overtone has thrice the fundamental frequency and is therefore the third harmonic. The next higher mode of vibration will correspond to the second overtone shown in Fig 1-27 iii). Here the length of the pipe will correspond to five quarter wavelengths. Hence

$$l = \frac{5}{4} \lambda_3$$

$$\therefore \lambda_3 = \frac{4}{5} l$$

$$\text{Hence } n_3 = \frac{5v}{4l} = 5\frac{v}{4l} = 5n_1$$

Hence the second overtone has a frequency five times that of the fundamental and therefore it is the fifth harmonic. Thus the different notes emitted have frequencies in the ratio 1 : 3 : 5...

(ii) Open organ Pipe

Open pipes like the nadaswaram have both ends open. When it emits a fundamental note, antinodes are formed at the ends

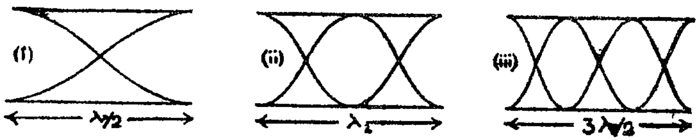


Fig. 1-28
Open organ pipe

and a node is formed at the middle (Fig. 1-28 i). The length of the pipe corresponds to one loop or half a wavelength.

$$l = \frac{\lambda_1}{2}$$

$$\text{or} \quad \lambda_1 = 2\,l$$

The fundamental frequency $n_1 = \frac{v}{\lambda_1} = \frac{v}{2l}$

For the first overtone two nodes are formed inside the pipe so that the length of the pipe corresponds to one wavelength (Fig. 1-28 ii). The frequency of the overtone is given by

$$n_2 = \frac{v}{\lambda_2} = \frac{v}{l} = 2 \times \frac{v}{2l} = 2n_1$$

The first overtone has a frequency twice that of the fundamental. For the next mode of vibration, three nodes are formed inside the pipe so that the length of the pipe corresponds to one wavelength and a half (Fig. 1-28 iii). The frequency of the overtone.

$$n_3 = \frac{v}{\lambda_3}$$

$$\text{But } l = \frac{3}{2}\lambda_3$$

$$\therefore \lambda = 2 \times \frac{2}{3} \times l$$

$$\therefore n_3 = \frac{3v}{2l} = 3n_1.$$

The frequency of the second overtone is thrice that of the fundamental.

Thus the frequencies of the harmonic series are in the ratio 1 · 2 : 3 : ......

In the case of open organ pipes, both odd and even harmonics are present. In the closed organ pipes only odd harmonics are present.

1. 5. 9. FORCED VIBRATION

Each vibrating body has a natural frequency depending upon its elastic constants, dimension and mode of vibration. Such vibrations are called free vibrations and the frequency of the vibrating body is called its *free* or *natural frequency*. A tuning fork set into vibration and a pendulum oscillating under the action of gravity are examples of free vibrations.

We shall now consider a body which is maintained in a state of vibration by a periodic force which may or may not have the same period as the free vibration. In this case the body ultimately vibrates with the same period as that of the periodic force. Vibrations of such a body are called forced vibrations since their frequency is determined by that of the applied force rather than the characteristics of the body. If a tuning fork set into vibration is held in the hand, the sound produced from it is feeble. But if the stem of the vibrating tuning fork is kept pressed on a table, the sound is intensified whatever may be the frequency of the tuning fork. Here the top of the table is forced to vibrate in response to the periodic force applied. The vibrations of the table are called *forced vibrations*. The amplitude of such vibrations depends upon the difference in frequencies of the external periodic force and the natural frequency of the vibrating body.

1. 5. 10. RESONANCE

If the natural frequency of the table is the same as that of the tuning fork, a very loud sound is heard. In this special case, when the natural frequency of the body is equal to the frequency of the applied force, the amplitude of the forced vibration becomes very large and the body is said to be in *resonance* with the periodic force applied.

Consider a swing which is swinging with its natural period. If at every time the swing starts forward in its periodic motion, a gentle push is given in the direction in which it tends to move, a very large motion of the swing will be produced. In this case the frequency of the pushes given (the frequency of the forcing body) is equal to the natural frequency of the swing and so there is resonance.

The amplitudes of the swing will certainly diminish if the pushes are not given at the right moment ; i.e. if the frequency of the forcing body is not equal to the natural frequency of the swing.

Suppose the length of the air column enclosed in a tube is adjusted to have the same frequency as that of a given tuning fork. If the tuning fork is set in vibration and held near the mouth of the tube, a booming sound is heard. This is due to the air column being driven into resonant vibrations.

1. 5. 11. DAMPED VIBRATIONS

When a freely vibrating body has no resistance offered against its motion, its amplitude remains constant. In general there is always some resistance to be overcome which is called *damping*. The effect of the damping causes a decrease in amplitude and slight change in frequency. Usually there will be a slight increase in the periodic time. Experiments have shown that the effect of damping is small in the case of a tuning fork. In the case of air column, the effect of damping is so large that the vibration dies out after a few oscillations.

1. 5. 12. ULTRASONICS

The normal human ear is sensitive to sounds of frequencies between 20 and 20000 hertz. Ultrasonics are sound waves whose frequencies are above 20000 hertz. Certain animals like bat and insects like cricket are capable of producing and detecting sounds of this type. By using special type of oscillators ultrasonic waves of frequencies of the order of 10 megahertz can be produced in the laboratory. These ultrasonic waves also obey the ordinary laws of sound. In addition, they have useful applications for which ordinary sound waves cannot usually be used.

USES OF ULTRASONICS

MATERIAL TESTING

Ultrasonic waves can penetrate through metals and other materials which are sometimes opaque to electromagnetic radiations. They are reflected at the surface of separation between two media. These facts are used to detect cracks in welding and flaws in metals.

ECHO SOUNDING

Echo sounding methods are used to locate icebergs and to measure the depths of oceans. A pulse of ultrasonic wave is sent and the reflected pulse is received using a suitable receiver. Knowing the time interval between the instant of transmission and the instant of reception, the pulse and the velocity of the waves in the medium, the depth of the sea can be determined. A narrow beam is generally preferred. Frequencies from 14 kHz to 80 kHz are generally used.

SONAR

One of the important applications of ultrasonics is the **SONAR**. This means Sound Navigation And Ranging. This is based on the principle of echo sounding. Ultrasonic waves can travel large distances through water without losing their energy. An ultrasonic generator and a receiver are placed inside a dome provided with stainless steel windows. The dome is lowered into water and rotated at a constant speed. Knowing the time that elapses between the transmitted pulse and the received pulse, the range is determined. The angular position of the transmitter at which it receives the echo gives the direction of the submarine or any other obstacle

MEDICINE

Ultrasonics can be used instead of X-rays to detect tumours in the human body In surgery, some cell structures can be destroyed by the heat generated by the absorption of ultrasonic waves. Ultrasonic waves are used in the treatment of arthritis, rheumatism, and abscess. It is used to lame small animals like rat, fish and frog and to kill bacteria.

Exercise 1.5

1. *What is 'simple harmonic motion'? Illustrate with examples.*
2. *What are the two types of progressive waves? Explain how these waves advance in a medium.*
3. *Define the wavelength and frequency of a wave. Obtain an expression for the velocity in terms of these.*
4. *What are stationary waves? How are they formed?*
5. *What is meant by 'harmonic series?' What are overtones?*
6. *Describe an open organ pipe and a closed organ pipe. How do the overtones in these pipes differ?*
7. *What is meant by forced vibrations? Explain the phenomenon of resonance.*
8. *What is an ultrasonic wave? Explain.*
9. *What are the uses of ultrasonics?*

2. THERMAL PHYSICS

2. 1. THERMOMETRY

2. 1. 1. INTRODUCTION

Consider a motor car brought to rest by applying brakes. The car loses kinetic energy but there is an increase in the temperature of the brake linings. A meteorite entering the earth's atmosphere loses energy as it does work against the frictional forces of the earth's atmosphere. Here, the loss of energy results in an increase of temperature and usually the meteorite melts before it reaches the earth. The rise of temperature in these examples indicate that something has happened to the body and this something is called the gain of *heat energy*. These are examples of the conversion of kinetic energy into heat energy. Similarly, electrical energy can be converted into heat. Such conversions indicate that "Heat is a form of energy". Hence it is measured in joules, the joule being the unit of energy.

Heat and temperature are two related quantities but they are different from each other. In other words, heat is the *cause* and temperature is the *effect*. If a body is heated, its temperature rises. On the other hand if a body is cooled, its temperature falls. Heat as mentioned above is energy and therefore measured in joules. But the temperature is measured in degrees.

Temperature may be measured with the help of thermometers. The centigrade thermometer has 100 degrees between ice and steam points and the room temperature is about $30°C$. The temperature of the melting ice under normal pressure is taken as $0°C$ by convention. The temperature of boiling water under normal pressure is taken as $100°C$, again by convention.

There is another scale of temperature called the absolute scale of temperature or the Kelvin scale of temperature. The interval of a degree in this scale is equal to the interval in the centigrade scale. But the zero in this scale namely $0° A$ or $0\ K$ corresponds to $-273°C$ which is the real zero of temperature. Thus the temperature of the melting ice is taken to be $273° A$ or $273\ K$ Similarly the temperature of the boiling water is $373° A$ or $373\ K$ under normal pressure.

Measurement of temperature is important in everyday life. The temperature of the cold-storage plants must be maintained at a particular value to store foodstuffs in good condition over a long period. Furnaces making glass must be operated at a paticular high temperature. The sense of touch is not a reliable method to measure temperature as, for example, a metallic object appears cooler than others when touched by hand even though all of them may be at the same temperature. The thermometer is a very reliable instrument to measure temperature.

Any property of a substance which changes continuously with temperature may be used for measurement of temperature. In the clinical thermometer, which is used to measure the temperature of the body, for example, the volume of mercury changes when its temperature changes.

2. 1. 2. SCALE OF TEMPERATURE

It is most essential to have two fixed temperatures to define the standard temperature interval. These temperatures which are called "Fixed Points" should be constant and capable of being easily reproduced at all times. The *lower fixed point* is the temperature at which ice melts under one standard atmospheric pressure and the *upper fixed point* is the temperature at which pure water boils under one standard atmospheric pressure. Celsius suggested that (1) the temperature of the melting point of ice should be given a number 0' and the temperature of the steam at a pressure of 760 *mm* mercury should be given number '100' and (2) the interval between the two fixed points should be divided into 100 equal parts. Hence this scale of temperature is called the centigrade scale. It is now called the **Celsius** Scale.

We have seen that heat is a form of energy. If the gas is heated the kinetic energy of the molecules increases. The gas molecules have on an average, greater velocity than the earlier value. If the gas is cooled the velocity is lowered. The molecules will have zero speed at some extremely low temperature and this temperature is called the "Absolute Zero" The gas pressure would then be zero theoretically. The scale of temperature in which the absolute zero is defined as "0 degree" is named as the Absolute Scale. It is also called the Kelvin Scale after Lord Kelvin who defined it. The absolute zero is denoted by 0 K. In this scale also, the interval between the ice point and the steam point is taken as 100 degrees so that 1 degree interval is the same in both the Celsius and the Kelvin Scales of temperature. This Kelvin or Absolute scale is used in the SI system of units. From experiments it is found that $0\ K = -273°C$ and $273K = 0°C$ and $373\ K = 100°C$.

Fig. 2·1
The Centigrade Thermometer

2. 1. 3. LIQUID THERMOMETERS

The property of expansion of liquids is utilised to construct the familiar *liquid-in-glass* thermometers. The change in volume of a liquid is used as a means of measuring the temperature. Mercury, alcohol, and toluene are used in thermometers. The choice of a liquid depends on the temperature range over which the thermometer is to operate and the sensitivity required. The range of the mercury thermometer is from its freezing point $-39°C$ to its boiling point $357°C$ For low temperatures, alcohol thermometers are used. Alcohol has a freezing point $-112°C$ and a boiling point $78°C$. In modern thermometers, toluene is preferred to alcohol, since its freezing point is $-90°C$ and its boiling point is $111°C$.

2. 1. 4. GAS THERMOMETERS

Liquid thermometers are not suitable for measuring very high or very low temperatures. Also, the reading on a liquid-in-glass thermometer depends on the length of the exposed column. A further disadvantage of liquid-in-glass thermometers is that the scale becomes inaccurate over a period of time, because of a slight contraction of the glass bulb. Gas thermometers depend on the changes in the pressure or volume of a gas when heated.

The use of gas as a thermometric substance is based on the two gas laws. **Boyle's law** states that the volume of a given mass of a gas varies inversely as its pressure, the temperature remaining constant. **Charles' law** states that the pressure of a given mass of a gas at constant volume varies directly as its absolute temperature, or the volume of a given mass of gas at constant pressure varies directly as its absolute temperature. Hence any one of the two principles may be used in the construction of gas thermometers. A gas thermometer possesses the following advantages ·

1. Gases expand much more than liquids when heated and this means that temperatures may be measured to a higher degree of accuracy.

2. A large change of volume of a gas occurs when its temperature is altered, so that the expansion of the container is negligible.

3 Gases also expand much more regularly with temperature than mercury.

4. A permanent gas as a thermometric substance maintains its state and behaves in the same way at very high as well as very low temperatures.

In the constant pressure gas thermometer the pressure is kept constant and changes in volume are observed with temperature. In the constant volume thermometers, the volume is kept constant and changes in pressure are observed with temperature.

2. 1. 5. THE CONSTANT VOLUME HYDROGEN THERMOMETER

Permanent gases like hydrogen, nitrogen and helium are ;ually used as thermometric substances in the constant volume gas thermometer. Since the changes in pressure at constant volume of hydrogen at ordinary temperatures are far more regular than in other gases, hydrogen is preferred as a thermo metric substance.

THE CONSTANT VOLUME HYDROGEN THERMOMETER

It consists of a bulb 'B' made up of platinum iridium alloy of one litre capacity, about 18 *mm* radius. It communicates with a mercury manometer by means of a capillary tube. The manometer consists of two limbs C and D which are connected to the mercury reservoir.

An index I_1 is fixed above the level of mercury in the tube D. The mercury reservoir M has to be raised or lowered such that the surface of mercury in the tube 'D' just touches the tip of the index fixed in the tube D.

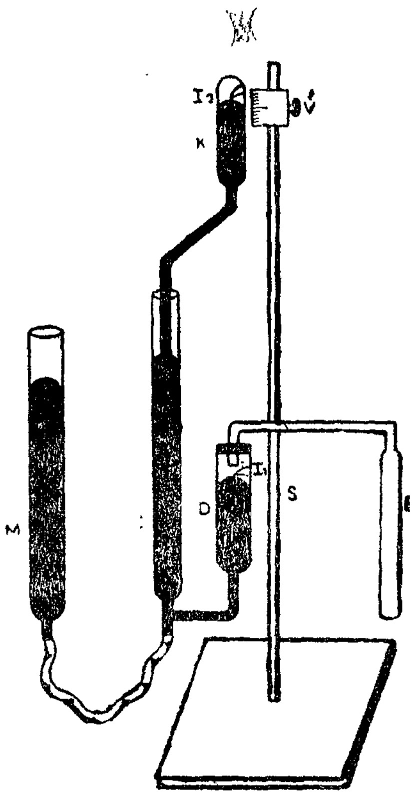


Fig. 2·2
The Hydrogen Thermometer

A barometer 'K' dipping in mercury kept in C is bent such that the surface of mercury in the wide closed end is vertically over the tube D. There is another pointer 'I_2' fixed near the closed end of the barometer tube.

The bulb 'B' is maintained at the ice point and the reservoir M is adjusted such that the surface of mercury in D is just tangential to the tip of the index I_1. The barometer tube is adjusted until the mercury surface in it is just tangential to the tip of the index I_2. The distance I_1 I_2 gives the pressure of the gas, which is read from the scale 'S' using a vernier 'V'. Let the pressure be P_0.

The experiment is repeated by exposing the bulb to steam (100°C) and the pressure is found as before. Let the pressure be P_{100}.

The bulb is exposed to an unknown temperature t°C and the pressure is found out as before. Let it be P_t. Since the variation of pressure at constant volume is uniform,

$$\frac{P_{100}-P_0}{100} = \frac{P_t-P_0}{t}$$

$$\therefore \quad t = \frac{P_t-P_0}{P_{100}-P_0} \times 100$$

Hence the unknown temperature can be determined. This thermometer can be used from — 200°C to 500°C.

2. 1. 6. RESISTANCE THERMOMETER

The electrical resistance of a metal wire generally increases with temperature, the increase being fairly uniform over large ranges. Callendar has shown that in the case of pure platinums the variation is represented by the equation $R_t = R_0\ (1+\alpha t+\beta t^2)$ where R_t and R_0 are the resitances of the coil of platinum wire at temperatures $t°\ C$ and $0°C$ respectively, and α and β are constants for the given specimen. These two constants can be experimentally found out and the equation is used to calculate the unknown temperature.

If R_0, R_{100} and R_t be the resistances of the platinum wire at 0°C, 100°C and an unknown temperature t°C respectively, we have

$$\frac{R_{100}-R_0}{100} = \frac{R_t-R_0}{t}$$

$$\therefore \quad t = \frac{R_t - R_0}{R_{100}-R_0} \times 100$$

Thus the temperature 't' is calculated from a knowledge of R_t, R_{100} and R_0.

2. 1. 7. PLATINUM RESISISTANCE THERMOMETER

It consists of a tube 'T' made up of glass or glazed porcelein. A wire of pure platinum free from silicon or other impurities is selected. It is doubled on itself to avoid induction effects and is wound on a thin grooved mica plate 'M' to ensure good insulation. It is kept at the bottom of the tube having an ebonite cap E.

The terminals of the platinum wire are connected to two long leads passing through mica discs m, m and connected to two binding screws A_1 and A_2 on the ebonite cap. A wire of the same material, length and diameter as the leads, is bent and passed through the mica discs m, m so as to insulate it from other leads and connected to the binding screws C_1 C_2 on the ebonite cap. This wire is provided to equalise the resistance due to the connecting leads to platinum wire and is known as compensating leads. The tube is exhausted and sealed to avoid possible oxidation of platinum at high temperatures.

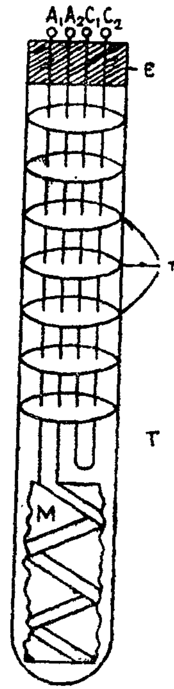


Fig. 2-3
Platinum Resistance Thermometer

The resistance of the platinum wire at different temperatures can be accurately measured. As discussed in the previous section,

$$t = \frac{R_t - R_0}{R_{100} - R_0} \times 100$$

Keeping the resistance thermometer in a bath along with the bulb of a constant volume hydrogen thermometer, the resistances are measured for various temperatures given by the hydrogen thermometer. A graph is drawn taking temperature on the X-axis and resistance on the Y-axis. The graph is a straight line when the temperatures are not high

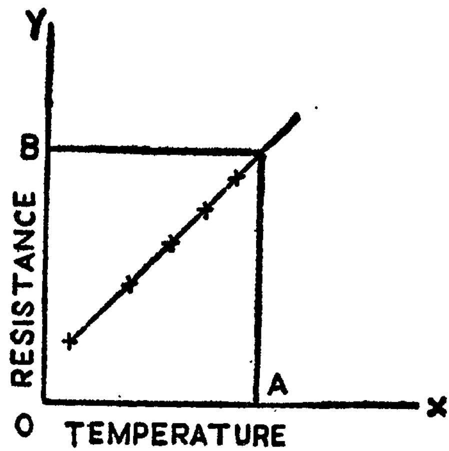


Fig. 2-4
Resistance—Temperature graph

From the graph the unknown temperature may be determined.

Advantages

1. The Platinum resistance thermometer can be used to measure temperatures over a wide range, from about 70 *K* to 1470 *K*.

2. The temperature can be measured correct to 0·01 *K* and the accuracy is 0·1 *K* if the temperature is higher than 800 *K*.

3. The resistance of pure platinum and annealed platinum wire is constant at a particular temperature. Hence the temperature measured is fairly constant.

4. When it is standardised once with a constant volume hydrogen thermometer, it can be used to measure even small differences in temperature fairly accurately.

Disadvantages

1. The platinum resistance thermometer has a large thermal capacity and the low conductivity of the material of the sheath covering the resistance wire will not allow it to reach the temperature of the bath quickly. Therefore this cannot be used to measure rapidly varying temperatures.

2. Also the presence of the impurities in the platinum wire will affect the accuracy of measurements.

3. Further this themometer cannot be used above 1300 K since there is the danger of contamination by the insulating materials.

Exercise 2. 1.

1. *What is meant by 'heat' and 'temperature'? Explain.*
2. *Why is a gas thermometer superior to a liquid thermometer ?*
3. *What is the necessity for a scale of temperature ? Explain the absolute scale of temperature.*
4. *Describe with a neat diagram,the constant volume hydrogen thermometer and explain its working.*
5. *What is the principle of a resistance thermometer ?*
6. *Draw a neat sketch of a platinum resistance thermometer and label its parts. Explain how the temperature is measured using it. What are its merits and demerits?*

2. 2. THERMAL EXPANSION

2. 2. 1. INTRODUCTION

We are familiar with the fact that nearly all substances increase in size when heated. The clinical thermometer shows the rise of temperature of a patient suffering from fever, as the mercury in the thermometer expands. Telephone cables sag more during summer than during winter. Gases expand more than liquids and liquids expand more than solids for the same changes in temperature.

2. 2. 2. EXPANSION OF SOLIDS

When a solid is heated, the changes in length, area and volume have to be considered. Most solids expand in the same ratio in all directions simultaneously when the temperature rises. They are called *isotropic* substances. In some solids the expansion may be different in different directions and such solids are called *anisotropic* substances. Here we are concerned only with isotropic expansion of solids.

2. 2. 3. COEFFICIENT OF LINEAR EXPANSION

When a rod is heated, its length increases. The increase in length per unit length per degree rise of temperature is called the *coefficient of linear expansion* of the material of the solid.

Let 'l_0' be the length of the rod at 273 K and 'l' its length at higher temperature $(273+\theta)$ K. Then from the above definition the coefficient of linear expansion α is given by

$$\alpha = \frac{\text{increase in length}}{\text{Original length} \times \text{rise of temperature}}$$

$$= \frac{l-l_0}{l_0 \times \theta}$$

$$\therefore (l - l_0) = l_0 \alpha \theta$$

$$\therefore l = l_0 + l_0 \alpha \theta = l_0 [1 + \alpha \theta]$$

If l_1 and l_2 are the lengths at θ_1 K and θ_2 K respectively,

then $l_1 = l_0 (1 + \alpha \theta_1)$

$l_2 = l_0 (1 + \alpha \theta_2)$

$$\therefore \frac{l_2}{l_1} = \frac{(1 + \alpha \theta_2)}{(1 + \alpha \theta_1)}$$

$$(l_2 + l_2\, \alpha\, \theta_1) = (l_1 + l_1\, \alpha\, \theta_2)$$

$$\therefore\ (l_2 - l_1) = \alpha\, (l_1\, \theta_2 - l_2\, \theta_1)$$

$$\therefore\ \alpha = \frac{(l_2 - l_1)}{(l_1\, \theta_2 - l_2\, \theta_1)} \doteq \frac{(l_2 - l_1)}{l_1\, (\theta_2 - \theta_1)}$$

if $l_2 = l_1$

Thus the co-efficient of linear expansion of the material of a rod may be determined by finding 'l_1,' and 'l_2' at 'θ_1' and 'θ_2' respectively.

The coefficients of linear expansion for some metals are given in the following table :

S. No.	*Metal*	α *in* K^{-1} $\times 10^{-6}$
1.	Lead	28
2.	Aluminium	25
3.	Brass	19
4.	Copper	19
5.	Iron	12

For example when a steel rod is heated through 1 *K*, it expands by 0·000012 m for every metre. Therefore a length 1 metre becomes 1·000012 metre when heated through 1 *K*.

2. 2. 4. DETERMINATION OF THE COEFFICIENT OF LINEAR EXPANSION OF A ROD (LEVER METHOD)

A metal rod of about 1 metre long is taken. It is sharpened at the top in the form of a knife edge and pointed at the bottom. This rod is enclosed inside a steam jacket with an inlet and an outlet for steam. It is provided with a side tube for inserting a thermometer. The steam jacket is fixed on a wooden stand. The lower end of the rod rests on a metal platform and its upper end is slightly protruding outside the jacket and is free to expand. *PQR* is a horizontal lever of about 1 metre length with a fulcrum at the end *P*. It rests on the knife edge of the experimental rod '*Q*' and the other end '*R*' moves over a scale fixed vertically.

The experimental rod is taken out of the jacket and its length is measured. The rod is put back inside the jacket and the lever is placed on it. The initial temperature is noted as θ_1. The initial reading '*R*' of the lever on the vertical scale is taken. Steam is passed through the outer jacket for about 30 minutes.

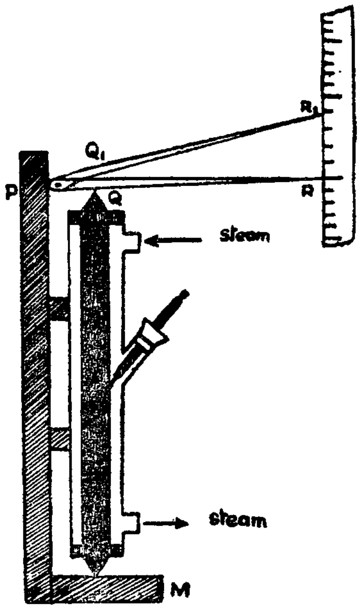


Fig. 2-5 (a)
Linear expansion apparatus

The rod expands and pushes the lever up from Q to Q_1. This causes the end of the lever to move from R to R_1. When the rod attains the temperature of the steam, the reading of the pointer remains steady showing thereby that the expansion of the rod is complete. Now the steady reading of the lever is noted as R_1. The final temperature of the rod is noted as θ_2.

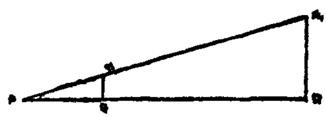


Fig. 2-5 (b)

Calculations

The initial length of the rod $= l_1\ m$

The initial temperature of the rod $= \theta_1\ K$

The initial reading of the end of the lever = R *metre*

The final temperature of the rod = θ_2 K

The final reading of the end of the lever = R_1 *metre*

The increase in length of the rod = QQ_1

Since $\triangle$ PQQ_1 and $\triangle$ PRR_1 are similar

$$\frac{QQ_1}{RR_1} = \frac{PQ}{PR} \qquad \therefore QQ_1 = RR_1 \frac{PQ}{PR}$$

RR_1 is the change in the reading of the end of the lever measured on the scale.

$\therefore$ The coefficient of linear expansion of the given rod

$$\alpha = \frac{\text{increase in length}}{\text{original length} \times \text{rise in temperature}}$$

$$\alpha = \frac{QQ_1}{l_1(\theta_2 - \theta_1)}$$

$$= \frac{RR_1 \cdot PQ}{PR} \cdot \frac{1}{l_1(\theta_2 - \theta_1)}$$

Thus the coefficient of linear expansion can be experimentally determined.

2. 2. 5. EFFECT OF TEMPERATURE ON CLOCKS

A pendulum makes a certain number of oscillations in a given time at a given place and this number is determined by the length of the pendulum. The period of oscillations of the pendulum controls the working of the clock. As the length of the pendulum is increased during summer, the clock loses time during summer. Similarly, the clock gains during the winter because the length of the pendulum gets reduced due to the fall of temperature.

2. 2. 6 COMPENSATED PENDULUM

Compensated pendulums are used to keep correct time. The principle of compensation consists in keeping the effective length of the pendulum constant. This is achieved by having two sets of rods of different materials. One set of rods expands in the downward direction, while the other set expands in the upward direction, keeping the effective length of the pendulum unaltered

2. 2. 7 HARRISON'S GRID-IRON PENDULUM

In order to keep the effective length constant, Harrison's grid-iron pendulum is built by a system of brass and iron rods. The expansion of brass is nearly one and a half times that of iron The total expansion of two brass rods of the same length is equal to the total expansion of three iron rods of the same length The

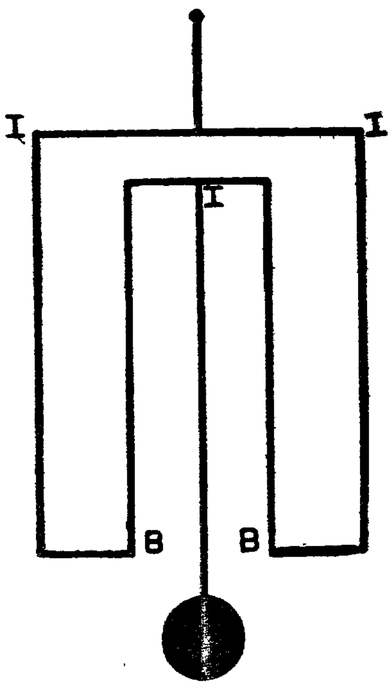


Fig. 2-6
Harrison's pendulum

grid-iron pendulum consists of five rods. The three iron-rods 'I' are so fixed that their expansions will be in the downward direction. The brass rods 'B' will expand in the upward direction The effective length of the pendulum is thus kept constant thereby keeping correct time.

2. 2. 8 EXPANSION OF CAVITY IN A SOLID

In order to consider the expansion of cavity in a solid (S), let us imagine that the cavity is filled with the same material as that of the solid, to form a compact mass with the rest of it. If this entire piece is now heated, the whole body will expand just as a homogeneous mass. The material in the space 'C' of the cavity fills it completely at all temperatures without pressing against its walls or leaving a gap.

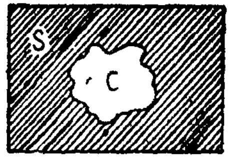


Fig. 2-7
Cavity in a solid

It follows, therefore, that the cavity must be expanding to exactly the same extent as the material filling it up. This means that the expansion of the cavity is exactly equal to the expansion of the material of the solid, having the same shape and size of the cavity. This principle is applied in finding the expansion of hollow tubes, hollow vessels and hollow spheres.

2. 2. 9. EXPANSION OF LIQUIDS

Since liquids have no definite shape of their own, we consider only their volume expansion. The liquid always requires a container which will also expand when heated.

If the expansion of a liquid is considered without considering the expansion of the container, it is called the *apparent expansion* of the liquid. If the expansion of the container also is taken into account it is called the *absolute expansion* of the liquid.

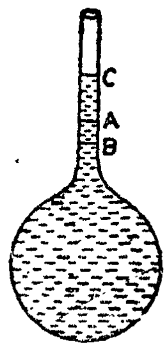


Fig. 2-8
Expansion of a liquid

The following illustration gives a clear idea of these two expansions : Consider a case in which some liquid is kept in a vessel having a narrow neck and the level of liquid is at A. Suppose the vessel is kept in a bigger vessel having some warm water just upto A. For the sake of argument consider that the vessel alone expands in the first stage. The level of liquid falls to the mark '*B*' just below *A*. Let the liquid now expand. As the expansion of the liquid is much greater than that of containing vessel, the level rises up to *C*, which is above *A*. The absolute expansion of the liquid is $BA+AC$. But such a variation in the level of the liquid will not be observed since the vessel and the liquid expand simultaneously. For an observer looking at it, the net expansion will be seen as *AC*. The actual expansion of the liquid is therefore the sum of the apparent expansion of the liquid and the cubical expansion of the container.

COEFFICIENT OF APPARENT EXPANSION OF A LIQUID

The coefficient of apparent expansion of a liquid is the observed increase in volume of the given mass of the liquid per unit volume per degree rise of temperature.

COEFFICIENT OF ABSOLUTE EXPANSION OF LIQUID

The coeffiicient of absolute expansion is the absolute or actual increase in the volume of a given mass of liquid per unit volume per degree rise of temperature.

EXPERIMENTAL DETERMINATION OF ABSOLUTE EXPANSION OF A LIQUID (DULONG AND PETIT METHOD)

The method is based on the principle of balancing of two columns of a liquid. The pressure exerted by a column of liquid depends only on the vertical height of the column and the density of the liquid and is independent of the shape or size of the containing vessel. Therefore, the measurement of the heights of two liquid columns, balanced, at different temperatures, enables us to measure the absolute expansion of the liquid directly.

The apparatus consists of a '*U*'-tube in which the experimental liquid is taken. The limbs of the '*U*' tube are surrounded by wide glass tubes. These tubes are provided with holes for inserting thermometers and also with inlet and outlet tubes. Steam from a steam heater is circulated around one limb while cold water is circulated around the other. This is continued till the levels of the liquid in the two limbs of the '*U*'

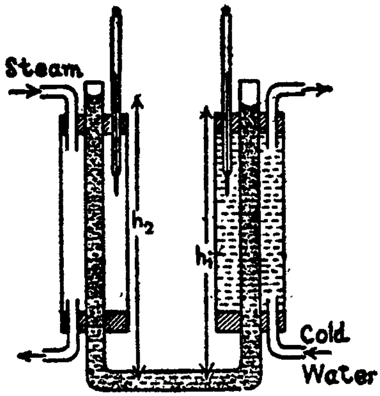


Fig. 2-9
Dulong - Petit method

tube are constant. The heights 'h_1' and 'h_2' of the cold and hot columns of the liquid are read. Let θ_1 and θ_2 be the temperatures of the two columns of the liquid respectively.

It can be shown that

$$\left.\begin{array}{r}\text{the coefficient of absolute}\\ \text{expansion of the liquid}\end{array}\right\} \gamma = \frac{h_2 - h_1}{h_1\theta_2 - h_2\theta_1}$$

Therefore by measuring h_1, h_2, θ_1 and θ_2, the coefficient of absolute expansion of the liquid can be calculated.

Exercise 2. 2

1. *Explain the thermal expansion of solids with suitable examples*
2. *Define linear expansion of a solid. Obtain an expression for it.*
3. *Explain how the linear expansion of a rod can be determined experimentally in the laboratory.*

4. *Why are pendulums of clocks compensated? What is Harrison's pendulum?*
5. *Explain the expansion of a cavity inside a solid.*
6. *Define the apparent expansion and absolute expansion of a liquid. How are they related?*
7. *Explain how the coefficient of absolute expansion of a liquid is determined experimentally in the laboratory.*

2. 3. CALORIMETRY

2. 3. 1. INTRODUCTION

Joule made an experimental study on the relation between heat and work and proved that heat produced is directly proportional to the work done. The ratio of the mechanical work spent to the heat energy generated is a constant known as *Joules constant* (J) or *Mechanical equivalent of heat.* i.e.,

$$\frac{\text{Work spent}}{\text{Heat produced}} = \frac{W}{H} = J = \text{a constant.}$$

This is known as the first law of thermodynamics.

Hence the mechanical equivalent of heat is the amount of work that is to be done in order to produce one calorie of heat and is measured in joules/calorie (J = 4·2 joules/calorie)

2. 3. 2. SPECIFIC HEAT CAPACITY

The specific heat capacity of a substance is the amount of mechanical work that has to be done to raise the temperature of 1 kilogram of the substance through 1 K. It is measured in oules/kg/K.

The specific heat capacity of water is 4180 *joules/kg/K*. The following table gives the specific heat capacities of some common substances.

S. No.	*Substance*	*Specific heat capacity in $J\ kg^{-1}\ K^{-1}$*
1.	Water	4180·00
2.	Kerosine	2090·00
3.	Cocoanut oil	1964·00
4.	Mercury	137·94
5.	Copper	384·60
6.	Brass	372·00
7.	Aluminium	903·10
8.	Lead	125·40
9.	Iron	480·70
10.	Glass	668·80

Thermal capacity : *The thermal capacity or heat capacity of a substance is the amount of work that has to be done in order to raise the temperature of the substance through 1 K.*

∴ Thermal capacity = mass × specific heat capacity. It is measured in JK^{-1}.

Water equivalent : *The water equivalent of a body is the mass of water having the same thermal capacity as the body. If a body of mass 'm' kg has specific heat capacity of 's' $J\ kg^{-1}\ K^{-1}$, then its water equivalent is*

$$W = ms/4180\ kg.$$

2. 3. 3. NEWTON'S LAW OF COOLING

It states that the rate of loss of heat (due to radiation) of a hot body is proportional to its excess of temperature over that of the surroundings.

If a body which is at a mean temperature θ, loses a quantity of heat energy H in a small interval of time due to radiation, then the rate of cooling,

$$\left(\frac{dH}{dt}\right) \propto (\theta - \theta_0)$$

where θ_0 is the temperature of the surroundings. $(\theta - \theta_0)$ is the mean excess of temperature over the surroundings.

SPECIFIC HEAT CAPACITY OF A LIQUID BY THE METHOD OF COOLING

An empty copper calorimeter with the stirrer is weighed. It is then filled almost fully with water. It is heated to about 80°C (353 K) in a water bath without altering the nature of outer surface during the experiment. It is then removed, dried and then suspended in a constant temperature enclosure. The water is kept stirred all the time. The temperatures are recorded at regular intervals till the water cools to about 50°C (323 K). The calorimeter with water is now weighed.

The calorimeter is then emptied, dried and filled with an equal volume of liquid whose specific heat capacity is to be found and the experiment is repeated. The calorimeter with liquid is weighed. The cooling curves for water and liquid are drawn on the same graph sheet. From the graph, the time taken for water and liquid to cool through the same range of temperature is noted. The readings are recorded as given below.

Mass of empty calorimeter and stirrer	$= m_1\ kg$
Mass of calorimeter + stirrer + water	$= m_2\ kg$
Mass of calorimeter + stirrer + liquid	$= m_3\ kg$
Mass of water taken	$= (m_2 - m_1)\ kg$
Mass of liquid taken	$= (m_3 - m_1)\ kg$

Since the calorimeter is filled with the same volume of water and liquid the area of the cooling surface is the same in both cases. The same range of temperature is taken for both. Therefore the difference in temperature between the calorimeter and the surrounding is the same. Therefore the rate of loss of heat must be the same according to the Newton's law of cooling.

Let t_1 be the time taken for water to cool from θ_1°C to θ_2°C.

Let t_2 be the time taken for liquid to cool from θ_1°C to θ_2°C.

The rate of loss of heat energy by calorimeter and water at the mean temperature $\left(\frac{\theta_1 + \theta_2}{2}\right)$°C is $= \frac{(m_1 s + MS)(\theta_1 - \theta_2)}{t_1}$

where 's' is the specific heat capacity of the material of the calorimeter and 'S' the specific heat capacity of water and $M = (m_2 - m_1)$ is the mass of water taken.

The rate of loss of heat by calorimeter and liquid at the mean temperature for the same range $(\theta_1 - \theta_2) = \dfrac{(m_1 s + mX)\,(\theta_1 - \theta_2)}{t_2}$

where 'X' is the specific heat capacity of the liquid that is to be found out and $m = (m_3 - m_1)$, the mass of liquid taken.

Since the rate of loss of heat is the same in both cases

$$\frac{(m_1 s + MS)\,(\theta_1 - \theta_2)}{t_1} = \frac{(m_1 s + mX)\,(\theta_1 - \theta_2)}{t_2}$$

$$\text{i.e.}\quad \frac{(m_1 s + MS)}{t_1} = \frac{m_1 s + mX}{t_2}$$

from which 'X' is calculated.

Thus, using Newton's law, the specifiic heat capacity of a liquid can be determined.

2 3. 4 SPECIFIC HEAT CAPACITIES OF A GAS

A gas is characterised by its pressure and volume. According to Boyle's law when the temperature remains constant, the pressure of a given mass of a gas varies inversely as the volume. From Charles law we know that the volume of a gas at constant pressure varies directly with absolute temperature or the pressure of a given mass of a gas at constant volume varies directly with absolute temperature. This shows that gases are highly compressible and therefore changes of temperature cause enormous changes in both the pressure and the volume of a gas. Since the gas can expand with volume remaining constant or with pressure remaining constant, we have two types of expansions and hence two specific heats.

SPECIFIC HEAT CAPACITY AT CONSTANT VOLUME(C_v)

The amount of work that is to be done on one kilogram of a gas at constant volume for a rise of temperature is known as the specific heat capacity of the gas at constant volume (C_V). In this case, the work done is utilised in increasing the internal energy of the gas in whatever manner the energy may be distributed among the molecules and the atoms of the gas.

SPECIFIC HEAT CAPACITY AT CONSTANT PRESSURE(C_p)

The amount of work that has to be done on one kilogram of the gas at constant pressure for a rise of temperature of 1 K is known as the specific heat capacity of the gas at constant pressure (C_p). In this case since the pressure is kept constant, there is an increase in the volume of the gas as the temperature rises. The gas is allowed to expand against external pressure. This expansion of the gas causes external work to be done by the gas and this energy expansion has to be taken from the heat supplied to the gas. In addition to this, there is an increase in internal energy also. Hence C_p is greater than C_v.

2. 3. 5. MEYER'S RELATION

The specific heat capacity at constant pressure 'C_p' and the specific heat capacity at constant volume C_v are related by the following relation:

$(C_p - C_v) = R$ where 'R' is the universal gas constant. This is known as Meyer's relation connecting the two specific heat capacities of a gas.

Exercise 2. 3

1. *Define the specific heat capacity of a substance. What is the unit in which it is measured?*
2. *Explain the thermal capacity and water equivalent of a substance.*
3. *State and explain Newton's law of cooling.*
4. *How is the specific heat capacity of a liquid determined experimentally by the method of cooling?*
5. *Explain why a gas has got two specific heat capacities.*
6. *Define the two specific heat capacities of a gas What is the relation connecting them?*

2. 4. CHANGE OF STATE

2.4.1 EVAPORATION

It is a common phenomenon to find that a liquid exposed to atmosphere slowly disappears on account of evaporation. Liquids such as alcohol, ether etc, evaporate quickly whereas liquids like water, oil etc., take a long time to evaporate. Evaporation is therefore the process by which a liquid is slowly or quickly converted into its vapour at normal temperatures (below the boiling point). Evaporation takes place from the surface of the liquid exposed.

According to the kinetic theory of matter, a liquid consists of molecules which are in constant random motion. The velocities of these molecules are quite large. Some of these molecules near the surface of the liquid having velocities large enough to escape from the liquid, go out of the liquid into the space in the form of vapour. Similarly, some of the vapour molecules enter the liquid also. But the number of molecules leaving the liquid is usually more than the number returning into the liquid. (These numbers will be equal if the vapour is saturated.) Therefore the volume of the liquid gradually decreases.

2. 4. 2. VAPOUR PRESSURE

If the liquid is kept in an enclosed space and if the number of molecules leaving the surface is equal to the number of molecules returning to the liquid, the vapour formed is said to be *saturated*. The pressure due to this number of molecules of the liquid is called *saturated vapour pressure*. If sufficient liquid is not kept in the enclosed space, all the liquid would have evaporated forming unsaturated vapour and the pressure of the vapour inside the enclosed space is called unsaturated vapour pressure.

Consider three barometers in a wide vessel containing mercury. The space above mercury in a barometer will be vacuum. The height of mercury in all the three tubes would be the same. Let it be H. The atmospheric pressure is equal to the height 'H' of mercury.

In one barometer, introduce about 0·5 cc of alcohol by means of a bent pippette. In the second barometer introduce about 10 cc of alcohol. Alcohol goes to the top of the mercury column and evaporates. No liquid is introduced in the third barometer. The height of mercury in the third barometer is H (Fig.2-10).

There is some alcohol left over the mercury in the second barometer. This means that the vapour is saturated. There is no alcohol left over mercury in the first barometer. This means that the vapour is unsaturated.

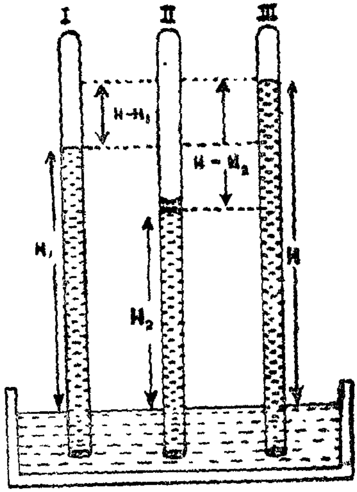


Fig. 2.10

Saturated and unsaturated vapour pressures

Let the height of mercury in the first barometer be H_1 and that in the second be H_2 and that in the third be H. The pressure of unsaturated vapour in the first barometer is $(H—H_1)$ The pressure of saturated vapour in the second barometer is $(H—H_2)$.

$(H—H_2)$ is greater than $(H—H_1)$. This indicates that the pressure of saturated vapour is greater than that of the unsaturated vapour.

The saturated vapour pressure of a liquid increases with increase in temperature Since the velocity of the molecules increases with rise of temperature, a large number of molecules leave the surface of the liquid forming vapour at higher tempera ture In equilibrium condition a large number of molecules will be found in the enclosed space.

2.4.3 VAPOUR PRESSURE AND BOILING POINT

When a liquid in an enclosed space is slowly heated and as the temperature of the liquid increases, a large number of molecules get out of the free surface of the liquid forming the vapour So the saturated vapour pressure of the liquid also increases After some time vapour bubbles begin to form within the liquid itself and move up to the surface. Soon the entire liquid attains the state of boiling at a definite temperature called "Boiling Point" and the liquid gets gradually converted into vapour The constant temperature at which the change of state from liquid to vapour takes place is known as the **boiling point of the liquid**. The supply of heat is not able to increase the temperature of the liquid but converts the liquid into vapour. Thus the heat supplied to every kilogram of the liquid is hidden in the vapour and is known as *latent heat of vapourization*. If each kilogram of the vapour is allowed to condense into liquid, it gives out the same amount of heat and it is termed as *latent heat of condensation*. When a liquid begins to boil, the saturated vapour pressure of the liquid becomes equal to the external pressure acting on it. So the *boiling point of a liquid can be defined as the temperature at which its saturated vapour pressure becomes equal to the external pressure acting on it.*

2.4.4. DETERMINATION OF THE BOILING POINT OF A LIQUID BY THE J-TUBE METHOD

A 'J' tube with a longer limb of height 20 cm and a shorter limb of height 5 cm, closed at the top is taken (Fig. 2-11). It is filled with mercury such that the whole of the shorter limb and a portion of the longer limb to a height of about 2 cms is filled with mercury. A few drops of a liquid whose boiling point is to be found is introduced into the shorter limb. It rises to the top and a portion of it is vapourised and the rest of it remains on the surface of mercury. The 'J' tube is attached to a scale and placed in a water bath and heated. As the temperature rises, the mercury is pushed down by the liquid vapour in the shorter limb. At a temperature θ_1 °C the levels of mercury in the limbs become equal. The temperature is noted. The heating is continued till there is a difference of about 1 cm between the levels of mercury in both the limbs. The bath is then allowed to cool.

As the temperature falls, at another temperature θ_2 °C once again the mercury levels in both the limbs becomes equal.

The average temperature $\left(\frac{\theta_1 + \theta_2}{2}\right)$ is the temperature at which the saturated vapour pressure of the liqu becomes equal to the external pressure. By this experiment the boiling point of a liquid which is available in small quantities can be ascertained.

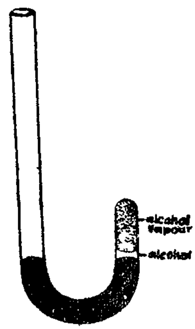


Fig. 2-11 Boiling Point by using a J-tube

2. 4. 5. ELEVATION OF BOILING POINT WITH INCREASE OF PRESSURE

The boiling point of water is 100°C at normal pressure. If the pressure is increased by keeping the water in an enclosed vessel and not allowing the vapour to escape, the pressure increases and thereby the boiling point of water rises when heated.

At higher altitudes like those of mountain tops, water starts boiling even before it reaches 100°C (373*K*) due to lower atmospheric pres sure prevailing there. Therefore cooking becomes difficult. Similarly when pressure increases there is an increase in the boiling point. This principle is used in pressure cookers. In an ordinary cooker the temperature will not go above the steam point (373 *K*) and hence the foodstuff will be maintained at this temperature and it takes a long time for the the substance to be cooked In a pressure cooker the increased pressure. increases the boiling point so that vegetables and other foodstuffs are raised to a much higher temperature than the normal boiling point and hence the substance is cooked quicker than in an ordinary cooker.

Exercise 2. 4

1. *Explain the phenomenon of evaporation.*
2. *What is meant by saturated vapour and saturated vapour pressure?*

3. *Describe an experiment to demonstrate that the saturated vapour pressure is the maximum pressure that could be exerted by a vapour at a particular temperature.*

4. *What is 'boiling point' of a liquid? How does the boiling point depend on the vapour pressure?*

5. *Deseribe an experimental method of determing the boiling point of a liquid, using the property of saturated vapour.*

6. *Explain the working of a pressure eooker.*

3. ELECTRICITY AND MAGNETISM

3.1. ELECTROSTATICS

3.1.1. ELECTRIC CHARGE

Ancient Greeks, as long ago as 600 B.C. observed that amber (a kind of hardened resin), when rubbed with fur, attracted very light objects like bits of straw. Today it is known that any solid, rubbed with a suitable material under suitable conditions will attract light objects to a certain extent. A familiar example is a rubber or ebonite comb after it has been used to comb dry hair, attracting bits of paper. This property is described by saying that the solids have been electrified or have acquired electric charge. The terms 'electric', 'electrified' etc., are derived from 'electron', the Greek word for amber.

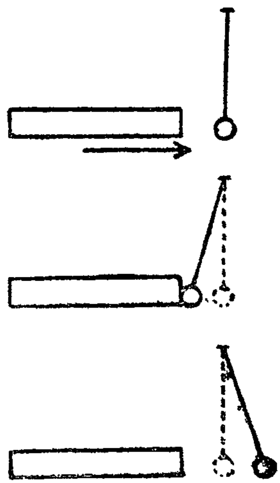


Fig. 3-1
Electrified rod attracts and then repels a pith ball

An ebonite rod can easily be electrified by rubbing it with a piece of fur and a glass rod by rubbing it with a piece of silk and both rods will attract small bits of paper or suspended pith balls. It will also be noticed that if the attracted bits of paper or pith ball make contact with the rods, they will cling to the rods momentarily and then will be repelled away. The reason for this will be evident later.

Now, if an electrified ebonite rod is brought near another electrified ebonite rod suspended by a silk thread, they are found to repel each other. Similarly two electrified glass rods will repel each other, but an electrified glass rod will attract an electrified ebonite rod. This leads to the idea that there are two kinds of electric charges and that like charges repel each other and unlike charges attract each other. The charge on the electrified ebonite rod is referred to as *negative charge* and that on the electrified glass rod as *positive charge*. If we examine the fur with which the ebonite rod was rubbed, it will be found to have an amount of positive charge, equal to the amount of negative charge on the ebonite rod. Similarly the silk with which the glass rod was rubbed will have an amount of negative charge. equal to the amount of positive charge on the glass rod. This can easily be explained, if we look at the structure of atoms. Every atom has a positively charged nucleus, consisting either of a single positively charged proton as in the case of hydrogen, or of protons and uncharged neutrons, and negatively charged electrons which orbit round the nucleus. The amount of positive charge on a proton is exactly equal to the amount of negative charge on an electron and in any atom, the number of protons in the nucleus is exactly equal to the number of orbiting electrons. Hence an atom as a whole is always neutral. When a glass rod is rubbed with silk, some of the electrons from the glass rod are transferred to the silk, so that the glass-rod now has a deficiency of electrons or negative charge, which makes it positively charged. The silk on the other hand has now an excess of electrons and this makes it negatively charged. Similarly when an ebonite rod is rubbed with fur, the ebonite rod gets some extra electrons from the fur, thus getting negatively charged and the fur which has lost some electrons becomes positively charged. So when bodies are electrified by contact—rubbing is only to facilitate contact—there is no "creation" of electric charge, but only a redistribution of the negatively charged

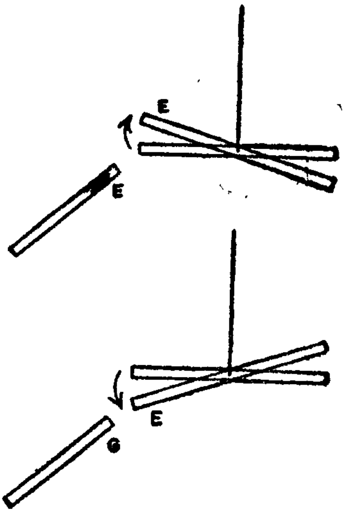


Fig. 3-2
Repulsion and attraction between electrified rods

electrons. It should be remembered that it is always the light negatively charged electrons which are transferred from one body to another and never the heavier positively charged protons. The amount of charge on a single electron or a single proton is the smallest amount found on any particle This quantity is denoted by '*e*' and has a value equal to $1{\cdot}6029 \times 10^{-19}$ coulombs, the coulomb being the unit of charge in the *S. I.* system of units. Since eletrification is due to redistribution of electrons, a body can have charge only in multiples of '*e*'. Hence the electric charge on a body cannot be increased continuously, but only in steps of '*e*'. This phenomenon is referred to as 'quantisation' in Modern Physics, the quantum (or small) unit in this case being the charge '*e*' on an electron.

3.1.2. CONDUCTORS AND INSULATORS

We have seen that an ebonite rod held in hand and rubbed with fur attracts bits of paper. On the other hand, a brass rod held in hand and rubbed with fur will not show this property. But if the brass rod is provided with a glass handle and then it is rubbed with fur, holding it by the handle, it will attract bits of paper. But if the rod is now touched by hand or electrically connected to the earth, it will immediately lose its attracting property. This is because, the charge acquired by the brass rod is conducted away to the earth, through the hand. In the case of an ebonite rod the charge acquired by it remains stationary on it. Substances like brass which allow eletric charge to move through them are known as **conductors** and substances like ebonite, glass etc, which do not allow charge to move through them are known as **insulators**. All metals and their alloys are good conductors. In the atoms of a metal, the orbital electrons (valence electrons) are not very rigidly bound to their nuclei and they are capable of drifting around, whereas it is not so in the case of insulators.

3.1.3. ELECTRIC INDUCTION

If a negatively charged body *A* is brought near the end *B* of a conductor *BC* mounted on an insulating stand, it will be found that the end *B* acquires positive charge and the end *C*, negative charge. It can be shown that the amounts of positive and negative charges at *B* and *C* are exactly equal in magnitude and equal to the amount of charge on A. If A is now moved away, the positive and negative charges will redistribute in BC and BC will be neutral again. But if in the presence of A, the end C is touched, or connected to earth, the negative charge will escape to earth and if now A is removed, the positive charge at B will redistribute itself uniformly throughout BC, thus making BC a positively charged body, having the same amount of charge on it as on A. This method of charging a body is known as *charging by electrostatic induction*.

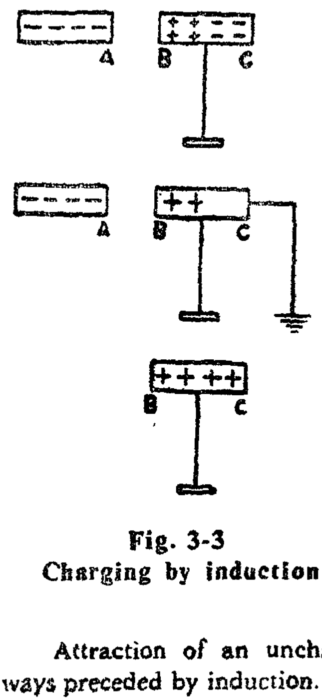


Fig. 3-3
Charging by induction

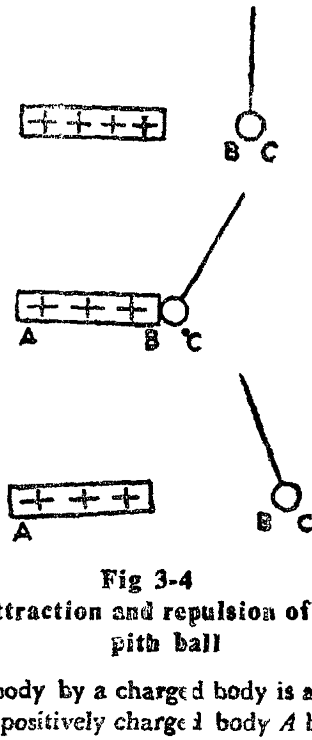


Fig 3-4
Attraction and repulsion of a pith ball

Attraction of an uncharged body by a charged body is always preceded by induction. Let a positively charged body A be brought near a suspended pith ball BC. The end B of the ball acquires induced negative charge and the end C, induced positive charge. Since A is nearer to B than to C, the force of attraction between A and B will be greater than the force of repulsion between A and C. Hence the pith ball will be attracted towards A, but as soon as it makes contact with A, some of the electrons from B are transferred to A, neutralising part of the positive charge on A and leaving a net positive charge on the pith ball. Now since A and the pith ball have both positive charge the pith ball will be repelled by A

3.1.4. FORCE BETWEEN CHARGES

A French scientist by name Charles Coulomb— the unit of electric charge is called coulomb in his honour—formulated the law of forces between point charges. He stated that **the force, whether attraction or repulsion, is directly proportional to the product of the magnitudes of the charges and inversely proportional to the square of the distance between them and is along the line joining the charges.** i.e. the force F between two charges q_1 and q_2 separated by a distance r is given by $F \propto \frac{q_1 q_2}{r^2}$ or $F = K\frac{q_1 q_2}{r^2}$ where K is constant of proportionality. The value of K depends on the choice of units of F, r and q. Since the units of F and r are already fixed as newton and metre in the $S.I.$ system of units the value of K will depend only on the choice of unit of q. The unit of charge, coulomb in the rationalised $M.K.S.$ system and $S.I.$ system is fixed from considerations of force between electric currents and hence the value of K is found experimentally. But instead of K, the constant usually used is $\frac{1}{4\pi 3_0}$ for free space (or vacuum) and Coulomb's law is written as

$$F = \frac{1}{4\pi\varepsilon_0}\frac{q_1 q_2}{r^2} \quad \dots \quad (3\text{-}1)$$

ε_0 is a constant known as the permitivity of free space. In any other medium,

$$F = \frac{1}{4\pi\varepsilon}\frac{q_1 q_2}{r^2} \quad \dots \quad (3\text{-}1)$$

where ε $(=\varepsilon_0\varepsilon_r)$ is the absolute permitivity of that particular medium where ε_r is the relative permitivity or specific inductive capacity of the medium. ε for air is very nearly equal to ε_0

The value of $\frac{1}{4\pi\varepsilon_0}$ works out to $9 \times 10^9 \frac{\text{newton-metre}^2}{\text{coulomb}^2}$

Hence the force between two charges, each of 1 coulomb separated by 1 metre in vacuum is 9×10^9 newtons. Compared to gravitational attraction between bodies of comparable size, this force is very very large—about 10^{36} times and hence when calculating electrical force between charges, the gravitational force which exists between all bodies can be completely neglected. It should also be noted that the gravitational force which exists between all bodies can be completely neglected. It should also be noted that the gravitational force is always one of attraction, since there is only one kind of mass unlike two kinds of electrical charges.

3.1.5. ELECTRIC FIELD

The fact that when two electric charges are kept near each other, there is a force acting on each other even though there is no actual contact between them, leads to the idea that the presence of a charge modifies the space surrounding it. This modified space is referred to as an *electric field.* (compare gravitational field). The **intensity** of the electric field at any point is a vector quantity and is measured in magnitude and direction by the force experienced by a unit positive charge kept at that point A unit charge placed at a point distant r metres in vacuum from a charge q coulombs will experience a force

$$\frac{1}{4\pi\varepsilon_0} \cdot \frac{q \times 1}{r^2} \text{ newtons}$$

and hence the intensity E of the field at a distance r from charge q is given by

$$E = \frac{1}{4\pi\varepsilon_0} \cdot \frac{q}{r^2} \quad \text{newtons/coulomb} \qquad \dots \quad (3.3)$$

In a field of intensity E newtons/coulomb a charge Q coulomb will experience a force $F = EQ$ newtons.

3.1.6. ELECTRIC LINES OF FORCE

It was Michael Faraday who introduced the concept of lines of force. *A line of force in an electic field is an imaginary line drawn in such a way that the direction of the tangent drawn to it at any point gives the direction of the field at that point.* The lines of force are generally curved.

Even though a line of force can be drawn through every point in an electric field, usually the number of lines are restricted, so that the number of lines crossing unit area taken at right angles to them, gives numerically, the intensity of the electric field

In a region where the intensity is large, the lines will be closer together than in a region, where the intensity is small. If E is the uniform electrical intensity in a direction perpendicular to A, the total number of lines of force passing through A will be $E A$.

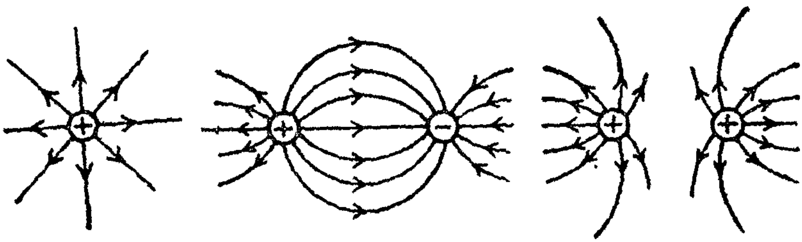


Fig. 3-5
Lines of force

To calculate the number of lines of force emanating from a point charge q, consider a sphere of radius r concentric with the charge. The total number of lines of force passing over the whole surface area of the sphere $= E.\ A.$

$$= \frac{1}{4\pi\varepsilon_0} \cdot \frac{q}{r^2} \quad 4\pi r^2$$

$$= q/\varepsilon_0$$

i.e. the total number of lines of force emanating from the point charge kept in vacuum is equal to q/ε_0. Lines of force are always continuous lines starting from a positive charge and ending on a negative charge.

3.1.7. ELECTRIC POTENTIAL

If a unit positive charge is kept near a positively charged body, it experiences a force of repulsion which measures the electric intensity at that point. If the unit positive charge is to be moved against this force of repulsion, work will have to be done and this work will be stored in the system as potential energy. (Compare with the potential energy gained when a body is moved up from the surface of the earth against the gravitational attraction.) If the unit positive charge were perfectly free, it would move down the potential gradient, i.e. away from the positive charge due to which the potential exists, i.e. a positive charge will always move from a point of higher potential to a point of lower potential This is analogous to heat flowing from a point at a higer temperature to a point at a lower temperature and water flowing from a higer level to a lower level.

The electrical potential difference between two points A and B is measured by the work done in carrying unit charge from B to A. The electrical field as well as the potential at a point at an infinite distance from the charge can be taken to be zero and so the potential at the point A will be the work done in taking a unit positive charge from infinity upto A. Let (V_1-V_2) be the small potential difference between two points separated by a small distance (x_1-x_2) the electrical intensity E over the small distance (x_1-x_2) remaining constant. The force on unit positive charge is E and the work done in moving it through $(x_1-x_2) = -E(x_1-x_2)$ the negative sign showing that the displacement (x_1-x_2) is opposite to the direction of E. Hence

$$V_1-V_2 = -E(x_1-x_2)$$

$$\text{or } E = -\frac{V_1-V_2}{x_1-x_2} \quad \ldots\ldots \quad (3.4)$$

i.e. electric intensity at any point is the negative gradient of the potential at that point.

The S.I. unit of potential is the **volt** *which is defined as the difference in potential between two points so that one joule of work is done in carrying 1 coulomb of positive charge from one point to the other.*

$$\text{volt} = \frac{\text{joule}}{\text{coulomb}}$$

The electrical intensity $E = \dfrac{\text{volt}}{\text{metre}}$ which can be shown to be the same as newtons/coulomb as defined earlier.

$$E = \frac{\text{volt}}{\text{metre}} = \frac{\text{joule}}{\text{coulomb} \times \text{metre}} = \frac{\text{newton} \times \text{metre}}{\text{coulomb} \times \text{metre}} = \frac{\text{newton}}{\text{coulomb}}$$

The electrical potential is a scalar quantity i.e. it has only magnitude and no direction whereas *the electrical intensity is a vector quantity having both magnitude and direction.* The potential difference between two points depend only on the position of the points and not on the path between them. The work done in taking a unit charge from A to B will be *the same,* whether the path taken is ACB or ADB.

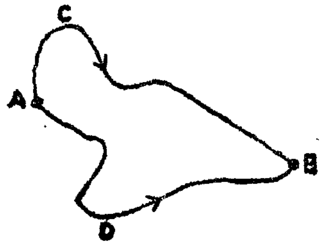


Fig. 3-6
P.D. between A and B independent of path

The potential at a point distant r metres from a positive point charge q coulombs in vacuum is

$$V = \frac{1}{4\pi\varepsilon_0} \frac{q}{r} \text{ volts} \quad \dots \quad (3.5)$$

3.1.8. EQUIPOTENTIAL LINES AND SURFACES

An equipotential line is a line joining points where the potentials are the same. Similarly an equipotential surface is one at all points of which the potential has the same value.

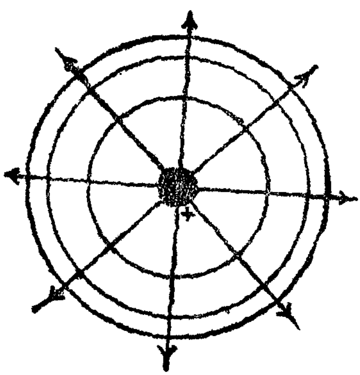


Fig. 3-7
Lines of force (arrows) and equipotential lines around a positive charge

Since the potential at every point on the equipotential surface is the same no work will be done in carrying a charge from one point to the other on the surface. Hence the direction of the electrical field will be normal to an equipotential surface at every point or *the lines of force* will be normal to an equipotential surface. The surfaces of all charged conductors are equipotential surfaces. The space inside a charged hollow sphere is an equipotential volume and all *surfaces* concentric with *the* surface of the sphere will be equipotential surfaces.

Exercise 3.1.

1. *How many kinds of charges are there?*
2. *What is the nature of force between charges ?*
3. *Is a normal atom neutral because of the absence of charges ? If not, what is the reason ?*
4. *After electrons are removed from a neutral body, what would be the nature of the body ?*
5. *What happens when a glass rod is rubbed with silk ?*
6. *What happens when an ebonite rod is rubbed with fur ?*
7. *When one body is rubbed with another, are positive charges ever transferred from one to the other ?*
8. *What is the difference between a conductor and an insulator ?*
9. *What happens when a negatively charged rod is brought near an uncharged, insulated metal sphere ?*
10. *What happens if in the presence of the negatively charged rod,the metal sphere is earthed ?*
11. *What happens when an insulated uncharged metal sphere is touched by negatively charged rod ?*
12. *Two positive charges of magnitudes 5 coulombs and 3 coulombs respectively are kept at a distance of 3 metres from each other in vacuum. What is the force between them ?*
13. *What is the electrical field at a point distant 2 metres in air from a point charge of 5 coulombs ? Assume* ε *for air to be equal to* ε_0.
14. *What is the force on a charge of 2 coulombs in an electric field of intensity 4 newtons/coulomb ?*
15. *How is the electrical potential difference between two points defined ?*
16. *How is the electrical potential at a point defined ?*
17. *Find the work done in taking a charge of 2 coulombs between two points which differ in potential by 2 volts.*
18. *What is the potential at a point distant 3 metres in air from a positive point charge 5 coulombs ?*
19. *What is the potential diffrence between 2 points distant 2 metres and 3 metres respectively in air from a point charge 6 coulombs ?*
20. *What is the work done in taking a charge of 3 coulombs from a point 4 metres distant to a point 2 metres distant in vacuum from a point charge of 5 coulombs ?*

3.2 CURRENT ELECTRICITY

3.2.1. ELECTRIC CURRENT

As we have seen earlier, in certain substances, charged atomic particles can move freely through the material. In such a substance, an electric charge can be transferred from one point to another by a *general drift of the* charged particles within it. Such a movement of an electric charge is called an electric current and the current is equal to the quantity of charge passing a given point in *unit time*.

As mentioned earlier, materials through which charged particles can move freely or through which an electric current will pass are *known* as conductors. Metals with free valence electrons are good conductors. Another important type *of conductors* is *the class of* solutions known as electrolytes. When certain chemical *compounds go into a* solution, *they* break up into positively and negatively charged parts or ions. These move independently in the solution and are therefore available to conduct electric current through it. When there is a potential difference between two points along a conductor, charges will flow from one point to the other, immediately equalising the potential. This will constitute a transient current. If a steady current is to be maintained between two points along a conductor, a steady potential difference has to be maintained between these two points. This is done with the help of devices such as an electric battery or a dynamo. The positive direction of the current is taken as the direction in which positive charge would flow. But since in a metallic conductor it is always the negative electrons which move, the conventional direction of current is opposite to that of the flow of electrons. In an electrolyte, both positive and negative ions move, but in opposite directions, thus contributing to the current in the same direction.

Although we cannot see the movement of electric charges when a current flows, the presence of the current is detected usually by one of the three effects (i) heating effect (ii) chemical effect or (iii) *magnetic* effect.

3.2.2. MAGNETIC EFFECT OF A CURRENT

It is found that a magnetic field is produced around a conductor carrying current as shown by the deflection of a pivoted

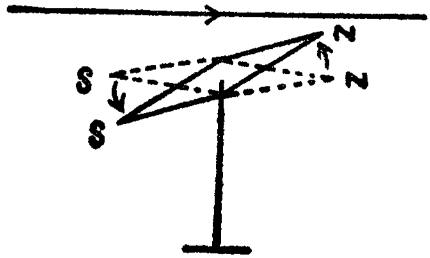


Fig 3-8
Magnetic needle NS deflected by a current

magnetic needle. The direction of the magnetic field is given by Ampere's swimming rule. Imagine a man swimming along the conductor in the direction of the current with his face towards the magnetic needle. The north pole of the needle is deflected towards his left.

Also a conductor carrying current experiences a force when placed in a magnetic field, except when the field is in the same direction as the current. Hence a conductor carrying

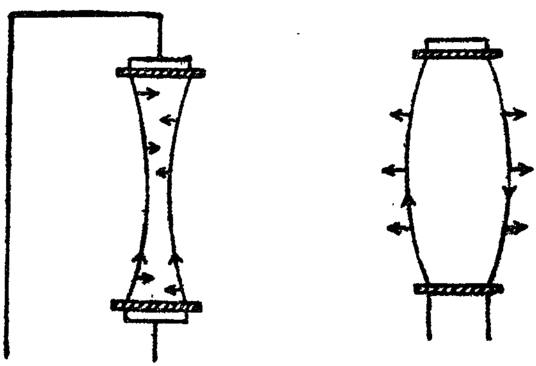


Fig. 3-9
Force between currents

current kept parallel to another conductor carrying current will experience a force. If a large current is passed through two flexible conductors, it will be seen that when the currents are in the same direction, there is a force of attraction between the conductors drawing them together and when the currents are in the opposite direction, there is a force of repulsion between them.

The *SI* unit of current, called the *ampere*, in honour of the French scientist who established the laws governing the magnetic forces between currents is based on the force between conductors carrying currents

The ampere is that steady current, which flowing in two infinitely long straight parallel conductors of negligible circular cross section, placed a metre apart in vacuum produces a force between them of 2×10^{-7} newtons per metre length of conductor.

The unit of charge is coulomb and is derived from that of the ampere.

The coulomb is that quantity of electric charge that passes a given point in a circuit when a current of one ampere flows for one second.

If a current of I ampere flows in a circuit for t seconds, the quantity Q of electric charge that passes is given by $Q = It$ coulombs.

The charge on one electron is $-1{\cdot}6029 \times 10^{-19}$ coulombs so that one coulomb is numerically the charge on nearly $6{\cdot}24 \times 10^{-18}$ electrons.

As stated earlier, potential difference between two points acts so as to drive an electric current from the point at higher potential to the point at the lower potential. It is equal to the work done or energy converted from electrical to other forms per unit charge passed.

The volt, named in honour of the Italian scientist Alessandro Volta, *is the potential difference between two points if the work done in carrying 1 coulomb of positive charge from one point to another is 1 joule.*

If a charge Q coulombs flows between two points which are at a potential difference of V volts, the electrical energy used is given by

$$W = Q\,V \text{ joules.}$$

or $\quad W = I\,t.\,V$ *joules, since* $Q = I\,t$

i.e. $\quad W = I\,V,\,t$ *joules* $\qquad (3.6)$

The unit of energy which is used in atomic and nuclear calculations is the **electron volt**. It is the quantity of energy gained by an electron in falling through a potential difference of 1 volt. Since the charge on one electron is $1{\cdot}6029 \times 10^{-19}$ coulombs, one electron volt (eV) $= 1{\cdot}6029 \times 10^{-19}$ joules.

3.2.3. OHM'S LAW

A current flows in a conductor, between two points when there is a potential difference between them. The relationship between the current and the potential difference is given by Ohm's law, which states that *a steady current flowing through a metallic conductor is proportional to the potential difference between its ends provided the temperature remains constant.*

i.e. $I \propto V$ or $V/I =$ constant for a given conductor at constant temperature. This constant, given by the ratio of potential difference to current is known as the resistance of the conductor.

i.e. $V/I = R$ when temperature is constant. The $S.I.$ unit of resistance is the ohm. *The* **ohm** *is the resistance of the conductor through which a steady current of* 1 *ampere passes when a potential difference of* 1 *volt exists across it.*

3.2.4. RESISTANCES IN SERIES

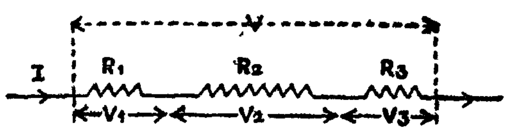


Fig. 3-10
Resistances in series

Let R_1, R_2, R_3 be three resistances, connected in series as shown in Fig. 3-10, so that the same current i flows through them. Then the combined effective resistance R is given by

$R = V/I$ where V is the total potential difference across the combination. V is the sum of the separate potential differences across the individual resistances.

$$V = V_1 + V_2 + V_3$$

$$\text{and } V_1 = IR_1, \quad V_2 = IR_2 \text{ and } V_3 = IR_3$$

$$V = IR = IR_1 + IR_2 + IR_3$$

$$\text{or } R = R_1 + R_2 + R_3 \quad \ldots \quad (3.7)$$

This relationship will hold good for any number of resistances connected in series.

3.2.5. RESISTANCES IN PARALLEL

In this arrangement as shown in Fig. 3-11, the same potential difference V exists across all the three resistances, R_1,

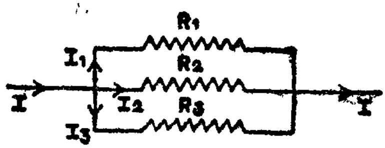


Fig. 3-11
Resistances in paralell

R_2, and R_3. The current I in the main circuit, divides into three portions, I_1, I_2, I_3 passing through R_1, R_2 and R_3

$$\text{and} \quad I = I_1 + I_2 + I_3.$$

Now $i_1 = V/R_1$, $I_2 = V/R_2$ and $I_3 = V/R_3$ and if R be the combined effective resistance,

$$R = \frac{V}{I} \text{ or } I = \frac{V}{R}$$

$$\text{Hence } \frac{V}{R} = \frac{V}{R_1} + \frac{V}{R_2} + \frac{V}{R_3}$$

$$\text{or } \frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} \quad \ldots \quad (3.8)$$

This relationship also will hold good for any number of resistances connected in parallel. If there are n resistances, each of them equal to R, the combined effective resistance will be R/n.

If there are only two resistances R_1 and R_2 connected in parallel

$$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2}$$

$$\text{or } R = \frac{R_1 R_2}{R_1 + R_2} \qquad (3.9)$$

and if I_1 and I_2 are the currents through R_1 and R_2 respectively,

$$I_1 R_1 = I_2 R_2 = V = IR = I \frac{R_1 R_2}{R_1 + R_2}$$

$$\text{Hence } I_1 = \left(\frac{R_2}{R_1 + R_2}\right) I \text{ and } I_2 = \left(\frac{R_1}{R_1 + R_2}\right) I \ldots \quad (3.10)$$

i. e. the current in either of the two branches depends on the resistance in the other.

3.2.6. RESISTIVITY OR SPECIFIC RESISTANCE

The resistance of a conductor at a given temperature depends on its length and cross sectional area and the material of which it is made. It is seen that the resistance of a uniform conductor is directly proportional to its length l, and inversely proportional to its area of cross section A, i.e. as the length of a conductor increases, its resistance increases, but as its thickness increases, its resistance will decrease.

$$R \propto \frac{l}{A} \text{ or } R = \rho . \frac{l}{A}$$

where ρ is a constant for the material of the conductor known as its resistivity or specific resistance. The value of ρ is given by

$$\rho = \frac{R \times A}{l}$$

and it follows from this equation, that the dimensions of ρ is resistance $\times$ length and hence its unit is ohm-metre in the $S.I.$ system.

Exercise 3.2

1. What is an electric current ?
2. What is the type of force between two parallel conductors carrying current in the same direction ?
3. Define the unit of current in the S.I
4. What is the conventional direction of current in a conductor ?
5. Do electrons in a conductor move in the direction of the conventional current ?
6. Define coulomb, the unit of charge in the S.I
7. Define volt, the unit of potential difference.
8. Calculate the amount of work done when a charge of 3 coulombs is passed between two points which differ in potential by 4 volts.
9. What is the amount of electrical energy expended when a current of 5 amperes passes for 2 hours through a conductor, the potential difference between the ends of which is 20 volts ?
10. What is an electron volt and what is the relationship between an electron volt and a joule ?
11. State Ohm's law.
12. When two resistances are connected in parallel, will the equivalent resistance be greater or smaller than when they are connected in series ?
13. Two resistances R_1 and R_2 when connected in series have an equivalent resistance of 10 ohms and when connected in parallel, the equivalent resistance is 2·4 ohms. Find R_1 and R_2.
14. A resistance is marked as 2 ohms has actually a value of 2·05 ohms. Calculate the resistance to be connected in parallel to it so as to bring the effective resistance to the marked value.
15. The resistance of a conductor of length 2 metres and radius 0·2 mm is 0·6 ohms. Find the specific resistance of the material of the conductor.

3.3. MAGNETISM

3.3.1. INTRODUCTION

It was known for a very long time that pieces of iron ore found in Magnesia, which came to be known as natural magnets could attract unmagnetised iron and also set themselves approximately in the north-south direction, when freely suspended about a vertical axis. But it was not until early nineteenth century that it was discovered that there was any connection between electrical and magnetic phenomena. The Danish scientist, Hans Christein Oersted in 1819 showed that a pivoted magnetic needle deflected in the presence of a wire carrying electric current i.e magnetic effects could be produced by an electric current or moving electric charge. Now it is established, that all magnetic forces arise because of charged particles in motion. Since all atoms, contain circulating electrons, they are inherently magnetic. To a large extent, the orbits of electrons, form equal and opposite pairs, cancelling out their magnetic effect, but with many atoms, there is some residual magnetic effect, caused usually by the spin of the electrons. (The electrons in an atom, apart from orbiting round the nucleus, also spin about their own axes.) With most substances, the effects observed are very small, but in substances like iron, nickel, cobalt etc., the effects are very intense and these substances are called *ferro magnetic* materials. It is from these ferro magnetic materials that artificial magnets are made.

3.3.2. MAGNETIC FIELD AND MAGNETIC INDUCTION

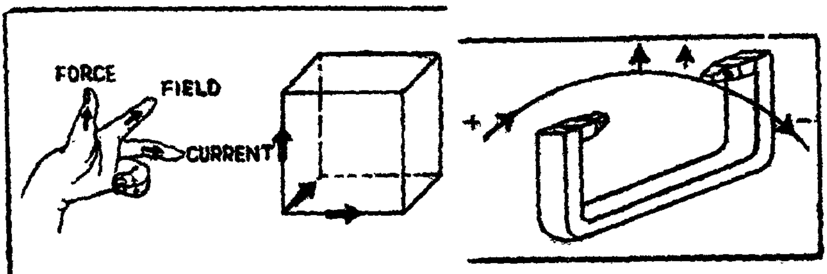
The space around a magnet or a current carrying conductor is known as a **magnetic field** just as the space around an electric charge is known as an electric field. And a magnet or a current carrying conductor will experience a force in an electric field. The electric intensity E at a point in space is measured in terms of the force F acting on a positive charge Q kept at that point

or $E = F/Q$

The corresponding quantity that describes a magnetic field is called its magnetic induction and is represented by the letter B. One way of defining magnetic induction at a point is in terms of the magnetic force F exerted on a small length of wire carrying an electric current. If the length of the wire is l and the value of the current I, the magnitude of B is given by the relation

$$B = \frac{F}{i\,l\,\sin\theta}$$

where θ is the angle between the directions of I and B. It follows from Eqn. (3-11) that (i) when the force is zero, θ is zero or I and B are in the same direction and (ii) when the force is maximum, I and B are at right angles to each other. It has also been found that the direction of force is at right angles to the directions of



(a) (b) (c)

Fig. 3-12
(a), (b)-Left hand rule (c) force on a current in a magnetic field

both I and B. This is given by Fleming's *left hand rule*. If the thumb and the first two fingers of the left hand are held mutually at right angles and the first finger is pointed in the direction of the field while the second finger is in the direction of the current, then the direction of the thumb gives the direction of the force.

From Eqn (3-11) it follows that *the magnetic induction B at a point in a magnetic field is the force per unit length on a conductor carrying unit current, lying at right angles to the magnetic field at that point.*

In the *S.I.* system since force is measured in newtons, current in amperes and length in metres, the unit of magnetic induction is newton/ampere-metre. This unit is now called a Tesla.

Just as electric intensity at any point is represented by the number of lines of force per unit area, magnetic induction at any point can also be represented as magnetic lines of induction per unit area, the area being perpendicular to the lines of induction. The total number of lines of induction passing across any area normal to it is known as the total magnetic flux linked with that area and hence magnetic induction B is also known as magnetic flux per unit area or **magnetic flux density**. The unit of flux is the weber and hence the unit of magnetic induction or flux density B is also given as weber/metre^2. Dimensionally, this can be shown to be the same as newton/ampere-metre.

Like the electric intensity or field strength E, magnetic induction B is also a vector quantity and the direction of B at any point in a magnetic field is given by the direction of the tangent drawn to the line of induction at that point.

Another vector qnantity associated with a magnetic field is the **magnetic field** intensity or magentising force denoted by the letter H. H at a point in free space is related to the value of B.

$$\therefore \quad H = B/\mu_0 \tag{3.12}$$

where μ_0 is a constant known as the absolute *permeability* of free space and has a value of $4\pi \times 10^{-7}$ units. H at a point in any other medium is given by the relation

$$H = B/\mu$$

where μ is the absolute permeability of that medium. $\mu = \mu_0 . \mu_r$ where μ_r is the relative permeability of the medium. For air, μ is very nearly equal to 1 so that μ can be taken to almost equal to μ_0. The unit of H, the magnetic field intensity is given as ampere-turn/metre. The reason for this will be evident later.

The magnetic field intensity or magnetising force H is also represented by lines drawn in the field. These lines are called magnetic *lines of force* and the number of lines of force per unit area represents the magnetic field intensity. In a given area, the number of lines of force will be $1/\mu$ times the number of lines of induction in the same area where μ is the absolute permeability of the medium. The direction of the magnetic field at any point is given by the direction of tangent drawn to the line of force at that point.

3.3.3. MAGNETIC POLES

In a magnet, the magnetic properties appear to be concentrated in certain regions only. These regions are called the **poles of the magnet**. Magnetic poles are of two kinds. In a magnetic field, the force experienced by the two kinds of poles, kept in turn at the same point will be in opposite directions. It is also seen that like poles repel each other and unlike poles attract each other. In any magnet, the poles always occur as equal and opposite pairs. Experiments are being conducted to see whether single poles exist, but so far, their existence has not been conclusively proved.

In a uniform magnetic field, the two poles of a magnet will experience equal and opposite forces in parallel directions and these will constitute a couple which will tend to rotate the magnet, until its axis is parallel to the direction of the field. In the earth's field, which can be considered to be uniform over a small area a pivoted magnetic needle will come to rest in the north-south direction and the pole turned towards the north is called the *north seeking pole or north pole* for short. Similarly the pole turned towards the south is called the *south seeking pole or south pole.*

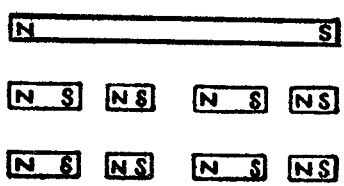


Fig. 3-13
Cutting a magnet produces new poles

When a piece of magnetised material is broken into two, fresh poles appear on either side of the break and this will continue however large number of bits the material is broken into. Even when the material is broken down into its constituent atoms, each atom can be assumed to be individually magnetised, with north and south poles. As was stated earlier, this is due to the moving electrons in the atom. A coil of wire carrying a current also acts as if it is a short magnet, with one face of the coil acting as a north pole and the other a south pole, depending on the direction of the current This again confirms the idea that all magnetic effects are due to electric currents.

3.3.4. FORCE BETWEEN POLES

Coulbmb's inverse square law states that the *force between poles is directly proportional to the product of pole strengths and inversely proportional to the square of the distance between them.*

i.e. $$F \propto \frac{m_1 m_2}{r^2}$$

where m_1 and m_2 are two poles separated by a distance r.

or $$F = C\frac{m_1 m_2}{r^2}$$

where C is a constant. In the *S.I* the value of C is taken as $\frac{\mu_0}{4\pi}$ for free space or vacuum

Hence $$F = \frac{\mu_0}{4\pi} \frac{m_1 m_2}{r^2} \qquad (3.13)$$

where μ_0 is the permeability for free space. μ_0 has a value $4\pi \times 10^{-7}$ henry/metre. (Henry is the unit of inductance and will be defined later.) The unit of μ_0 is also given as weber ampere-metre and it can be shown that dimensionally both units are the same.

In a medium other than free space, the force between two poles will be given as

$$F = \frac{\mu}{4\pi} \frac{m_1 m_2}{r^2} \qquad (3.13a)$$

where μ is the absolute permeability of that medium. As was given earlier, μ for air is very nearly equal to μ_0.

On the basis of Eqn (3.13), unit pole can be defined.

It is the strength of the pole which when kept at a distance of 1 metre from an equal and similar pole in vacuum repels it with a force of $\frac{\mu_0}{4\pi} \times 10^{-7}$ newtons. The unit of pole strength is called ampere-metre, It can also be defined as that pole strength which when placed in a field of unit induction experiences a force of one newton. Thus a pole of strength m ampere-metre placed in a magnetie field of induction B tesla will experience a force $F = mB$ newtons. (3.14)

If a unit pole is kept at a distance of r metres from a pole of m ampere-metre in vacuum, it will experience a force

$$F = \frac{\mu_0}{4\pi} \cdot \frac{m \times 1}{r^2} \text{ newtons.} \qquad (3.15)$$

Hence the magnetic induction B at a point distant r metres from a pole of strength m ampere-metre is given by

$$B = \frac{\mu_0}{4\pi} \cdot \frac{m}{r^2} \text{ tesla} \qquad (3.16)$$

The intensity of magnetic field (H) is defined as the ratio of the magnetic flux per unit area, B_0, in empty space to the permeability μ_0 of the space.

$$H = \frac{B_0}{\mu_0} \qquad (3.17)$$

It is expressed in *ampere turns per metre* (Am^{-1})

3.3.5. MOMENT OF A MAGNET

Let a bar magnet of length $2l$ and pole strength m for each pole be placed in a uniform magnetic field of induction B, making an angle θ with the direction of B (Fig. 3-14). Now the north pole and the south pole will each experience a force mB, the force on the north pole being in the direction of the field and the force on the south pole being in the opposite direction. These two forces will therefore constitute a couple and the moment of the couple will be $mB \times 2l \sin\theta$. If the magnet is kept at right angles to a field of unit induction, the moment will be equal to

$m \times 2l = 2\,ml$. This quantity $2ml = M$ given by the product of the pole strength and the length of the magnet is known

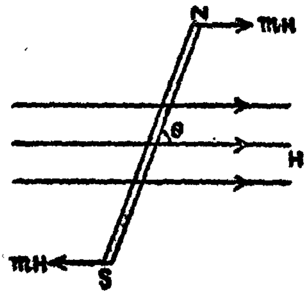


Fig. 3-14
Couple on a magnet in a magnetic field

as the moment of the magnet and is equal to the moment of the couple necessary to keep the magnet at right angles to a field of unit induction. The unit of magnetic moment is ampere metre² (A-m^2)

3.3.6. MAGNETIC INDUCTION AT A POINT ON THE AXIAL LINE OF A BAR MAGNET

Let NS be a bar magnet of length $2l$ metres, pole strength m Ampere-metres and moment M Ampere-meters². Let P be a point on the axial line distant d metres from the mid point O of the magnet. The magnetic induction at P due to the north pole at N

$$= \frac{\mu_0}{4\pi} \cdot \frac{m}{(d-l)^2} \text{ tesla in the direction } OP$$

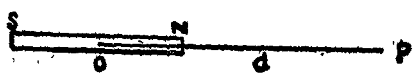


Fig. 3-15
Field along the axial line

The magnetic induction at P due to the south pole at S

$$= \frac{\mu_0}{4\pi} \cdot \frac{m}{(d+l)^2} \text{ tesla in the direction } PO$$

Resultant induction at P due to NS

$$B = \frac{\mu_0}{4\pi}\frac{m}{(d-l)^2} - \frac{\mu_0}{4\pi}\frac{m}{(d+l)^2}$$

$$= \frac{\mu_0 m}{4\pi}\left[\frac{4dl}{(d^2-l^2)^2}\right]$$

$$= \frac{\mu_0 \times 2ml \times 2d}{4\pi\,(d^2-l^2)^2}$$

$$B = \frac{\mu_0\, 2M\, d}{4\pi(d^2-l^2)^2} \text{ tesla} \qquad (3.18)$$

in the direction QP, where M is the magnetic moment of the magnet.

If the length $2l$ of the magnet is very small compared to the distance d, l^2 can be neglected compared to d^2 and hence the field intensity at P

$$= \frac{\mu_0}{4\pi}\cdot\frac{2Md}{d^4}$$

$$\text{or} \quad B = \frac{\mu_0}{4\pi}\cdot\frac{2M}{d^3} \text{ tes.} \qquad (3.19)$$

3.3.7. MAGNETIC INDUCTION AT A POINT ON THE EQUATORIAL LINE OF A BAR MAGNET

Let P be a point on the equatorial line of a bar magnet NS of length $2l$ metres, pole strength m Ampere-metres and magnetic moment M Ampere-metres^2, distant d metres from O the midpoint of the magnet.

Induction at P due to N pole

$$= \frac{\mu_0}{4\pi}\cdot\frac{m}{NP^2} \text{ tesla along } NP.$$

Induction at P due to S pole.

$$= \frac{\mu_0}{4\pi}\cdot\frac{m}{SP^2} \text{ tesla along } PS.$$

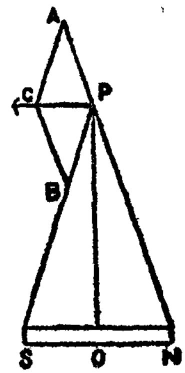


Fig. 3-16 Field along the equatorial line

Let these two inductions be represented in magnitude and direction by the adjacent sides PA and PB of a parallelogram.

Complete the parallelogram and draw the diagonal PC. Then PC represents the resultant induction at P due to NS, in magnitude and direction.

The two triangles PAC and NPS are similar.

$$\therefore \quad \frac{PC}{NS} = \frac{PA}{NP}$$

$$\text{or} \quad PC = NS.\ \frac{PA}{NP}$$

i.e., the resultant induction at $P = PC = NS.\ \dfrac{PA}{NP}$

$$\text{i.e.,} \quad B = \frac{2\,lm}{NP} \cdot \frac{\mu_0}{4\pi\ NP^2} \quad \text{tesla}$$

$$= \frac{\mu_0\ 2lm}{4\pi\ NP^3} \text{ tesla}$$

$$B = \frac{\mu_0}{4\pi} \cdot \frac{M}{(d^2+l^2)^{3/2}} \quad \text{tesla} \tag{3.20}$$

along PC parallel to NS.

If $2l$ is small compared to d, the induction at P,

$$B = \frac{\mu_0}{4\pi} \cdot \frac{M}{d^3} \quad \text{tesla} \tag{3.21}$$

direction parallel to NS.

Exercise 3. 3

1. Define magnetic induction.
2. What is the unit of magnetic induction in the $S.\ I.$ system?
3. How is the magnetic induction defined in terms of lines of induction ?
4. What is the expression for the force on a conductor carrying current kept in a magnetic field ?
5. State Fleming's left hand rule.
6. A straight wire 100 cms. long carries a current of 100 amp. at right angles to a uniform magnetic field of 1 Ampere/metre^2 Find the mechanical force on the wire. What will be the direction of the force ?
7. What is the law of force between two magnetic poles ?

8. What is the value of the permeability of free space ?
9. Define unit magnetic pole.
10. What is the value of magnetic induction at a point distant r metres from a pole of strength m Ampere metre ?
11. What is the total magnetic flux emanating from a pole of strength m Ampere meters ?
12. What is the relation between magnetic induction and magnetic field intensity ?
13. Define magnetic moment of a magnet.
14. What is the magnetic induction at a point distant d metres from the midpoint along the axis of a bar magnet length $2l$ metres and pole strength m Ampere metres ?
15. What is the magnetic induction at a point distant d metres from the midpoint along the equatorial line of a bar magnet of length $2l$ metres and pole strength m Ampere metres ?
16. Calculate the moment of the couple acting on a freely suspended magnet of moment 0·2 Ampere $(\text{Metre})^2$ when it is placed at an angle of 30° with the direction of a uniform magnetic field of induction 0·5 tesla.
17. The pole strength of a bar mognet is $9{\cdot}872 \times 10^{-6}$ Ampere metres and its length is 10 cms. Calculate the magnetic induction at a point on its axis at a distance of 20 cms. from its midpoint.
18. Calculate the intensity of the magnetie field at a point distant 20 cms. from the midpoint on the equatorial line of a bar magnet of pole strength 10^{-5} Ampere metres and length 10 cms.

3. 4. ELECTROMAGNETISM

3.4.1. MAGNETIC FIELD OF A CURRENT

Oersted's experiments in 1819 showed that a magnetic field is produced in the region surrounding a conductor carrying current. A picture of the magnetic field due to circuits of various shapes can be obtained by plotting lines of force with a small compass needle. The diagrams in the case of a straight conductor carrying current and a coil carrying current are shown in Fig. 3-17.

It is seen that the magnetic lines of force near a straight wire are circles concentric with the wire. The direction of the magnetic lines of force is given by Ampere's *swimming rule* or Maxwell's *corkscrew rule*. The corkscrew rule states that the

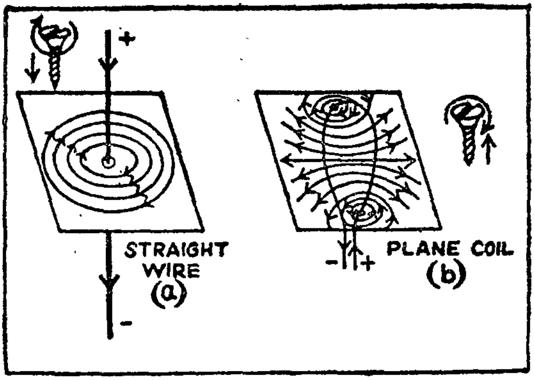


Fig. 3-17
Field due to a current

direction of the lines of force around a circuit is that in which a right handed screw should be turned in order to advance it in the direction of the current. In the case of a coil of wire, when the direction of the field around each section of the coil is found, it will be seen that the lines of force are circular near the wire, but within the space enclosed by the coil, the lines are all in the same direction and near the centre, they are parallel to each other. So the field at the centre is uniform and perpendicular to the plane of the coil. Thus it is seen that the lines of force are similar to that due to a very thin disc shaped magnet, with one face as a north pole and the other a south pole.

3.4.2. BIOT-SAVART LAW

Experiments done by Biot and Savart led to a relation, by means of which, the magnetic induction or flux density at any point of space around a circuit in which there is a current can be found. The circuit can be imagined to be divided into short elements of length dl. Each element will produce a field at all points of space and the total field produced at any point by the entire circuit is got by finding the resultant of the fields produced by all the elements. Let the length of the short element at O be dl and let it carry a current i. The induction dB at P or P' distant x from O is given by the Biot-Savart law

i.e., $dB = \frac{i\,dl \sin\theta}{r^2}$

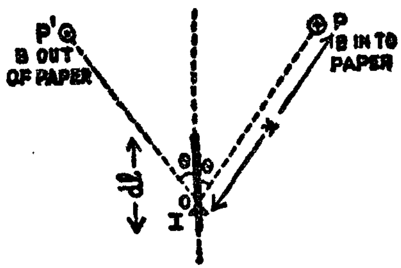


Fig. 3-18
Biot-Savart Law for finding field due to a current

where θ is the angle which the axis of the element dl makes with the direction OP. The direction of dB is at right angles to the plane containing dl and OP and is in the direction given by the right hand screw rule.

The constant of proportionality used in the rationalised S.I system is $\frac{\mu_0}{4\pi}$ for free space and hence the Biot-Savart law becomes

$$dB = \frac{\mu_0}{4\pi} \frac{i\,dl \sin\theta}{x^2}$$

From this equation, it is seen that μ_0, the permeability of free space has the dimensions of $\frac{B \times x^2}{i\,l}$. Since B is expressed as weber/metre², $\mu_0 = \frac{\text{weber}}{\text{metre}^2} \times \frac{\text{metre}^2}{\text{ampere-metre}}$

= weber / ampere-metre. As mentioned earlier μ_0 is also expressed in henry per metre, where henry is the unit of inductance and the magnitude of $\mu_0 = 4\pi \times 10^{-7}$. This value is derived from the definition of ampere.

3.4.3. THE TANGENT GALVANOMETER

The tangent galvanometer is an instrument for measuring current. It is an absolute instrument since it does not need any other instrument to calibrate it.

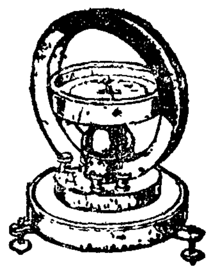


Fig. 3-19
Tangent
Galvanometer

It consists of a coil of wire wound on a non-magnetic circular frame of brass or wood. (Fig 3-19) The frame is fixed with its plane perpendicular to a circular turn table provided with levelling screws. The levelling screws can be adjusted so that the circular table is horizontal. The plane of the coil will then be vertical. At the centre of the circular table is a small upright projection on which is supported a compass box. The compass box consists of a small pivoted magnet, fixed at right angles to which is a long pointer which moves over a graduated circular scale. The scale is graduated in 4 segments each reading from 0° to 90°. The compass box is supported in such a manner thatthe centre of the pivoted magnet coincides exactly with the centre of the coil of wire. The length of the pivoted magnet is small, so that the magnetic field over its entire length can be assumed to be uniform. The coil of wire is usually in three sections, of 2, 5 and 50 turns each, and are connected to separate terminals. The 2 turn coil is made of thick wire and is used for measuring currents of a few amperes, the 5 turn coil is of thinner wire and is used to measure currents of the order of a tenth of an ampere and the fifty turn coil is of still thinner wire, and is used to measure currents of the order of milliamperes.

The tangent galvanometer as the name indicates, works on the *tangent law*. A magnetic needle suspended at a point where there are two crossed fields will come to rest in the direction of the resultant of the two fields. The plane of the coil of the tangent galvanometer is always adjusted to be exactly in the magnetic meridian, so that when a current is passed through the coil, the magnetic field produced at the centre of the coil, due to the current will now be at right angles to the magnetic meridian. The pivoted magnetic needle will now be under the influence of two magnetic fields, the horizontal component H of the earth's magnetic field along the magnetic meridian and the field of intensity F due to the current, in the coil, the direction of F being at right angles to that of H. The pivoted magnetic needle which will lie along the direction of H in the absence of any other magnetic field, will be deflected through an angle θ when the current is switched on, such that

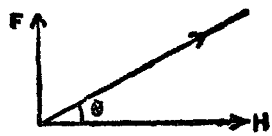


Fig. 3-20
Tangent Law

$$\tan\theta = \frac{F}{H}$$

If a current i amp. flows through the coil of n turns and radius r metres, it can be shown that the field at the centre of the coil

$$F = \frac{n\,I}{2r} \text{ amp. turn/metre. } (Am^{-1})$$

This field acts in a direction perpendicular to the plane of the coil.

$$\therefore F = \frac{n\,I}{2r} = H\tan\theta \text{ where } H \text{ is also in amp. turn/metre}$$

$$\therefore I = \frac{2\,rH}{n}\tan\theta \quad \text{amperes}$$

$$\text{or } I = K\tan\theta \quad \text{amperes}$$

where $K = 2\,rH/n$ is a constant for a given coil at a given place and is known as the **reduction factor** of the tangent galvanometer.

When using a tangent galvanometer, some initial adjustments have to be done carefully before passing the current through it.

(i) The plane of the coil should be adjusted to be in the magnetic meridian. This is done in two steps.

(a) The levelling screws on the circular table is adjusted so that the table is horizontal and hence the plane of the coil, vertical. (b) The coil is then turned so that its plane is parallel to the axis of the magnetic needle.

This adjustment ensure that the direction of F due to the current in the coil is perpendicular to that of H, due to the earth's magnetic field.

(ii) The pointer in the compass box should be made to read zero-zero.

The tangent galvanometer is the most sensitive when the deflection is around 45° and therefore the deflections are adjusted to be between 30° and 60°.

3.4.4. THE MOVING COIL GALVANOMETER

The moving coil galvanometer was first devised by Lord Kelvin and later modified by D'Arsonval. The D'Arsonval type consists of a circular or retangular coil of fine copper wire suspended between the poles of a very powerful horse shoe shaped magnet by a very thin strip of phosphor bronze (Fig. 3-21). The strip provides the controlling couple for the moving coil and also serves as the lead for the current to enter the coil. The current is led out of the coil by a fine phosphor bronze spring attached to the bottom of the coil. The pole pieces of the horse shoe magnets are curved inside and a soft iron piece is placed inside the coil without touching it. This makes the lines of magnetic force concentrate towards the centre of the space between the poles and hence renders the magnetic field radial. As a result for deflections upto 60°, the plane of the coil will be parallel to the magnetic field. A small mirror is attached to the suspension above the coil and the deflections of the coil can be determined using a lamp and scale arrangement.

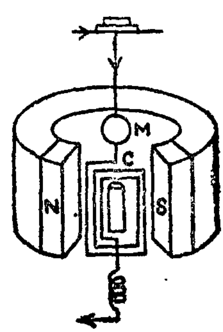


Fig. 3-21
Moving coil
galvanometer

Consider a single turn of a rectangular coil of wire $ABCD$ carrying a current i ampere in a magnetic field of induction B as shown in Fig. 3-22. The field is parallel to BC and AD of the coil and hence there will not be any force on these portions of the coil. But there will be forces each equal to Bil on the portions AB and DC of the coil where l is the length of AB or DC. These forces will be parallel but opposite in direction and so will constitute a couple, the moment of which will be equal to $Bil \times b$ where 'b' is the distance BC or the breadth of the coil.

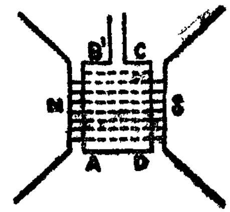


Fig. 3-22 Rectangular current loop in a magnetic field

i.e. the moment of the couple $= BiA$

where A is the area of the coil. If there are n turns of the coil, there will be a couple BiA on each turn and hence the total couple acting on the coil, tending to turn it will be $n\,BiA$. The torsional couple of the suspension will oppose this deflecting couple on the coil and the coil will come to rest in the position where the two couples balance. If θ is the angle through which the plane of the coil turns $nBiA = C\theta$ where C is the couple per unit twist of the suspension. Hence $i = \frac{C}{nBA}\,\theta$.
Thus the current is directly proportional to the deflection. This is so because the magnetic field is radial and the plane of the coil remains parallel to the magnetic field even when it is deflected.

3.4.5. AMMETER

An ammeter is an instrument for measuring currents and it is calibrated to read the values of currents directly. An ammeter is always connected in series in the circuit in which the current is to be measured and hence, in order that the introduction of the ammeter does not alter the value of the current the ammeter should be an instrument of very low resistence. A moving coil galvanometer can easily be converted into an ammeter by connecting a low resistance shunt in parallel with the galvanometer (Fig. 3-23). If G is the resistance of the galvanometer and S that of the shunt, the resistance R of the combination will be given by the relationship

$$\frac{1}{R} = \frac{1}{G} + \frac{1}{S}$$

or $$R = \frac{SG}{S+G} \qquad \dots \quad (3.9)$$

i.e, the resistence R will be less than even the small resistance of the shunt. Hence by using suitable shunts, the resistance of the combination can be made as low as necessary.

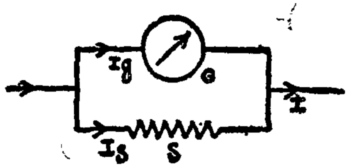


Fig. 3-23

Conversion of a galvanometer into an ammeter

Also when the main current is large, the shunt protects the galvanometer coil by allowing only a small fraction of the main current to pass through the galvanometer coil, If i is the current in the main circuit, the current i_g through the galvanometer is given by

$$i_g = i \times \frac{S}{S+G} \qquad \dots \quad (3.10)$$

S can be so chosen that

$$\frac{S}{S+G} = \frac{1}{10}, \frac{1}{20} \text{ etc.}$$

Thus the current through the galvanometer will be a small but known fraction of the main current i, and the deflections of the galvanometer are marked to read directly the current in the main circuit. The range of the ammeter can be varied by varying the resistance of the shunt.

3.4.6. VOLTMETER

The voltmeter is an instrument to measure potential differences between two points and it is always connected across the two points, parallel to the main circuit. In order that the voltmeter does not draw any appreciable current from the main circuit, the voltmeter should be an instrument of very high

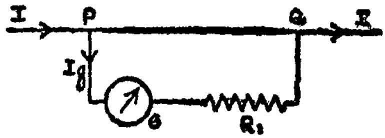


Fig. 3-24

Conversion of a galvanometer into a voltmeter

resistance. A moving coil galvanometer is converted into a voltmeter by connecting a very high reistance in series with it. Let R_1 be high resistance connected in series with the galvanometer of resistance G. Now

$$i_g\,(R_1 + G) = V_p - V_q$$

where $V_p - V_q$ is the potential difference between the points P and Q

$$i_g = \frac{V_p - V_q}{R_1 + G}$$

If the total resistance $(R_1 + G)$ of the voltmeter, which is the combination of the galvanometer and the high resistance is kept constant, i_g is proportional to $(V_p - V_q)$ and the deflections of the galvanometer are made to read directly the potential difference in volts. The range of the voltmeter can be varied by varying R_1.

Exerice 3.4

1. *State the corkscrew rule relating to the direction of the lines of force around a conductor carrying current.*
2. *State Biot Savart Law.*
3. *What is the value of the magnetic field intensity at the centre of a coil of n turns of radius r metres carrying a current i amps? What is the direction of the field?*

4. *Find the field at the centre of a circular coil of radius 16 cms having 10 turns when it carries a current of 8 amperes.*
5. *What is the principle of the tangent galvanometer ?*
6. *Why should the magnet pivoted at the centre of the coil in the tangent galvanometer be very small ?*
7. *What is the reduction factor of a tangent galvanometer ?*
8. *When using a tangent galvanometer, what are the initial adjustments to be made, before passing the current through it ?*
9. *A current of 1·5 amperes flows through the 5 turn coil of a tangent galvanometer having a diameter of 0·3 metres. If the deflection of the neddle at its centre is 45°, calculate the horizontal intensity of the earth's field at that place.*
10. *What is the principle of the moving coil galvanometer ?*
11. *How is the ammeter connected in a circuit in order to measure the current in it ?*
12. *Why is the ammeter a low resistance instrument ?*
13. *How is a galvanometer converted into an ammeter ?*
14. *A moving coil galvanometer of resistance 100 ohms, shows a full scale deflection when a current of 1/1000 amperes flows through it. What resistance should be connected in parallel to the galvanometer in order to convert it into an ammeter reading upto 5 amperes ?*
15. *How is the voltmeter connected in a circuit to measure the potential difference between two points ?*
16. *Why is the voltmeter a high resistance instrument ?*
17. *How is the moving coil galvanometer converted into a voltmeter ?*
18. *How can the galvanometer in question (14) be converted to a voltmeter reading upto 5 volts ?*
19. *How will you increase the range of a voltmeter of 500 ohms resistance, from 3 volts to 30 volts ?*
20. *A galvanometer of 10 ohms resistance shows full scale deflection, when a curmet of 1/100 amp. is passed through*

it. How can this be converted into (i) an ammeter reading a maximum current of 2 amperes and (ii) a voltmeter reading a maximum of 5 volts ?

3.5. ELECTROMAGNETIC INDUCTION

3.5.1. INTRODUCTION

Oersted's discovery that a magnetic field is produced around a conductor carrying current, led Michael Faraday to look for the converse effect. In 1831, he showed that currents can be produced in a closed conductor whenever there is a change in the magnetic flux passing through the conductor and that the current lasted only so long as the change lasted. The current produced in the conductor in this way is called an *induced* current. The electromotive force producing the current an *induced* e.m.f. and the whole phenomenon is known as **electromagnetic induction.**

3.5.2. FARADAY'S EXPERIMENTS

A strong bar magnet is moved with its N pole towards a closed coil consisting of a large number of turns of insulated copper wire connected in series with a galvanometer. There is a deflection in the galvanometer as long as there is relative motion between the coil and the magnet. If now the magnet is withdrawn,

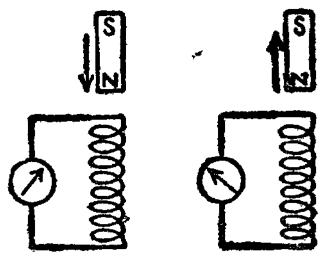


Fig. 3-25
Faraday's experiments

there will be a momentary deflection in the opposite direction. It will also be noticed that if the magnet is moved fast, the momentary deflections produced will be greater than when the magnet is moved slowly. Similar results will be produced when the magnet is moved with its south pole facing the coil, but the deflections produced will be in directions opposite to those produced when the north pole was facing the coil. It can be shown that in every case, the direction of the current produced in the coil will be such as to oppose the change that is causing the current. When a north pole is brought towards the coil, the direction of the current induced in it will be such that the end of the coil facing the north pole will behave like another north pole trying to repel the approaching north pole. When the north pole is withdrawn, the current induced will be in such a direction as to produce a south pole at the end of the coil facing the north pole, which will therefore attract the receding north pole and thus oppose the change that is taking place. It can easily be seen from this, why the direction of the induced current, when a south pole is brought towards the coil, is exactly opposite to that when a north pole is brought towards it. Essentially an e.m.f. is induced in a coil, whenever there is a change of magnetic flux through it and the direction of the induced current is always such that it will oppose the change that causes it. Also it is seen that if the magnet is brought to the coil or taken away from it at a faster rate, the magnitude of the induced current is greater. Hence it is the rate at which the magnetic flux passing through the conductor is changing which determines the magnitude of the induced current.

Change of flux and hence induced currents can be produced in a closed circuit, even by switching on or off a current in a neighbouring circuit. The direction of the induced current when switching on the current in the neighbouring circuit will be opposite to that when it is switched off and also the induced current in the circuit will last only till the steady current is established in the second coil, when the current is switched on and till the current falls to zero when it is switched off.

Currents can be induced in a coil, even by relative motion between it and another coil carrying current, because here again there is a change in the magnetic flux passing through the first coil, the flux being due to the magnetic field produced by the current in the second coil.

3.5.3. LAWS OF ELECTROMAGNETIC INDUCTION

When the magnetic flux linked with a circuit is changing

(i) *an e.m f. is induced in it, which is proportional to the rate of change of flux* (Faraday's law).

(ii) *the direction of the e.m.f. is such that the effects of any current it produces tend to oppose the change of flux.* (Lenz's law).

These two laws together are mathematically expressed as

induced e.m.f. $e = -\left(\frac{\phi_1 - \phi_2}{t}\right)$

where $\frac{\phi_1 - \phi_2}{t}$ is the rate of change of the magnetic flux ϕ. [It must be remembered that if ϕ is the flux linked with one turn of a coil of n turns, the total flux linked with the coil is $n\phi$.]

3.5.4. MUTUAL INDUCTION

We have seen earlier that two coils electrically unconnected can be linked together by magnetic flux and any change of current in either coil will cause an induced e.m.f. to be produced in the other. This particular form of induction is known as mutual induction. In D.C. circuits or direct current circuits, where the magnitude of the current remains steady, effects due to mutual inductance will arise only when the current in one circuit is switched on or off and pulses of current will appear in the other. But if an A.C. or alternating current, the magnitude of which is continually varying, flows through one coil, it produces a continually varying magnetic flux through the other and it will cause an alternating e.m.f. of the same frequency to be produced in it. This is the principle of the **transformer**.

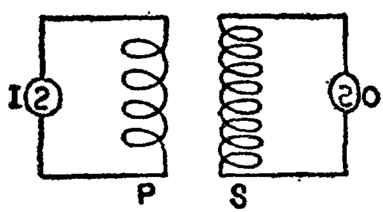


Fig. 3-26
Transformer

Let P and S be two coils placed close to each other as in Fig. 3-26. A current flowing through P will produce a flux, part of which will pass through the neighbouring coil S. For fixed positions of P and S, the flux linked with S will be proportional to the current i flowing through P.

i. e. $\phi \propto i$ or $\phi = M i$

where M is a constant known as *the coefficient of mutual induction*, and depends on the number of turns, relative positions etc. of the coils.

$$\text{If } i = 1, \phi = M$$

i.e. the coefficient of mutual inductance between two coils is numerically equal to the flux linked with one of them due to the flow of unit current through the other.

The coefficient of mutual inductance between a pair of coils is also numerically equal to the e.m.f. induced in one coil, due to unit rate of change of current through the other. On this basis S. I. unit of mutual inductance can be defined as the mutual inductance between two coils when an e. m. f. of 1 volt is induced in one coil when the current through the other changes at the rate of 1 amp. per sec. It is called a henry.

$$1 \text{ henry} = \frac{1 \text{ volt}}{1 \text{ amp. per sec.}}$$

3.5.5. SELF INDUCTION

In the process of mutual induction, e.m.fs are induced in a circuit. due to changes of current in another circuit. But e.m.fs can also be induced in a circuit by the changes in its own current since this also produces changes of flux through the circuit. This process is known as *self induction*.

When a steady current is flowing in a circuit, no induced e.m.f. can arise from self inductance and so in D. C. circuits, effects of self inductance can occur only at switching on or switching off of the current.

At switching on, the induced e.m.f. acts in opposition to the applied potential difference and delays the growth of current so that the maximum steady value of the current is reached only gradually. At switching off, the current suddenly falls to zero so that the rate of change of current is very large and the induced *back e. m. f.* tending to keep the current going will be very large. The peak value of this can be several hundred times larger than the original applied *e.m.f.* If ϕ be the magnetic flux linked with a coil when the current through it is i

$$\phi \propto i \text{ or } \phi = Li$$

where L is a constant known as the *coefficient of self inductance of the coil If* $i = 1$, $L = \phi$ *numerically i.e. the coefficient of self inductance of a coil is numerically equal to the magnetic flux linked with it when unit current flows through it.*

The coefficient of self inductance of a coil can also be defined as the e.m.f. induced in it when the rate of change of current through it is unity. The unit of self inductance is also henry. The coefficient of self inductance of a coil is one henry if an induced e.m.f. of one volt is set up in it, when the current through it changes at the rate of 1 amp. per sec.

3.5.6. EDDY CURRENTS

We saw that when the magnetic flux linked with a conductor varies, an e.m.f. is induced in it. But it is not necessary that the conductor is in the form of a wire or coil. Even when the conductor is a large lump of metal, significant e.m.fs. can act round closed paths inside the lump of metal itself. Although the e.m.fs. induced are not very large, since the resistance of the current path is negligible, very large currents will flow. These induced currents circulating inside a piece of metal are known as **Eddy currents** or **Foucault's currents**

Foucault in 1895 discovered that when a mass of metal is moved in a magnetic field, induced currents were produced, resulting in the dissipation of energy in the form of heat. Since according to Lenz's law, the direction of the induced currents will be to oppose the motion of the metal in the magnetic field, the induced currents will act as an effective brake on the motion of the body. The mechanical energy is turned into electrical energy of the eddy currents, which in turn is converted into heat inside the metal.

The effect of eddy currents can be demonstarted by swinging a pendulum with a thick copper plate as bob, between the poles of an electromagnet (Fig. 3-27). As soon as the electromagnet is switched on, the oscillations will slow down and the pendulum will rapidly come to rest. This is due to the braking effect of the eddy currents.

If now a series of slots are cut in the bob, very little braking effect will be noticed. This is because, the currents are now free to circulate only inside the relatively narrow teeth left between the slots.

Arago showed that if a copper disc, placed below a pivoted magnetic needle, is rotated rapidly about an axis passing through

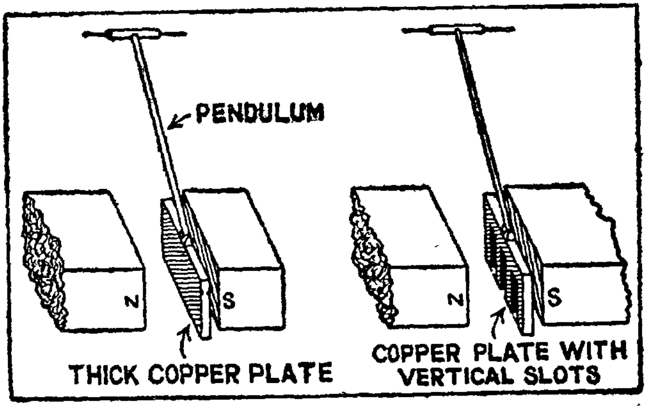


Fig. 3-27

**Effect due to Eddy currents-
heavy damping of pendulum avoided
by cutting slots**

the pivot, the magnetic needle will tend to follow the movement of the disc. This is because the relative motion between the magnetic needle and disc induces currents in the disc, which will tend to oppose the relative motion and hence the magnetic needle is dragged along.

In a dead beat moving coil galvanometer, eddy currents are made use of for damping the oscillations of the galvanometer. But in electrical machinery, eddy currents are undesirable because considerable energy will be wasted as heat. So in dynamos, transformers etc., iron parts are made out of stacks of iron sheets or laminations. The laminations are insulated from one another by varnish or enamel. The eddy currents are greatly reduced in this case, since they can circulate only in closed paths within the individual laminations.

3.5.7. TRANSFORMERS

The transformer is an electrical device based on the principle of mutual inductance between two coils. It is used to transfer electrical energy from one coil to another Energy at high voltage and low current in one coil can be transferred to another coil as energy at low voltage and high current. In this case, it is called a **step down** (voltage) **transformer.** Or it can be used to transfer energy at low voltage and high current in one coil to energy at high voltage and low current in another coil. In this case, it is called a **step-up** (voltage) **transformer.**

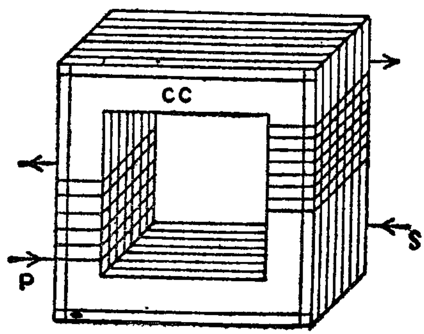


Fig. 3-28
Transformer

The transformer consists of two coils P and S wound separately on a laminated iron core CC (Fig. 3-28). P is the primary oil to which the voltage, which is to be converted is applied and S is the secondary which delivers the altered voltage.

In an ideal transformer,

$$\frac{\text{Secondary voltage}}{\text{Primary voltage}} = \frac{\text{no. of turns of secondary coil}}{\text{no. of turns of primary coil}}$$

$$\text{or} \quad \frac{E_2}{E_1} = \frac{n_2}{n_1}$$

If $n_2 > n_1$, the secondary e.m.f. will be greater than the primary e.m.f. and the transformer is a step up transformer. In this case, the primary is made up a few turns of thick wire and the secondary will have a large number of turns of thin wire.

If $n_2 < n_1$, the secondary e.m.f. will be less than the primary e m.f. and the transformer is a step down transformer.

In all ideal transformers, whether they are step up or step down, the energy fed into the primary will be equal to the energy delivered at the secondary. If i_1 and i_2 are the currents in the primary and the secondary circuits when E_1 and E_2 are the primary and secondary voltages, the power input $= E_1 i_1$ and the power output $= E_2 i_2$., and in an ideal transformer, where there is no loss of energy,

$$E_1 i_1 = E_2 i_2$$

$$\text{or } \frac{i_2}{i_1} = \frac{E_1}{E_2} = \frac{n_1}{n_2}$$

Thus when the voltage is stepped up in a certain ratio, current is stepped down in the same ratio and *vice versa.*

3.5.8. A.C. GENERATOR

Consider a flat coil rotating about an axis in its own plane at right angles to a uniform magnetic field. As the coil rotates, the flux through it varies as shown in Fig. 3-29.

When the plane of the coil is perpendicular to the magnetic field as in (i) and (iii), the magnetic flux passing through the coil

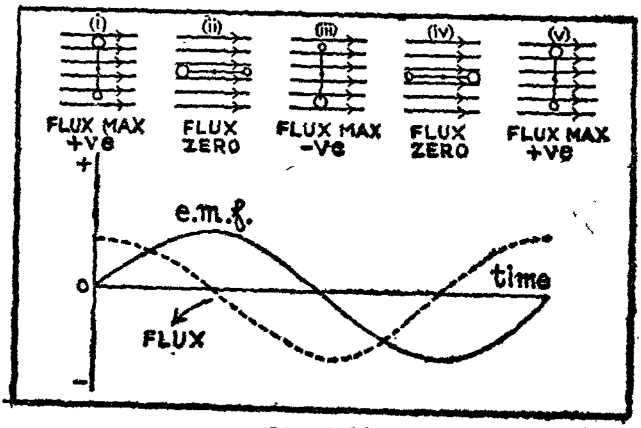


Fig. 3-29
Variation of flux through a rotating coil and the e.m.f. induced

is maximum, but it is momentarily unchanging whereas when the plane of the coil is parallel to the magnetic field as in (ii) and (iv), the flux passing through the coil is momentarily zero, but it is changing at a maximum rate, passing through zero, from a positive value to a negative value in position (ii) and passing through zero from a negative value to a positive value in position (iv). The induced e.m.f. in the coil which is equal to the rate of change of flux through it will therefore vary as shown in the diagram. The e.m.f. is zero when the flux is maximum as in positions (i) and (iii) and is maximum in magnitude when the flux is zero as in positions (ii) and (iv). And according to Lenz's law, which states that the induced e.m.f. acts so that the current produced by it opposes the change of flux, the induced e.m.f. is positive at position (ii) and negative at position (iv). The value of the induced e.m.f. can be derived mathematically as follows:

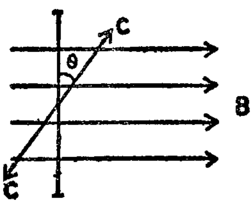


Fig. 3-30
Flux through a rotating coil

Let the plane of the coil make an angle θ with the direction of the normal to the magnetic field at any instant (Fig.3-30). The total flux passing through the coil in this porition is $AB \cos \theta$ where A is the area of the coil and B, the magnetic induction or flux density. If there are n turns in the coil, the total flux ϕ passing through the coil is given by $\phi = n\,AB \cos \theta$

Induced e.m.f. $E = \dfrac{d\phi}{dt} = n\,AB \sin \theta \dfrac{d\theta}{dt}$.

Now $\dfrac{d\theta}{dt}$ is the angular velocity, ω, of the coil and $\theta = \omega t$.

Hence $E = n\,AB\,\omega \sin \omega t$.

This shows that the e.m.f. generated in the coil is alternating and varies sinusoidally. The maximum value of the e.m.f. is $n\,AB\,\omega$ and if this is represented as E_0, the e.m.f. at any instant is given by

$$E = E_0 \sin \omega t$$

or

$$E = E_0 \sin 2\pi n$$

where $\omega = 2\pi n$, n being the number of rotations of the coil per second. In most generators, n is equal to 50.

Fig. 3-31 shows the essential parts of a simple form of A.C. generator. A rectangular coil rotates between the pole pieces *NS* of a powerful electromagnet, the current for which is supplied by a separate D.C. source. The ends of the coil are connected to two copper rings known as slip rings which are insulated from each other. These rings are concentric with the axis of rotation of the coil and rotate with the coil. Two carbon brushes which press against the slip rings connect the coil to the external circuit. The coil usually consists of a large number of turns of insulated copper wire wound around a core of laminated iron, the iron serving to concentrate the magnetic lines of force. The coil is known as the *armature* and the mechanical energy for its rotation is derived from steam turbines or water turbines or internal combusion engines.

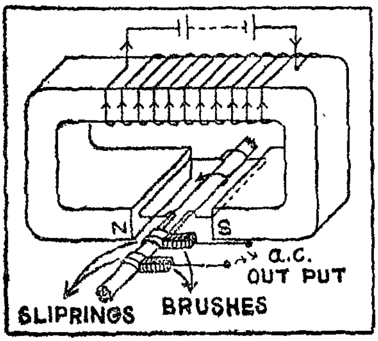


Fig. 3-31
Production of A.C. by rotation of a coil in a magnetic field

However in practice, it is found more satisfactory to have the armature fixed and the magnetic field rotate relative to it. The armature coil is is wound in two parts in series, between which the electromagnet rotates (Fig. 3-32). The direct current for the electromagnet is supplied through the slip rings. The advantage is that the slip rings and brushes are now required to carry only the relatively small current needed to magnetise the rotating electromagnet and not the very large currents developed in the generator. It is much easier to have the large currents flow through fixed connections. In this case, the rotating electromagnet is the *rotor* and the armature coils, the *stator*

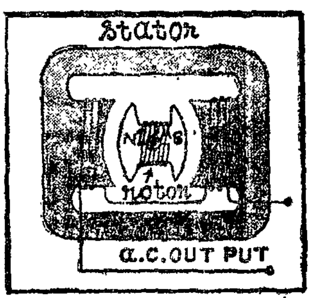


Fig. 3-32
Simple form of A.C. generator (2 - pole constant current)

In most generators, the armature is split into 4, 6 or 8 parts, spaced equally inside the stator. The rotor is then arranged to have the same number of poles, north and south poles being arranged alternately. Fig. 3-33 shows a four pole alternating current generator. In this, the speed of rotation needs to be only

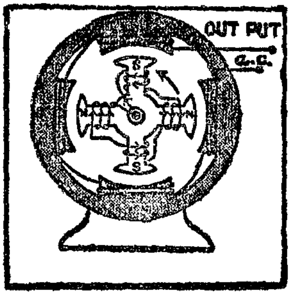


Fig. 3-33
4 - pole constant current
A.C. generator

half that is necessary to produce the same frequency in a generator with only one pair of coils.

Exercise 3.5

1. *If the north pole of a magnet is brought towards a closed coil of wire, a current is induced in it. What happens if a south pole is brought towards it ?*
2. *What happens if the movement of the magnet is stopped ?*
3. *What is the essential condition for an e.m.f. to be induced in a closed circuit ?*
4. *What are the ways in which the magnetic flux passing through a coil can be changed ?*
5. *State Faraday's law of electromagnetic induction.*
6. *State Lenz's law of electromagnetic induction.*
7. *What is the mathematical expression for the laws of electromagnetic induction ?*
8. *Calculate the e.m.f. induced between the wing tips of an aeroplane of wing span 30 metres flying horizontally at 250 metres/sec. if the flux density of the vertical component of the earth's magnetic field is 4×10^{-5} tesla (weber/metre^{-2}). [When the plane is flying horizontally, its wing will cut the vertical lines of induction due to the earth's magnetic field. When the plane with its wing span*

of 30 metres, covers 250 metres in one sec. the total area swept by the wing in 1 sec = 30 × 250 Sq. metres

∴ *Total lines of induction cut by the wing in 1 sec.*

$= 30 \times 250 \times 4 \times 10^{-5}$

i.e. rate of change of magnetic flux

$= 30 \times 250 \times 4 \times 10^{-5}$ *webers/sec.*

$= 0{\cdot}3$ *webers/sec.*

Induced e.m.f. $e = -\dfrac{d\theta}{dt} = 0{\cdot}3$ *volt.*

The negative sign only shows that the induced e.m.f. opposes the change that is producing it.]

9. *A horizontal gramaphone turn table made of brass and of diameter 30 cms. rotates at* $33\frac{1}{3}$ *revolutions per minute in a uniform vertical magnetic field of flux density* 0·01 *tesla. Calculate the e.m.f. induced between the centre and rim of the turn table.*

[*The area swept by any radius in one second*

$= \pi r^2 \times$ *no of rev. per second*

$$= \pi \times \frac{30}{2 \times 100} \times \frac{30}{2 \times 100} \times \frac{100}{3 \times 60}$$

$$= \frac{\pi}{80} \text{ Sq. metres } s^{-1}$$

∴ *No. of lines of induction cut per sec.*

$\dfrac{\pi}{80} \times 0{\cdot}01$ *webers per sec.* $= 0{\cdot}00039$ *webers/sec.*

∴ *Induced e.m.f.* $= -0{\cdot}00039$ *volts*

$= -0{\cdot}39$ *milli volts.*]

10. *Explain what is meant by mutual inductance.*
11. *Define the coefficient of mutual inductance.*
12. *What is self inductance?*
13. *What is the unit of inductance?*
14. *Calculate the magnitude of the back e.m.f. due to self inductance in a coil of self inductance 5 millihenries in which the current increases from zero to a steady value of 1 ampere in 1/100 second.*

    [$e = L \times$ *rate of change of current*]

Rate of charge of current = $\frac{1\text{-}0}{1/100}$ *amp/sec* = 100 *amp./sec.*

$e = -5 \times 10^{-3} \times 100 = -0.5$ volts.]

15. *What are eddy currents ?*

16. *Why are iron parts of electrical machinery made of laminated sheets ?*

17. *What is a transformer used for ?*

18. *Can D. C. voltages be used in a transformer ?*

19. *In an ideal transformer, how are the voltages in the secondary and primary related to their number of turns ?*

20. *If the ratio of the secondary to primary voltage in an ideal transformer is* 100, *what is the ratio of secondary to primary current ?*

21. *In an ideal transformer, if 1000 watts of energy is supplied at the primary, what is the energy delivered at the secondary ?*

22. *How is an alternating e.m.f. generated ?*

23. *Calculate the maximum e.m.f. generated in a rectangular coil of length 12 cms. and breadth 7 cms. and of 1000 turns when rotated at the rate of 3000 revolutions per minute about an axis in its own plane, at right angles to a magnetic field of flux density 0·02 tesla.*

4. ELECTROMAGNETIC RADIATIONS

4.1. ELECTROMAGNETIC WAVES

4.1.1. INTRODUCTION

While investigating the properties of electric and magnetic fields, James Clerk Maxwell found the possibility that electromagnetic disturbance may be propagated through space as a wave. His calculations showed that the velocity of these waves is the same as the velocity of light. He then suggested that light may be *electromagnetic waves*.

We have seen that a wave is energy propagated from one place to another. We have studied about sound waves which are waves in air (or liquid or solid) and we are familiar with waves in water. The characteristics of any wave motion are its velocity 'v' frequency 'ν' and wave length 'λ' which are related as follows :

$$v = \nu \lambda$$

Sound waves and water waves need a material medium for their transmission. But the electromagnetic wave predicted by Maxwell can pass through empty space as well. For example the light from the sun reaches us through the empty space between the sun and the earth.

Maxwell's prediction was experimentally verified later by Hertz in 1888 when he produced and detected electromagnetic waves. Electromagnetic waves are called so because they consist of electric and magnetic fields which vary in time and from point to point in space just as sound waves in air are characterised by variation of air pressure in time and space.

Electromagnetic waves can have wavelengths ranging from xtremely small magnitudes of the order of $10^{-14} m$ to very large values of the order of hundreds or thousands of metres. Our eyes are sensitive to a very small range of the spectrum of wavelength ranging from $4 \times 10^{-7}\ m$ to $7 \times 10^{-7}\ m$. Wavelengths

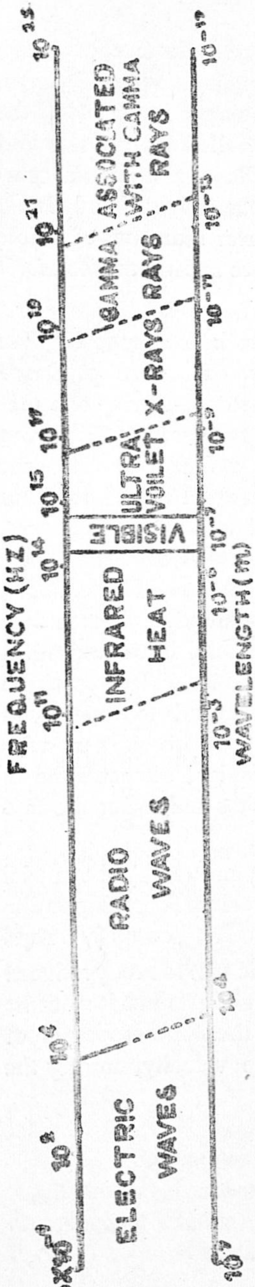


Fig. 4-1 Electromagnetic spectrum

in this region are usually measured in nano(10^{-9}) metres (*nm*). They are also measured in terms of a unit called the *Angstrom*($A°$)which is equal to $10^{-10}m$ Thus the visible range extends from 400 *nm* (violet) to 700 *nm* (red) or 4000 A° to 7000 A°. The first measure ments on the wavelengths of light waves were made around 1827 by Young, Fresnel and Fraunhofer.

Electromagnetic waves of any frequency can be produced with an appropriate source or generator. However, the means of production of different ranges of wavelengths are widely different, so are their properties and therefore their methods of detection and study. In fact, because of these wide differences, it was not realised that they all were basically the same phenomenon until long after they had been individually discovered and studied by different people and at different times.

4.1.2. THE ELECTROMAGNETIC SPECTRUM

Fig. 4-1. shows the frequency and wavelength ranges of various regions of the electromagnetic spectrum studied so far.

4.1.3. SOURCES AND GENERAL CHARACTERISTICS

On the long wavelength side we have the *radio waves* which can be produced by making use of suitable electronic circuits. The range of wavelengths used in broadcasting (radio and T. V.) is roughly from 10^5 *m* to 0.1 *m*. Electromagnetic waves of the order of 10^{-2} *m* in wavelength are usually called *micro waves* and are used in radar apart from other scientific and industrial purposes. Of still shorter wavelengths are the *infra red* and *visible radiations*. Radiations of wavelengths shorter than that of visible light ranging from 400 *nm* to 100 *nm* are known as *ultraviolet radiations*.

Electromagnetic waves of wavelengths in the range 10^{-10} *m* to 10^{-12} *m* correspond to those of *X-rays*. They are produced when fast moving electrons are suddenly stopped by a solid target. The chance discovery of X rays by Roentgen in 1895 proved to be one of the most important events in modern science, not only because of the practical applications it has found but for its role in the development of physical and biological sciences. X-rays are used in the diagnosis of a variety of diseases. They are also used in the treatment of certain diseases. For example, malignant cells are somewhat more readily killed by X-rays than are normal ones and therefore they are useful in the treatment of cancer Hidden defects in objects and concealed cracks in metals may be revealed by X-rays and this leads to a variety of applications in industry. X-ray spectroscopy provides us with a powerful method of identification of chemical elements and in the unravelling of the arrangements of atoms and molecules in a crystal.

Waves of even shorter wavelengths are emitted by the nuclei of some substances such as uranium and thorium following radioactive disintegration. These waves are called *gamma* (γ) rays.

The different ranges of electromagnetic waves are produced in different ways and have quite different effects but it must be stressed that they all consist of waves of the same fundamental nature and travel in vacuum with the same velocity, namely the velocity of light.

Exercise 4-1

1. *What is the velocity of electromagnetic waves ?*
2. *Name the different kinds of electromagnetic radiations.*
3. *Compare the electromagnetic waves with the sound waves.*
4. *In what unit are wavelengths measured ?*

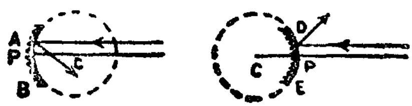
5. *What are the minimum and maximum wavelengths of the visible region of the electromagnetic spectrum ?*
6. *Mention the order of the wavelength for the following*
   (i) Microwaves
   (ii) Infra red
   (iii) Ultra violet
   (iv) X-rays
7. *What is the principle based on which X-rays are produced ?*
8. *Mention a few medical, scientific and industrial uses of X-rays.*

4.2. REFLECTION OF LIGHT

4.2.1. REFLECTION AT SPHERICAL SURFACES

A highly polished smooth surface will act as a mirror. If the reflecting surface is plane, it is called a plane mirror, If it is spherical, it is called a spherical mirror. The centre 'C' of the sphere, of which the mirror forms a part, is called the **centre of curvature,** the radius '*r*' of this sphere is called the **radius of curvature** of the mirror.

Spherical mirrors are divided into two classes, *concave* and *convex*. In the case of the concave mirror, the light falls on the surface from the same side as the centre of curvature. In the case of the convex mirror the light falls on the surface from the side opposite to the centre of curvature (Fig. 4-2).



(a) Concave (b) Convex
Fig. 4-2
Spherical mirrors

The geometric centre of the reflecting surface is **called the** pole (*P*) of the spherical mirror. The straight line, **passing** through the pole and the centre of curvature of the **mirror, is** called the **principal axis** of the spherical mirror.

PRINCIPAL FOCUS

A narrow beam of parallel rays, parallel and close to the principal axis, would, after reflection, converge to a fixed point on the axis, in the case of a concave mirror or would appear to diverge from a fixed point on the axis in the case of a convex mirror. This point is known as the principal focus of the mirror.

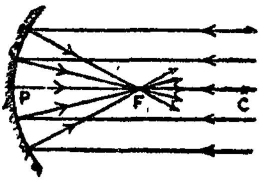


Fig. 4-3 (a)
Concave mirror :
Parallel rays
converge to focus

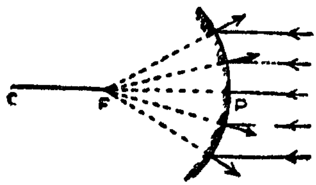
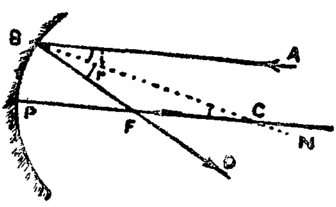


Fig. 4-3 (b)
Convex mirror:
Parallel rays
appear to diverge from F

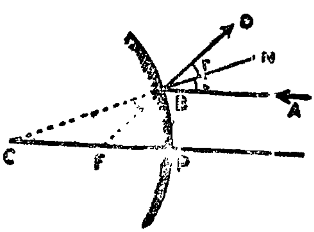
The distance of the focus from the pole is the **focal length** of the spherical mirror.

4.2.2. RELATION BETWEEN THE FOCAL LENGTH AND RADIUS OF CURVATURE OF A SPHERICAL MIRROR

Let P be the pole of a spherical mirror, C its centre of curvature, and F its focus (Fig. 4-4).



(a) Concave mirror



(b) Convex mirror

Fig. 4-4 Relation between f and r

Let the radius of curvature PC be 'r' and the focal length PF be 'f'

Let AB be a ray of right parallel and close to the principal axis CP.

BCN is the normal to the mirror at A and BD is the reflected ray.

$$\therefore \quad \angle ABN = \angle NBD \qquad (i = R)$$

$$\text{But} \quad \angle ABN = \angle BCF \; (\because AB \text{ and } CP \text{ are parallel})$$

$$\therefore \quad \angle NBD = \angle BCF \quad \text{(Fig. 4-4a)}$$

$$= \angle CBF \quad \text{(Fig. 4-4b)}$$

$$\text{In the } \triangle BCF, \quad \angle CBF = \angle BCF$$

$$\therefore \quad FC = FB$$

$$= FP \; (\because B \text{ is close to } P)$$

$$\text{Now,} \quad CP = CF + FP = 2\,FP = 2f$$

$$\text{i.e.,} \quad r = 2f$$

Hence the radius of curvature of a spherical mirror is twice the focal length.

4.2.3. IMAGE FORMATION BY A CONCAVE MIRROR (Graphical Method)

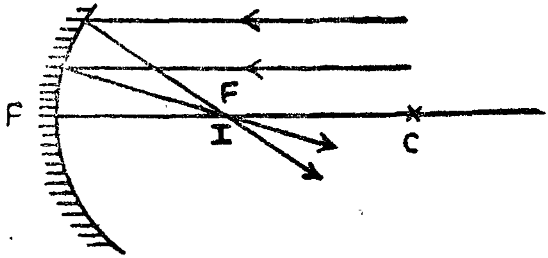
To locate the position of an object graphically, we make use of the following properties or the rays :

(1) A ray incident parallel to the principal axis, after reflection from the mirror passes or appears to pass through the principal focus 'F'.

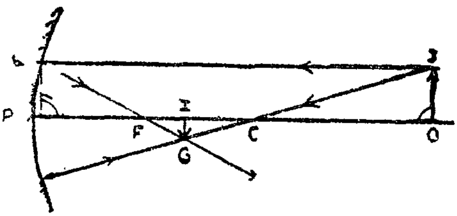
(2) A ray through the centre of curvature (C), being normal to the mirror, is reflected back along the same line.

(3) A ray through the principal focus (F) after reflection from the mirror, passes parallel to the principal axis.

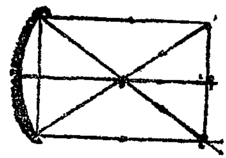
In general the first two rays are sufficient. The positions and characteristics of the images formed when the object is placed at various distances from a concave mirror are indicated in Fig. 4-5.



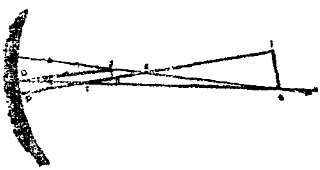
(a) Object at infinity



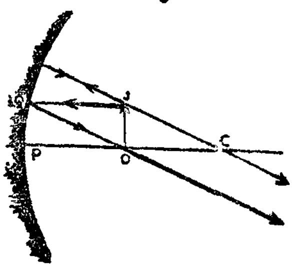
(b) Object beyond C



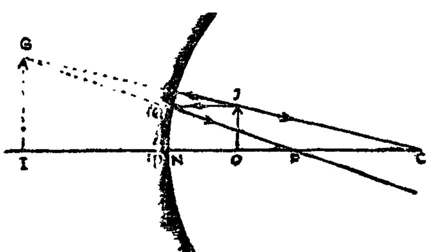
(c) Object at C



(d) Object between F and C



(e) Object at F



(f) Object between F and P

Fig. 4-5 Image formation by a concave mirror

The following table summarises the results obtained with reference to a concave mirror :

Position of object	Image			Fig
	Position	Nature	Size	
At infinity	At F	Real	Point size	4·5 *a*
At definite distance beyond C	Between C and F	Real and inverted	Diminished	4·5 *b*
At C	At C	Real and inverted	Same size	4·5 *c*
Between C and F	Beyond C	Real and inverted	Enlarged	4·5 *d*
At F	At infinity	—	—	4·5 *e*
Between F and P	Behind mirror	Virtual and erect	Enlarged	4·5 *f*

4.2.4. IMAGE FORMED BY A CONVEX MIRROR

In this case, the image is always formed behind the mirror between the pole and the focus. The image is always virtual, erect and diminished (Fig. 4-6).

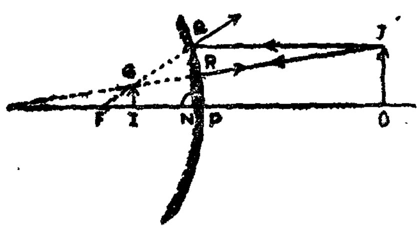


Fig. 4-6
Image formed by a Convex mirror

When the object is at an infinite distance, the image is formed at the focus 'F'. As the object moves from infinity towards the pole, the image moves from the focus towards the pole. The movement of the image is very small; it appears almost stationary. A convex mirror is therefore used by motorists to see the vehicles coming behind them. The vehicles, at a very great distance behind are seen as very tiny images. They become gradually bigger as they approach nearer.

4.2.5. RELATION BETWEEN u, v AND f

In the theory of mirrors, it is found convenient to use certain conventions of signs.

(1) All distances are measured from the pole of the mirror.

(2) Distances of real objects and real images are positive ; whereas distances of all virtual objects and virtual images are negative.

(3) Since the focus of a concave mirror is real, its focal length is positive, and since the focus of a convex mirror is virtual, its focal length is negative.

The distances of the object from the pole is denoted by the letter 'u', that of the image from the pole by 'v'.

Let OJ be an object perpendicular to the axis PC at a distance u from the mirror and IG its image formed by a spherical mirror, at a distance v from the mirror, the image being constructed in the usual manner. JQ is a ray parallel and close to the axis, incident on the mirror at Q.

From Q draw QN perpendicular to the axis. If Q is near the pole, P then N will be very close to P and the distances

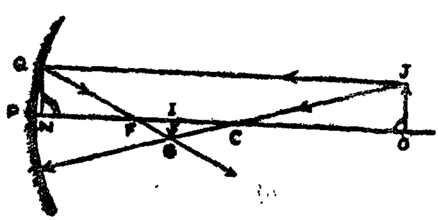


Fig. 4-7 (a)
Concave mirror forming real image

measured from N will differ very little from distances measured from P.

In all cases of image formation by spherical mirrors (Figs. 4-7a, 4-7b, and 4-7c) $\triangle$ *OJC* and $\triangle$ *IGC* are similar

$$\therefore \quad \frac{OJ}{IG} = \frac{CO}{CI}$$

Similarly, in the two $\triangle s$ *QNF* and *GIF*

$$\frac{NQ}{IG} = \frac{NF}{FI}$$

But $NQ = OJ$ ($\because$ *JQ* is parallel to *ON*)

$$\frac{CO}{CI} = \frac{NF}{FI} = \frac{PF}{FI} \quad (1) \quad (\because N \text{ is close to } P)$$

Case I: Real image formed by a concave mirror

In this case u, v and f are all positive.

$CO = PO - PC = u - 2f$

$CI = CP - IP = 2f - v$

$NF = PF = f$

$FI = PI - PF = v - f$

Hence equation (1) becomes

$$\frac{u-2f}{2f-v} = \frac{f}{v-f}$$

i.e. $2f^2 - vf = uv - uf - 2vf + 2f^2$

$vf + uf = uv$

$uf + vf = uv$

Dividing throughout by uvf, we get

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

Case II: Virtual image formed by a concave mirror

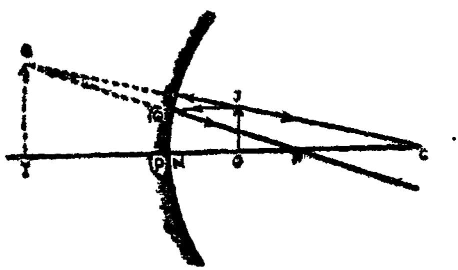


Fig. 4-7 (b)
Concave mirror forming virtual image

In this case u and f are positive and v is negative. In Fig. 4-7*b* we find

$$CO = CP - OP = 2f - u$$
$$CI = CP + PI = 2f + (-v) \quad (\because v \text{ is negative})$$
$$= 2f - v$$
$$PF = f$$
$$FI = FP + PI = f + (-v)$$
$$= f - v$$

Hence equation (1) becomes

$$\frac{2f-u}{2f-v} = \frac{f}{f-v}$$

i e $\quad 2f^2 - vf = 2f^2 - 2vf - uf + uv$

$$uf + vf = uv$$

Dividing throughout by uvf, we get

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

Case III : Virtual image with a convex mirror

In this case u is positive and v and f are negative.

From Fig. 4-7 c. $OC = OP + PC = u + (-2f)$
$$= u - 2f$$

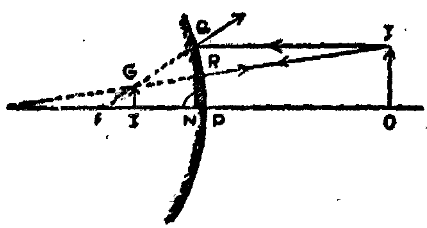


Fig. 4-7 (c)
Image formed by a convex mirror

$$IC = PC - IP = -2f - (-v)$$
$$= v - 2f$$
$$PF = -f$$

and $IF = PF - IP = -f - (-v) = v - f$

Hence equation (1) gives

$$\frac{u-2f}{v-2f} = \frac{-f}{v-f}$$

$$-vf + 2f^2 = uv - uf - 2vf + 2f^2$$

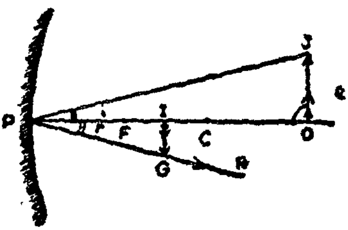
$$uf + vf = uv$$

Dividing throughout by uvf, we get

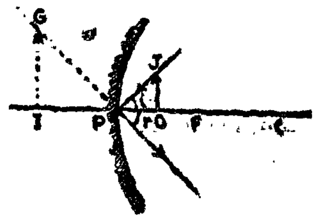
$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

4.2.6. MAGNIFICATION

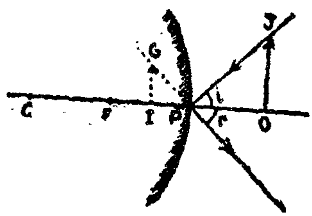
The ratio of any linear dimension (length and breadth) of the image to the corresponding linear dimension of the object is called magnification



(a)
Concave mirror forming real image



(b)
Concave mirror forming virtual image



(c)
Fig. 4-8
Convex mirror forming an image

In Fig. 4-8 let OJ be an object and IG its image. A ray from the top J of an object after reflection at P passes through the point G or appears to come from G, of the image. Since OP is the normal to the mirror at the point P.

$\angle IPG = \angle OPJ$ (being the angle of reflection and angle of incidence respectively at P)

But $\angle PIG = \angle POJ = 90°$

$\therefore$ $\triangle$ *IPG* and $\triangle$ *OPJ* are similar.

$$\therefore m = \frac{\text{image height}}{\text{object height}} = \frac{IG}{OJ} = \frac{PI}{PO} = \frac{v}{u}$$

Hence magnification $m = \dfrac{v}{u} = \dfrac{\text{Distance of image from mirror}}{\text{Distance of object from mirror}}$

4.2.7. DETERMINATION OF FOCAL LENGTH OF A CONCAVE MIRROR

(a) Distant Object Method

The concave mirror is fixed on a vertical stand and directed towards a distant object. A clear well-defined image of the object is made to fall on a white screen. The distance between the mirror and the screen gives the value of the focal length of the concave mirror.

(b) u-v Method

In this method, a circular wire-gauze illuminated by a candle or by an electric lamp serves as the object. The concave mirror and the screen are arranged as shown in Fig. 4-9.

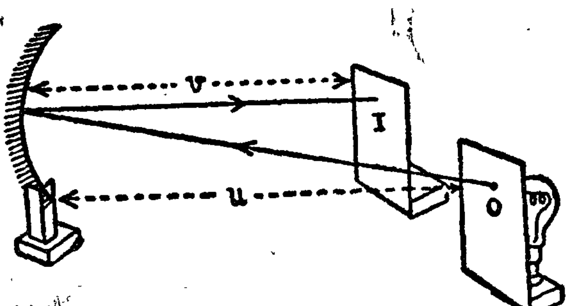


Fig. 4-9
Focal length of concave mirror : u-v method

The pole of the mirror and the centre of the wire-gauze must be arranged to be at the same height from the table.

With the object at a fixed distance ($u > f$) from the mirror, the distance between the mirror and the screen (v) is adjusted to get the clear image of the object. The distances between the mirror and the source (u) and the mirror and the screen (v) are measured and tabulated as shown below. The focal length of the given mirror is calculated in each case using the formula $f = \frac{uv}{u+v}$

Tabular column

No.	u	v	$f = \frac{uv}{u+v}$

A graph can be drawn representing u on the X-axis and v on the Y-axis. Both the axes should have the same scale and the origin should have the same value on both the axes. The bisector of $\angle XOY$ is drawn to intersect the graph at Q. Corresponding

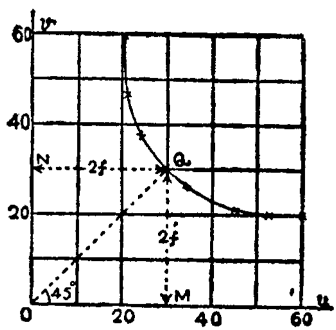


Fig. 4-10
u - v graph

to the point Q on the graph, the values of u and v are equal. In a concave mirror, this is possible only if $u = v = r = 2f$. Hence OM or ON in the graph gives the radius of curvature r and $f = \frac{r}{2}$.

(c) Normal Reflection Method

The concave mirror is mounted on a stand facing the illuminated wire-gauze. The distance of the mirror is adjusted so that a sharp image of the object is obtained very close to the object (Fig. 4-11) itself. It will also de found that the size of the image is nearly equal to the size of the object.

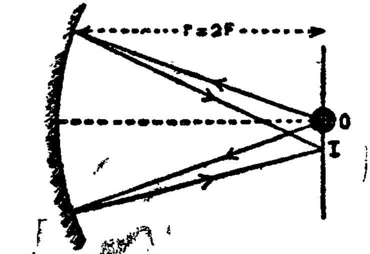


Fig. 4-11

Concave mirror : normal reflection method

The distance between the object and the pole of the mirror gives the radius of curvature 'r' of the mirror. The focal length of the concave mirror is half of the radius of curvature.

$$\text{i.e. } f = \frac{r}{2}$$

4.2.8. DETERMINATION OF THE FOCAL LENGTH OF A CONVEX MIRROR

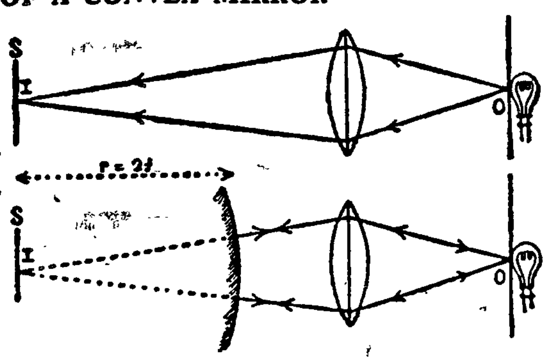


Fig. 4-12

Focal length of convex mirror

The object 'O' is an illuminated wire-gauze. A short focus convex lens is used to get an image I on the screen S. The position of the screen is adjusted to get a clear image.

The given convex mirror whose focal length is to be determined is now placed between the lens and the screen as shown in Fig. 4-12 and is moved to and fro to get the clear image of the object close to O itself. The path of the rays is shown in Fig.4-12. It is easy to see that the rays incident on the mirror should be normal at places of incidence. Hence I will give the position of the centre of curvature. The distance between the pole of the mirror and the screen gives the radius of curvature of the convex mirror. The focal length of the mirror is half of this value.

Exercise 4.2

1. What is a spherical mirror ?
2. Define the following terms with reference to a spherical mirror.
   pole, centre of curvature, radius of curvature, principal axis, principal focus, focal length.
3. Draw diagrams showing the formation of real images diminished and enlarged with a concave mirror.
4. Draw diagrams showing the formation of virtual image (i) with a concave mirror (ii) with a convex mirror
5. Describe an approximate method of determining the focal length of a concave mirror.
6. Describe an experiment to determine accurately the focal length of a concave mirror.
7. Mention any two rays proceeding from an object which are sufficient to locate the position of its image formed by a spherical mirror.
8. Derive the relation connecting u, v and f of a concave mirror when it forms a real image.
9. Derive the relation connecting u, v and f of a concave mirror when it forms a virtual image.
10. How do you distinguish between a plane mirror and a concave mirror ?
11. How do you distinguish petween a concave mirror and a convex mirror ?
12. At what distance should a man hold his shaving mirror in order to see an enlarged image of his face ?

4.3. REFRACTION OF LIGHT

4.3.1. REFRACTION AT A PLANE SURFACE

When light passes from one transparent medium to another, the ray suffers a deviation in its path. This is called refraction. The ray in the first medium is usually referred to as the incident ray and the ray in the second medium the refracted ray. The angle between the incident ray and the normal to the surface of separation at the point of incidence is called the **angle of incidence** (i) and the angle between the refracted ray and the normal to the surface of separation at the point of incidence is called the **angle of refraction** (r)

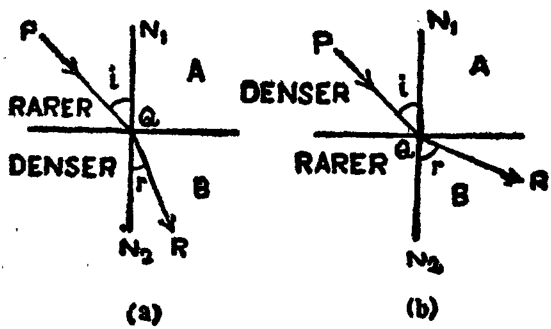


Fig. 4-13
Refraction at a plane surface

If a ray travelling from medium A to medium B is bent towards the normal on refraction, i.e. if the angle of refraction is smaller than the angle of incidence, it is said that B is *optically denser* than A (Fig. 4-13 a). But if the angle of refraction is larger than the angle of incidence (Fig. 4-13 b) the medium B is said to be *optically rarer* than the first medium A.

4.3.2. LAWS OF REFRACTION

(1) *The incident ray, the refracted ray and the normal to the surface at the point of incidence all lie in one plane.*

(2) *The ratio of the sine of the angle of incidence to the sine of the angle of refraction is a constant for a given pair of media and for light of a given colour* (This is known as *Snell's law*)

According to the second law $\frac{\sin i}{\sin r} = \mu$

where μ is a constant for a given pair of media and for light of a given colour. This constant is called the refractive index of the second medium with respect to the first. When a ray undergoes refraction from air (or a vacuum) to a second medium, the ratio $\frac{\sin i}{\sin r}$ is called the refractive index of the second medium.

If the incident ray is normal to the surface of separation, the angle of incidence i is zero. Then by the equation $\sin i = \mu \sin r$, the angle of refraction r is also zero. In this case the ray does not undergo any deviation.

4.3.3. REFRACTION THROUGH A PRISM

A *prism* is a *portion of a refracting medium bounded by two plane surfaces which are inclined to each other at an angle*. The two plane surfaces are called the *refracting faces* of the prism, the line along which the faces meet is called the *edge* of the prism. The angle between the refracting faces is called the *angle* of the prism. The surface opposite to the edge of the prism is called the *base* of the prism. Any section of the prism by a plane perpendicular to the faces of the prism is called its principal section.

In Fig. 4-14, *ABC* represents a principal section of the prism. One of the angles, say *A*, is chosen as the refracting angle of the prism.

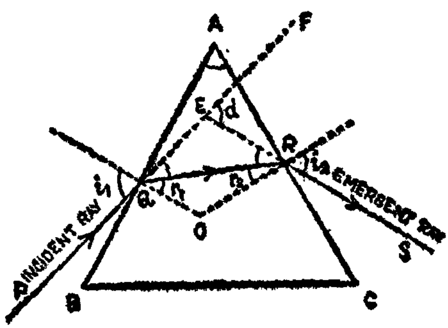


Fig. 4-14
Refraction through a prism

Let *PQ* be a ray of light incident at *Q* on the face *AB* of the prism at an angle of incidence i_1. It is called the incident ray.

On entering the prism the ray passes from air (a rarer medium) to glass (a denser medium). Hence it bends towards the normal drawn at Q. QR shows the path of the ray through the prism. At R the ray suffers refraction once again and emerges into air. RS is called the emergent ray. Here it bends further towards the base. Why?

Let i_1 and r_1 be the angles of incidence and refraction at Q, and let the angles of incidence and refraction at R be r_2 and i_2.

Let A be the angle of the prism.

The angle $\angle FES$ between the incident ray PQ and the emergent ray RS in Fig. 4·14. is called the *angle of deviation d*.

In the $\triangle QER$,

the exterior angle $\angle FES = d = \angle EQR + \angle ERQ$

$$= (i_1 - r_1) + (i_2 - r_2)$$
$$= (i_1 + i_2) - (r_1 + r_2)$$

In the quadrilateral $AQOR$, since the angles at Q and R are right angles,

$$A + \angle QOR = 180°$$

But in $\triangle QOR$, $\quad r_1 + r_2 + \angle QOR = 180°$

$$\therefore \quad r_1 + r_2 = A$$
$$\therefore \quad d = i_1 + i_2 - A$$
$$\text{or} \quad A + d = i_1 + i_2$$

4.3.4. EXPRESSION FOR REFRACTIVE INDEX

The angle of deviation 'd' depends upon

1. the angle of incidence
2. the angle of prism
3. the refractive index μ, of the material of the prism with respect to the medium that surrounds it and
4. the wavelength λ or the colour of the incident light.

For a given prism A and μ, fixed and for light of a given wavelength, d depends only on the angle of incidence i.

As the angle of incidence i gradually increases, the angle of deviation 'd' decreases, reaches a minimum value (D) and then increases. D is called the *angle of minimum deviation.* It will be seen from the graph that there is only one angle of incidence for which the deviation is a minimum.

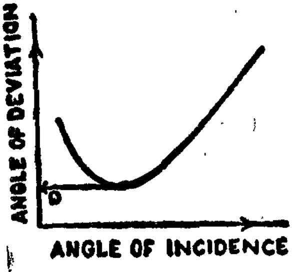


Fig. 4-15
i-d curve for a prism

It can be shown easily that rays of light pass through a prism symmetrically in the minimum deviation position i. e. in this position

$$i_1 = i_2 \text{ and } r_1 = r_2$$

and the refracted ray QR lies parallel to the base of the prism. If the path of the ray of light $PQRS$ be reversed by placing a plane mirror perpendicular to the rays RS, it will retrace its course and the angle of deviation will still be a minimum for an angle of incidence i_2,

At the angle of minimum deviation

$$i_1 = i_2 = i$$

and $\quad r_1 = r_2 = r$

$$2r = A \text{ or } r = \frac{A}{2}$$

and $\quad 2i = A+D \text{ or } i = \frac{A+D}{2}$

$$\therefore \mu = \frac{\sin i}{\sin r} = \frac{\sin \dfrac{A+D}{2}}{\sin \dfrac{A}{2}}$$

4.3.5. LENSES

A lens is a transparent medium bounded by curved surfaces. If the surfaces are parts of spheres, they are called spherical lenses. As shown in Fig. 4-16, there is a great variety of lenses.

However they can be divided into two broad groups (i) those that are thicker in the middle than at the edges, called **convex**

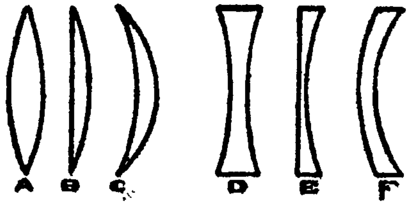
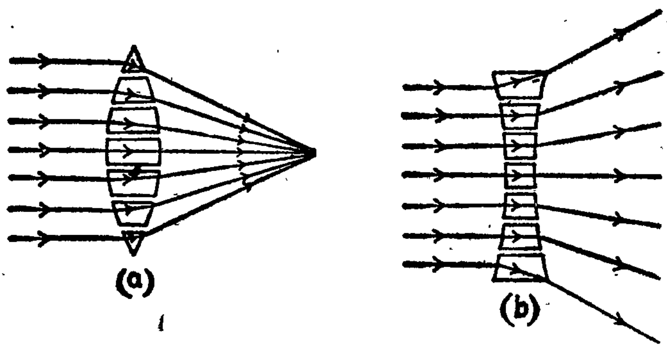


Fig. 4-16
Convex and concave lenses

(converging) lenses and (ii) those that are thinner in the middle than at the edges called **concave** (diverging) lenses.

When a lens is bounded by two spherical surfaces, the line joining the centres of curvature is the **principal axis** of the lens.

To understand the function of a lens, it can be supposed to be made up of a set of truncated prisms in contact, as shown in Fig. 4-17



Refraction through a convex lens

Refraction through a concave lens

Fig. 4-17

We have seen that light entering a prism from a rare medium is always bent towards the base (or the thicker side) of the prism after refraction through it. It is therefore clear that rays of light starting from an object will be bent towards the axis and converge to a point in the case of a convex lens while they will be bent away from the axis or diverage, in the case of a concave lens. This explains why a concave lens is a *converging* lens while a concave lens is a *diverging* lens.

4.3.6. OPTIC CENTRE

If a ray of light passes though a lens in such a way that it emerges parallel to its original direction, the path of the ray in the lens (or the path produced) intersects the axis in a fixed point; this point is called the **optic centre** *of the lens* (Fig. 4-18) If the lens is thin, the lateral displacement between the incident and

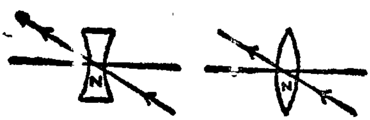


Fig. 4-18
Optic centre

emergent rays which are parallel, will be negligible and so the ray passing through the optic centre may pe considered to emerge along the same line as the incident ray. In our study, in this section, we shall consider only *thin* lenses. The optic centre in a thin lens may be assumed to be at the centre of symmetry.

4.3.8. PRINCIPAL FOCUS

If a narrow beam of parallel rays, parellel to the axis, fall on a

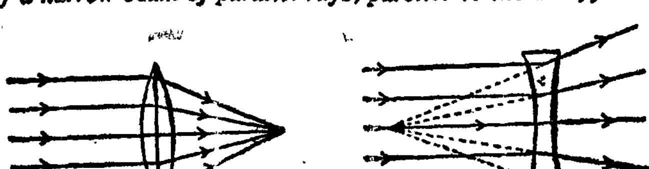


Fig. 4-19
(a) Principal focus of convex lens
(b) Principal focus of concave lens

lens centrally, it will after refraction through the lens, converge to a point on the axis in the case of a convex lens, or appear to diverge *from a point on the axis in the case of a concave lens. This point is called the principal focus of the lens* (Fig. 4-19). The distance between the focus and the optic centre of the lens is called the **focal length** of the lens.

It must be noted that each lens has two principal foci, one on either side, since light may fall on a lens from either side to get refracted.

4.3.8. IMAGE FORMATION BY A LENS (Graphical Method)

The following properties of rays are made use of to locate the position of the image graphically:

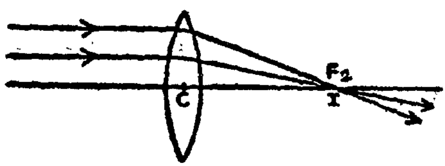
(1) A ray of light from the object travelling parallel to the principal axis passes though the principal focus in the case of a convex lens, and appears to proceed from the principal focus in the case of a concave lens, after transmission through the lens.

(2) A ray through the optic centre of a lens emerges out undeviated, if the lens were thin.

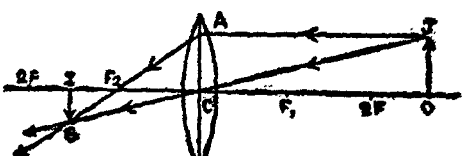
(3) A ray though the principal focus, after refraction through the lens, emerges parallel to the axis.

In general, the first two rays are sufficient.

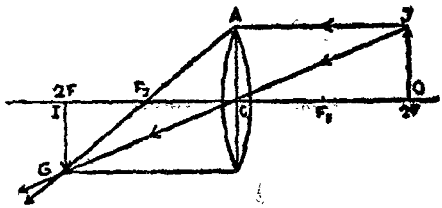
The following diagrams (Fig. 4-20) illustrate the formation of the image when the object is placed at different positions along the axis of the lens.



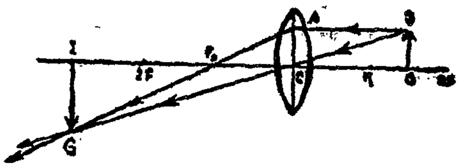
(a)
Object at infinity



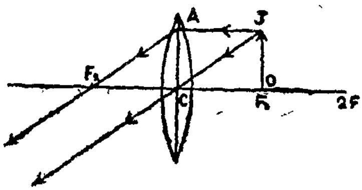
(b)
Object beyond 2F
Fig. 4-20

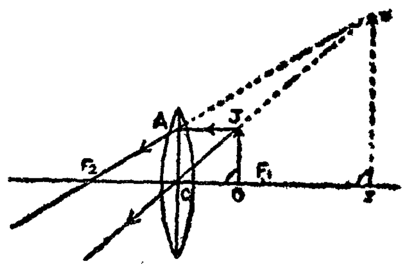


(c)
Object at 2F



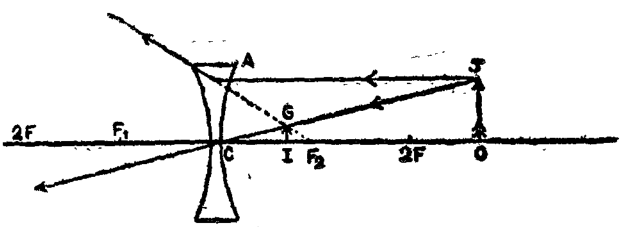
(d)
Object between F and 2F





(e)
Object at F

(f)
Object between F and C



(g)
Image formed by a concave lens
Fig. 4-20

The following table summarises the results obtained.

Type of lens	*Position of object*	*Position of image*	*Description of image*		*Fig.*
			Nature	*Size*	
Convex	At infinity	At F_2 (Other side)	Real & inverted	Point	4-20 *a*
,,	Beyond $2F$	Between F_2 and $2F$ (Other side)	,,	Diminished	4-20 *b*
,,	At $2F$	At $2F$ (Other side)	,,	Same size	4-20 *c*
,,	Between F_1 and $2F$	Beyond $2F$ (Other side)	,,	Enlarged	4-20 *d*
,,	At F_1	Rays parallel. No image—At infinity	—	—	4-20 *e*
	Between F_1 and C	Same side as object	Virtual & erect	Enlarged	4-20 *f*
Concave	Any position	,,	,,	Diminished	4-20 *g*

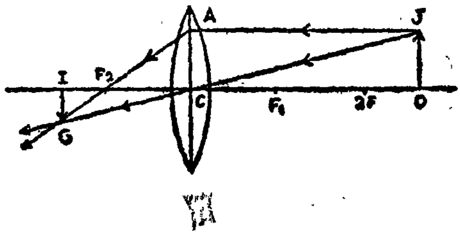
4.3.9. RELATION BETWEEN u, v and f

The following sign conventions are used :

1. u and v are distances of the object and the image measured from the optic centre of the lens.
2. Distances of real objects and images are considered positive and those of virtual objects and images, negative.
3. The focal length of a convex lens is positive and that of a concave lens is negative.

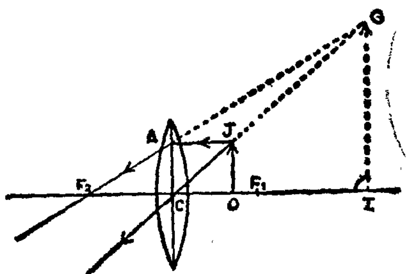
In Fig. 4-20 *g* and 4-21 *OJ* is the object and *IG* is the image (the image being constructed in the usual manner)

Let u be the distance of the object and v the distance of the image from the optic centre of the lens.



(a)

Convex lens forming real image



(b)

Fig. 4-21

Convex lens forming virtual image

Let F_1 and F_2 be the two principal foci of the lens and f its focall ength

In all the three cases, Fig. 4-21 a, b and 4-20 g triangles JCO and GCI are similar.

$$\frac{OJ}{IG} = \frac{CO}{CI}$$

In the two similar triangles $AF_2\,C$ and GF_2I

$$\frac{CA}{IG} = \frac{F_2C}{F_2I}$$

But $CA = OJ$ in the parallelogram $CAJO$.

$$\therefore \frac{CO}{IC} = \frac{F_2\,C}{F_2I} \qquad (1)$$

Case I : Real image formed by a convex lens (Fig. 4-21 a)

In this case u, v and f are all positive

Now $OC = u$
$IC = v$
$F_2C = f$
$F_2I = CI - CF_2 = u - f$

$\therefore$ Equation (1) becomes

$$\frac{u}{v} = \frac{f}{v-f}$$

$$uv - fu = vf$$
$$uf + vf = uv$$

Dividing throughout by uvf, we have

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

Case II : Virtual Image formed by a convex lens (Fig. 4-21b)

Since the object is real and the image is virtual, u and f are positive and v is negative.

Now $OC = u$
$IC = -v$
$F_2C = f$
$F_2I = CI + CF_2 = -v+f = f-v$

$\therefore$ Equation (1) can be written as

$$\frac{u}{-v} = \frac{f}{f-v}$$

ie, $uf - uv = -vf$
$uf + vf = uv$

Dividing throughout by uvf, we get

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

Cas III : Virtual image formed by a concave lens (Fig.4-20g)

In this case u is positive since the object is real, v is negative since the image is virtual and f is negative since it is a concave lens.

Now $OC = u$

$IC = -v$

$F_2C = -f$

$F_2I = CF_2 - CI = -f - (-v) = v - f$

Hence equation (1) becomes

$$\frac{u}{-v} = \frac{-f}{v-f}$$

$$uv - uf = vf$$

$$uf + vf = uv$$

Dividing throughout by uvf, we have

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

4.3.10. MAGNIFICATION

Magnification is the ratio of the linear size of the image formed to the linear size of the object.

In Figs. 4-21 *a, b* and 4-20 *g*, the ratio $\frac{IG}{OI}$ gives the magnification.

But $\frac{IG}{OJ} = \frac{CI}{CO} = \frac{v}{u}$

Thus the magnification due to a lens is equal to the ratio of the distance of the image to that of the object from the lens.

4.3.11. DETERMINATION OF FOCAL LENGTH OF A CONVEX LENS

(*a*) DISTANT OBJECT METHOD

The image of a distant object like a building or a tree may be formed by the convex lens on a screen kept on the other side. By moving it to and fro, a sharp image may be obtained. The distance from the centre of the lens to the screen gives the focal length of the lens.

This method is a convenient one for determining approximately the focal length of a convex lens.

(*b*) u-v METHOD

The optic bench may be conveniently used for many experiments on lenses. A scale is fixed along the length of the bench to enable the positions of the supports to be noted.

A circular wire-gauze is illuminated with an electric lamp. This constitutes the object. The convex lens is arranged with the object on its principal axis. The screen is brought into position on the other side of the lens and a sharp image of the object is obtained. The values of u and v may be measured with the scale for several sets of readings. The readings may be tabulated as shown below :

TABULAR COLUMN

No.	u	v	$f = \frac{uv}{u+v}$

The mean focal length of the convex lens is thus determined.

(c) u-v GRAPH METHOD

With the readings taken in the previous experiment, a graph is drawn with the value of u on the X-axis and v on the Y-axis It is necessary that the scale starts from the same value at the

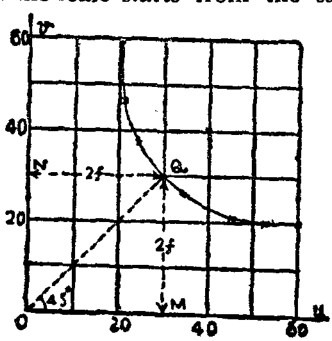


Fig. 4-22
u-v graph

origin and that the same scale is taken on both axes. The bisector of the angle XOY cuts the graph at the point Q. At this point $u = v$, so that from the equation,

$$\frac{1}{v}+\frac{1}{u}=\frac{1}{f} \text{ we get}$$

$$\frac{2}{u}=\frac{1}{f}$$

$$\therefore f=\frac{u}{2}$$

Hence OM and ON corresponding to the point Q are measured, and their average is found out. Half of this value gives the focal length of the lens.

(d) PLANE MIRROR METHOD

Fig. 4-23 illustrates the method adopted. The rays of light starting from O pass through the lens L and are incident on a plane mirror M placed almost normally to the principal axis of the lens.

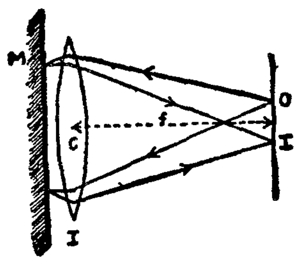


Fig. 4-23
F of Convex lens using a plane mirror

The object is arranged to lie on the axis. The lens is moved to and fro. At one position, a clear image of the object may be seen very close to it. Slight tilting of the mirror is enough to bring the image to any desired position. The image is found to have the same size as that of the object. Hence the rays of light from the object retrace their path.

The rays of light after passing through the lens should from a parallel pencil, parallel to the principal axis. This parallel beam is reflected by the mirror. The object should lie at the principal focus of the lens in order that this may happen. The distance of the object from the centre of the lens gives its focal length.

Exercise 4.3

1. What happens to a ray when it passes
   (a) from a rarer to a denser medium
   (b) from a denser to a rarer medium.
2. State the laws of refraction of light.
3. Define the terms : angle of incidence, angle of refraction and angle of deviation.
4. Define : Refractive index of a medium.
5. Draw a diagram showing the course of rays through a prism
6. Explain the term "angle of minium deviation."
7. Derive the relation $\mu = \dfrac{\sin \dfrac{A+D}{2}}{\sin \dfrac{A}{2}}$ for a prism.
8. What is a lens ?
9. Define the following terms with reference to a lens : Principal axis, Optic centre, Principal focus, Focal length.
10. Mention any two rays proceeding from an object, which are sufficient to locate its image,formed by a lens.
11. Draw diagrams showing the formation of real images formed by a convex lens.
12. Draw diagrams showing the formation of virtual image (i) by a convex lens (ii) by a concave lens.
13. State the laws of distances of a lens.
14. Describe an approximate method of determining the focal length of a convex lens.
15. Describe an experiment, to determine accurately the focal length of a convex lens.
16. Derive the relation connecting u, v and f of a convex lens when it forms a real image.
17. Derive the relation connecting u, v and f of a convex lens when it forms a virtual image.
18. Derive the relation connectiug u, v and f of a concave lens.
19. State the uses of convex and concave lenses.
20. How do you distinguish between a convex lens and a concave lens ?
21. How do you distinguish between a convex lens and a plane glass plate ?

4.4. DISPERSION OF LIGHT

4.4.1. INTRODUCTION

When a beam of white light falls on a prism, it splits up into its constituent colours. *The phenomenon of the splitting up of composite light into its constituent colours is called dispersion.* It was Newton who using a beam of sunlight first obtained a spectrum. He interpreted and identified the seven colours in it—violet indigo, blue, green, yellow, orange and red. Normally there is an overlapping of different colours of the spectrum so that the pectrum is not clear. This is known as *impure* spectrum. A *pure* spectrum is one in which the overlapping of the different colours does not occur and in which each of the different colours is clearly separated. In a spectrum formed by a prism,it is found that red has a smaller deviation than violet.

4.4.2. FORMATION OF PURE SPECTRUM

The best arrangement for producing a pure spectrum is shown in Fig. 4-25.

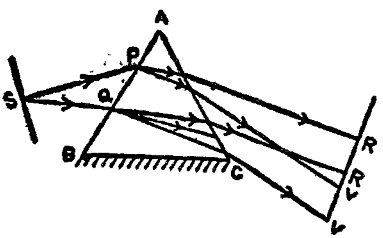


Fig. 4-24
Spectrum formed by a prism

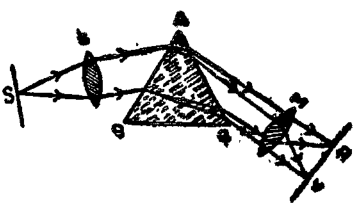


Fig. 4-25
Formation of pure spectrum

Light from the given source is made to pass through a narrow slit and fall on a lens L. The slit is adjusted to be at the focal plane of the lens, so that the rays of light after refraction through the lens are rendered parallel. Thus a narrow beam of parallel rays is incident on the prism. Considering any one of these rays, the prism produces a definite deviation for each of the colours. Since the incident beam is parallel, the deviation produced for any one colour of light will be the same for all the rays of the incident light, but will be different for the different colours. Hence in the emergent beam, rays of the same colour are parallel-for example all red rays are parallel, all violet rays are parallel and so on. If now a second convergent lens, *M* is placed in the path of the emergent beam all the red rays are brought to focus at the same point, all the yellow rays at a different point and so on. Again it is advantageous to have the prism adjusted to be in the minimum deviation position. Usually the prism is adjusted to be in the minimum deviation position for yellow, the mean wave-length. Then it will be very nearly in the minimum deviation position for rays of other colours.

Thus, the following conditions should be fulfilled to obtain a pure spectrum:

(i) The incident light must pass though a narrow slit

(ii) The rays incident on the prism must be parallel.

(iii) The emergent beam of light should be brought to focus by a converging lens and

(iv) The prism must be in the minimum deviation position.

4.4.3. CONTINUOUS SPECTRA

A spectrum in which the various colours are formed without any break is known as a continuous spectrum. In general, a continuous spectrum results when a solid or a liquid or a gas under high pressure is heated to incandescence. White hot platinum, lime, carbon etc. give continuous spectra. Candle flame also emits a continuous spectrum due to the incandescent particles of carbon present in it.

In such a spectrum though only seven colours can be recognized by the eye, there are various shades in each colour patch. One colour does not abruptly change to the next, but gradually shades into the next. Thus for example, starting from the centre of the yellow patch, the colour very gradually changes from yellow to orange at the centre of the orange patch on one side and shades very gradually to green on the other side. This means that we must consider each line in the spectrum as a separate colour and therefore the number of colours in the continuous spectrum is infinite.

4.4.4. LINE SPECTRA

If the spectrum of a gas at low pressure is examined, the appearance is entirely different. In place of a continuous band of colours, a characteristic set of distinct and separate coloured lines is seen against a dark background. In contrast with the continuous spectrum, this is called a *line* spectrum. The occurence of a line spectrum signifies the fact that the source sends out a limited set of wavelengths of light and no others.

The gas may be made to radiate light by maintaining it at high temperature, by passing a spark between metal rods or by sending a high voltage electric current through it (tubes filled with neon or other gases).

Each element has its own characteristic line spectrum. Their wavelengths will be helpful in identifying the element. Thus, for example, if sodium chloride is introduced into a flame and its spectrum examined with a powerful spectroscope, it will be found that there are two close yellow lines.

The wavelengths in nanometres of the principal lines of some elements are given below :

Sodium (Na)	Mercury (Hg)	Copper (Cu)
589·0	404·7	578·2
589·6	407·8	521·8
530·3	434·8	453·1
[illegible]	435·8	437·8
515·4	546·1	

4.4.5. SPECTROMETER

The spectrometer is one of the most useful instruments used in the study of the spectra of different sources of light and the refractive indices of materials.

The essential parts of the spectrometer are (i) the collimator (ii) the telescope (iii) the prism table and (iv) the circular scale and vernier.

(i) The **collimator** is an arrangement intended to secure parallel rays of light. It consists of a vertical slit fixed at one end of a hollow cylindrical tube which slides in another hollow cylindrical tube. The width of the slit can be varied by means of a screw. A convex lens is fixed at the opposite end of the second tube. The lens of the collimator is near the prism table. The distance between the slit and the lens can be adjusted by sliding the first tube inside the second by rotating a knob. If the slit is adjusted to be in the focal plane of the lens, rays of light diverging from any point in the slit will be rendered parallel after emerging from the lens. The collimator is fixed to the instrument with its axis perpendicular to the axis of rotation of the prism table, and cannot be moved.

(ii) The **telescope** is an astronomical one, with its axis arranged to be in the same horizontal plane as that of the collimator. It consists of an eye piece and an objective, both being convergent lenses. In front of the eye piece is fixed a cross wire consisting of two fine fibres at right angles to each other. The distance between the cross wire and the eye piece can be varied. The distance between the objective and the eye piece can also be varied to secure clear focussing. The telescope can rotate about a vertical axis coinciding with the axis of rotation of the prism table and can be clamped in any position. By means of a tangential screw the telescope, after being clamped, can be slightly moved either way for fine adjustment.

(iii) The **prism table** consists of a metal platform on which the prism is to be mounted. It is provided with three levelling

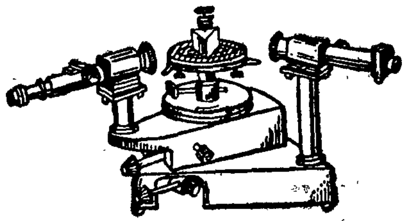


Fig. 4-26
Spectrometer

screws. It may be raised or lowered and clamped in any position by a screw and it can also be turned about a vertical axis passing through its centre and is provided with clamping and tangential screws.

(iv) The **circular scale** is attached to the telescope and the scale rotates with it. It is commonly graduated into half degrees. The prism table carries with it a vernier and thus the relative rotation between the telescope and the prism table can be measured accurately.

Adjustments of the Spectrometer

The following adjustments must be made before commencing an experiment with a spectrometer.

(*i*) *Adjustment of the eye piece* : The eye piece is adjusted so that the cross wires are distinctly seen. This can be done by turning the telescope towards an illuminated surface and sliding the eye piece to and fro until the cross wires are clearly seen. The cross wires are normally adjusted to be vertical and horizontal.

(*ii*) *Adjustment of the telescope* : The telescope is adjusted for parallel rays by turning it towards a very distant object and focussing it to get a clear image on the cross wires without parallax.

(*iii*) *Adjustment of the collimator* : The telescope is then brought in line with the collimator. The slit of the collimator is illuminated by a source of light. The distance between the slit and the lens of the collimator is adjusted until a clear image of the slit is seen at the cross wires of the telescope without parallax. Since the telescope is already adjusted for parallel rays, the well-defined image of the slit can be formed only when the light rays emerging from the collimator are parallel.

(*iv*) *Levelling the prism table* : This is done by means of a spirit level using the levelling screws.

4.4.6. DETERMINATION OF THE REFRACTIVE INDEX OF A PRISM

Angle of the prism

After making the necessary initial adjustments of the spectrometer, the prism whose refractive index is to be determined is placed on the prism table with its base perpendicular to the axis of the collimator and its edge bisecting the beam, so that the parallel rays from the collimator fall on both the faces of the

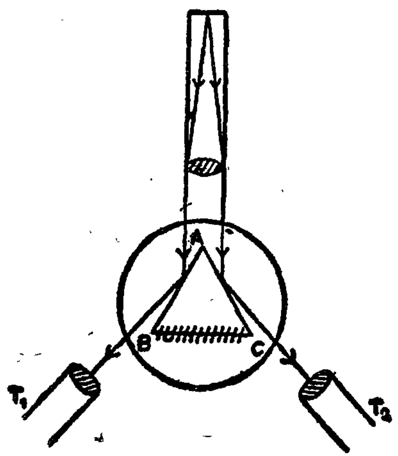


Fig. 4-27 (a)
Finding angle of the prism

prism forming the edge. The two images reflected from the two faces are first located with the naked eye. Now the telescope is brought to receive the first image in the field of view and clamped. The centre of the cross wire is made to coincide with the image by adjusting the tangetial screw of the telescope in the final position. The reading of the circular scale is read by using the vernier.

The telescope is then rotated to receive the second reflected image and the reading is taken in a similar way.

The difference between these two readings gives the angle between the reflected rays from the two faces. Half of this will be the angle of the prism (A).

Angle of minimum deviation

The prism is placed on the prism table with one refracting face facing the collimator and the refracted image is observed

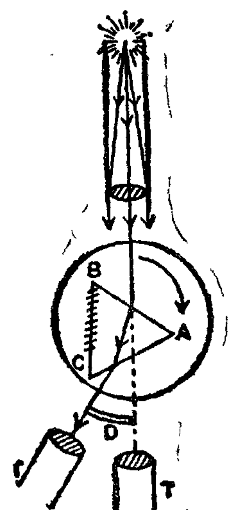


Fig. 4-27 (b)
Angle of deviation

through the telescope. The prism table is now rotated so that the refracted image moves towards the direct ray and the angle of deviation is decreased. If necessary the telescope is also rotated so as to follow the image.

It will be found that, as the prism table is rotated in the same direction, the image moves towards the direct ray upto a point and then turns back. The position of the image where it turns back is the minimum deviation position and the prism table is fixed in this position. The telescope is now adjusted so that its vertical cross wire coincides with the image and the corresponding reading is taken on the circular scale and vernier. Now the prism is removed and the telescope turned to receive the direct ray and the vertical cross wire is adjusted to coincide with the image. The reading of the scale and vernier is now noted. The difference between the two readings gives the angle of minimum deviation (D).

The refractive index of the material of the prism (μ) is calculated using the formula

$$\mu = \frac{\sin\left(\frac{A+D}{2}\right)}{\sin\frac{A}{2}}$$

Refractive index of a liquid

A hollow prism made of thin glass plate may be used to find the refractive index of a liquid. The prism is filled with the liquid, without air bubbles. The experiment is performed in the same manner as with the solid prism and the refractive index of the liquid determined.

Exercise 4-4

1. *What is dispersion ?*
2. *What is a spectrum ?*
3. *What is the cause of dispersion of white light ?*
4. *Name the colours of the spectrum. How are they arranged?*
5. *How will you synthesise the colours of the spectrum to produce white light ?*
6. *Distinguish between pure spectrum and impure spectrum.*
7. *How will you get a line spectrum in the laboratory ?*
8. *Write notes on :*

   (i) *continuous spectrum.*

   (ii) *line spectrum.*

9. *Describe the construction of a spectrometer.*
10. *What are the preliminary adjustments to be made in a spectrometer before using it.*
11. *Narrate briefly the procedure for determining the angle of the prism using a spectrometer*
12. *Narrate briefly the procedure for determining the angle of minimum deviation of a prism using a spectrometer.*
13. *Narrate briefly the procedure for determining the refractive index of the given liquid using spectrometer and a hollow prism.*

4.5. OPTICAL INSTRUMENTS

4.5.1. REFLECTING TELESCOPES

It is a common observation that when an object moves away from us it appears smaller and smaller, so that very distant objects cannot be seen clearly. A telescope is an optical instrument which helps us see distant objects. The first telescope was probably constructed in Holland in 1608 by an obscure spectacle grinder, **Hans Lippershey.** A few months later **Galileo** upon hearing that objects at a distance may be made to appear close at hand by means of two lenses, designed and made with his own hands the first authentic telescope.

The telescope is one of the most important instruments used by astronomers for direct observation of celestial objects. The most important part of a telescope is the **objective,** which collects the light from the object and produces its image which can be observed through a magnifying system which is called the **eye piece.**

Today there are two general kinds of astronomical telescopes in use. They are (1) *refracting telescopes,* which utilize lenses to produce images, and (2) *reflecting telescopes,* which utilize mirrors to produce images.

In the refracting telescope an image of the object to be observed is first produced by means of the objective (a convex lens or lenses) by refraction. The image produced is inverted. The image is observed through a smaller convex lens system called the eye piece (Fig. 4-28).

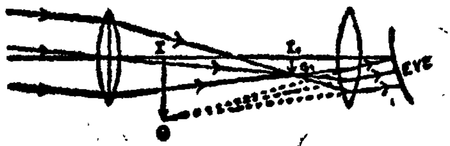


Fig. 4-28
Refracting Telescope

In astronomical work, the objective diameter is made as large as possible. This gives the telescope a good light collecting power, so that faint objects may be more easily seen. The world's largest refracting objective, at Yerkes Observatory, U.S.A., has a diameter of 1 m and a focal length of 19·8 m. Lenses bigger than this would by very expensive to construct. Large objective lenses are difficult to grind and are not easily supported without sagging, which changes the shape of the lens. The glass from which the objective is ground must be of high quality, free of bubbles and other imperfections. For these reasons all the world's largest telescopes have reflecting objectives. An objective of this sort need only have one perfect surface, and may be easily supported from behind. We also attain greater stability of the telescope by installing the large and heavier optical part (in this case the concave mirror) at the bottom of the instrument (as against the refracting type where the large objective lens has to be at the top).

The first reflecting telescope was built in 1668 by Newton. A concave mirror is used as the objective in the *Newtonian telescope*, and light incident on the mirror is reflected towards its focus. A concave mirror produces an image in front of it and therefore, in the path of the incident light. This means that any device for observing the image, will have to be in the path of the incident beam. Therefore, if the image is to be inspected visually the observer blocks out some incoming light. To avoid this the reflected light is usually diverted before it comes to a focus so that the image is formed at a more convenient location.

There are different types of reflecting telescopes. In the Newtonian type of reflecting telescopes, a small mirror is mounted diagonally to intercept the reflected light just before it reaches the focus and deflects it sideways to the eye piece (Fig. 4-29).

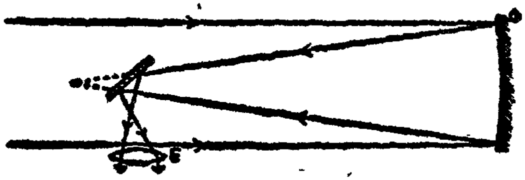


Fig. 4-29
Reflecting Telescope—Newton's Type

Another arrangement for a reflecting telescope is the *Cassegrain* system, in which a small convex mirror is used to reflect light back to a focus at the pole of the large concave mirror (which is the objective) and the eye piece is situated beyond the pole. In many reflecting telescopes a small hole is provided at the centre of the objective mirror so that light reflected from the convex mirror can form an image behind the objective (Fig. 4-30). The hole cuts out only a small percentage of the mirror's total area.

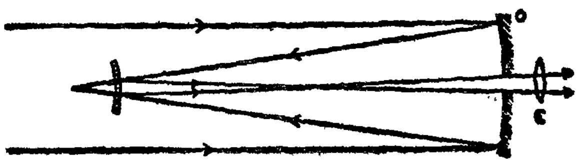


Fig. 4-30
Reflecting Telescope—Cassegrain type

Some of the world's largest telescopes in operation are the 2·54 *m*. (100 inches) diameter reflector at Mount Wilson, U.S.A. and the 5·08 *m* (200 inches) reflector at Mount Palomar, U.S.A.

4.5.2. PHOTOGRAPHIC CAMERA

For the photographic camera (Fig. 4-31) the converging lens is supported at one end of a light-proof box, with the sensitive material at the other end, at a distance just greater than the focal length of the lens. There is usually a provision for varying the distance between the lens and the film slightly, so that objects at different distances can be focussed sharply. A scale of distance will be marked on the lens mount or elsewhere.

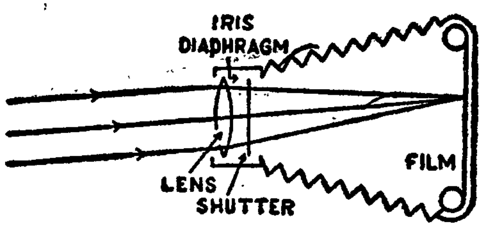


Fig. 4-31
Photographic Camera

Behind the lens, there is a shutter which can be arranged to open, for varying lengths of time, to make the exposure, perhaps 1/25, 1/50 or 1/100 of a second. A diaphragm of variable aperture, called the '*iris*' is also provided so that the amount of light entering the camera can be controlled. At large apertures, only a short exposure is needed. but the focussing must be accurate and will, in general, be sharp at one distance only. Greater depth of focus can be obtained by reducing the aperture but longer exposures will be required. The aperture is usually denoted by a fraction of the focal length of the lens, for example an aperture of 1/8 of the focal length is known as f /8, 1/n. This is called the '*f number*'.

A large '*f number*' therefore means a smaller aperture and the 'stops' are usually arranged in such a way that moving from one to the next higher '*f number*' allows about half as much light through. The exposure is proportional to the area of the aperture, while the '*f number*' is inversely proportional to its diameter. To cite an example, the light admitted at f /8 will be related to that at f /11 in the ratio 11^2 to 8^2 or 2 to 1 approximately.

It follows that an exposure of 1/50 second at f/8 will be equivalent to one of 1/25 second at f /11. Assuming that either of these would be sufficient to give a satisfactory negative with the plate or film being used, the choice would depend on other factors For example, if movement of the object is likely to take place, a shorter exposure would be preferable, while if depth of focus is the primary consideration, a smaller aperture would be used in conjunction with a longer exposure.

In recent cameras, there are automatic arrangements which will select the appropriate '*f number*' and the exposure time when a particular object is focussed.

Exercise 4.5

1. *What is the main difference between a reflecting telescope and a refracting telescope ?*
2. *Draw the diagram of a reflecting telescope of Newtonian type.*
3. *What is the main difference between the Newtonian type and Cassegrain type of reflection telescopes ?*
4. *Draw a diagram to show how an image is formed by a photographic camera.*
5. *Explain the construction of a photographic camera.*
6. *Explain* (i) *f number*
   (ii) *Exposure or exposure time.*

4.6. INFRARED AND ULTRAVIOLET RADIATIONS

4.6.1. INFRARED RADIATIONS

In 1800, Sir **William Herschel** investigated the heating effect of the different parts of the solar spectrum with a sensitive thermometer. He found that the maximum heating effect occured outside the visible spectrum, beyond the red end. He thus discovered a radiation, which is invisible, and has a longer wavelength than that of the visible light. This radiation was named as the *infrared* radiation.

The wavelength of the infrared (IR) region ranges from 10^{-6}m to 10^{-3}m and is divided into two classes, the near IR and the far IR. The near IR is called prism IR ranges from 3×10^{-6} to 25×10^{-6}m and the far IR extends from 25×10^{-6}m to 10^{-3}m.

The natural source of near IR is the sun. In IR experiments, the two most widely used sources of IR radiations are the Nernst glower and the globar. The glower consists of a filament composed of a mixture of various rare earth oxides, the most important being those of zirconium, thorium and cerium. The filament is maintained incandescent by electrical heating and run at a temperature higher than 1800 K. The globar is a rod of silicon carbide which is maintained at a constant temperature by the passage of electricity. It is more useful at longer wavelengths ($>10^{-6}$m). Carbon arc, with crater temperature of 3000 K, may also be used as a source of IR. Incandescent solids at temperatures in the range 1000 K to 1500 K will also emit IR.

Most of the IR detectors designed are based on the heating effect produced by the radiation. Some of the most important detectors of this type are the vacuum *thermocouple* and the *bolometer*. The thermocouple is based on the thermoelectric phenomenon of generation of e.m.f., which occurs when there is a temperature difference between the two junctions of two wires of different materials. The bolometer works on the principle of variation of electrical resistance of materials with temperature.

In the IR spectographs, either a prism or grating or both may be used as the dispersing medium. Prisms can be used for the near IR and this is the origin of the name *prism IR* for this region. The prism material has to be selected depending on their transparencies in different regions. For example, glass transmits to abour $2{\cdot}5\mu$ ($1\mu=10^{-6}$m) and quartz to about 4μ. For longer wavelengths, they are normally replaced (as material for the prism) by rock salt, which transmits to about 20μ.

Infrared radiations penetrate atmospheric haze (but not fog or mist) better than visible light. They are found to be most suitable for long distance photography. Infrared photography is, therefore, very convenient for aerial mapping. Water absorbs

IR radiation quite strongly, so that lakes and rivers appear black on a photograph and can be detected clearly.

The study of the *IR* absorption spectrum has been used in the analysis of chemical compounds, determination of molecular structure and the detection and identification of organic molecules. This is because many chemical compounds exhibit characteristic *IR* absorption spectra. In fact the characteristic *IR* spectra have been called the 'finger prints' of the molecule.

4.6.2. THE ULTRAVIOLET RADIATION

At about the same time when Herschel noted the extension of the spectrum beyond the red, Ritter and others found that photographic materials placed beyond the spectrum at the violet end were affected. This suggested that there was some radiation present in this region also. This radiation with wavelengths just shorter than that of the violet end of the visible spectrum is called *ultraviolet* (*UV*) radiation. It ranges from 400 nm to about 10 nm. The ultra violet radiation is emitted by any object that is hot enough (any object that is white hot for example). Normal glass is practically opaque to *UV* radiation. Glass is transparent to wavelengths down to about 300 nm. Quartz will transmit *UV* radiation down to about 180 nm and fluorspar (calcium fluoride) will allow upto 100 nm. *UV* radiation of very short wavelength is absorbed even by air, and therefore has to be studied using a vacuum spectrograph. The incandescent filament of an electric light bulb emits *UV* radiation, but this does not get out of the bulb envelope unless it is made of quartz. Mercury vapour lamps and hydrogen discharge tubes, usually in quartz. envelope, are some of the intense sources of *UV* radiation.

Ultraviolet light affects photographic plates and this property serves as a method of detecting and studying these radiations. When *UV* radiation falls on certain substances, they emit visible light and this is called *fluorescence*. In general when a radiation falling on a substance causes emission by the substance of radiation of wavelength longer than that of the incident radiation, fluorescence is said to occur. Many substances fluoresce with a characteristic visible light under *UV* radiation. For example, a solution of quinine sulphate in dilute sulphuric acid gives blue light crystal of fluorespar gives out violet light, uranium oxide gives out green light and so on. Observations of the nature and intensity of the fluorescent radiation enable tests for many purposes to be made very accurately and very quickly. For example the shells of newly laid eggs emit a rosy fluorescence, those of stale eggs blue or violet. Normal teeth look bright under ultraviolet light, whereas artificial teeth look black. Identity of fluorescence has been used to establish that samples of material are from identical sources. Different inks used in forged documents have been distinguished. In some cases the emitted light persists even after the *UV* radiation has been cut off. For example, if we place soms calcium sulphate in a beam of the *UV* radiation, it will continue to glow with a bright light for as long as an hour after the radiation is switched off. This phenomenon is called ***phosphorescence***. Another property of ultraviolet radiation is its ability to cause electrons to be emitted when it falls on some metals, a phenomenon known as the *photoelectric* effect. The above mentioned properties may be made use of to detect and study *UV* radiation.

Ultraviolet radiations have biological effects, beneficial as well as harmful. Short wavelength *UV* radiation kills bacteria. Low intensity radiation can be used to assist in the healing of wounds. When absorbed by the skin, the radiation causes the formation of vitamin D (necessary for the prevention of rickets) *UV* radiation can be harmful in large doses or high intensities. Fortunately the large amount of *UV* radiation from the sun is absorbed by the atmosphere and so at ground level the intensity is not too high to be harmful but the intensity at great heights above the ground can be dangerous.

Exercise 4.6

1. ***What is the wavelength range of infrared region ?***
2. ***Name the most commonly used sources of infrared.***
3. ***Mention the detectors used in the study of infrared.***
4. *Write a note on infrared spectrographs.*
5. *Mention any three applications of infrared.*
6. ***What is the wavelength range of the ultraviolet region ?***

7. *What are the prism materials used in the study of ultra-violet ?*

8. *What properties of the ultraviolet rays lead to their detection ?*

9. *Mention a few applications of ultraviolet rays.*

5. ATOMIC PHYSICS

5.1. DISCHARGE OF ELECTRICITY THROUGH GASES

5.1.1. INTRODUCTION

At normal pressure, air and most of the gases do not conduct electricity. However, it has been observed that the gases become conductors of electricity at low pressure, under the action of a strong electric field. In this case, an electric *discharge* takes place between two electrodes (between which a strong electric field is applied) in air or gas.

The study of electric discharge through gases gives valuable information regarding the structure of atoms. This has led to the discovery of electrons by **J.J. Thomson** and later on, to the discovery of X-rays by **Roentgen.** When we walk along the streets of a modern city in the night, we see the beautiful display of neon (bright red) and helium (pale green) advertising signs. Modern offices and homes are illuminated by light tubes. In all these cases we deal with the passage of electricity through a rarified gas. This phenomenon was the subject of life long studies by J. J. Thomson.

5.1.2. THE DISCHARGE TUBE

A simple experimental arrangement required to study the passage of electric current through a gas is known as a discharge tube. This is shown schematically in Fig. 5-1.

This consists of a strong glass tube of length about 1 to 1·5 metres and diameter about 3 to 4 cms, provided with two aluminium electrodes at the two ends. They are connected to the secondary terminals of a powerful induction coil such that A is the anode and C, the cathode. There is a side tube T which is connected to a high vacuum pump and a low pressure gauge. By reducing the pressure of the gas in the discharge tube with the pump and by applying a high potential of about 50,000 volts

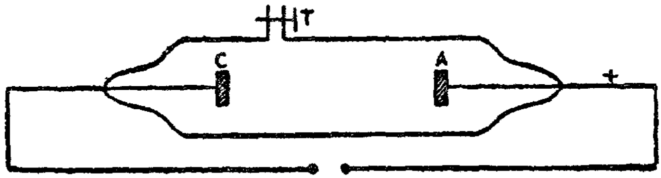


Fig. 5-1
Discharge Tube

from the induction coil, an electric discharge can be produced between the electrodes A and C.

5.1.3. DISCHARGE OF ELECTRICITY THROUGH GASES—VARIOUS STAGES

As the pressure of the gas in the discharge tube is gradually reduced by working the exhaust pump, a series of phenomena is found to take place in succession. They are discussed below and illustrated in Fig. 5-2.

(1) At pressures above 10 cm of mercury, no discharge passes through the tube. If an ammeter is iucluded in series with the electrodes, it will indicate zero current.

(2) At a pressure of about 10 cm of mercury, a cracking noise is heard which is an indication that the air has started conducting electricity. Soon irregular streaks of light appear in the tube as shown Fig. 5-2 *a*.

(3) As the pressure is reduced to the order of 10 mm of mercury, the crackling striates broaden out into a luminous column starting from the anode. This is known as the *positive column*. The colour of this column depends upon the nature of the enclosed gas. This is shown in Fig. 5-2 *b*.

(4) At a still lower pressure of about 1 mm of mercury, a dark region called the **Faraday dark space** appears near the cathode which divides the bright discharge into two parts—a long section called the positive column and a short bluish section called the *negative glow*. This is indicated in Fig. 5-2 *c*.

(5) As the pressure drops still further, the Faraday dark space grows in size and the negative glow moves away from the cathode, another dark space between it and the cathode called the

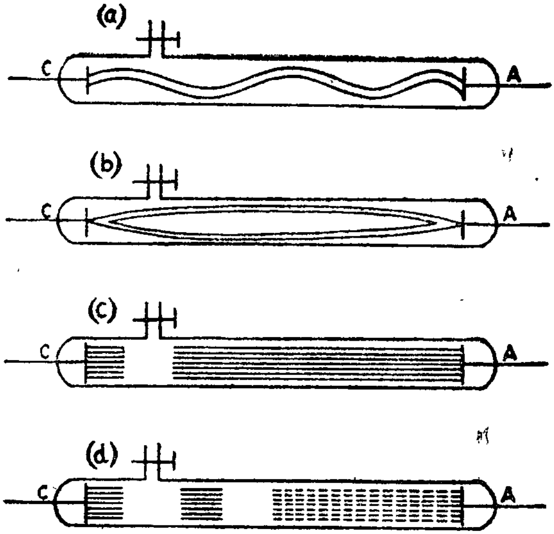


Fig. 5-2

(a) Pressure ≈ 10 mm. of Hg; Irregular streaks
(b) Pressure ≈ 20 mm. of Hg; Positive columns
(c) Pressure ≈ 1 mm. Faraday dark space of Hg
(d) Pressure ≈ 0·1 mm. of Hg; Crookes dark space and striations

Crookes dark space appears. At the cathode, a glow called the cathode glow is visible. At the same time the positive column is broken into a number of regularly spaced layers (striations) This condition is obtained at a pressure of about 0·1 mm. of mercury. This stage is given in Fig 5-2 *d*.

(6) If the pressure is reduced further, the striations and the negative glow are no longer present. At a pressure of about 0·01 mm. of mercury, the Crookes dark space widens and fills the entire tube. This state is known as the black discharge. At this stage, a new feature appears viz., the glass tube opposite to the cathode glows with a faint greenish light.

5.1.4. CATHODE RAYS

The green glow in the final stage of the gaseous discharge just explained was soon found to be a fluorescence of the glass produced by some invisible rays emanating from the cathode. These rays were called the **Cathode rays.** These cathode rays were systematically studied by many outstanding investigators like Crookes, J. J. Thomson, Perrin and others. It has been finally established that the particles constituting the cathode rays are *negatively* charged particles.

5.1.5. PROPERTIES OF CATHODE RAYS

Cathode rays were originally believed to be an "ultra gaseous state" by Crookes and to be a "fourth state" of matter by Hittorf. Later, several ingeneous experiments were carried out and the following properties of cathode rays have been established :

(i) The cathode rays travel in straight lines. This was first demonstrated by Hittorf, whose simple experiment is shown

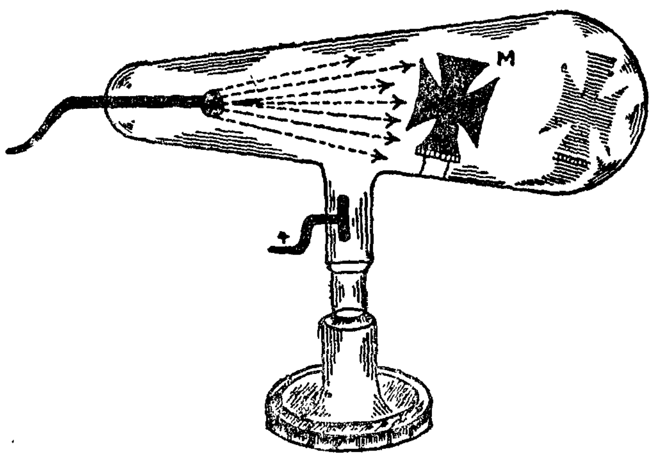


Fig. 5-3

Maltese cross in the path of cathode ra.

in Fig. 5-3. He interposed an obstacle M in the shape of a *Maltese cross* in the path of the cathode rays. A sharp shadow of the cross is cast on the screen.

(ii) Crookes proved that the cathode rays possess momentum and mechanical energy by means of a simple apparatus as shown in Fig. 5-4. Mica vanes are attached to a small wheel capable of rotating over two parallel rails. When cathode rays are incident on the vanes, because of mechanical pressure, the wheel is rotated towards the anode. This shows that cathode rays possess momentum and energy.

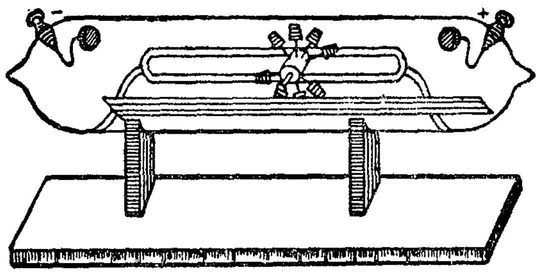


Fig. 5-4

Cathode rays possess momentum

(iii) The rays produce heat when they fall upon matter. If the cathode is concave in shape and a small piece of platinium is placed at the centre of curvature, it soon becomes white hot. This experiment also shows that the rays are emitted at right angles to the cathode surface, their direction being independent of the position of the anode.

(iv) The rays produce chemical reaction similar to the reactions produced by light. For example, certain silver salts change colour when hit by these rays. Like light, therefore, these rays can affect the photographic plate.

(v) Many minerals, glass and salts fluoresce brilliantly with characteristic colour when placed in the beam of cathode rays. **Examples :** zinc sulphide, potassium platino-cyanide etc.

(vi) These rays can penetrate through thin sheets of matter such as aluminium and gold foil.

(vii) These rays ionise gases.

(viii) When the fast-moving cathode rays are suddenly stopped by a solid obstacle, the obstacle becomes the source of X-rays.

(ix) These rays are deflected by an electric field towards the positively charged plate. This is shown in Fig. 5-5. A pair of parallel metal plates is introduced in the discharge tube. A potential is applied betw,een them such that the direction of the electric field is at right angles to the direction of the cathode rays.

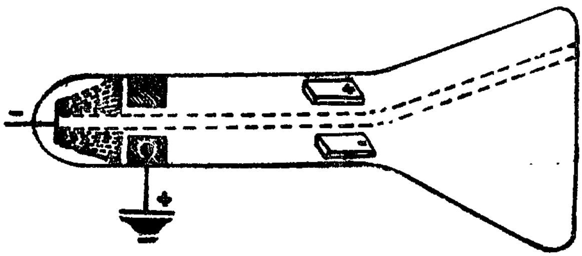


Fig. 5-5
Deflection of cathode rays by an Electric field

It is found that the rays are deflected towards the positive plates. This demonstrates that the cathode rays carry negative electric charge. Since these rays are invisible, the deflection can be observed only if the end of the discharge tube opposite to the cathode be coated with a fluorescent material. A luminious spot made by these rays will be seen on the discharge tube.

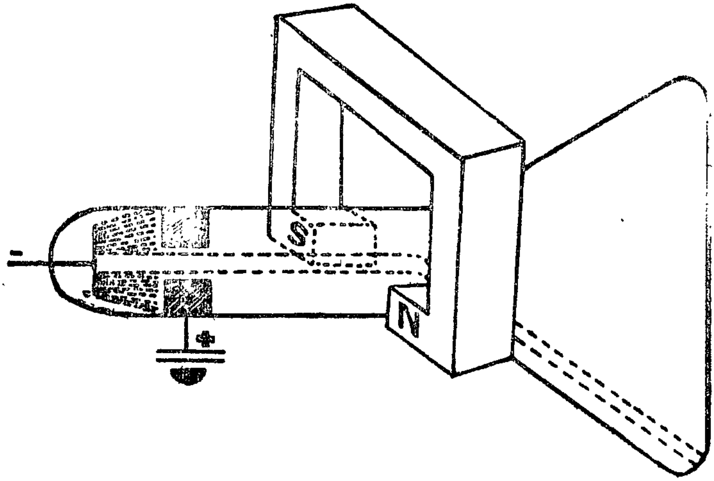


Fig.
Deflection of cathode rays by a magnetic field

(x) The cathode rays are deflectedb y a magnetic field. This is shown in Fig. 5-6. A magnetic field is applied by placing the discharge tube between the pole-pieces of an electro-magnet. The uorth pole is in front and south pole at the back. The direction of the field is at right angles to the path of the cathode rays. The cathode beam is found to be shifted vertically downwards. If the poles are reversed, the beam shifts upwards. This effect also demonstrates that the cathode rays carry negative electric charge.

5.1.6. DISCOVERY OF THE ELECTRON

An extremely important observation about cathode rays was that their properties were quite independent of the nature of the cathode material and of the gas in the discharge tube. This fact led to the conclusion that cathode rays must consist of a particular kind of particles which are present in all kinds of atoms. These particles should of course have negative charge since as already mentioned, the cathode rays carry negative charge. It became then a matter of great importance to determine the mass (m) of the particle and the amount of charge (e) carried by it. J. J. Thomson himself took the first step by determining the ratio of charge to mass of the particle from observations of the extent to which cathode rays are deflected by known electric and magnetic fields. H. A. Millikan devised a clear experiment to mea sure e. This together with the known (e/m) gave the mass m. Incidentally, it was J. J. Thomson who gave the name *electron* to the particle.

The values of m and e of the electron have been determined with great accuracy by later experiments.

$$e = 1{\cdot}6011 \times 10^{-19} \textit{ coulomb}$$
$$m = 9{\cdot}1096 \times 10^{-31} \textit{ kg}$$

Each electron is identical to every other electron; all carry exactly the same charge and have exactly the same mass.

5.1.7. ORIGIN OF CATHODE RAYS

Under the action of the electric field, some of the gas molecules in the discharge tube get ionised. *Ionisation* is simply the process of removal of one (or more) electrons from the neutral atoms or molecules. The molecules which have been deprived of any electrons become positively charged and contribute the positive ions. In the cathode-ray tube, the positive ions are accelerated towards the negative electrodes (cathode) and the electrons towards the positive electrode (anode). The current passing through the gas in the discharge tube is constituted by the motion of positive and negative charges in opposite directions.

Normally the expelled electrons soon attach themselves to neutral molecules and form negative ions. But, owing to the special conditions of ionisation at every low pressure in the discharge tube, the electrons expelled from the neutral molecules have no chance of getting attached to neutral molecules. These isolated electrons constitute cathode rays. The positive ions also give rise to a contribution to the stream of cathode rays, in the following way. As they are accelerated towards the cathode by the electric field, they gain considerable kinetic energy and strike the cathode surface with much force. The result of this bombardment by the positive ions is that the cathode emits electrons, which then get accelerated towards the anode. Thus the cathode rays or stream of electrons in the discharge tube are principally produced by (i) the process of ionisation under special conditions and (ii) the bombardment of the cathode by fast-moving positive ions.

5.1.8. APPLICATION OF THE DISCHARGE PHENO MENON IN ILLUMINATION

The phenomenon of the low pressure electric discharge has been put to use in different branches of science and industry. Perhaps, the most important of these is the use of discharge tubes as sources of light in spectral analysis and of fluorescent tubes for lighting purposes. We shall discuss here only the modern lamps of great luminosity such as (i) the sodium lamp, (ii) the mercury vapour lamp and (iii) the fluorescent lamp.

(i) SODIUM VAPOUR LAMP

This is a lamp based on 'hot cathode positive column' discharge.

The sodium vapour lamp is shown in Fig. 5-7. The discharge tube is bent in the form of a U-tube with electrodes E_1 and E_2 fused at the two ends. Some specks of metallic sodium are deposited on the inner walls of the tube. The tube also contains a very small quantity of neon gas at a pressure of about 10mms of mer-

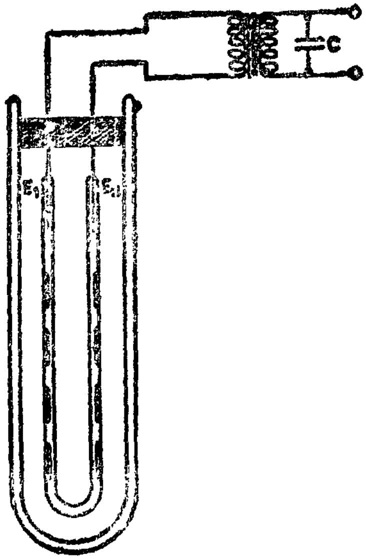


Fig. 5-7
Sodium vapour lamp

cury. This is used as a catalyst to start the discharge. The lamp requires an operating temperature of about $300°C$ for good luminosity. To minimise heat losses, the discharge tube is surrounded by a double - walled vacuum tube. For the discharge to start a voltage of about 400 volts is required which is supplied by a leak transformer.

Initially the discharge passes through neon gas and the colour of the light is red. The heat of the discharge is sufficient to vapourise the metallic sodium and thereafter the discharge passes preferentially through sodium vapour. This results in the familiar brilliant yellow light. The operating voltage of the lamp is then only 230 volts.

The sodium vapour lamp is commonly used in the laboratory as a source of monochromatic (single colour) light. It is used for lighting up show-cases in shops and public places. Sodium vapour lamps together with mercury lamps have been used for street lighting,as the two together give a more pleasant and intense light at comparatively low cost.

(ii) MERCURY VAPOUR LAMP

The mercury vapour lamp is also of the "hot cathode" type and operates at a temperature of about 600°C. The construction is similar to that of the sodium vapour lamp and shown in Fig. 5-8.

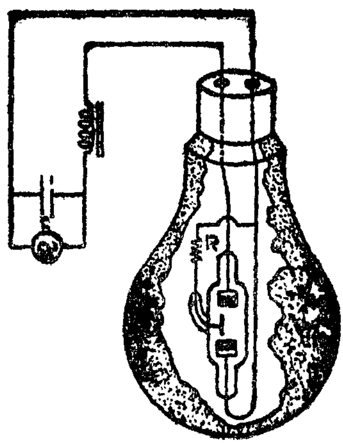


Fig. 5-8
Mercury vapour lamp

The inner vessel contains a small amount of mercury and argon gas at a pressure of 10 mm. mercury. This has two main electrodes sealed at its two ends, and also a starting electrode. This is surrounded by an outer vacuum jacket. The starting electrode helps to start the discharge through argon. As time elapses, mercury vapourises and maintains the discharge, giving a very brilliant light. It works at a lower voltage and proper choke is provided for use with it.

The light from the lamp is blue and is rich in ultraviolet, violet and green. A person looks ashen in such a light. Mercury lamps are now very commonly used for street lighting purposes,

(iii) THE FLUORESCENT LAMP

In the mercury vapour lamp, the invisible ultravoilet light goes waste. But if the outer cover of these lamps is coated on the inside with a fluorescent material such as zinc sulphate, cadmium borate etc (also called phosphors), they would convert the invisible ultraviolet light into useful visible light. Thus the fluorescent lamps are extensions of the mercury vapour lamp. By suitably varying the phosphors, any desired colour or even a

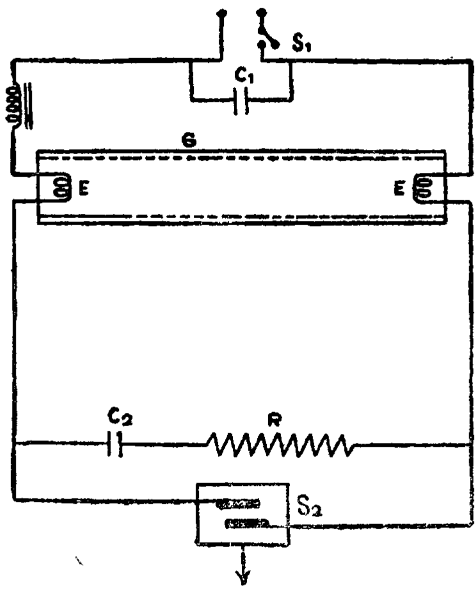


Fig. 5-9
Fluorescent lamp

mixture of colours can be obtained. The common form of the fluorescent lamp is shown in Fig. 5-9. This consists of a long glass tube *G* with a coating of fluorescent materials on its inner side.

A small quantity of mercury is introduced in the tube. The tube is then sealed off at the two ends at a pressure of about 1 mm, after introducing two electrodes *EE* made of thick tungsten wire. These electrodes are preheated by means of a starter circuit for about 2 to 3 seconds. During this time a current flows though the electrodes and makes them red-hot. Under these conditions, the discharge occurs between the main electodes.

The preheating starter S_2 is of glow-type. This consists of two metallic strips inside a glass bulb filled with helium. The pressing of the main switch S_1 causes a glow discharge to occur across the strips. This discharge heats up the strips causing them to bend towards each other and finally touch. This enables the main tungsten electrodes to be heated. The touching of the strips destroys the glow discharge. Hence the strips get cooled and separate again. However, the heating of the main tungsten electrodes results in the start of the main discharge and the lamp starts operating. The capacitor C_2 in the starter circuit is for bypassing radio interferences. The resistance R (about 100 ohms) in series with C_2 prevents the starter strips from welding together. The choke acts as a current limiter while the capacitor across the main corrects the power factor.

The fluorescent lamps are characterised by higer luminosity than ordinary filament bulbs for a given power consumption and are being increasingly used for domestic lighting purposes. They also minimise glare and thich shadows and produce uniformly diffused light. The average life of these lamps is also high. These advantages are made use of in the design of lamps for some specific purposes such as work in mines, surgical operations etc.

Exercise 5.1

1. *What is meant by discharge of electricity through gases ?*
2. (*i*) *Who discovered the electrons ?*

   (*ii*) *Who discovered the X-rays ?*
3. *Describe a discharge tube with the help of a diagram.*
4. *Describe with suitable diagrams, the behaviour of discharge through a gas, as the pressure of the gas is gradually decreased.*
5. *What are cathode rays ? Mention their important properties.*

6. *How can you experimentally demonstrate that cathode rays travel in straight lines ?*
7. *Describe the experiment to demonstrate that cathode rays possess momentum and energy.*
8. *Describe an experiment to prove that cathode rays consist of negative charges.*
9. *What is the charge of an electron ? What is its mass ?*
10. *How are cathode rays produced ?*
11. *Explain the construction and action of a sodium vapour lamp.*
12. *Describe a mercury vapour lamp with a neat diagram. Explain its uses.*
13. *Describe a flourescent lamp with a neat sketch. What are its merits and uses ?*

5.2. SOURCES OF ELECTRONS

5.2.1. INTRODUCTION

Electrons are the building blocks of all metals and in fact of all forms of matter. Even a very small quantity of matter contains enormous number of electrons. Hence it is expected that any kind of matter can be made to emit electrons under suitable conditions. For instance, cathode rays, which consist of electrons, are emitted in a discharge tube as we have already seen. The other important processes of electron emission are (*a*) *thermionic emission* and (*b*) *photo-electric emission.* The study of these processes is very useful as they have very wide practical applications.

5.2.2. THERMIONIC EMISSION

The emission of electric charges by hot metals was extensively studied by O. W. Richardson. When metals like tungsten are heated to a high temperature, by passing an electric current, they supply dense streams of electrons. It reminds one of the evolution of steam from water when heated to the boiling point. This emission is due to higher temperature and not due to electric current. Heating a metal by any other means will also produce the same effect. The emission of electrons by a hot body is known as thermionic emission.

Prof. O. W. Richardson gave the name of thermions to this subject. The ions (electrons or negative ions) emitted from hot metals are called *thermions* or *thermo electrons*. The flow of these thermions constitutes an electric current and is known as the *thermionic current*.

5.2.3. DISCOVERY

In 1885, Thomas Alva Edison, while trying to invent a satisfactory electric bulb made an accidental discovery, known as the *Edison Effect*. The arrangement to demonstrate this effect is shown in Fig. 5-10.

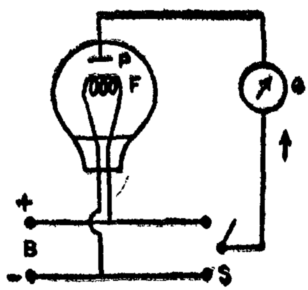


Fig. 5-10
Edison Effect

F is a carbon filament and *P* is a metal electrode; both are placed inside a highly evacuated glass bulb. *G* is a galvanometer. The filament is raised to incandescence by connecting its terminals to a battery *B*. When the switch *S* is connected to the positive end of the battery, *P* is made positive relative to the filament. The galvanometer shows a current. When *P* is made negative by connecting *S* to the negative end of the battery, the galvanometer does not show any current. This clearly suggests that the incandescent filament emits negative charges. When *P* is positive, these negative charges are attracted by the plate, resulting in a flow of current. When P is negative, the negative charges from the filament are repelled, and there is no current. However, it was only after the discovery of the electron in 1895 that it could be recognised that these charges are carried by particles (the electrons).

5.2.4. RICHARDSON'S EXPERIMENT

Although several earlier workers investigated the phenomenon of thermionic emission, Prof. O. W. Richardson was the first to offer a satisfactory explanation of this process based on his experimental results. His experiments have firmly established, not only the nature of thermions but the dependence of thermionic emission on the temperature, nature and area of the emitter, on the potential difference applied and the nature and pressure of the surrounding gas. The apparatus used in the study of the thermionic emission is shown in Fig. 5-11.

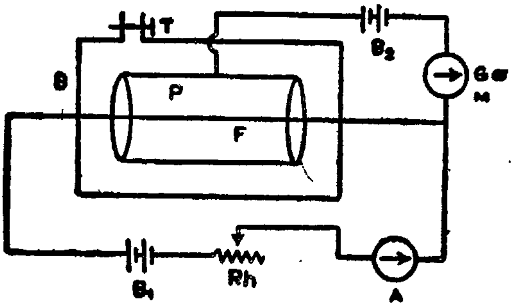


Fig. 5-11
Richardson's experiments on thermionic emission

The metal to be heated is taken in the form of a thin wire called the filament F. This filament is stretched along the axis of a hollow metallic cylinder called the plate P. The filament and the plate are enclosed in a glass bulb B which is connected to a vacuum pump through a side tube T. A good vacuum can be maintained in the bulb or a gas can be filled at any desired pressure. The filament is heated by passing an electric current from the battery B_1, connected in series with an ammeter and a rhoestat. The ammeter measures the current that flows through the filament, called the filament current I_f. The plate P is connected to another battery B_2 which enables a p.d. to be applied between the plate and the filament. A sensitive galvanometer G or a milliammeter M is included in the plate circuit to measure the thermionic current or plate current I_p.

To begin with, the positive terminal of the battery B_2 is connected to the plate P. When the filament is at ordinary temperatures, there is no deflection in the galvanometer indicating that there is no current in the B_2 battery circuit. This means that there is no emission of charged particles from the filament. But if we raise the temperature of the filament by passing more current to the filament, the galvanometer will register a current. This indicates that the charged particles emitted from the filament pass through the empty space between the filament and the plate. As we raise the temperature of the filament further going from orange hot to white hot, this current increases very rapidly.

Now the experiment is repeated by connecting the negative of the battery B_2 to the plate P. In this case the galvanometer shows no current even when the filament is white hot. Hence we conclude that all the electric particles emitted by a hot metal are negatively charged. These particles are in fact electrons.

5.2.5. IMPORTANT RESULTS OF THE RICHARDSON EXPERIMENTS

(i) Nature of Ions

The negative ions emitted by the hot filament are electrons. Thomson confirmed this by measuring the value of e/m (also called specific charge) of these ions and finding the value to be the same as for an electron.

(ii) Saturation current

The temperature of the filament can be evaluated from its resistance. By keeping the temperature T constant, the thermionic current I_p is measured for various voltages applied to the plate (V_p). When the data are plotted, a curve of the form shown in Fig. 5-12 is obtained. It is seen that the thermionic current does not obey Ohm's law. It increases very slowly first, then

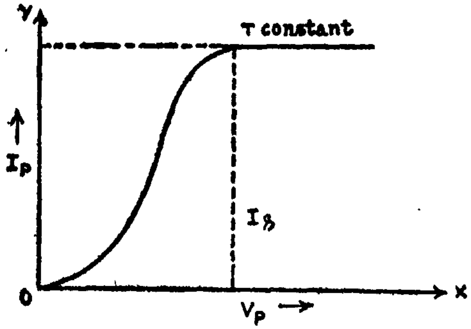


Fig. 5-12
Variation of theromionic current with plate voltage

more rapidly and finally becomes a constant. This is known as saturation current I_s. The voltage corresponding to this current is called saturation voltage.

(iii) Space - Charge Saturation

The plate potential V_p is kept constant and the temperature

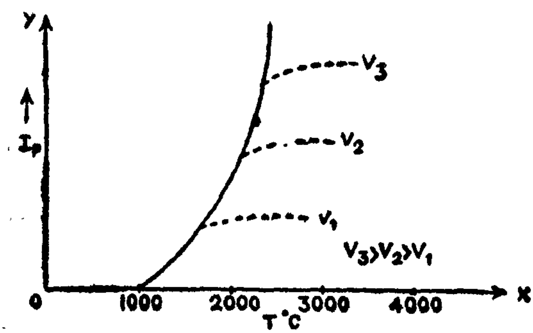


Fig. 5-13
Space-charge saturation

T of the filament is gradually increased. The current I_p begins to flow at about 1000°C. As the temperature is further increased, the I_p increases very rapidly until a saturation current is reached as shown by the curve in Fig. 5-13. Similar curves are obtained for different constant voltages V_1, V_2 and V_3 in the increasing order. The current is limited to a certain maximum value for each of these voltages because of what is called *space-charge* saturation. This is explained as follows:

The rate of emission of electrons is small at low temperatures so that the applied voltage (V_1) is able to attract all the electrons towards the plate. As the temperature increases the rate of emission of electrons also increases. But the applied voltage is not sufficient enough to attract all the electrons towards the plate with the result that the unattracted electrons remain in vicinity of the filament. This is called *space-charge*, i.e. electrons will be lingering in the space, between the filament and the plate. These electrons will make it difficult for any more electrons to escape from the filament. When the temperature is still further increased, a situation develops in which only a constant number of electrons per second are allowed to reach the plate. Thus saturation is reached due to space-charge effects. This limitation of current due to insufficiency of the applied potential is called space charge saturation. The higher the applied potential, the greater is the saturation current.

5.2.6. PHOTO ELECTRIC EFFECT

A series of experiments conducted during the nineteenth century revealed that electrons are emitted from a good number of metals when ultraviolet or visible light falls upon them. This phenomenon is called *photo electric effect* since an electric current is obtained from photo (light). The electrons so emitted are called photo-electrons and the resulting current is known as *photo electric current*.

5.2.7. DISCOVERY OF PHOTO ELECTRIC EFFECT

A telegraph operator W. Smith was the first to observe the phenomenon in 1873. He was measuring the resistance of trans-atlantic cables using selenuim resistors. He observed that when sun light fell upon these resistors, the current in the circuit varied considerably. In 1887, Hertz found that when ultraviolet rays fell on a spark gap, the sparks passed more easily. Later work has shown that all substances, solids, liquids and gases exhibit photo electric effect under appropriate conditions. It is however, convenient to study the effect with metallic surfaces.

5.2.8. EXPERIMENTAL STUDY

Fig. 5-14 illustrates the type of apparatus employed to study the photo electric effect.

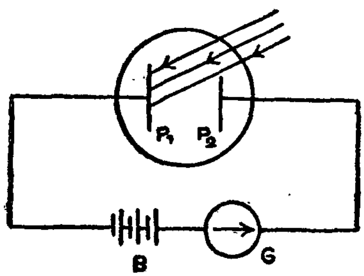


Fig. 5-14
Photo electric effect

Two zinc plates P_1 and P_2 were kept inside an evacuated quartz bulb. They were connected to a battery B and a galvanometer G. When ultraviolet light fell on the plate P_1 which was connected to the negative terminal of the battery, a current was found to flow as indicated by the galvanometer. But when light fell on the positive plate P_2 there was no flow of current. These observations led to the conclusion that the particles emitted in the photo electric effect must be negatively charged; for the observed current can be explained only by the fact that the negative ions emitted by the negative plate are attracted towards the positive plate and thus cause a flow of electricity. Lenard determined the value of e/m for the emitted ions and found it to be the same as for electrons. Hence it is clear that the particles emitted by the photo-emissive substance are electrons.

5.2.9. RESULTS OF THE PHOTO ELECTRIC EFFECT

From the various experimental studies of the photo-emission phenomenon the following results have been obtained:

(*a*) A metal shows photo electric effect only if the incident radiation has a frequency above a certain value called the *threshold frequency.*

This frequency is characteristic of the substance on which the radiation falls.

(*b*) The photo electric current or the total number of electrons released is directly proportional to the intensity of the incident light provided its frequency is greater than the threshold frequency.

(*c*) The velocity of the photo electron and consequently, its kinetic energy cannot exceed certain maximum values which depend upon the frequency of the incident light and not upon the intensity of the incident light. The maximum kinetic energy is proportional to $(\nu - \nu_0)$ where ν_0 is the threshold frequency.

(*d*) If light of a given frequency can liberate electrons, the emission of photo electrons is instantaneous. In fact, even for the lowest light intensities, the time interval between the incident light on the metallic surface and the appearance of electrons is not more than 10^{-9} second.

The second and the third conclusions are also called the laws of photo electric emissions.

Exercise 5.2

1. *Mention the important sources of electrons.*
2. *What is 'thermionic emission ?*
3. *What are 'thermions' ?*
4. *What is 'thermionic current ?*
5. *What is 'Edison effect' ? Describe an experiment to demonstrate 'Edison effect'.*
6. *Describe Richardson's experiment and show how it explains thermionic emission.*
7. *What are the important conclusions of Richardson's experiment ?*
8. *Explain space charge saturation.*
9. *What is photo electric effect ?*
10. *What are photo electrons ?*
11. *What is meant by photo electric current ?*

12. *Write a note on the discovery of photo electric effect.*
13. *Describe an experiment to study the photo electric effect.*
14. *What are the important conclusions drawn from photo electric effect ?*

5.3. ATOMIC STRUCTURE

5.3.1. POSITIVE RAYS

When many physicists were investigating the properties of the cathode rays, Goldstein designed a special dicharge tube and discovered in 1886, a new type of rays called canal rays. A tube designed to illustrate this discovery is shown in Fig. 5-15.

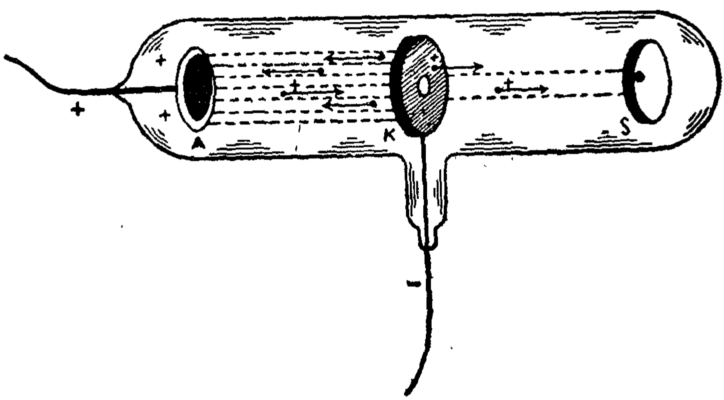


Fig. 5-15
Canal rays

In the figure, the cathode K is perforated and the anode A is a solid disc of aluminium. S is a fluorescent screen. At a pressure of about 1*mm*. of *mercury*, a luminous stream behind the cathode proceeding in a direction opposite to that of the cathode rays, was observed. Goldstein called them **canal rays** since these rays, travelling in straight lines, in the opposite direction to the cathode rays, pass through and emerge from a hole or canal in the cathode.

When a beam of canal rays was deflected in a magnetic field, he came to the conclusion that the rays consisted of positively charged particles. Hence canal rays are more commonly known as *positive rays.*

5.3.2. PROPERTIES OF POSITIVE RAYS

The positive rays are found to possess the following properties :

(i) They are composed of positively charged particles. They move in straight lines with a velocity which is high but less than that of cathode rays in the same discharge tube.

(ii) They are deflected by electric and magnetic fields but their deflection is opposite to that of the cathode rays.

(iii) They penetrate through some thickness of matter such as aluminium foils.

(iv) They ionise the gas through which they pass.

(v) They affect photographic plates.

(vi) When these rays fall on fluorescent screens, they produce tiny flashes called scintillations which can be observed by a microscope and

(vii) The colour of their luminosity depends upon the nature of the gas inside the discharge tube.

5.3.3. ORIGIN OF POSITIVE RAYS

The existence of positive rays can be easily understood from the general mechanism of discharge. We have already seen that, in a discharge tube, the neutral atoms or molecules lose one or more electrons (this process is called ionisation) and the resulting atom or molecule is called a positive ion. There exist, therefore, between the anode and the cathode, two streams of particles, electrons attracted towards the anode (cathode rays) and the positively charged atoms or molecules (positive ions) attracted towards the cathode. Of these positive ions, the one which pass through the perforations in the cathode constitute the canal rays. As each atom or molecule (+ ve ion) strikes the fluorescent screen, a tiny flash of light is produced. These tiny flashes which can be seen individually in the field of view of a microscope, are called scintillations. If the cathode is not perforated, they simply strike the cathode and give rise to a faint glow at its surface.

The condition for the production of positive rays is the ionization at a low pressure in a strong electric field. The low pressure prevents the positive ions from colliding too frequently with other atoms or molecules. The strong field imparts the required high speed to the positive ions inspite of their relatively heavy mass.

5.3.4. POSITIVE RAY ANALYSIS

Positive rays can be produced from molecules as well as atoms, and can be used to obtain direct information as to the masses of individual atoms and molecules (not merely the average mass of a large aggregate of atoms as in chemistry). Thomson's investigations on positive rays led to the development of a mass spectrograph for the separation and detection of atoms of different masses. Among other things, this led to the discovery of various isotopes. The method is generally called *positive ray analysis.*

5.3.5. RUTHERFORD'S EXPERIMENT (CONCEPT OF THE NUCLEUS OF THE ATOM)

(i) Introduction

The scientists of the nineteenth century aecepted the idea that the chemical elements consist of atoms. But they never knew anything about the atoms themselves. The discovery of the electron and the realization that all atoms contain electrons provided the first imporant insight into the atomic structure. Electrons carry negative electrical charges, while atoms themselves are electrically neutral. Hence, every atom must contain enough positively charged matter to balance the negative charge of its electrons. Further, electrons are thousands of times lighter than whole atoms; this suggests that almost all of the mass must be associated with the positively charged part of the atom.

(ii) Thomson's Atomic Model

Following his experiments on positive ray, J.J. Thomson was the first to give a model of the atom in 1904. According to this model, an atom consists of a positively charged substance (positive electric field) distributed uniformly in a sphere of atomic dimension, with negative electrons embedded in this continuous positive charge, like seeds in a watermelon or raisins in a pudding as shown in Fig. 5-16.

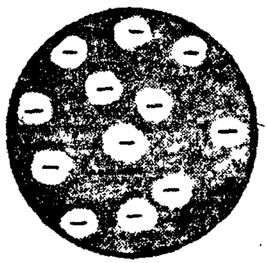


Fig. 5-16
Thomson's atom model

The electrons were assumed to be arranged inside the sphere in such a manner that their mutual repulsion was balanced by the attraction between them and the positive charge. The number of units of positive charge was supposed to be equal to the number of electrons so that the atom could be considered as electrically neutral. Though there are certain arrangements of electrons in this model which could be in equilibrium (as shown by Thomson), it is in unstable equilibrium. And of course this model could give no indication for the regularity in the properties of elements which is displayed through the periodic table.

(iii) Rutherford's Alpha-ray Scattering Experiment

The most direct way to find out what is inside the atom was suggested by Ernest Rutherford, a New Zealand - born physicist. He gave the first correct description as to how the positive and negative charges are distributed within the atoms. An experiment on this aspect was performed by Geiger and Marsden in 1911, at the suggestion of Rutherford. They employed as probes, the fast alpha particles spontaneously emitted by certain radioactive elements.

(Alpha particles carry charge equal to twice the electron charge in magnitude but of positive sign, and have about 4 times the mass of a hydrogen atom. We shall examine their origin and properties in more detail later).

The experimental set up is shown in Fig. 5-17.

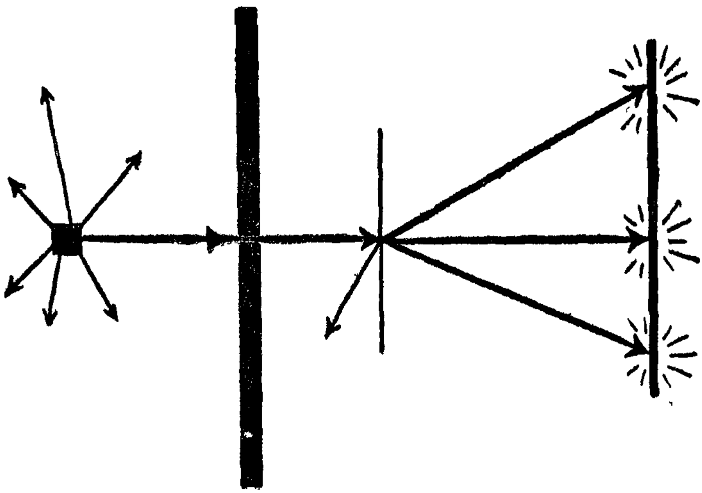


Fig. 5-17
α - particle scattering

Geiger and Mardsen placed a speck of alpha-emitting radioactive material behind a lead screen. The screen had a small hole which allowed a narrow beam of alpha particles to pass through. This beam was directed at a thin gold foil. A movable zinc sulphide screen which gives off a visible flash of light (called scintillation) when struck by an alpha-particle was placed on the other side of the foil.

It was observed that most of the alpha particles went through the foil and most of them suffered only a small deflection (small angle scattering). But some particles were deflected through large angles, some of them even being thrown almost directly backwards (large-angle scattering).

The scattering (or deflection) of alpha particles is caused by the repulsive electrical forces existing between the alpha particles and the positive charges of the atoms in the scattering material. There is also an electrical force of attraction between the electrons (negative charged) of the scattering material and the alpha particles (positively charged). But, since the mass of electrons is extremely small compared with that of alpha particles, the electrons do not cause appreciable deflection of the alpha-particles during their passage through the scatterer. Hence the deflection of alpha particles is caused mainly by the action of the positive charges of the large number of atoms of the scatterer.

(iv) Rutherford's Nuclear Model of the Atom

Now if the positive charge of the atom is spread over a spherical volume of radius R, it exerts the maximum repulsive force on the α-particle, equal to $2\,Ze^2/R$, when the latter just reaches the surface of the sphere. (Once the α-particles penetrate into the sphere the force begins to decrease). Rutherford was able to see that if the charged sphere has the same size as the atom itself (i.e. $R =$ atomic radius $\approx 10^{-10} m$) then this force is not sufficient at all to produce the large angle scattering seen in the Geiger-Marsden experiments. To make the force strong enough to produce such numbers of large-angle scatterings as observed, it was found necessary to assume that R is extremely small compared to the radius of the atom itself. That is to say, the whole positive charge (and mass) of the atom must be concentrated in an extremely small central region. Rutherford named this central region the "*atomicn nucleus*". Later experiments have shown that the radius of the nucleus is of the order of femtometre (femto-$10^{-15} m$)

If all the positive charge is concentrated in a small region (about $10^{-15}\ m$ radius), then the atom as a whole, having a radius of about $10^{-10} m$ must be practically empty (space occupied only by the atomic electrons—these electrons being mere specks in this vast space). It is now easy to see why most of the alpha particles went straight through the thin foil. Since many of the alpha particles do not get too near the nucleus, they are deflected only by small angles. Only a small number of particles approach the core of the atom (nucleus) and they are deflected at large angles. Thus Rutherford was able to explain both the large angle scattering and small angle scattering on

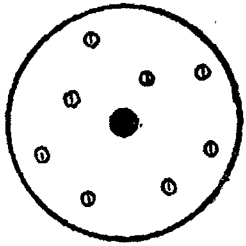


Fig. 5-18
Rutherford's nuclear atom model

the basis of his model, known as Rutherford's nuclear atom model. This is shown in Fig. 5-18. The deflection of the alpha particles on this model of the atom is sketched in Fig. 5-19.

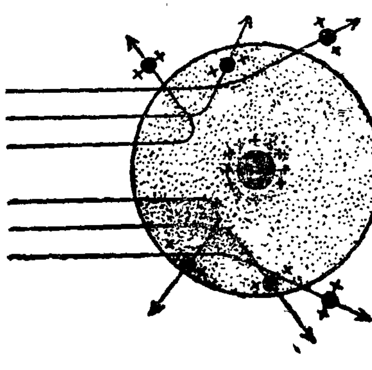


Fig. 5-19
Scattering of α - particles by a nucleus

On the basis of his atomic model, Rutherford arrived at a formula, describing the scattering of alpha-particles by thin foils that agreed with the experimental results. He is therefore credited with the discovery of the nucleus. Though the model was unanimously accepted, this was not free from drawbacks. One serious drawback concerns the stability of the atom. Since the electrons are negatively charged and the nucleus is positively charged, there is an electrostatic force of attraction between the electrons and the nucleus, with the result that the electrons will fall into the nucleus, thus destroying the stable structure of the atom. The difficulty was overcome by Neils Bohr who suggested that the electrons revolve round the nucleus in specified orbits.

5.3.6. CONSTITUENTS OF NUCLEUS

The composition of the atomic nucleus itself was finally understood in 1932, when **James Chadwick**, an associate of Rutherford, discovered a new particle, which was named "*neutron*". It was then established that the central nucleus is composed of two kinds of elementary particles called the protons and the neutrons. The proton is the nucleus of the hydrogen atom, and carries a single unit of positive electric charge. The atomic electrons move round this nucleus in specified orbits.

For example, the ordinary hydrogen nucleus is composed of only one proton. Since the mass of the electron is negligibly small, the mass of the hydrogen atom is very nearly the same as that of the hydrogen nucleus. Consider now a helium atom. Its mass is close to four times the mass of the hydrogen atom.

Hence the nucleus of helium has 2 protons and 2 neutrons. The helium atom has 2 electrons and the negative charge of these electrons is balanced by the positive charge possessed by the 2 protons inside the nucleus. This makes the atom neutral since neutrons are electrically neutral. In general since the atom as a whole is electrically neutral in its normal state and the charge carried by the proton is numerically equal to that of an electron, the number of protons in the nucleus must be equal to the number of extra-nuclear electrons, since the neutrons are electrically neutral. The number of protons in the nucleus is called the atomic number Z. This is also equal to the number of extra-nuclear electrons in the neutral atoms. The sum of the masses of the protons and the neutrons in the nucleus is called the mass number A. These two numbers, viz., Z and A characterise a nucleus. The number of neutrons in the nucleus is equal to $(A-)Z$.

9.3.7. ISOTOPES

The atomic weights of most elements are very close to integers and this fact led **Prout** to suggest that all atoms are built of hydrogen atoms since the atomic weight of hydrogen is very nearly 1. However, it was shown later that the atomic weight of certain elements like neon (20·2) and chlorine (35·5) are not whole numbers (integers). The reason for the departure from the general rule in these cases was discovered by Thomson.

In 1912, Thomson in comparing the mass of neon atoms with the known masses of other elements, discovered that there were two types of neon atoms, one of mass 20 and another of mass 22 (measured in units equal to one sixteenth of the mass of an oxygen atom). Because these two kinds of atoms exist as a mixture (90% of which have a mass 20 and 10% a mass of 22 in naturally occurring neon), their atomic weight, when measured by chemical methods, is found to be their average value, 20·2. It was impossible to separate these two kinds of atoms by any chemical means, as they have identical chemical properties. Soddy gave the name "*isotopes*" to such atoms which have identical chemical properties but differ in their atomic masses. The atoms of different isotopes of the same element have the same atomic number Z but different mass number A, i.e. the isotopes of an element have the same number of protons in the nucleus but different numbers of neutrons. They also have the same number of electrons (equal to Z) revolving round their respective nuclei. This is necessary for the atom to be neutral. Since the chemical properties are determined by the number of electrons in the atom, all isotopes of a given element have the same chemical properties. The equipment used to distinguish the different isotopes is known as a *mass spectrograph.*

It separates the different isotopes by making use of the difference in their masses caused by the different numbers of neutrons which they contain.

The atomic weight of any element as determined by chemical means is the average weight of all the isotopes of the element weighted according to their relative abundance in nature and expressed in *atomic mass units (a.m.u.)* The *atomic mass unit* is defined as one sixteenth the mass of one atom of the most abundant oxygen atom ${}_8O^{16}$ (1 *a.m.u.* = $1{\cdot}66 \times 10^{-27}$ *kg*).

Recent developments in mass spectroscopy have made it possible to detect exceptionally rare isotopes. In neon, for example, an isotope of mass number 21 has been found making three in all, with their relative abundance (the percentage availability of different isotopes in nature) given in the following table.

Isotope	*Abundanee*
Ne^{20}	90·92
Ne^{21}	0·26
Ne^{22}	8·82

They are represented as follows :

${}_{10}Ne^{20}$ isotope consists of 10 protons and 10 neutrons.

${}_{10}Ne^{21}$ isotope consists of 10 protons and 11 neutrons.

${}_{10}Ne^{22}$ isotope consists of 10 protons and 12 neutrons.

Similarly Aston used his mass spectrograph for chlorine (atomic weight 35·46 and showed that there were two kinds of chlorine atoms of mass 35 (75·4%) and 37 (24·6%). Since the atomic number of chlorine is 17 for both, it follows that:

${}_{17}Cl^{35}$ consists of 17 protons and 18 neutrons.

${}_{17}Cl^{37}$ consists of 17 protons and 20 neutrons.

Hydrogen has 3 isotopes of mass number 1,2 and 3, The ordinary hydrogen nucleus consists of just 1 proton. The hydrogen with nucleus of mass number 2 (i.e. 1 proton + 1 neutron) is called heavy hydrogen or deuterium and was discovered by Urey in 1932. The water made from heavy hydrogen is called heavy water. A thirdi sotope of hydrogen having a mass number 3

(1 proton + 2 neutrons) has been discovered and is called tritium. They are represented as H^1, H^2 and H^3 respectively.

Thus many of the elements of the periodic table consists of several isotopes. Mercury and xenon have 9 isotopes each. Tin has the largest of 11 isotopes. Fluorine, aluminium, gold etc. possess only one isotope each.

Exerise 5-3

1. *What are canal rays ?*
2. *How are canal rays produced in a canal ray tube ?*
3. *Mention the properties of canal rays.*
4. *How are positive rays produced ?*
5. *Mention the circumstances under which positive rays will be produced.*
6. *What is a mass spectrograph ?*
7. *What important discovery was made using a mass spectro graph ?*
8. *Explain Thomson's atom model. What are its defects ?*
9. *Describe Rutherford's experiment on the scattering of Alpha rays.*
10. *Explain how Rutherford,f atom model satisfactorily explains small and large angle scattering of alpha particles.*
11. *What is "nucleus" ? Who discovered it ?*
12. *What is the defect of the Rutherford's atom model ? How was it overcome ?*
13. *What are the main constituents of the nucleus ?*
14. *In normal state , explain how an atom is neutral.*
15. *What is meant by (i) atomic number (ii) mass number (iii) atomic weight ?*
16. *What are isotopes ? Who discovered them ?*
17. *What is meant by 'relative abundance' ?*
18. *Describe the isotopes of neon chlorine and hydrogen.*

19. *Which element has the largest number of isotopes ? How many isotopes does it have ?*
20. *What is heavy water ?*
21. *Who suggested the name 'isotope' ?*
22. *Define the unit of 'atomic mass.'*
23. *Mention some elements which have only one isotope.*

5.4. RADIO-ACTIVITY

5.4.1. INTRODUCTION

Radio-activity is an important nuclear phenomenon. It is the spontaneous disintegration or decomposition of atomic nuclei and is almost entirely confined to the heaviest elements of atomic number 83 and above. In 1896, Becquerel discovered that a heavy element like Uranium (atomic number 92) emitted some kind of highly penetrating rays. This phenomenon came to be known as *radio-activity*. The emission of these rays is spontaneous in the sense that it occurs naturally (in its own accord) and is unaffected by external agencies, whether physical or chemical.

Perhaps no single phenomenon has played so significant a role in the development of both atomic and nuclear physics as radio-activity. A close analysis of natural radio-activity has also given very precious information about the structure and stability of the core of the atom viz. the nucleus. From the study of these unstable and self-disintegrating nuclei, our knowledge of the nucleus regarding its constituents, their relative disposition and interaction has greatly improved.

5.4.2. DISCOVERY

The discovery of radio-activity, like many other wonderful discoveries, was purely accidental. It was discovered in 1896 by the French Physicist, A. Henri Becquerel. He attended a meeting of the French Academy of Science, in January 1896, when the effects of the newly discovered X-rays by Roentgen were demonstrated. It was stated at the meeting that X-rays appeared to originate in the luminescent spot produced where the cathode rays impinge on the discharge tube. He, at once, set himself to the task of discovering some connection between X-rays and the luminescence which is the ability of certain substances to transform ultraviolet radiation that falls on them into visible light. Becquerel collected various minerals that he was going to use for his studies. He left them wrapped up in a paper in a drawer. It so happened that in the drawer there were some photographic plates covered by black papers that did not allow sunlight to pass through it.

One day, Becquerel used some of the photographic plates kept in the drawer in order to photograph something. When he developed the plates, he was disappointed to find that they were badly fogged as if previously exposed to light. A check on the other plates in the drawer also revealed the same condition. It became very difficult for him to understand the cause for the mishap of these photographic plates inspite of the fact that all the plates were wrapped in thick black paper and were kept in a sealed box. He guessed that it could have something to do with one of the minerals in the drawer.

Being of inquistive mind, Becquerel investigated the situation by performing a series of experiments. He finally showed that the uranium ore which he kept in the drawer emitted, even in the dark, some kind of radiations, which penetrated the paper wrappers and affected the photographic plates. These radiations were called *Becquerel rays.* The ability of uranium compounds to emit Becquerel rays, rapidly brought this element to a prominent position in Physics.

Becquerel's discovery attracted the attention of the Polish-born **Madame Curie** herself and her husband, **Pierre Curie,** who tried to discover some more substances which could give out Becquerel rays. In July 1898, the Curies succeeded in isolating from bismuth a strongly radio-active element which they named *Polonium,* in honour of Poland, the mother country Madame Curie. In December 1898, the Curies discovered the most important of all the radio-active elements, which they named *radium.* Radium was found to be a million times more radio-active than uranium. The study of radio-activity carried on by many investigators led to the discovery of many other radio-active elements such as raden, thorium, actinium, etc.

5.4.3. SEPARATION OF BECQUEREL RAYS

The radiations emanating from a radio-active element can be easily separated by making use of their different penetrating powers; but deflection methods are much better for their separation. Rutherford and his co-workers discovered that the Becquerel rays were composed of three quite different kinds of radiation. The experimental arrangement is shown in Fig. 5-20.

A small hole is drilled in a lead block. A small speck of radio-active material is placed at the bottom of the hole. This produces a narrow beam of rays emerging from the top of the

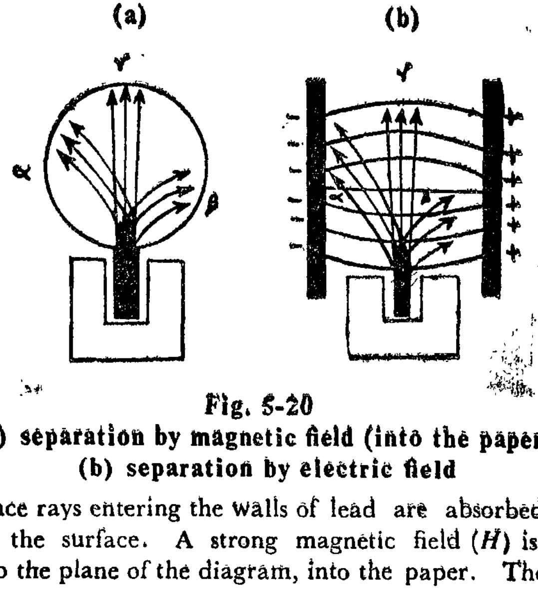


Fig. 5-20
(a) separation by magnetic field (into the paper)
(b) separation by electric field

block, since rays entering the walls of lead are absorbed before reaching the surface. A strong magnetic field (H) is applied normal to the plane of the diagram, into the paper. The single beam will be separated into three groups as shown in Fig. 5-20(a).

The paths of some rays are bent to the left, some to the right, and some not bent at all. The same effect is exhibited when the rays are subjected to an electric field between a pair of plates, as shown in Fig. 5-20 (b). Paths bending to the left indicate positively charged particles called alpha particles or alpha rays (α rays or particles); those bending to the right indicate negatively charged particles, called beta rays or beta particles (β rays or particles) and those going straight ahead possess no charge and are called gamma rays (γ-rays).

5.4.4. PROPERTIES OF ALPHA, BETA AND GAMMA RAYS

A. PROPERTIES OF ALPHA RAYS

(i) Alpha rays are composed of doubly ionised helium atoms i.e., helium atoms with both their electrons removed. Such particles are nothing but helium nuclei. They carry two units of positive charge ($+2e$) and have a mass four times that of hydrogen.

(ii) They are ejected with a speed from one-tenth to one-hundredth of the velocity of light.

(iii) They have great ionising power.

(iv) The penetrating power of these rays is small; even a thin cardboard is able to stop most of these rays.

(v) The range of an α-particle in air is defined as the distance it will travel through dry air at normal atmospheric pressure and temperature. The range of α-particles varies from about 2·6 cm. to about 8·6 cm. for different radio-active elements.

(vi) Glasses, crystals and minerals change colour and glass becomes brittle when subjected to irradiation by an intense beam of alpha rays.

(vii) They affect photographic plates.

(viii) They are deflected both by magnetic and electric fields.

(ix) They can produce artificial disintegration (induced radio-activity) of nuclei.

(x) They produce scintillations (tiny flashes of light) when they fall on certain materials like zinc sulphide.

B. PROPERTIES OF BETA RAYS

(i) β-rays are composed of negatively charged particles; actually they are electrons.

(ii) They move faster than alpha particles, some of them travelling with a velocity as high as 97% of the velocity of light.

(iii) They have a smaller ionising power than the alpha particles, smaller by a factor of about 100.

(iv) They are more penetrating than alpha particles, and their range in air is usually quite large.

(v) They are deflected by both electric and magnetic fields.

C. PROPERTIES OF GAMMA RAYS

(i) γ-rays are electromagnetic waves like X-rays and light, but have wavelengths even shorter than those of X-rays. They travel with the velocity of light, of course.

(ii) They are unaffected by electric and magnetic fields.

(iii) They have very little ionising power, the reason being that they carry no mass and charge.

(iv) Their penetrating power is very great. They can penetrate several centimetres of lead without being absorbed.

(v) They produce fluorescence.

(vi) They affect photographic plates.

(vii) They can be reflected from the surface of a crystal at certain special angles, in the same way as X-rays are reflected.

5.4.5. LAW OF RADIO-ACTIVE DISINTEGRATION

A. DECAY CONSTANT

It has been observed that the nuclei of the atoms of any radio-active element are unstable and that radio-activity consists in the expulsion of a particle (α or β) or electromagnetic energy (γ) from the nucleus. Radio-active nuclei are like loaded guns ready to go off at any time, but unlike the case of the gun, there is no way of controlling or predicting the time at which an individual nucleus will disintegrate. The nuclei of the various atoms of a sample of radio-active substance undergo decay at random times. However, *for a given radio-active substance, the rate of decay (i.e. the number of nuclei which disintegrate per second) is proportional to the number of radio-active nuclei present at that instant.*

If N be the number of radio-active atoms at any instant then the rate of disintegration is proportional to N.

$$\text{i. e.} \quad -\frac{dN}{dt} \propto N$$

$$\text{or} \qquad \frac{dN}{dt} = -\lambda N$$

where λ the constant of proportionality is called the *decay constant or disintegration constant.* (Its value depends upon the disintegrating element).

Let the number of radio-active atoms present initially (at time $t=0$) be N_0. Then, the number of atoms which remain intact (not decayed) at some later time t is given by

$$N = N_0 e^{-\lambda t}.$$

Here e is the base of natural logarithms ($e=2{\cdot}718$) and the function $e^{-\lambda t}$ is the exponential function,

$$e^x = 1 + \frac{x}{1!} + \frac{x^2}{2!} + \frac{x^3}{3!} + \ldots\ldots$$

It has the property that its derivative is equal to itself.

Using this property it may be verified that equation $N = N_0 e^{-\lambda t}$ leads to the equation $\frac{dN}{dt} = -\lambda N$

The equation $N = N_0 e^{-\lambda t}$ gives the *exponential law* of radio active disintegration. The function $e^{-\lambda t} = \frac{1}{e^{\lambda t}}$ is a rapidly decreasing function, whose value goes down as in the following table :

t	0	$\frac{1}{\lambda}$	$\frac{2}{\lambda}$	$\frac{3}{\lambda}$	$\frac{4}{\lambda}$
$e^{-\lambda t}$	1	0·368	0·135	0·050	0·018

It follows then from the equation $N = N_0e^{-\lambda t}$ *that*

at $t = 1/\lambda$, $N = N_0e^{-1} = \frac{1}{e}N_0$

So the decay constant, λ is seen to be the reciprocal of the time taken for the radio-active atoms to get reduced to a fraction (1/e) of their initial number. Hence, λ is expressed in the unit of sec^{-1}

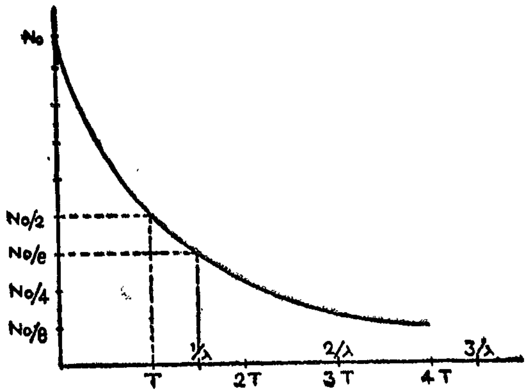


Fig. 5-21
Variation of number of radio-active nuclei with time

The manner in which N decreases with time is shown graphically in Fig. 5-21.

B. HALF LIFE PERIOD

According to the exponential law of disintegration an infinite time is required for the activity to disappear completely. This is true of all the radio-active elements. Hence, in order to compare the activity of one radio-active substance with another, a qnantity known as the *half-life period* or simply *period* (T) is often used.

The half-life period (T) is defined as the time in which the number of undecayed radio-active nuclei, and hence the rate of radio-active disintegration, are reduced to half their original values.

We know $N = N_0e^{-\lambda t}$

When $t = T$, $N = \frac{1}{2}N_0$,

$$\frac{N_0}{2} = N_0 e^{-\lambda t}$$

or $$\frac{1}{2} = e^{-\lambda T}$$

i.e. $$e^{\lambda T} = 2$$

Hence $\lambda T = \log_e 2 = 2{\cdot}303 \log_{10} 2 = 0{\cdot}693$

$$\therefore T = \frac{0.693}{\lambda}$$

T is therefore inversely proportional to λ i.e. if λ is very small, *T* is very large and vice versa. The period *T* is a characteristic constant of the disintegrating substance, like λ. Hence it can be used to differentiate the various radio-active elements. For example, *T* for radium is 1590 years while for radon, it is only 3·8 days.

The statement that a certain radio-active element has a half life of 5 hours, signifies that every nucleus of this element has a 50% chance of decaying in any 5 hour period. This does not mean a 100% probability of decaying in 10 hours. A half life of 5 hours implies a 75% probability of decay in 10 hours (50% in first 5 hours + 50% of the remaining nuclei in another 5 hour period i.e. 25%=total 50%+25%=75%). This increases to 87·5% in 15 hours (i.e. 50%+25% + 12·5% = 87·5%) to 93·75% in 20 hours (50%+25%+ 12·5% + 6·25% = 93·75%) and so on since in every 5 hours interval, the probability is 50%.

C. UNIT OF RADIO-ACTIVITY

The degree of radio-activity of a sample of a substance is evidently determined by the number of radio-active atoms which disintegrate per unit time. It is customary to express the radio-activity in terms of the *curie* and its sub-multiples, the *millicurie* (*mCi*) and *microcurie* (μCi). They are defined as follows :

1 curie = $3{\cdot}7 \times 10^{10}$ disintegrations/sec.

1 *mCi* = 10^{-3} curie = $3{\cdot}7 \times 10^{7}$ disintegrations/sec.

1 μCi = 10^{-6} curie = $3{\cdot}7 \times 10^{4}$ disintegrations/sec.

The curie is a unit of radio-activity and is defined as the quantity of any radio-active element in which the number of disintegrations per second is $3{\cdot}7 \times 10^{10}$.

5.4.6. DISPLACEMENT LAW AND RADIO-ACTIVE SERIES

A. INTRODUCTION

Rutherford and his colleagues discovered that when a radio-active nucleus disintegrates into an alpha (or beta) particle and a new nucleus, (the latter is often radio-active) sooner or later, ejects a particle and thus gets converted into the nucleus of yet another element. This process is continued through a series of elements, ending up finally with a type of atom that is stable and non-radio-active. Such a chain of disintegration process is known as a *radio-active series*.

In each decay, the original atom is known as the *parent* and the new atom formed after the emission of an α or β particle is called the *daughter*. It is now known that nearly all natural disintegration processes finally end up with stable lead atoms of atomic number 82. A law which governs these radio-active changes was discovered by Soddy and Fajans in 1913. This is known as the *displacement law* in radio-activity.

B. DISPLACEMENT LAW

The law may be stated as follows : (*i*) *In all known radio-active transformations, either an alpha or a beta particle is emitted by the atom. Both are not emitted simultaneously nor are two particles of the same kind emitted simultaneously by the same atom.*

(*ii*) *When an alpha particle is emitted, a new atom is formed whose mass number is less by four units and atomic number less by two units than in the case of the parent atom.*

(*iii*) *When a beta particle is emitted, the new atom formed has the same mass number but the atomic number is increased by one unit.*

This law can be readily understood since the alpha particle, identified with the helium nucleus has mass number 4 and atomic number 2, while the beta particle, identified with the electron has unit negative charge but has negligible mass.

For example, uranium with atomic number ($Z=92$) and mass number 238 ($A=238$) emits an alpha particle (atomic number 2 and mass number 4) and the daughter must therefore have atomic number 90 and mass number 234.

This element with $Z=90$ is thorium. This is summarised in the following equation:

$${}_{92}U^{238} \longrightarrow {}_{90}Th^{234} + {}_{2}He^{4} \text{ (alpha particle)}$$

If the parent is a beta emitter, the atomic number of the daughter must be one higher than that of the parent since the beta particle is an electron possessing negative charge (Loss of a negative charge is equivalent to gain of a positive charge). Further, the electron is so light that the mass number of the parent and the daughter are the same. For example. ${}_{90}Th^{234}$ the daughter of ${}_{92}U^{238}$, is a beta emitter. The daughter of ${}_{90}Th^{234}$ must have atomic number 91 and mass number 234. This new element is called protoactinium *(Pa)*. This is summarised as follows:

$${}_{90}Th^{234} \longrightarrow {}_{91}Pa^{234} + {}_{-1}e^{0} \text{ (beta particle)}.$$

It may be noted that in the above equation, the following rules known as *Rutherford-Soddy rules*, are satisfied.

(*i*) *The total electric charge (atomic number)* or *the algebraic sum of the charges before disintegration is equal to the total electric charge after the disintegration.*

(*ii*) *The sum of the mass nnmbers of the initial particles is equal to the sum of the mass numbers of the final particles.*

C. RADIO-ACTIVE SERIES

We have seen that in natural radio-activity, a succession of radioactive transformation takes place forming radio-active series. With the help of the displacement law, one can determine easily the masses and the atomic numbers of the different elements in the series, provided the mass and the atomic numbers of the original parent are known. Most of the radio-active elements found in nature are members of four **radio-active series.** They are the **uranium, thorium, actinium** and **neptunium** series. In all the series, a parent element of large atomic mass and very long life gives rise to a series of radio-active elements as a result of successive emission of alpha or beta particles. The transformation may be represented, in general, by expressions of the form ;

$$_{Z}X^{A} \longrightarrow {}_{Z-2}Y^{A-4} + \alpha$$

$$_{Z}P^{A} \longrightarrow {}_{Z+1}Q^{A} + \beta$$

Here Z is the atomic number and A is the mass number. X is an alpha emitter and P is a beta emitter. Y and Q are daughters. They are themselves radio-active if they have atomic numbers greater than 82, the atomic number of lead.

The table gives a list of names of the four 'important radio-active series, the parents and their half-lives, and the stable daughters which are the end products of the series. The half life of neptunium is so short compared with the estimated age (about 10^{10} years) of the universe that the members of this series are not found in nature today. They have, however, been produced in the laboratory, by the neutron bombardment of other nuclei.

Table

FOUR RADIO-ACTIVE SERIES

Series	*Parent*	*Half-life in years*	*Stable end product*
Thorium	$_{90}Th^{232}$	1.39×10^{10}	$_{82}Pb^{208}$
Neptunium	$_{93}Np^{237}$	2.25×10^{6}	$_{83}Bi^{209}$
Uranium	$_{92}U^{238}$	4.50×10^{9}	$_{82}Pb^{206}$
Actinium	$_{92}U^{235}$	7.07×10^{8}	$_{82}Pb^{207}$

In thorium, uranium and actinium series, the end product is an isotope of lead. In neptunium series, the end product is $_{83}Bi^{209}$. The study of these series is important since it not only brings out the sequences in the formation of the different radio-active elements but also leads to the knowledge of the existance of isotopes. The discovery of isotopes among the natural radio-active elements was in fact the starting point in the search for the isotopic constitution of stable elements.

5.4.7. AGE OF THE EARTH

Among the many important applications of radio-activity, the one that we consider here is the estimation of the age of the earth. The decay of radio-active elements and its complete independence of physical and chemical conditions give us an extremely valuable method for estimating the age of the earth. This can be determined from the following considerations:

The uranium series of radio-active elements begins with $_{92}U^{238}$ whose half-life is 4·5 giga years (1 giga = 10^9 years) and whose decay constant is less than one thousandth of that of any other element in the series until we come to Pb^{206}. This lead is stable having an infinite half-life (decay constant zero). This means that, after a giga years or so, the only element present in any appreciable concentration will be uranium and lead. The decay of the first element (uranium) and the last (lead) is represented by the following equation:

$$N_{pb} = (N_0)_u \left(1-e^{-\lambda_u t}\right)$$

Here N_{pb} is the number of lead atoms present now, having been formed by the decay of uranium; $(N_0)_u$ is the number of uranium atoms present originally and λ_u is the decay constant of uranium.

U^{238} ore always contains Pb^{206} because the lead is the end element of the uranium radio-active series. It is reasonable to assume that this is the only reason for the presence of lead since there is no other compelling reason why uranium and lead should be found together. It follows that the present number of lead atoms plus the present number of uranium atoms must be equal to the number of uranium atoms originally present or

$$N_{pb}+N_u = (N_0)_u$$

The present concentrations of Pb^{206} and U^{238} can be measured experimentally. Substituting this in the previous equation we get

$$N_{pb} = (N_{pb}+N_u)\left(1-e^{-\lambda_u t}\right)$$

Knowing N_{pb}, N_u and λ_u the age of the earth 't' may be calculated. This 't' is the time that has passed since the earth solidified and the original packet of uranium was sealed in the rock.

There have been many assumptions in our argument, but they are justified by the fact that the above equation has been applied to many ore samples from different parts of the earth and the results have consistently shown its age to be about 4×10^{12} years.

WORKED EXAMPLES

1. Tritium $_1H^3$ has a half-life of 12·5 years against beta decay. What fraction of the sample will remain undecayed after 37·5 years ?

   To begin with, let there be 1 gram of $_1H^3$. After 12·5 years, the fraction left undecayed is $\frac{1}{2}$.

   After 25 years, the fraction left undecayed $=\frac{1}{2}\times\frac{1}{2}=\frac{1}{4}$.
   After 37·5 years the fraction left undecayed $=\frac{1}{2}\times\frac{1}{4}=\frac{1}{8}$
   ∴ The fraction of the sample left undecayed after 37·5 years is $\frac{1}{8}$.

2. The half-life of $_{11}Na^{24}$ is 15 hours. How long does it take for 93·75% of a sample of this isotope to decay?
   Let there de 100 gms. of $_{11}Na^{24}$ in the beginning.

The amount decayed in the first 15 hrs $=\frac{1}{2}\times100=50$ gms.

The amount decayed in the next 15 hrs $=\frac{1}{2}\times50=25$ gms.
The amount decayed in the next 15 hrs $=\frac{1}{2}\times25=12{\cdot}5$ gms.
The amount decaved in the next 15 hrs $=\frac{1}{2}\times12{\cdot}5=6{\cdot}25$gms.
Hence, the amount decayed in 60 hrs $=93{\cdot}75$ gms.

The time taken for 93·75% of the sample of the isotope $_{11}Na^{24}$ to decay is 60 hours.

3. The half-life of thorium is $13{\cdot}9\times10^{10}$ years. Calculate the decay constant of thorium.
   Given $T=1{\cdot}39\times10^{10}$ years

   $$\therefore T = 1{\cdot}39\times10^{10}\times365\times24\times60\times60 \text{ secs,}$$

We know $\lambda = \frac{0.693}{T}$

$$\therefore \lambda = \frac{0.693}{1.39 \times 10^{10} \times 365 \times 24 \times 60 \times 60}$$

$$\lambda = 1.581 \times 10^{-18} \text{ sec}^{-1}$$

$\therefore$ The decay constant of thorium is 1.581×10^{-18} sec^{-1}

Exercise 5.4

1. *What is radio-activity ? Who discovered it ?*
2. *What is meant by spontaneous disintegration ?*
3. *Explain the circumstances under which radio-activity was discovered.*
4. *What are the radio-active elements discovered by Madame Curie and Pierre Curie ?*
5. *State the components of Becquerel rays.*
6. *Describe the deflection experiment for the separation of Bequeral rays.*
7. *What are alpha, beta and gamma rays.*
8. *State the properties of (i) alpha rays (ii) beta rays and (iii) gamma rays.*
9. *What are the relative ionising powers of the different rays ?*
10. *State and explain the exponential law of radio-active disintegration.*
11. *Define decay constant and half-life period. Derive the relation between them.*
12. *What is the meaning of the statement that a radio-active element has a half-life of 5 hours ?*
13. *State and define the unit of radio-activity.*
14. *What is the radio-active series ? Who are parents and daughters in a radio-active series ?*
15. *What is the stable end product of a radio-active series ? What is its atomic number ?*
16. *Enumerate the displacement law.*
17. *What change takes place in the nucleus of an atom when an alpha particle is emitted ? Give an example.*

18. *What change takes place in the nucleus of an atom when a beta particle is emitted? Give an example.*

19. *State Rutherford-Soddy rules to be satisfied in a radio-active transformation.*

20. *Give a brief account of the radio-active series bringing out its significance.*

21. *Explain how the age of the earth can be estimated from radio-activity studies.*

PROBLEMS

1. The half-life of radon is 3·8 days. Calculate its decay constant.
2. The half-life of radium is 1600 years. Calculate its decay constant.
3. The isotope $_{92}U^{23}$ undergoes two successive beta decays. What is the resulting isotope?
4. The isotope $_{92}U^{238}$ successively undergoes eight alpha decays and six beta decays. What is the resulting isotope?
5. The half-life of a radio-active sample is 20 minutes. What fraction of this sample will remain undecayed after 80 minutes?
6. The half-life of a radio-active sample is 30 hours. How long does it take for 87·5% of the sample to decay?
7. The activity of a radio-active sample is $11{\cdot}1 \times 10^{10}$ disintegrations per second. Express this in Curie, milli Curie, and micro Curie.
8. The disintegration constant of a radio-active sample is 1525 second. Calculate its half-life.

6. ELECTRONICS AND INSTRUMENTS

6.1. CONDUCTORS

6.1.1. INTRODUCTION

The rate of flow of electric charges is known as the *electric current*. To cause a flow of charges, a potential differenence has to be applied. The current caused by such a potential difference is governed by **Ohm's law** which states that *the current flowing in a given wire, at a given temperature, is proportional to the potential difference to which it is subjected.*

Some substances allow electricity to flow freely through them when a potential difference is applied while some others do not. The former substances are called *conductors* and the latter *insulators*. The ratio of the potential difference applied between two points of a conductor to the current caused by it, is called the *resistance* of the conductor. Good conductors like metals offer very little resistance while insulators like wood, rubber, paper, glass offer very high resistance.

6.1.2. RESISTANCE

The resistance of a conductor at a given temperature depends upon the material and its dimensions. The resistance is directly proportional to the length and inversely proportional to the area of cross section of the wire. For most conductors the resistance increases with temperature. For some materials the resistance decreases as temperature increases.

Exercise 6.1

1. *What is an electric current?*
2. *State Ohm's law.*
3. *Define the resistance of a conductor.*

6.2.1. RESISTORS

In electronic circuits, conductors with different specific values of resistances are needed. These resistors, as they are called, are of different types. Some are made by winding wires over a porcelain tube, the ends of the wire being fixed to two terminals (Fig. 6-1). Sometimes these are provided with some taps as shown in Fig. 6-2. So that it is possible to obtain different values of the resistance. Resistors with very high resistance are used in circuits to control the amount of current flowing in a circuit.

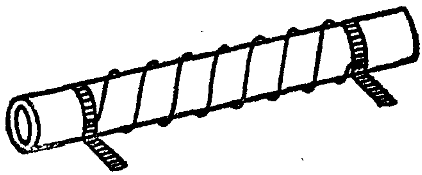


Fig. 6-1
Wire wound resistor

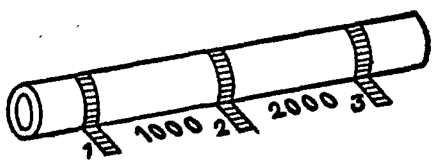


Fig. 6-2
Resistor with tapping terminals

Another kind of resistor is what is usually called the carbon resistor. It is a compressed mass of some form of carbon or graphite mixed with a binder such as clay, bakelite or sometimes rubber. The value of the resistance is determined by the size of the resistor as well as the proportion of carbon to the rest of the substance.

The flow of current through a resistor heats it up and unless this heat is dissipated, the resitor has a limit to the current it can carry. It is expressed in terms of the power (in watts) concerned and it is called the *wattage rating* of the resistor. For example a 10 watt resistor can carry approximately three times the current which a 1 watt resistor can. The wattage rating depends on the surface area of the resistor as a larger surface permits a faster dissipation of heat.

6.2.2. COLOUR CODE FOR RESISTOR

To enable the resistance of a resistor to be recognised quickly and easily, a *colour code* is generally marked on it. In the old convention the whole of the resistor and one of its ends were coloured with a coloured spot in the middle of its body. In the new convention there are three coloured rings marked from the left end of the resistor. The meaning of the conventions are shown in Table 6-1.

Table 6-1 Colour code for resistor

Convention	Value of the resistance		
	First digit	Second digit	No. of zeroes that follow
Old	Body colour	End colour	Dot colour
New	First ring from left	Second ring	Third ring

The numerical values corresponding to the different colours are given in Table 6-2.

Table 6-2 Values of colours

Colour	Value
Black	0
Brown	1
Red	2
Orange	3
Yellow	4
Green	5
Blue	6
Violet	7
Grey	8
White	9

It is very difficult to make carbon resistors of exact values. The specified values are, therefore, subject to variations, called the *tolerance* of the resistor. Manufacturers specify the amount of variation by means of a ring in addition to the coloured rings mentioned above. Thus, a gold ring indicates a two percent tolerance and a silver ring denotes a 5 percent tolerance. Absence of either of these generally indicates a 10 percent tolerance.

Exercise 6.2

1. *What is meant by 'wattage rating' of a resistor ?*
2. *What is meant by the 'tolerance' of a resistor ?*

6.3 CAPACITORS

A capacitor is a system of conductors, usually two, between which a potential difference can be maintained. Therefore, practically all pairs of conductors in existence must be capacitors. In practice, we usually mean two insulated conductors which can be arranged so close to each other that they can be charged to different potentials and thereby an electrostatic field established in the insulating space between them. The insulating material between the two conductors is generally a *dielectric*.

Capacitors commonly available today are of different kinds, sizes and shapes. The most common type is the *paper-capacitor* which is used commonly in most electronic circuits. Two long strips of tin foil are glued to the two faces of a strip of thin paper. This is then soaked in paraffin or oil and rolled up with another paraffin soaked strip of paper into a small compact unit. The tin foil is the conductor and the paper is the dielectric separating them.

Mica is one of the best materials used as a dielectric in a fixed capacitor. A familiar form consists of alternate plates of tin-foil and mica kept in a metal or bakelite case which is filled with wax. Alternate plates are connected together in two sets to form the two terminals (Fig. 6-3).

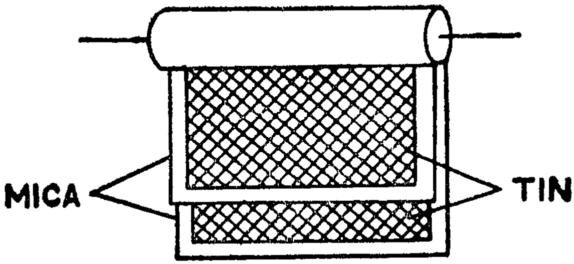


Fig. 6-3
Thin plates of tin and mica rolled into a capacitor

The *electrolytic capacitor* receives its name from the fact that the dielectric is a thin film of corroded aluminium formed on the surface of an aluminium foil conducting plate by electrolytic action. The value of a capacitance will be directly proportional to the area of the plates and inversely to the distance between them. In the case of electrolytic capacitors the thickness of the dielectric can be extremely small and so, large values of capacitances can be realised. The positive and negative terminals are marked on the capacitor and this has to be strictly adhered to. This also means that this capacitor should **not** be used in a circuit where alternating voltages are involved.

Variable capacitors are necessary in a tuning circuit as in radios. Several plates are kept separated by air and alternate plates are connected together in two separate sets. One set of plates can be moved between the other set by turning a knob and this changes the effective plate area and hence the capacitance of the capacitor.

A capacitor is generally used to allow only alternating currents to pass through and block direct current. It is also used in controlling alternating currents of different frequencies, and in oscillatory circuits.

Exercise 6.3

1. *What is a capacitor ?*
2. *Explain the construction of a paper capacitor.*
3. *Why is the electrolytic capacitor called so ?*
4. *What are the uses of capacitors ?*

6. 4. INDUCTORS

6.4.1. CHOKES AND TRANSFORMERS

We have already seen that an inductor offers resistance to alternating currents. A coil forming a part of a circuit primarily by virtue of its inductors is commonly referred to as an *inductor.* An inductor has the property that it offers a low resistance to the direct current and a high resistance to the alternating current. Since the power wasted in a circuit depends upon the resistance to d.c, an inductor has a low wastage of power. A choke is an inductor with a high inductance and low resistance. This is to control alternating currents with as little wastage of power as possible. The inductance of an inductor can be increased by winding it on a core of a magnetic substance like iron. Thus inductors of high inductance usually have magnetic cores.

A transformer is a device which converts a large alternating current at low voltage to a low current at high voltage or vice versa. They are called *step-up* and *step-down* transformers respectively. Two coils of wire are wound on the same core, one being called the primary coil and the other the secondary coil. A given a.c. voltage is applied to the primary coil and this causes an induced e.m.f in the secondary coil. The value of this e.m. is given by

$$\frac{\text{Secondary e.m.f.}}{\text{Primary e.m.f.}} = \frac{\text{Number of turns in the secondary}}{\text{Number of turns in the primary}}$$

$$= \frac{N_s}{N_p}$$

It is obvious that a d.c. applied to the primary coil will give no e.m.f. in the secondary coil and thus the transformer can also be used to block d.c. in a circuit.

Exercise 6.4

1. *What is an inductor ?*
2. *What is a choke ? What is its use ?*
3. *What is a transformer and where is it used ?*

6. 5. A. C. CIRCUITS

6.5.1. SERIES AND PARALLEL RESONANT CIRCUITS

In radio receivers, transmitters and many other applications, it is often necessary to pick out a particular signal. This is done with tuned circuits which will only respond to one frequency or more correctly, to a very narrow band of frequencies. There are many ways of doing this, but basically all of them depend on series or parallel resonances. In this section we shall deal with resonance in series and parallel circuits.

You must have heard that the electricity supply to our houses is at 250 volts, 50 cycles per second Hz. This is called an alternating voltage and it various between +250 volts and—250 volts and back to +250 volts and this happens 50 times per second. If we draw a graph connecting the voltage and the time. It will be as in Fig. 6-4.

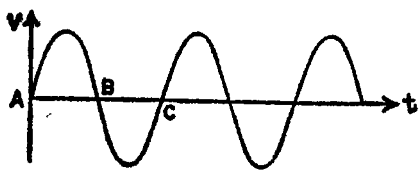


Fig. 6-4
Alternating voltage

Variations from A to C form one cycle and this happens in 1/50 of a second in the above case. If the frequency is 1000 Hz., then one cycle will take 1/1000 th of a second. It is possible to generate alternating voltages of different frequencies. For example freqnencies range from 10^3 to 10^6 Hz. in the case of radio transmission.

Generally the resistance of an ordinary conductor to an a. c. does not depend on the frequency. The resistance offered by an inductor to an a. c. increases with frequency, while the resistance offered by a capacitor decreases with frequency. In the case of an inductor the 'resistance' offered by it is directly proportional to the frequency whereas in the case of a capacitor it is inversely proportional to the frequency. The inductor offers a low resistance but a capacitor offers a high resistance to an a.c. of low frequency. A direct current has zero frequency and, accordingly, the capacitor offers infinite resistance to d. c., thus blocking it completely. At high frequencies, an inductor has a high resistance and a capacitor has a low resistance. Hence at high frequencies even an inductor with inductance can be used to block an a c. Thus, for example in the radio frequency circuits, chokes are usually air-cored.

6.5.1. (a) SERIES RESONANCE

A **series-resonance circuit** *consists* of a *resistance*, an *inductance and a capacitance in series as shown:*

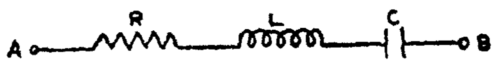


Fig. 6-5
Series resonance circuit

If an alternating voltage is applied between the **terminals** *A* **and** *B*, *the alternating current generated in the circuit is* **found** *to vary with the frequency of the applied voltage.*

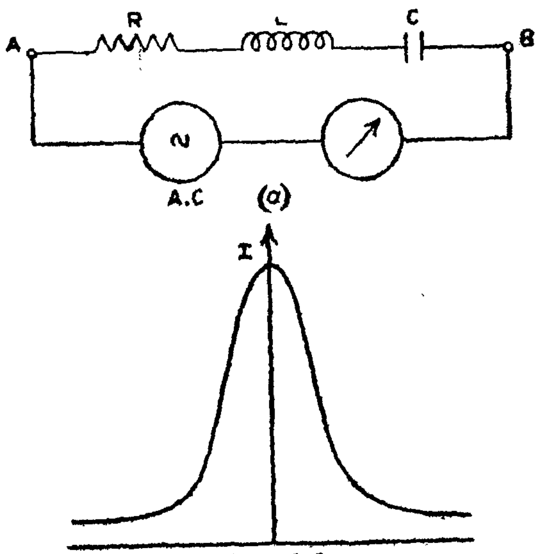


Fig. 6-6
(a) An a. c. voltage applied to a series resonance circuit
(b) Variation of current with frequency-series resonance.

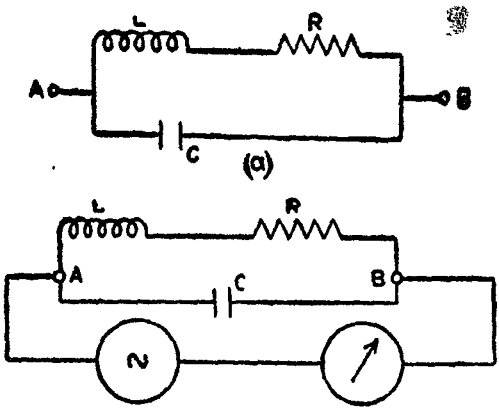
The variation of the current with the frequency is shown in Fig. 6-6*b*. It will be noticed that the current is low for most frequencies and is appreciable only for a very small range of frequencies. The centre of the small range gives the maximum current. We say that the circuit is resonant with the applied frequency at this point. This frequency is usually called the **resonant frequency** (f_r). Therefore, this circuit can be used to 'select' alternating currents of this particular frequency. The value of the resonant frequency depends upon the values of the inductance and the capacitance in the circuit. It is given by

$$f_r = \frac{1}{2\pi\sqrt{LC}}$$

where the L is the value of the inductance in henry and C the value of the capacitance in farad. The value of the maximum current at the resonant frequency will depend only on the resistance of the circuit.

6.5.1. (*b*) PARALLEL-RESONANCE CIRCUIT

Another type of resonance circuit is the **parallel resonance circuit** which is given below (Fig. 6-7 *a*).



(b)

Fig. 6-7

(a) Parallel resonance circuit

(b) An a.c. voltage applied to the parallel resonance circuit

If we apply an alternating voltage across A and B as shown in Fig. 6-7, it is found that the current registered by the ammeter varies with frequency again, except that, in this case, we get a minimum instead of a maximum at a particular frequency. For this reason it is sometimes called a **rejector circuit.** The variation of current with frequency is shown in Fig. 6-8. This means that the circuit offers maximum resistance to the applied voltage at the resonant frequency, that is, it blocks at this frequency as against the behaviour of the series resononance circuit. When the resistance is small, the resonant frequency is given by

$$f_r = \frac{1}{2\pi\sqrt{LC}}$$

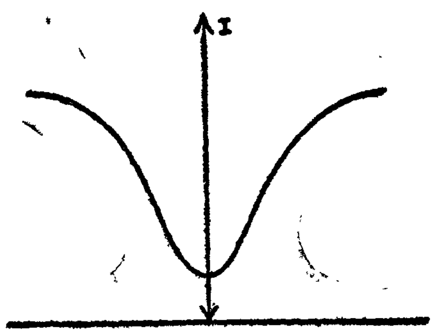


Fig. 6-8

Variation of current with frequency (parallel resonance)

Parallel resonance circuits can be used to reject unwanted signals.

6.5.2. IMPORTANCE OF TUNING

In radio circuits, it is necessary to select the particular frequency of the station desired. For this purpose, we must use a series resonance circuit. By varying the capacitance in the circuit, we can vary the resonance frequency of the circuit, and thus the circuit can be made to respond to any desired frequency. For this purpose, the variable capacitor is used.

Exercise 6.5

1. *What is meant by saying that the electric supply to our house is 250V, 50 Hz ?*
2. *How do the resistances offered to an alternating current by an inductor and a capacitor vary with the frequency of the voltage applied to them ?*
3. *What is a series resonance circuit?*
4. *Explain the action of a series resonance circuit with suitable diagrams.*
5. *Give the expression for the series resonance circuit.*
6. *Explain the action of a parallel resonace circuit with suitable diagrams.*
7. *What is the importance of tuning?*

6. 6. DIODES

6.6.1. Introduction

Electronic circuits use two types of circuit elements. The first type involves resistance, capacitance and inductance while the second type involves vacuum tubes and semiconductor devices. The latter type is known as the *active device* while the former type is called the *passive device*. We have so far discussed about the first type and we shall now consider the second type *viz*, active devices.

6.6.2. VACUUM TYPE DIODE

The simplest vacuum tube is a diode which has two electrodes, a *plate* and a *cathode*. The cathode is either a metal filament which is directly heated by the passage of a current, or a metal cylinder which is indirectly heated by a filament kept along its axis. The heated cathode emits electrons which occupy the space surrounding the cathode. This forms a '*space charge*' which tends to oppose further emission of electrons from the cathode. The loss of electrons will tend to make the cathode positive, and thus cause it to attract the electrons from the space charge. At some stage the electrons emitted by the cathode will be equal to the electrons coming back to it due to the attraction of the cathode as well as the repulsion of the space charge. A dynamic equilibrium will thus be established. This process is similar to the equilibrium of a liquid with its vapour above it.

Generally, the cathode is made up of a nickel alloy base on which there is a coating of oxides of barium and strontium. In the filament type, the substance used is tungsten or thorium coated tungsten. These materials are chosen because they are efficient emitters giving appreciable emission even at lower temperatures.

The plate, which is the anode, is a metal cylinder, surrounding the cathode. This is usually made of nickel, iron or molybdenum. The entire arrangement is enclosed in a glass envelope, which is completely evacuated.

If we give a positive potential to the plate with respect to the cathode, that is, when it is forward biased, the space-charge electrons will be attracted towards the plate, and this movement of electrons generates a current. In the circuit shown in Fig. 6-9., a variable positive potential is connected to the plate with respect to the cathode.

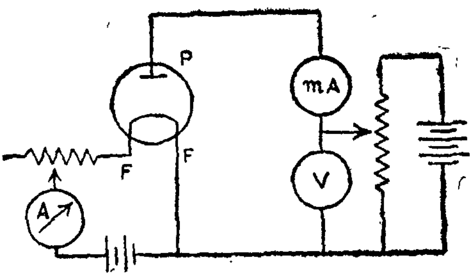


Fig. 6-9

A diode characteristic circuit

The electrons reaching the plate, as stated above, will continue through the outer circuit in the direction shown in

Fig. 6-10. The conventional current is in the opposite direction. The ammeter A in the circuit will measure this current, while

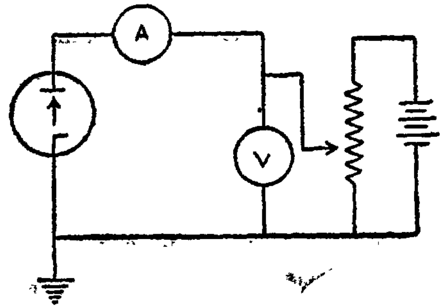


Fig. 6-10.
Circuit showing the direction of flow of electrons

the voltmeter V will measure the p. d. applied between the plate and the cathode. The current is called the plate current I_p while the p.d. is called the plate voltage V_p. When the plate voltage is increased, the plate current also increases and, after a particular stage, becomes constant. This maximum value of the current is called the saturation value. The variation of current with voltage is shown in the graph, for a particular temperature of the cathode (curve B). If the temperature of the cathode is increased, the saturation value goes up (curve A). It must be noted that, if the plate is made negative with respect to the cathode, that is when it is reverse biased, there will be no current, as the electrons will be repelled by the plate there can be no flow.

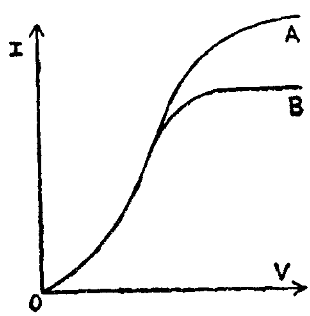


Fig. 6-11
Diode characterestic curve

The diode therefore allows current to pass through only in one direction. The curves which represent the behaviour of the plate current with respect to the plate voltage are called the characteristics of the diode.

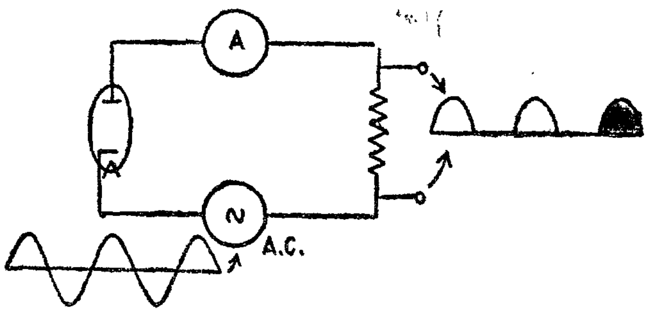


Fig. 6-12
An a.c. voltage applied to the diode and the current in the external circuit.

It is usual to say that when a diode is *forward biased,* it has a low resistance, and when it is reverse-biased, it has a very high resistance. This property is used for the conversion of a.c. into d.c. which is called '**rectification**'. In the circuit shown, when the applied voltage is such that the plate is negative, the diode will allow no current.

Therefore, an alternating voltage applied to the diode will allow current in the external circuit only during those half-cycles when the plate is made positive with respect to the cathode. The applied voltage and the current in the external circuit are shown

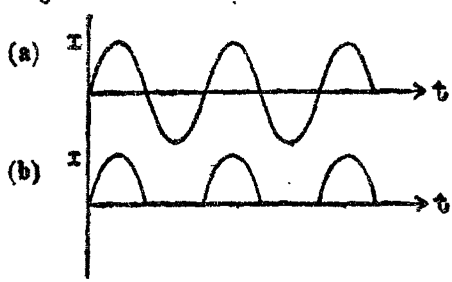


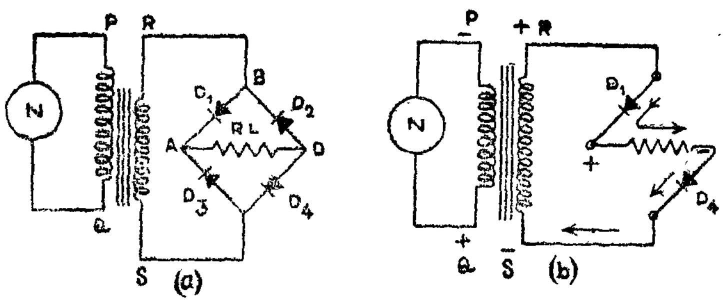
Fig. 6-13
The input and the output when an a.c. is applied to a diode
(a) applied voltage (b) current

in Fig. 6-13. The current in the external circuit is always in one direction but flows only for half a cycle. Therefore the circuit is known as the *half-wave rectifier*. The diode in this case functions simply as an automatic switch which connects the supply voltage to the external circuit for a half cycle. i.e. only when the plate is positive with respect to the cathode.

In the above case, one half of the supply voltage is not used. In the circuit given below (Fig. 6-14) *full wave rectifier* is shown in which the full cycle is used.

D_1,D_2,D_3, and D_4 are the diodes and the solid triangles represent the plates and actually show the directions in which the current will flow. When the generator voltage is such as to make the point R positive (Fig. 6-14 a), the end B is positive and

(Symbolic representation of a diode is used in the figures below.)



(*a*) **Full wave rectifier** (*b*) **Path of current when R is +ve**

Fig. 6-14

C is negative. Therefore D_1, and D_4 have the proper directions for conduction, and the current will flow through D_1 and D_4 through R_1, as shown in Fig. 6-14 (*a*). The path of the current is shown by the arrow.

During the next half of the generator voltage, S will be postive and R negative, which will make C positive and B negative. Now D_2 and D_3 are in the proper direction for conduction. Therefore the current will flow through D_2 and D_1 through R_L, as shown in Fig. 6-15.

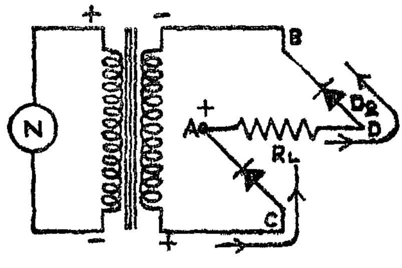


Fig. 6-15.

Path of current when S is + ve

It will be seen that the current through R_L is always in the same direction, and that the current flows through R_L during both halves of the generator voltage. The supply voltage and the electric current in the external circuit are shown below:

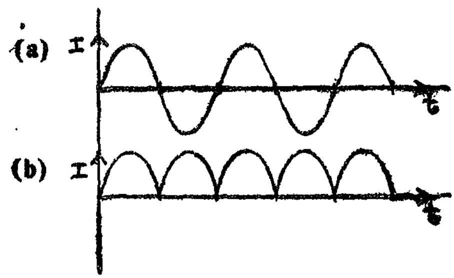


Fig. 6-16

The input and output of a full wave rectifier
(a) Supply voltage (b) Current in the external circuit

The voltage across the resistance R_L will have the same shape as the current.

6.6.3. SEMI CONDUCTORS

Electrical conductivity of solids depends on the current-carrying electrons which are free to move through the material. Good *conductors* like metals have approximately one free electron per each atom. In an *insulator* there are practically no free electrons. Germanium crystals are insulators where impurities are less than 1 part in a billion. Traces of impurities produce some conductivity due to the presence of as little as one free electron per every million atoms. Germanium of this type is called a **semi conductor**, since its conductivity lies between that of a conductor and that of an insulator. The conductivity of the semi-conductor can be controlled by adjusting the amount of impurity added to it.

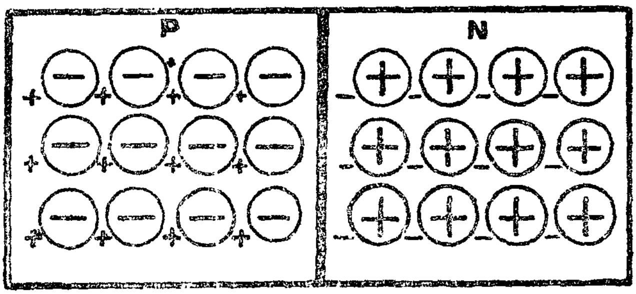
The materials generally used are germanium or silicon. They belong to the fourth group of the Periodic table and have four electrons in their outermost orbits. In the crystal structure these 'valence' electrons of each atom are shared by their four nearest neighbours. The bond formed by sharing of two electrons between two atoms is called a covalent bond. At temperatures close to the absolute zero, all electrons in the crystal are held strongly by these chemical bonds. When the crystal is raised to the room temperature, however, the thermal vibrations of the atoms cause some of the bonds to break thus freeing some of the electrons. Where an electron thus breaks free, a vacancy has been created, and this is called a '**hole**'. The part of the crystal which was neutral in the beginning now lacks an electron. The hole is thus equivalent to a net positive charge. Thermal vibrations again can cause a bound electron next to the hole to move across to fill the gap. The net motion of the negative charge from one bonded position to another is effectively equivalent to the motion of a hole in the opposite direction. The holes therefore act like mobile positive charge-carriers.

If the germanium crystal has an element from the fifth group, say antimony, introduced as impurity, then every atom of antimony is surrounded by four germanium atoms. There are five electrons in the outermost orbit of antimony, and four of these get engaged in establishing covalent bonds with germanium neighbours. Thus, one is left free, and becomes available for conduction, thus increasing the conductivity of the crystal. An impurity like antimony is called '*donor*' impurity, because they donate an electron for conductivity. The electrical conductivity is primarily due to these electrons. Unlike in the pure germanium mentioned earlier, the presence of the free electron does not involve a vacancy, i.e. there is no corresponding hole. Since the current carriers are primarily negative charges here, this is called an *N*-**type** semi-conductor

Consider the impurity to be introduced from the third group of elements from the periodic table, say aluminium. Aluminium has only three valence electrons, and therefore can form bonds only with three of the four germanium atoms. Therefore the fourth place has a vacancy, that is a hole. In this system, the conductivity is mainly due to these holes which are equivalent to positive charge carriers as explained above. A semi-conductor of this type is called a *P*-**type** semi-conductor. As each impurity atom (aluminium) has a vacancy which can accept an electron, it is called an '*acceptor*' type impurity.

6.6.4. SEMI CONDUCTOR DIODES

If the two semi-conductors *R* and *N* types respectively are put together as in Fig. 6-17 (a) we have a ***P-N junction diode.*** Roughly, the action of the junction diode may be described as follows: The '*P*' type crystal has a predominance of holes over free electrons, as was said earlier, while the '*N*' type crystal has



(a) *P-N* junction diode-potential barrier

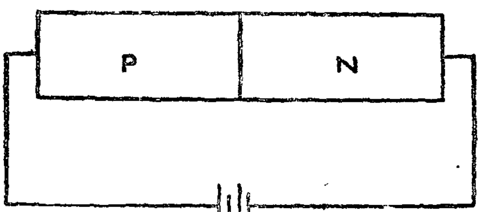
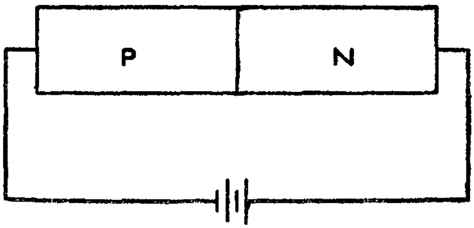


Fig. 6-17

(b) Diode-Forward biased

a predominance of free electrons over holes. When they are in contact, there will be a diffusion of the free electrons across the junction, which causes neutralisation of some of the holes on the '*P*' side. At the same time the removal of electrons from the *N* side means creation of some holes. The net result may be described as a flow of electrons to the '*P*' side from the '*N*' side and a flow of holes from '*P*' side to the '*N*' side. To begin with, both crystals ('*P*' and '*N*' types) were neutral, and this two-way



(c) Diode Reverse biased
Fig. 6-17

flow across the junction causes a net positive charge on the '*N*' side(due to the removal of negatively charged electrons, and the inflow of positively charged holes) and a negative charge on the '*P*' side (due to the inflow of electrons and removal of holes) in the vicinity of the junction. This creates a **potential barrier**, depending on the '*P*' and '*N*' materials used across the junction. When this stage is reached, the '*P*' side across the junction is negative with respect to the '*N*' side.

If a battery is connected to the terminals of the diode as shown in Fig. 6-17 (b) the potential barrier across the junction is reduced, and this will allow a net flow of current carriers viz. holes and electrons. If, however, the battery is connected as in Fig. 6-17 (*c*), then the potential barrier is actually increased, and no net flow charge carriers will be possible, i.e. there can be no current in this case. The action is the same as in the vacuum tube diode, and hence the name.

The *PN* junction diode allows the current to flow through when the '*P*' side is positive with respect to the '*N*' side, and not vice versa. In the former case, it is said to be *forward biased*, and in the latter case, *reverse biased*. These can be used for rectification in the place of vacuum tubes as explained in Fig. 6-13.

The variation of the current through the diode with applied voltage under forward and reverse bias conditions is shown in

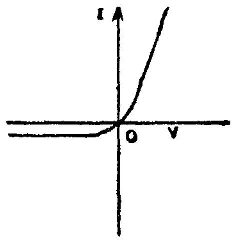


Fig. 6-18

Relation between applied voltage and current in a junction diode

Fig. 6-18. It will be seen that there is a small current under reverse bias also, unlike in the vacuum diode.

Exercise 6.6

1. *What are the two types of circuit elements?*
2. *Describe the construction of a diode.*
3. *Describe the action of a diode when an a. c. voltage is applied to it.*
4. *Explain a rectifier circuit with a suitable diagram.*
5. *What are semi-conductors? Why are they called so ?*
6. *What is a hole?*
7. *What is an N-type semi-conductor.*
8. *What is a P-type semi-conductor?*
9. *What is a P-N junction diode?*
10. *Explain the action of a P-N junction diode.*
11. *What is meant by saying a P-N junction diode is (i) forward biased and (ii) reverse biased ?*

7. APPLIED PHYSICS

7. 1. ATMOSPHERE

7.1.1. ATMOSPHERIC PRESSURE

As you know, the *atmospheric pressure* is the pressure exerted by the huge column of air which surrounds the earth. One of the main constituents of the atmospheric air is water vapour. The amount of water vapour contained in the atmosphere plays an important role in determining the conditions of the weather. As the water vapour has a density less than that of dry air, the atmospheric pressure decreases with the increase of water vapour in the atmosphere. Hence the weather may be predicted from observations relating to the atmospheric pressure.

The atmospheric pressure at a place undergoes hourly, daily and annual changes. Hence, to study the weather, the atmospheric pressure has to be measured simultaneously at different places. The changes in the atmospheric pressure have two origins, namely, (i) thermal origin and (ii) dynamic origin.

The **thermal variation** of the atmospheric pressure is due to the changes in the density of air with temperature. The density of air decreases with increase of temperature. Hence the variation of temperature causes vertital and horizontal displacements of air leading to changes in the atmospheric pressure. Therefore the atmospheric pressure will be high at regions of low temperature and vice versa.

The **dynamic variation** is due to friction and the centrifugal force developed in the atmospheric air. These forces are caused by the rotation of the earth. The rotation of the earth gives rise to a sliding motion of the surface of the air in contact with it.

7.1.2. HUMIDITY

Humidity is a measure of the content of water vapour in the atmosphere. You might have studied in the junior classes about humidity and how it affects the human comfort. The content of water vapour and hence the humidity of the atmosphere plays a major role in the formation of clouds, mist etc., and hence in controlling the conditions of the weather.

You may also know that the water vapour in a given space can have a maximum pressure at any temperature, known as the saturation pressure. When the atmosphere contains water vapour at this pressure, it is said to be *saturated* with water vapour. At this stage no more evaporation can take place. When the atmosphere is in the saturation state or in the near-saturation state, evaporation stops or is slowed down. The more the amount of water vapour contained in the atmosphere at a particular temperature,the nearer it is to the saturatian state and hence slower will be the evaporation. If the water vapour in the atmosphere (at a temperature) is increased beyond saturation, it is said to be *super saturated.* This super saturation will make the excess water vapour to condense. The super saturation of the atmosphere may be brought about by the decrease of its temperature. It is this super saturation and condensation of water vapour that bring about the formation of clouds. If the atmosphere contains more water vapour at a particular temperature, it will be nearer the saturation stage at that temperature; super saturation can be attained with ease.

7.1.3. FORMATION OF CLOUDS

We have seen that the condensation of water vapour in the atmosphere is the cause of formation of clouds, In order that water vapour may condense, it requires some nuclei. The smoke from factory chimneys, pollen grains from the flowers, dust particles, ions all form such nuclei. When air becomes cooler and cooler, water vapour condenses around these nuclei forming tiny droplets. Millions of such droplets cluster together in groups forming clouds. If these groups are formed at ground level, they are called *fog.* The clouds are formed at heights ranging from 1 km to about 12 kms. It is possible to predict a storm or heavy rain in advance by an appraisal of the nature of the clouds.

7.1.4. DEW, FROST, SNOW STORM AND HAILSTONES

When the soil rocks and plants cool during night, the layer of air which is close to the ground is also cooled. The water vapour in the cooled air condenses as tiny drops of water on blades of grass and other plants. These tiny drops are *dew* drops. On very cold nights the water vapour condenses directly into ice crystals. These crystals form a white film on the ground, known as *frost*.

When the air high in the sky cools very quickly, the water vapour condenses directly into ice crystals leading to a *snow storm*. *Hailstones* actually start as rain drops. But before the drops of water can fall very far, the wind blows them up to the colder regions where they freeze and begin to fall again. As they fall more water condenses on them. They are again blown up and the process will be repeated till the frozen drops of water become too heavy to be lifted by the wind and fall to earth. These frozen drops of water are known as hailstones. Hailstones are usually not much bigger than a pea, though some may be as large as cricket balls.

7.1.5. LIGHTNING

Lightning is nothing but an electrical discharge through the atmosphere due to the charges accumulated on the clouds. The discharge may take place between the clouds or between the cloud and the earth. The atmospheric air is continuously ionised by the action of the ultraviolet rays from the sun and the cosmic rays. As a result, the atmosphere always contains positive and negative ions which create an electric field in it. When the water droplets of the cloud which are initially neutral fall through this electric field, they get charged, some negatively and some positively. According to C.T.R.Wilson, the heavier droplets acquire negative charges while the lighter ones acquire postive charges. This leads to the clustering of negative charges at the base of the cloud and the positive charges at the top. Sometimes a small quantity of positive charges is also found at the base of the cloud. The accumulation of charges continues until the potential difference becomes so great that a discharge takes place between the charged surface of the same cloud or between two clouds. The flash produced by the discharge is the *'lightning'*. Further, the charges on the cloud induce charges of opposite kind on the surface of the earth, leading to a discharge between the cloud and the earth. The flashes are often very long sometimes several kilometres, the diameter being 10 to 15 cms. Often, many flashes of very short duration follow slightly different paths in quick succession. This produces an illusion of *forking*.

The potential differences causing the flash is very large ranging from about 100 million to 2000 million volts, the value of the current being as high as 20,000 to 200,000 amoeres. It has been estimated, that over the entire globe, lightning occurs at a rate of about 100 per second. This rate of electric discharge represents about 4 terra (10^{12}) watts of continuous power. It is this enormous potential difference and the current which produce the destructive effect of lightning. In spite of its destructive effects lightning is also beneficial to mankind. Owing to the enormous heat produced during lightning, nitrogen is separated from air. This nitrogen dissolves in rain water and reaches the earth which is the main cause of prolonged fertility of the soil.

7.4.6. THUNDER

About 75% of the electrical energy of the lightning discharge is used up in heating the atmospheric gases in and around the flash, raising the temperature to about 10,000 K in about a few microseconds. As a result, a pressure wave, on expansion, gives rise to compressions and rarefactions, producing a violent sound called *thunder*.

7.1.7. WEATHER FORECASTING

Weather forecasting is the forecast of the conditions of the atmosphere based on observations of factors such as the temperature, the humidity, the direction and the velocity of winds and the types and movements of clouds. Anticipation of future weather is of vital use in many areas.

In the field of agriculture, long range forecasting helps efficient planning for maximum productivity and reduction of waste.

Marine forecasts are made about the wind, the visibility, the storm and the general conditions of weather in coastal areas. Such forecasts help to preserve life and property at sea and along the coastal areas. The activities in shipping and fishing areas are also made safe.

Aviation forecasts are made for particular geographical areas, for one or more air terminals and for specific routes and flights. These forecasts are essential for safe and efficient operation of air navigation. The upper-air observations and forecasts play an important role in space flights.

7.1.8. PRINCIPLE OF WEATHER FORECASTING

Weather *forecasting, in general, consists of three stages.* In the *first stage*, the prevailing climatic conditions in a region and its surroundings are to be found. These observations are entered in a weather map which is called a *synoptic chart.* The next stage is to analyse the chart and estimate the weather corresponding to a later period. In the third stage, these estimates are compared with the developments that had taken place in the past under similar *conditions.* Then, *from a knowledge which* the forecaster has gained about the factors which change the climatic conditions, the forecast is made.

For efficient forecasting, an idea of the *conditions in the upper air is also essential. These are determined using balloons, radiosondes* and *airplanes.* Nowadays satellites equipped with cameras are being used to orbit the earth to obtain valuable data for weather forecasting.

7.1.9. CYCLONE TRACKING

Before going into the details of cyclone tracking, it is useful to know clearly about the cyclone and its nature.

The word cyclone is defined in the dictionary as a low pressure storm area into which air spirals, usually forming clouds and rains. The *word has its origin from* Greek words "Kyklos" and "Kykloein" meaning circle and whirling.

How does a cyclone develop ? *Owing to the* warming up *of* the atmospheric *air by the heat of the sun,* the lower levels of air are set in constant motion from the poles to the equator This motion is converted to a spiral motion by the earth's rotation. If, in addition, the air passes over large seas, it picks up the water vapour and heat; it moves slowly at first and picks up speed.

Pictures of the cyclonic storm that struck East Pakistan in 1970

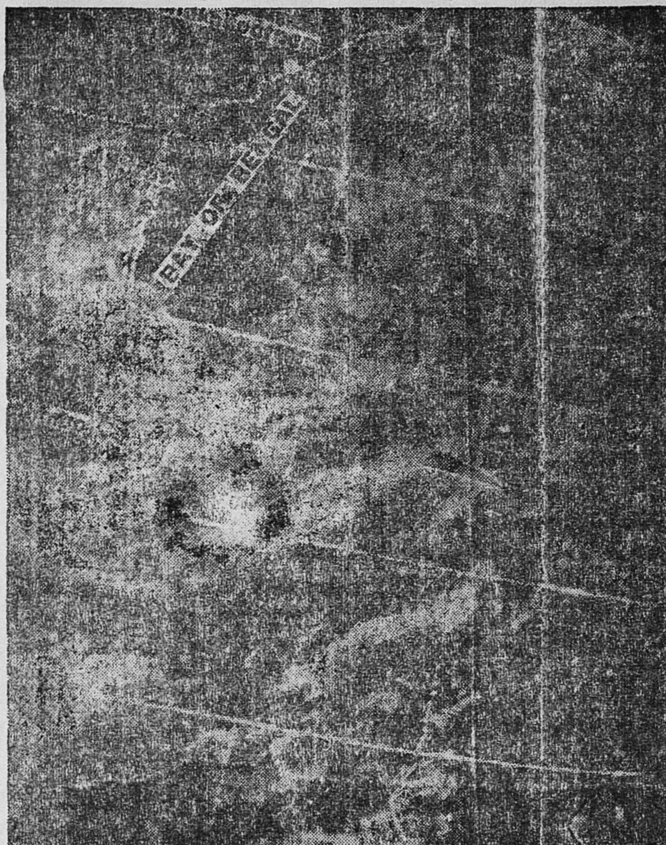


Fig. 7-1 *(a)*

Nov. 8 - time 15 *hrs.* 26 *min.* IST. Unsettled weather conditions in the Bay of Bengal –Well marked low pressure area: a cyclone is born

A glance at the pictures of the cyclonic storm **Fig. 7-1** (*a*, *b*, *c*, *d*, *e*, *f*) that struck **East Pakistan in 1970, taken by** TOS-1 will show how a cyclone can be tracked efficiently.

Fig. 7-1 (*b*)

Nov. 9- time 14 *hrs*. 39 *min*. IST. The cyclone takes shape and an 'eye' is seen forming. The clouds cover almost the whole of the Bay of Bengal

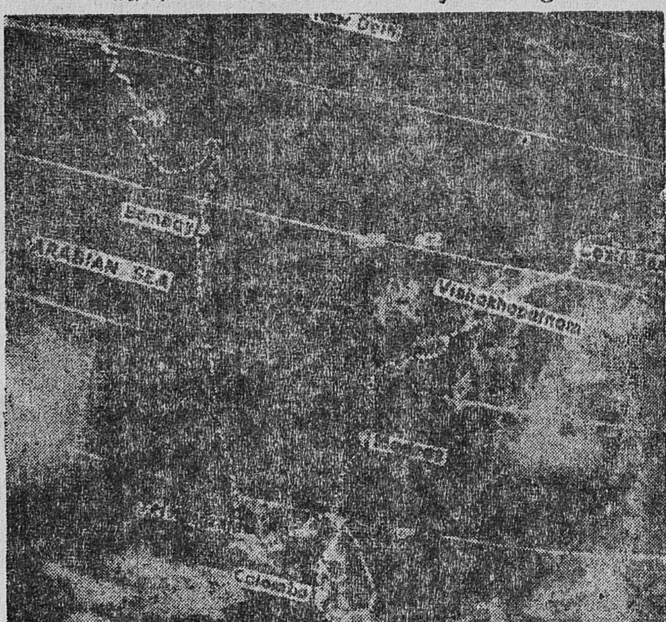


Fig. 7-1(*c*)

Nov. 10- time 15*hrs*. 27 *min*. IST. The clouds have formed into thick smaller circular bands and the 'eye' has become more prominent

Fig. 7-1 (*d*)

Nov. 11- time 14 *hrs.* 29 *min.* IST. The cyclone has intensified into a huge vortex with a prominent ,eye' which is about 400 km due.east of Vishakapatnam

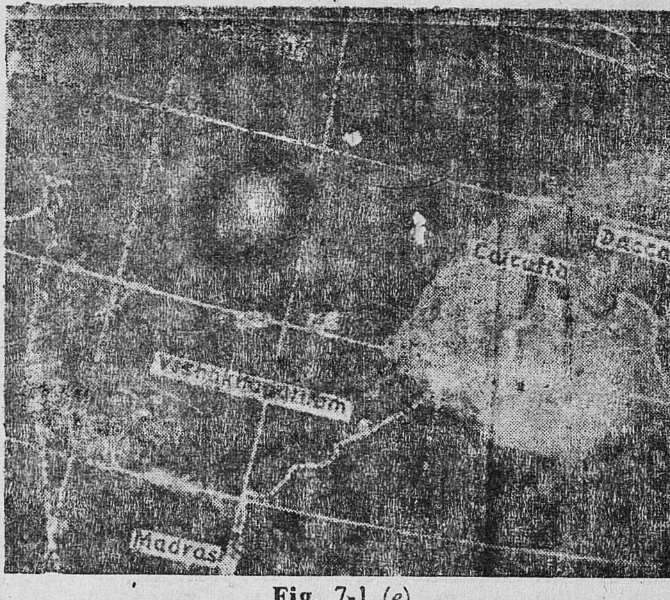


Fig. 7-1 (*e*)

Nov. 12- time 15 *hrs.* 24 *min.* IST. The violent cyclone crossed the coast of East Pakistan toward nightfall. Note the clear 'eye' on the cloud mass over sea-a black dot below the 'x' mark in the picture

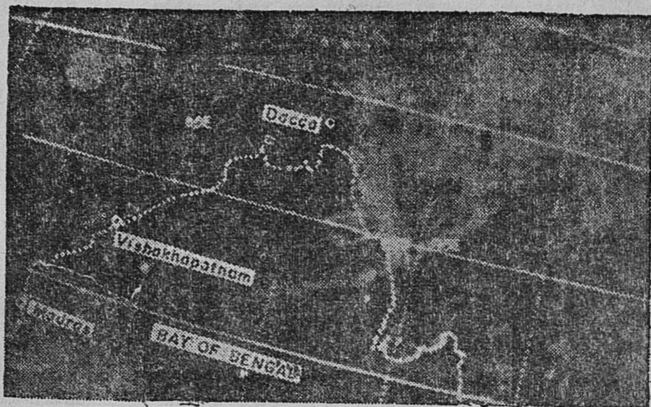


Fig 7-1 (*f*)

Nov. 13 - time 14 *hrs*. 26 *min*. IST. The cyclone has blown over the land virtually spending its fury. The cyclone is dead but remains as a huge cloud mass

The earth's rotation causes it to spiral inwards. As the air moves to the centre of the spiral, the warm air and the clouds pick up more speed. At the vortex of the spiral, they rise upward and the clouds spread out in great sheets. As a result the clouds are cooled resulting in heavy rains The spiral of air followed by the heavy rains is called the cyclone.

With the development of man's ability to place artificial satellite in orbit, it has now become possible to have a close watch over the birth and the movement of the cyclones to a high degree of accuracy and to forewarn about the impending cyclone. The satellites designed for this purpose are known as *meteorological satellites*. The first group of such satellites belong to the *TIROS* group (Television and Infra Red Observation Satellites). These satellites carry special types of TV cameras along with many meteorological instruments and automatic picture transmission system. Switching system is provided to start the cameras when the satellite approaches the sunlit side of the earth, and to shut them off on the night side so that the cloud patterns may be determined even during nights. Improved models of *TIROS* viz. *TOS* (Tiros Operation Satellite) are now in use.

Exercise 7.1

1. *What is meant by the thermal variation of atmospheric pressure ?*
2. *What is meant by the dynamic variation of atmospheric pressure ?*
3. *How do the clouds form ?*
4. *Explain the formation of frost and hailstones.*
5. *How does lightning take place ?*
6. *How is thunder produced ?*
7. *What are the uses of weather-forecast ?*
8. *How are the weather forecasts made ?*
9. *How is the storm formed ?*
10. *What are TIROS ?*

7.2 DEFENCE

7.2.1. SOUND RANGING IN LOCATING ENEMY CAMPS

A very important application of the acoustical principles in defence is that of finding the exact position of an enemy gun from the sound of its firing. The method, known as "*Multiple point sound ranging*" consists of arranging atleast three receivers of sound at known positions in a straight line and determining the instants at which the receivers receive the firing sound in succession.

The principle can be explained as follows: Suppose two observers situated at a distance receive the sound due to the explosion of a gun simultaneously, they may easily guess that the gun is situated on the perpendicular bisector of the line joining them. If, however, one of them receives the sound earlier than the other then the gun will be situated at a point away from the bisector and on the side of the observer who receives the sound earlier. From a knowledge of the time interval between the instants of the arrival of the sound by the two observers and the distance between them, it is possible to calculate the direction of the gun. From the observations made by another pair of observers, the position of the gun may be exactly located.

In practice the sound is received at more than three stations at known distances by microphones which are connected to a central station, where the instants of the arrival of sound at each station is recorded on a moving strip of paper. Thus knowing the time of intervals between different pairs of stations and the distances between them, the position of the gun will be located exactly. It may be added that the same principles are applied also to locate explosions of mines and torpedoes in the sea.

7 2.2. RADAR

The name "RADAR" has been derived from the initial letters of the words **"Radio Detection And Ranging"**. It is a system for locating the objects which reflect radio waves. Any object which reflects the radio waves may be detected and located in space. Since the radio waves penetrate darkness, fog and rain, this method is employed in locating planes, ships and other objects more accurately than by employing optical and acoustic methods.

Though the origin of the radar may be traced to 1923, the method was successfully used only in 1929 by Appleton. He utilised the principle of the radar to measure the altitude of the different layers of the atmosphere. But in 1935, Sir Robert Watson Watt found that the reflected radio wave from the object can be visually observed on a screen by means of a cathode-ray tube. After this discovery, the radar underwent many improvements and it played an important role in World War II.

7.2.3. PRINCIPLE OF THE RADAR

Though the working details of the various radar systems are different, the basic principles are the same. A series of radio frequency pulses are transmitted by a *'directional antenna'*. The pulses are very short and very accurately timed. The frequency range is approximately from one hundred *MHz* to several thousand *MHz*. If these pulses strike a conducting object, some of the energy is reflected. The reflected pulses may be picked up by the radar receiving system with the help of the same antenna. The direction of the antenna when a reflected pulse is detected gives the direction of the object. The duration between the transmission of the pulse and its reception gives a measure of the distance of the object.

Fig.7-2

A Radar Antenna

7.2.4. USES OF THE RADAR

The uses of the radar both in war and peace are numerous. It is useful in guiding the fighter aircraft, to intercept and engage the enemy bombers before they could reach their targets. For this it provides continuous positions of the approaching bombers at the greatest possible range from the radar station. It also tells the height of the bombers so that the fighters could go straight up to the right height as well as to the correct position. Besides, the radar provides information about the strength of the bomber force enabling to send the right number of fighters. Each aeroplane sends back a certain amount of energy and the total effect of the reflections from the group of enemy planes gives a clue to the number of aircrafts in the group. When the radar units are fitted in the bombers, they help the bombers to locate their targets exactly. Thus the radar is useful both in defence and in offence during war time.

Apart from these uses of the radar in war, it has got many applications in meteorology and astronomy. In meteorology, it is used for the detection of cyclones and in tracking them. In astronomy it is used to measure the distance and the nature of the surface of the moon. It may also be used to determine the distance of the planets.

7.2.5. SONAR

Sonar is an under-water device which is used to obtain information about objects or events below the surface of water with the help of sound waves. It is very much similar to the radar in principles. Radar is of no use under the water, since radio waves do not pass through water easily, whereas sound waves are easily transmitted through water. The term 'Sonar' is formed by using the first letters of the words **'Sound Navigation And Ranging'**.

The sonar equipment can produce, transmit and detect sonic or ultrasonic waves. A pulse of sonic or ultrasonic waves is sent through water from the sonar equipment of a ship. It this pulse strikes any solid object on its path, an echo is produced which will be received by the sonar equipment. The direction of echo indicates the position of the object. The distance of the object is calculated by finding the time taken between the transmission of the original pulse and the reception of its echo. This method may be adopted to hunt the submarines during the war.

Sonar has also many uses in peace time. It is helpful to navigators in obtaining the depth of water at a particular place.

The depth of water is realised by sending an ultrasonic beam down towards the ocean-bed and receiving its echo. The sonar equipment fitted to fishing boats helps to locate shoal of fish With the help of more sensitive sonar systems, it is possible even to identify the fish, finding the size of each shoal and the depth of swimming. Further the system is of great help in navigation through narrow and rocky channels.

7.2.6. INFRA-RED PHOTOGRAPHY

You have learnt about the production, properties and range of the infra-red rays under the chapter *"Electromagnetic Radiation"*. They can be detected in many ways of which the photographic method is one. Since the infra-red rays undergo negligible scattering, they can penetrate fog and haze through long distances Hence they provide a better method for long distance photography.

The infra-red photographs (*IR* photographs) may be taken using an ordinary camera fitted with a suitable *IR* filter. But the photographic plates should be specially sensitive to the *IR* region. These photographic plates are manufactured by Ilford Co. During war this method can be used for airplane photography, since there is a *"dense haze"* between the earth's surface and the high-flying plane. Also the landscape may be photographed by aircrafts flying at high altitutes covering larger areas Photographs taken with infra-red rays reveal more details.

7.2.7. IR TELESCOPES

The distant objects may also be seen visually with devices called "IR telescopes". Such devices are found to have much of military interest. Two simple factors render this possible viz. (i) they are invisible and (ii) all targets (including human beings) are emitters of infra-red rays. Hence it is possible to illuminate a target with the infra-red and observe it without the observer being aware of it.

The device uses a source of the infra-red which illuminates the target. The light reflected by the target is collected by an optical system. It is then focussed on the front surface of an image converter tube. The front surface consists of a caesium-oxide silver surface to which a high negative voltage is applied.

This high voltage drives the electrons from the surface and focusses them on the back surface which is a phosphorescent screen and to which the positive of the high voltage is applied. The focussed electrons produce a greenish image of the target which can be viewed by an eye-piece which magnifies it (Fig 7-3).

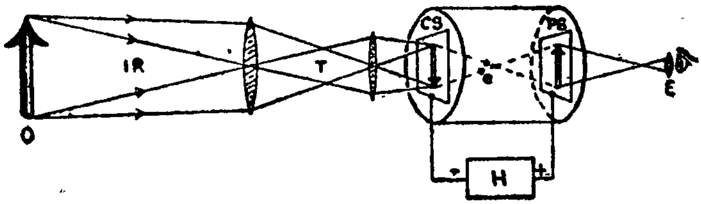


Fig. 7-3
IR Telescope

If this unit is fitted to a rifle, it is called a *sniperscope* which is used for shooting in the night. Similar type of binoculars enable vehicle drivers to drive the vehicles in the night.

The police can use such telescopes to observe the movements of the miscreants.

Exercise 7.2

1. *Explain sound ranging.*
2. *What is meant by Radar ?*
3. *Explain the principle of Radar.*
4. *What are the uses of Radar?*
5. *What is meant by Sonar ? What are its uses ?*
6. *Explain the infrared telescope with a neat sketch Mention some of its uses.*

7.3. CINEMATOGRAPHY

7.3.1. PHOTOCELLS

It was mentioned in the chapter on Atomic Physics that one mode of emission of electrons is the photo-electric emission. The devices which use the photo-electrons are called the *photocells*. There are three types of photocells. They are (i) *photo-electric* *cell, (ii) photo-voltaic cell and (iii) photo-conductive cell.* Though all of them are based on the same principle of the emission of electrons under the influence of light radiations, they differ slightly in principles of construction. The photo-electric cell is based on the *outer* photo-electric effect in which the liberated electrons are ejected out of the material exposed to light. The other two are based on the *inner* photo-elelectric effect in which the liberated electrons remain inside the material and increase the number of free electrons responsible for the conduction of electricity

7.3.2. PHOTO-ELECTRIC CELL

The photo-electric cell is also known as the emission cell or alkali-metal cell. It consists of a cylindrical surface *C* with its concave surface coated with an alkali metal such as potassium, rubidium or caesium, enclosed in a glass or quartz tube. The cylindrical surface acts as the cathode. A thin wire '*A*' stretched along the axis of the cylindrical surface serves as the anode (Fig 7-4). The tube may be highly evacuated or filled with an inert gas such as helium, argon or neon at a very low pressure.

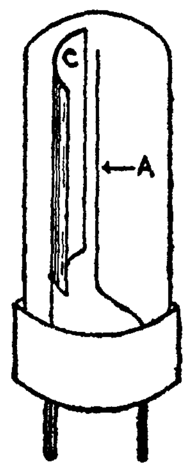


Fig. 7-4
Photo-electric cell

If an external source of d.c. supply is connected to the cell through a milliammeter to measure the current as shown in Fig 7-5, in the absence of any light radiation there will be no current (Fig. 7-5 *a*). If, however, a beam of light is made to be incident on the cylindrical surface, a current is found to flow through the circuit (Fig. 7-5 *b*). As the photo-electric current thus produced is very weak, it is usually amplified by means of suitable circuits.

The amplified current may be used to operate a loud speaker as in sound reproduction in cine films or a relay as in automatic switching of street lights and automatic door closers or a counting mechanism as in automatic counting of objects.

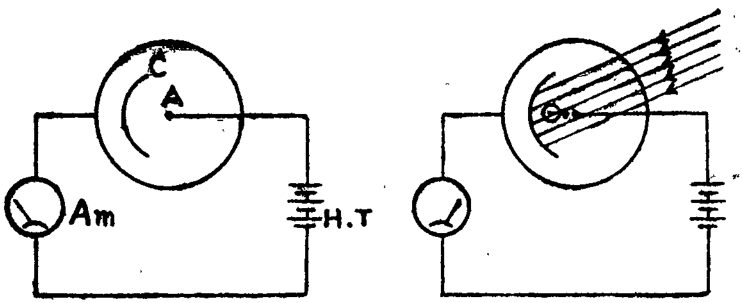


Fig. 7-5

(a) No current in the absence of light

(b) When light falls on C there is current in the circuit

7.3.3. PHOTO-VOLTAIC CELL

The photo-voltaic cell is a true cell in the sense that it generates an e.m.f and requires no external power supply for its operation. The cell in general consists of a layer of a semi-conducting material such as cuprous oxide or selenium formed on the surface of a metal plate. Over the semi-conductor is a thin semi-transparent film of metal which maintains electrical contact

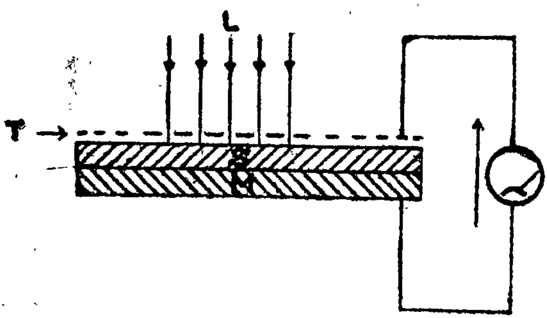


Fig. 7.6
Photo-voltaic cell

between the semi-conductor and the metal, and at the same time allows light to illuminate the semi-conductor layer. (Fig 7-6).

When light falls on the thin metal film and illuminates the layer of semi-conductor, an e.m.f is developed between the two metal layers. Such photo-voltaic cells are used in photographic exposure meters, illumination meters and other such meters used for the measurement of the intensity of illumination. The major application has been as a source of electrical energy in space-crafts in the form of solar cells. Thousands of such cells will be fixed to the outer surfaces of the space-crafts. These cells under the influence of sunlight produce enough power to operate the various instruments in the space-crafts. About 1800 such cells fitted to the Indian satellite **Aryabhatta** were capable of producing a power of about 46 watts.

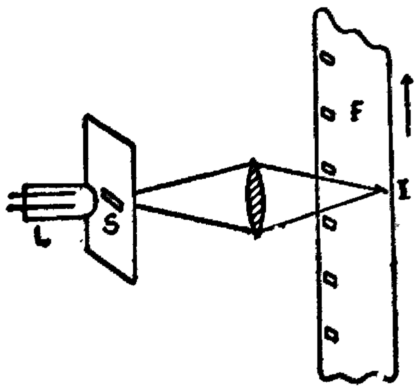
7.3.4. PHOTO-CONDUCTIVE CELL

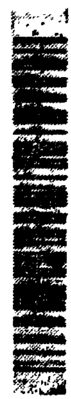
The photo-conductive cell is based on the photo-conductive effect exhibited by some substances like cadmium sulphide or cadmium selenide. Such substances when exposed to sunlight undergo large changes in their conductivity approximately in the ratio of $1:10^9$. These are used as detectors of radiation, and as switches which are sensitive to light and hence actuate relays. They are used for infrared detection, in xerography which is a dry process for photo copying and in T.V pick-up tubes.

7.3.5. FILM RECORDING (SOUND)

Attempts to produce a "*talking*" motion picture were made as early as 1925 when Thomas Alva Edison recorded the "*talk*" on a gramaphone record and synchronised it with the picture. In about three years it was possible to produce a sound record on the film itself so that perfect synchronization could be attained between the picture and the "talk"

To record the talk or the sound on the film, it is necessary to convert sound waves into variations of light to which the film could be exposed. For this, the sound vibrations are first converted into electrical variations by a microphone. The varying current is then amplified by suitable circuits and passed through a galvanometer or through a special lamp. The lamp or the galvanometer varies or modulates the light reaching a constantly moving film in such a way that either the amount of exposure or the area exposed varies in accordance with the sound vibrations produced before the microphone. When the amount of exposure is varied, the sound vibrations are recorded on the film as differences in the density of blackening of the film. Such a recording is known as *variable density recording* (Fig. 7-7*b*). When the area exposed is varied, the vibrations are recorded as differences in the width of the exposed area and the recordings appear as a wavy curve(Fig.7-8*b*). The recording is known as *variable width-recording*.

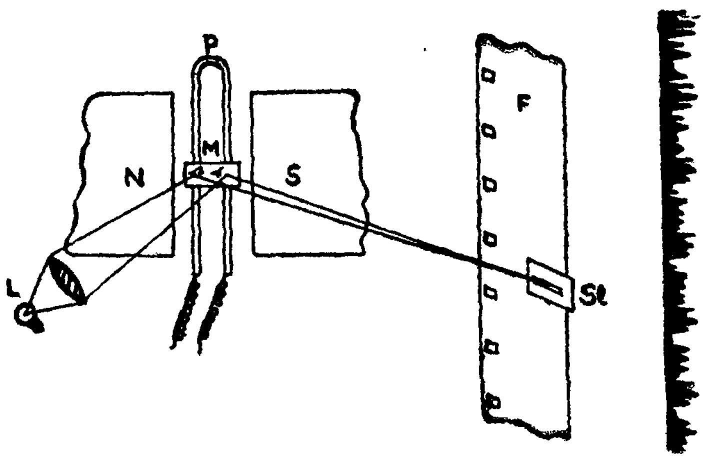




(a) Fig. 7-7 (b)
Variable density recording of sound

In the variable-density recording, a fine slit of dimensions 0·05 mm × 6·5mm is illuminated by a special type of discharge lamp whose intensity can be varied in as short an interval as a few micro seconds and hence in accordance with the sound vibrations. A reduced image of the slit (0·025mm × 2·5 mm in 35 mm film) is focussed on to the edge of the film which will be moved at a constant speed perpendicular to the length of the image of the slit (Fig. 7.7a). As the intensity of the lamp varies, along the edge of the film, different portions (in the form of stripes of 0·025 mm width) will be exposed to different amounts of illumination. The film when developed will contain the sound recordings in the form of stripes of same dimensions of different densities of blackening.

In the variable width recording, the amplified, varying current obtained from the microphone is passed through a fine loop of phosphor-bronze strip of a galvanometer. The phospher bronze loop, placed between the pole pieces of a strong magnet, undergoes a deflection which varies in accordance with the sound vibrations. Light from a lamp of constant intensity is focussed on to a mirror attached to the phosphor-bronze strip. The reflected light from the mirror is focussed onto a slit close behind which is run the photographic film at a constant speed. As the phosphor-bronze loop, along with the mirror, is deflected, the reflected light constitutes a line of light illuminating the slit to different lengths and hence exposing the film to different lengths along the slit. The film when developed will have a serrated black band as in Fig. 7-8 *b*.



(a) Fig. 7-8 (b)

Variable width recording of sound

7.3.6. REPRODUCTION OF FILM (SOUND)

For the reproduction of sound, the film has to move continuously. Further if the sound record of a particular scene is in the same frame as the scene, due to the difference in velocities of light and sound, the sound will be delayed in reaching the audience and will not be in synchronization with the motion of the lips. Hence usually the sound record pertaining to a frame is made to lie in advance at a distance of about 37 cms (in 35 mm film) from the corresponding frame. This is done by making the optical and sound records in separate films and then superposing them suitably.

Now coming to the process of reproduction of sound, light from a special type of lamp is narrowed down to a thin strip of the same width as the sound track and is focussed on to the sound track. The amount of light emerging from the other side of the film will vary due to the varying densities or varying areas of transparent portions of the sound track. The emerging light of varying amount is made to be incident on a photo-electric cell which produces a varying current. The varying current is then amplified by means of suitable amplifiers to actuate the loud speakers.

Exercise 7.3

1. *What are the kinds of photo-cells? Mention some of their uses?*
2. *Explain a photo-etectric cell with a neat sketch.*
3. *Explain the photo-voltaic cell with a neat sketch.*
4. *Explain a photo-conductive cell with a neat sketch.*
5. *What is meant by xerography?*
6. *What are two kinds of sound recording?*
7. *How is the variable density recording done?*
8. *How is the variable width recording done?*
9. *How is sound reproduced by a projector?*

NEW PHYSICS

8.1. EARTH'S MAGNETIC FIELD

8.1.1. ELEMENTS OF EARTH'S MAGNETIC FIELD

You might have learnt in your previous classes that, when a magnet is suspended freely, it comes to rest so that one of its ends always points towards the north and other end towards the south. This property of a magnet may be explained by considering the earth as a huge magnet. The earth acts like a huge magnet with its magnetic north pole somewhere near its geographic south pole and the magnetic south pole near its geographic north pole.

You know that there is a magnetic field in the space surrounding a magnet. As the earth behaves like a magnet it has also a magnetic field. The strength and the direction of the earth's magnetic field at a place may be found by measuring certain quantities. They are (i) **declination** (ii) **dip** and (iii) **the horizontal intensity** of earth's field. These quantities are known as the elements of the earth's magnetic field.

Now let us see briefly what these elements are and how they are measured.

A vertical plane drawn through a point on the surface of the earth so that it passes through the magnetic north and south poles is known as the **magnetic meridian** at that point. It is also the vertical plane containing the axis of a freely suspended magnet. Similarly a vertical plane drawn through a point on the surface of the earth and the geographic north and south poles is known as the **geographic meridian** at that point.

The angle between the magnetic and the geographic meridians drawn at that point is called the declination at that point.

If a magnetic needle is pivoted so that it can rotate freely in a vertical plane it will come to rest with its axis inclined to the horizontal i.e. one of its ends will dip down. *This angle of inclination of the axis of a magnet at a place is known as the dip at the place.*

We know that the direction of a magnetic field at a point will be indicated by the direction of the north pole of a magnetic needle placed at that point. It may, therefore, be seen that the direction of the earth's magnetic field at a place is inclined to the horizontal. Hence the earth's magnetic field at a place may be resolved into horizontal and vertical components. *This horizontal component of the earth's magnetic field at a place is called the horizontal intensity of the earth's magnetic field at the place.*

The declination and the horizontal intensity may be accurately determined by an instrument known as *Kew magnetometer*. The dip may be accurately determined by an instrument known as the *dip circle.*

8.1.2. VARIATIONS OF THE ELEMENTS OF THE EARTH'S MAGNETIC FIFLD

The elements of the earth's magnetic field at a place are not the same at all times. They are subjected to variations. The declination is found to vary by a few *minutes* from its average value once a day. Likewise the horizontal intensity also varies but by a very small amount i.e. about a few parts in 10000. These are *daily variations*. However, the daily variations are not constant. The elements are also found to undergo *annual variations i.e* once in a year.

Apart from these variations, there are violent and sudden changes in the elements recorded all over the world. These sudden changes follow a eleven-year cycle and are called *magnetic storms*. The occurence of the magnetic storms coincides with the occurence of the sunspot activities.

Exercise 8.1

1. *What are the elements of the earth's magnetic field ?*
2. *What are magnetic storms ?*

8.2. THE SUNSPOTS AND THE SOLAR FLARES

8.2.1. SUNSPOT ACTIVITIES

If we see the sun through a fine hole in a piece of cardboard, most of the time we observe some dark regions on the face of the sun (Fig. 8-1). These are popularly known as sunspots. These were first discovered by **Galileo Galilei** in 1611 These spots are among the most interesting features of the sun. They appear dark because they are cooler than the surrounding areas. Though the temperature (at the centre of the spots) in general is about

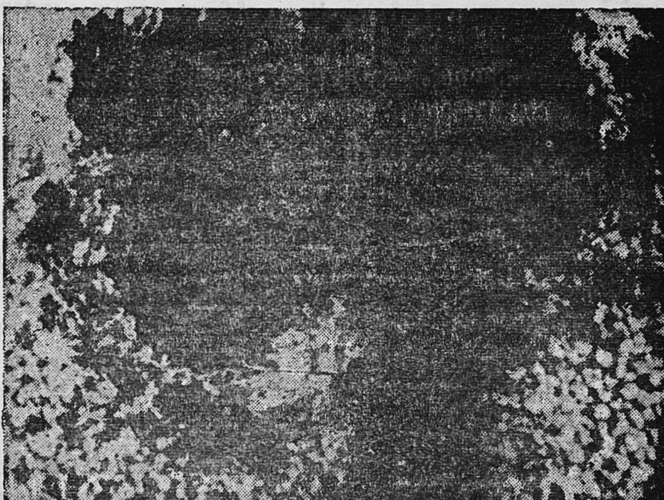


Fig. 8-1
Sunspots

4600 *K*, it may even be as low as 3700 *K*. The surface temperature of the sun is about 6000 *K*. The spots range in size from less than 150 *km*. to 100000 *km*. in diameter. The life of a spot is found to be related to its size. The smaller spots are short-lived lasting only a day or so, while the larger spots last for several days.

The sunspots do not occur at random on the sun's face. They show certain strange and unexplained regularities. Considering their number, it is found to undergo regular changes. The number of spots varies gradually from a minimum (when no spots may be visible for days or even weeks) to a maximum and then to a minimum. This cycle of variations between two minima is found to have an average period of about 11 years. This interval is popularly known as the *eleven year cycle.*

Further studies about the sunspots reveal the following facts:

(i) The spots are associated with large magnetic flux densities such as 0·4 tesla. The flux density at the centre is normal to the solar surface.

(ii) Material flows out from regions in and around the spots along the magnetic lines of force, as was indeed discovered at the **Kodaikanal** observatory in our country by **Evershed.**

8.2.3. SOLAR FLATE

Another important solar phenomenon which is clearly associated with the sunspot activity is the solar flare.

Tremendous local explosions are frequently observed near the sunspots. These explosions give rise to sudden increase in the intensity of light which then decreases slowly, lasting over an hour or so. Such a phenomenon is known as the solar flare. The solar flare is associated with sudden outbursts of matter in the form of charged particles of very high energy.

The solar flare and hence the sunspot activity causes certain changes in the earth's atmosphere disrupting even telegraphic services. The high-energy particles which rush out of the gravitational field of the sun during the maximum sunspot activity may be the cause of increased cosmic ray activities observed a few minutes after the solar flare. They also bring about spectacular displays in the earth's atmosphere at high altitudes in the polar regions. These particles and the charged particles from the sun's *corona* (outermost atmosphere) following them cause violent changes in the earth's magnetic field leading to *magnetic storms*. These magnetic storms affect the ionospheric regions of the earth which play a vital role in the radio and television broadcasts. Last but not the least, the solar flare and the sunpot activity also play a vital role in changing the weather and influencing the growth of plants.

Exercise 8.2

1. *What are sunpots ?*
2. *What is a solar flare ?*
3. *What are the effects of solar flares?*
4. *What is eleven year cycle ?*

8.3. BALLOON FLIGHTS

8.3.1. INTRODUCTION

The balloon plays an important role in the study of the atmosphere. The balloon usually is of spherical shape due to the pressure of the gas inside. In the early stages, air, coal gas and hydrogen were used as lifting media. Nowadays, helium is invariably used as the lifting medium. The balloons of to day are made up of cotton-rubber fabric, plastic and polythylene, whereas paper, oiled silk, goldbeater's skin and rubber were used in the early stages. However, the choice of the material depends upon the range of temperature to which the balloon may be subjected.

8.3.2 TYPES OF BALLOONS

Balloons are basically distinguished as (i) captive or tethered balloons and (ii) free balloons. The captive balloons are secured by a cable to the ground. But the 'free ballons' are not restrained in their ascent into the atmosphere.

The important applications of captive balloons are mainly in defence, navigation and communications. In defence, they are used to detect enemy troop movements, artillery motors or rocket firing sites. In navigation, they serve to locate air-borne and ground vehicles at extended ranges. In communications they have proved useful in spanning long distances as much as 800 *km* and large areas such as 260 000 square kilometres.

The free balloons are now used in the meteorological study of the upper atmosphere. Professor Piccard of Belgium who ascended to about 20 *km* was the pioneer in the field of ballooning.

In 1933, three Russians ascended to a height of about 24 *km*. Unfortunately they were killed during their descent. In spite of many such accidents, free balloons of very large capacities have ascended to a height of 27 *km*. These balloons provide useful information about the space, relating to Van Allen radiation, earth's magnetic field or cosmic rays. Many kinds of balloons are used for meteorological purposes of which the following are a few :

(i) **Ceiling balloons:** These are used to find the height of a cloud-base.

(ii) **Pilot balloons:** They are used in finding the speed and direction of winds in the upper atmosphere.

(iii) **Radiosonde balloons:** These are used to carry radiosonde. These are larger than ceiling or pilot balloons.

(iv) **Hurricane beacons:** These balloons are used to trace the course of the cyclones.

Apart from the above types of balloons used for meteorological purposes, there are two other types of balloons used for scientific research. They are (*i*) *zero-pressure balloons* and (*ii*) *super-pressure balloons*. They are designed to float at constant altitudes.

The zero pressure balloons derive their names from the fact that the internal and external pressures are always in equilibrium. These are now extensively used to carry scientific instruments almost to the top of the atmosphere for astrophysical studies such as cosmic rays. X-rays, gamma rays and other radiations emitted by celestial bodies, and for upper atmospheric studies such as measurements on atmospheric constituents, winds, temperature, pressure, amount of water vapour, ozone and nitric oxide concentration They are also used for biological investigations such as the effect of low pressure on living organisms, from the smallest microbes to monkeys.

Zero-pressure balloons suited to atmospheric conditions above our country have been developed by the *Tata Institute of Fundamental Research* (TIFR). A *National Balloon Facility* (NBF) is now in operation at Hyderabad and it may be noted that it is of international significance.

In the super-pressure balloons, as the name itself indicates, when they reach the desired altitudes, the pressure inside it is in excess over the external pressure with their volumes always remaining the same The advantage of this type of balloons over the zero-pressure balloons is that they can be maintained at the desired altitude for days together at a lower cost. The longest recorded flight of such a balloon is 411 days.

Exercise 8.3

1 *What are the different types of balloons?*

2. *What are the uses of balloon flights?*

8.4. ATMOSPHERIC REGIONS

8.4.1. INTRODUCTION

The various regions of the earth's atmosphere are characterised according to their thermal, compositional and electrical features. The heights of the lower and upper limits of the atmospheric regions are estimated depending upon certain properties The conventional way of distinguishing the atmospheric layers is as follows:

Region	*Lower limit (km)*	*upper limit (km)*
Troposphere	Earth's surface	8 - 20
Stratosphere	8 - 20	16 - 64
Mesosphere	16 - 64	50 - 80
Ionosphere	50 - 80	480 - 660
Exosphere	480 - 660	Outer space

The region extending from 25 *km* to the outer atmosphere is called the *thermosphere*

Of these, the first three regions and the thermosphere are distinguished by their temperature distributions. Hence they are called the thermal regions, The ionosphere is characterised by its electrical nature while the exosphere is characterised by the composition, such as air density and the mean free path of the atmospheric particles.

8.4.2. NATURE OF THE VARIOUS REGIONS

The following section gives a brief account of the nature of the individual regions :

Troposphere : In this region the temperature decreases as the height increases. Starting from a mean temperature of 288 K at the earth's surface the temperature decreases at the rate of 6·5 K *per km* till it reaches a minimum value of 218 K. This is the region where the cloud, mist and rain are formed.

Stratosphere : The atmospheric region above the troposphere is known as the stratosphere. This is more or less a calm region where the temperature remains almost constant at 218 K, But at an average altitude of about 48 *km* the temperature starts increasing at the rate of about 5 K *per km* till it reaches a maximum value of about 263 K at about 57 *km*.

Mesosphere : The region of increasing temperature is sometimes called the lower part of the mesosphere. Thereafter the temperature decreases to a minimum of about 183 K. This region of decreasing temperature is referred to as the upper part of the mesosphere.

Thermosphere : This is above the mesosphere. The temperature of the atmosphere after reaching a minimum value at the top of the mesosphere continues to increase upto a height of about 320 *km*. to 500 *km*. Beyond this altitude the temperature seems to be independent of the altitude for several hundred kilometres. The temperature of the thermosphere may reach even about 1970 K at the time of some solar flares. However, the average temperature ranges from 1023 K to 1523 K.

Ionosphere : This is a region of charged particles high above the earth's surface. It extends indefinitely into space. This plays an important role in the long distance radio communication. Owing to the ultra-violet radiations from the sun and cosmic rays from space, the gases in this region are ionised giving rise to free electrons, positive and negative ions.

The structure of the ionosphere varies continuously during the day, the seasons of the year, solar activity and also according to the latitudes. However, on the basis of the electron density (number of electrons per unit volume) several regions such as *D*, *E*, *F* and *G* regions are distinguished in the structure of the ionosphere.

The lower portion of the ionosphere is called the *D* region. It extends from about 55 *kms*. to 60 *kms*. It is characterised by a low electron density. Hence in the absence of the solar radiation it disappears completely during the night time. It reflects low frequency (20 *kHz* to 100 *kHz*) and hence long wavelength radio waves which are usually used for very long distance communications. The strength of such waves change only very gradually even during long distances.

The next region which extends from about 60 *kms* to 145 *kms* is called the *E* region. Though the electron density of this region also decreases during nights, it never falls to zero and hence this region does not completely disappear during nights. This reflects the medium-length radio waves, in the frequency range of 550 *kHz* to 1500 *kHz*. This was formerly known as *Kennelly-Heaviside layer*. Arthur E. Kennelly of USA and Oliver Heaviside of England were the first to discover independently the existence of the ionised layer of the atmosphere in the year 1902.

The *F* region starts from about 145 *km* to more or less indefinite altitudes. It is also known as the Appleton layer. It is the most useful for long range radio transmission (frequencies above 1·5 *MHz*). It exhibits a further division into F_1 and F_2 regions during the day, which however merge into a single region during nights. F_1 region is the lower of them and extends from about 145 *km*. to about 240 *km*. The F_2 region begins from 240 *km*.

The *G* layer is supposed to exist at an altitude of about 400 *km*. It has been found only in the equatorial region and that too only rarely.

Exosphere : This is the outermost region of the atmosphere. The air density in this region is very low resulting in very long mean free path of the atmospheric particles. Hence the particles may escape from the earth's environment; However, the charged particles are not allowed to escape due to the earth's magnetic field. The temperature of the lower part of the region is about 1750 K and that of the upper part is about 10 000 K.

Exercise 8.4

1. *What are the different regions of the atmosphere ?*
2. *What are the different parts of the ionosphere ?*

8.5. AURORA BOREALIS AND AURORA AUSTRALIS

8.5.1. AURORAS

It has been mentioned in article 8.2.2. that the charged particles thrown out from the sun's surface during the solar flare bring about spectacular displays at high altitudes in the polar regions. Such displays are in general called the auroras. At places very far north or south they can be seen on almost any night. The auroras of the northern region are known as *Aurora Borealis* while those of southern hemisphere are called *Aurora Australis*. They appear in a variety of shapes (Fig. 8-2). An aurora sometimes appears with long rays like search lights. It may appear like a curtain of many colours. Though aurora takes many colours like flaring red, green, yellow or rose, often it takes a luminous greenish white. The appearance of aurora cannot be predicted in advance. Usually they are more likely to appear before midnight than after midnight.

8.5.2. CAUSE OF AURORAS

As already mentioned, the aurora is a result of solar flares. As the charged particles, chiefly the slow-moving protons emitted from the sun during the solar flares enter the earth's atmosphere, they are accelerated by the earth's magnetic field. These fast-moving charged particles collide with the atoms of the atmospheric gas and make them glow giving rise to auroras.

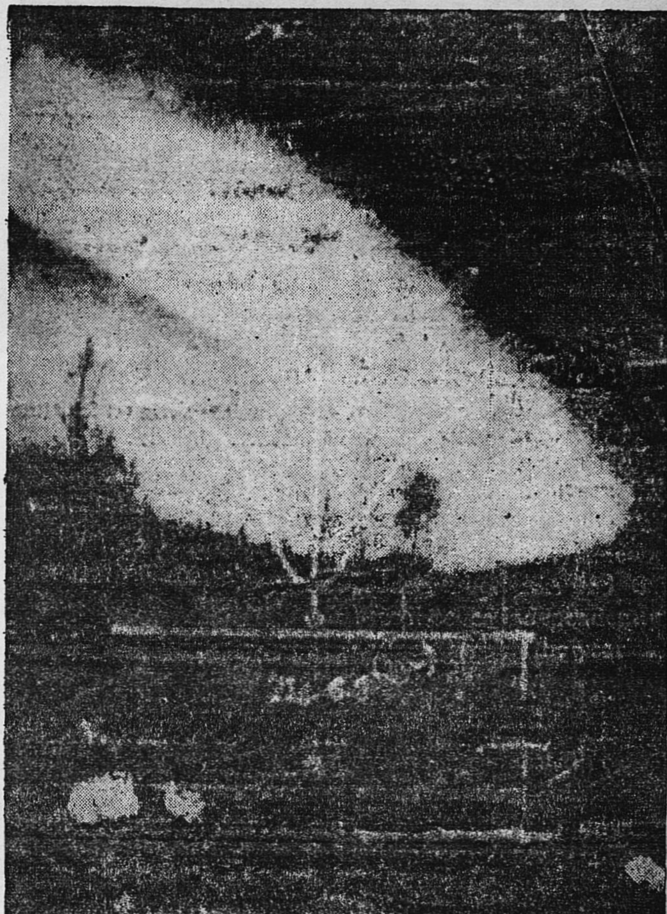


Fig. 8-2 (a)
Auroras (some views)

Fig. 8-2 (b)
Auroras (some more views)

Fig. 8-2 (c)
Auroras (some other views)

Exercise 8.5

1. *What are auroras ? How are they caused ?*

8.6 THE UNIVERSE

8.6.1. INTRODUCTION

The vast space that surrounds us is called the universe. The earth, its satellite, the moon, the planets in our solar system, the sun and the millions of tiny stars are all tiny little specks of the universe. When we look high up, in trying to peep through the earth's atmosphere during clear nights we see millions and millions of stars always twinkling as if smiling at us. Do we have only that much of stars as our unaided eyes are able to show us ? Are those tiny little sources of light really as tiny as we observe them ? Not at all. A powerful telescope will show that there are actually many more stars besides those visible to the unaided eyes. Again the telescope will also show that they are really very large, many times larger than the sun itself. Some of them are actually clusters of millions and millions of individual stars, extending to millions and millions of kilometres in space. Such clusters of stars are called the ***galaxies***. Their apparent smallness is due to their very large distances from us.

8.6.2. SIZE OF THE UNIVERSE

The sizes of the galaxies and their distances as well as the stars are so large that units such as kilometres used in terrestrial measurements cannot be used and new units are to be involved. One such convenient unit is known as the *light year*. One light year is the distance that light with its incredible velocity of 300 million metres per sec. (300×10^6 m/s or 3×10^5 km/s) would travel in one year. It is equivalent to 9460×10^{12}m. Our telescopes can reach into the universe upto a distance of 3,000 million light years, i. e., our telescopes can show us the stars from which light has already started 3000 million years ago. The extent to which we can reach into the space with our instruments of limited capacities is known as the observable universe. The actual universe is really limitless.

The observable universe thus consists of a large number of aggregations of matter in large and small scales. The largest aggregations are called the *super galaxies* which are really groups of independent galaxies spread out to enormous distances such as thousands of light years.

Exercise 8.6

1. *What are galaxies ?*
2. *What is a light year ?*

8.7 THE GALAXIES

8.7.1. INTRODUCTION

As already mentioned, though the galaxies are clusters or groups of stars, they are not so simple in structure. They are really complex organizations consisting of millions and millions of stars, ***nebulae*** i.e., gas and dust particles of the average size between 10^{-7}m. to 10^{-6} m. There are also galaxies observed with no gas and dust.

From a consideration of the appearances, the galaxies are broadly divided into three classes viz. (i) *elliptical galaxies,* (ii) *spiral galaxies* and (iii) *irregular galaxies.*

The elliptical galaxies are in fact spherical in appearance. A spiral galaxy consists of a bright circular or elliptical nucleus from which two or more arms of approximately spiral shape emerge coiling round the nucleus. The third class of galaxies, i.e. the irregular galaxies, as the name itself indicates, does not

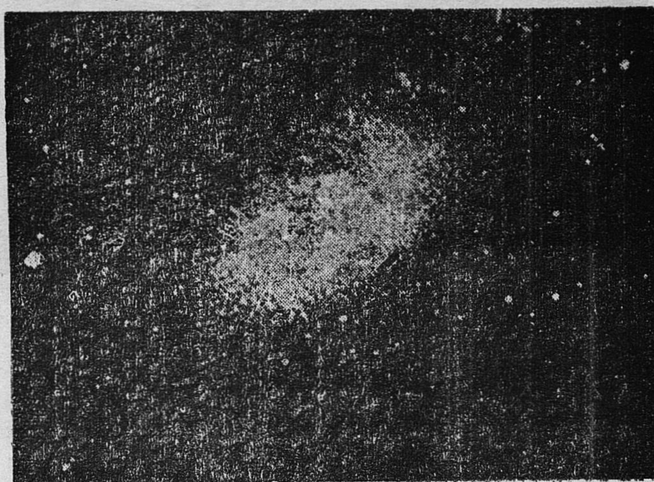


Fig. 8-3 (*a*)
Elliptical galaxy NGC 147 in Andromeda constellation. This galaxy does not contain any gas or dust.

Fig. 8-3 (*b*) (i)
M 101 in Ursa Major constellation-spiral galaxies

possess any regular shape. Fig. 8-3 shows some of the examples of the different classes of galaxies.

Due to the limited power of our instruments it is impossible to say anything about the number of galaxies in the universe.

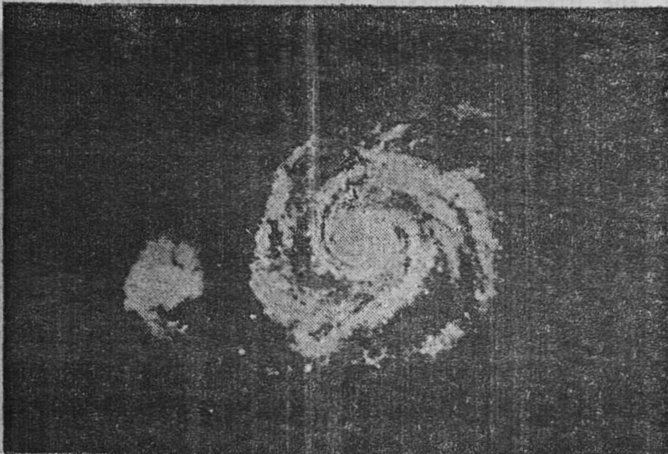


Fig. 8-3 (*b*) (ii)

M 51 in Canes Venatici constellation–spiral galaxies

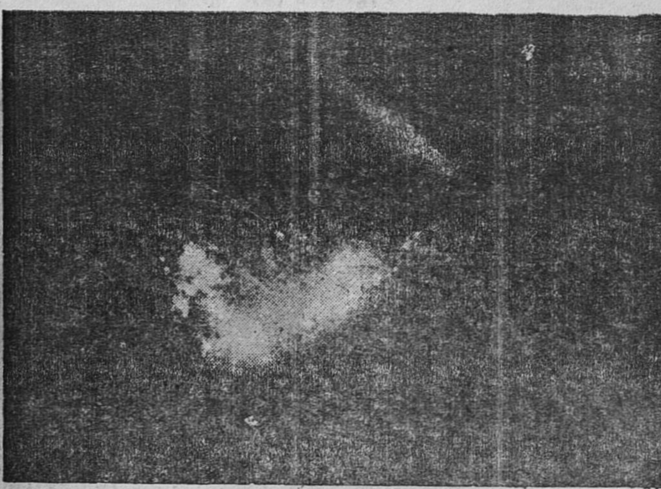


Fig. 8-3 (*c*)

Irregular galaxy-the large Magellanic cloud

However, it is estimated that with the 200 inch telescope at Mt. Palomar in California, it is possible to record about a billion galaxies.

Our sun is actually a star belonging to the galaxy called the *Milky way*. The other galaxies are very far from us. The large Magellanic cloud is one of the nearest galaxies to our own galaxy and is at a distance of about 72000 light years from us. Compare this distance with the distance of 8 light minutes which separates us from our sun !

8.7.2. THE MILKY WAY—THE GALAXY TO WHICH WE BELONG

Looking at the sky on a clear moonless night, we will see a broad band of luminous haze extending from horizon to horizon across the sky. This is popularly known as the *Milky way* on account of its appearance. Since we are situated in this galaxy it is impossible to know its shape by optical telescopes. However observations with the help of radio telescopes have revealed that it belongs to the class of spiral galaxies like M 101 [Fig. 8-3(*b*)(i)] It has been estimated that it has a diameter of about 150000 light years and a maximum thickness of about 20000 light years at the centre. Our sun is in one of the spiral arms at a distance of about 30 000 light years from the centre.

8.7.3. CONSTITUENTS OF GALAXIES

As already mentioned the galaxies in general consist of a number of stars. A galaxy of the size of our galaxy contains about a hundred thousand million (10^{11}) stars. As can be seen from the pictures of the galaxies, these stars are not distributed uniformly. They have a tendency to collect together Such collections are called *clusters*.

And then what about the space in which the stars are found? Do they just exist in vacuum or is the space filled with some form of matter ? Observations show that the stars are situated in an extremely thin gaseous medium. Sometimes the gaseous medium also contains extremely minute particles of solid matter. Such particles are called the '*Interstellar dust*'.

Observations through radio telescopes indicate that the gaseous medium is generally composed of neutral atoms of sodium, calcium, potassium and iron, ionised calcium and titanium, molecules of cynogen and hydrogen. Hydrogen is the most abundant in the interstellar gas. The density of the gas is very low. The average density lies between 10^{-21} to 10^{-19} kg/m^3 or in terms of atoms 10^6 to 10^8 atom/m^3 i.e. 1 to 100 atom/c.c. These densities are about a ten thousand millionth of the densities obtainable in the most perfect vacuum which can be produced in the laboratory. Hence the interstellar space may be taken to be more or less a vacuum. However, taking into consideration the enormous space the interstellar gas occupies, its total mass in a galaxy is comparable to the total mass of stars contained in it. It is estimated that nearly half of the matter of a galaxy is in the form of such gas and dust while the other half is in the form of stars. We have already seen that a galaxy contains about hundred thousand million stars. Hence the total mass of matter contained in a galaxy will be equal to the total mass of two hundred thousand million stars. It is learnt that the mass of the sun is 2×10^{27} tons. We can now imagine what enormous amount of matter a galaxy contains.

8.7.4. EXPANDING UNIVERSE

Are these galaxies at rest ? Observations indicate that all the observed galaxies are in rotation. However, they are not rotating as a whole like the wheel of a cart. The different parts rotate with different speeds. Those parts near the centre move with greater speeds than those farther away from the centre. It is estimated that the stars in our galaxy near the sun make one complete round about the centre of the galaxy in about 225 million years.

An analysis of the spectra of the light emitted by the galaxies also indicates that the galaxies are receding from us. This leads to the concept of what is called an '*expanding universe*'.

Exercise 8.7

1. *State the different types of galaxies with examples.*
2. *Write a note about our galaxy.*
3. *What are the constituents of a galaxy ?*

8.8. THE STARS

8.8.1. TYPES OF STARS

A star may be described as a huge, more or less spherical mass of glowing gas held together by gravity. Some stars are large, and some small, some hot and some comparatively cooler. Some produce a large amount of radiant energy which pours out into space.

There are three important types of stars viz. (i) double and multiple stars, (ii) intrinsically variable stars and (iii) *novae* and *supernovae*.

In a galaxy there are only a few single stars like the sun. Recent estimates have shown that about 80 percent of the stars are either double stars (binaries) or multiple stars. The binary stars are pairs of stars moving round their common centre of gravity in stable equilibrium. An intrinsically variable star shows variation in its apparent brightness. This is due to the fluctuations in the energy radiated by it per second.

It is observed that occasionally some stars suddenly attain extremely large brightness that they may be seen even during day time and then slowly fade away. Such stars are called novae. Supernova is a large nova. So far about 100 novae have been recorded. The most brilliant nova was observed by Tycho Brahe in 1572.

8.8.2. CLASSES OF STARS

We see a star by the light it emits. A close observation of the stars will reveal that the stars appear in different colours. Based on their colours and hence their temperatures, the stars are divided into 10 broad classes as O, B, A, F, G, K, M, R, N, and S. The temperature decreases progressively from O to S. Only very few stars belong to R, N and S classes. The colours and temperatures of the other classes of stars are listed in the tabular coloumn.

The sun belongs to class G and has an effective temperature of 6000 K.

Having seen something about the universe and its contents, one may wonder as to how the universe and the stars are formed. A brief answer to this question may be found in the following sections.

Table 8.1. Colours and Temperatures of Stars

Class	*Colour*	*Effective Temperature in K*
O	Blue	25000
B	Blue white	12000—25000
A	White	8000—11000
F	Straw	6200 – 7800
G	Yellow	4600— 6000
K	Orange	3500— 4600
M	Red	2600— 3400

Exercise 8.8

1. *State the classes of stars. Give their temperatures and colours.*

8.9. THE EVOLUTION OF THE UNIVERSE

8.9.1. INTRODUCTION

Considering the origin of the universe there are at present two theories viz. (i) the *Evolutionary theory* or the *Big Bang theory* and (ii) the *Steady State theory.*

8.9.2 THE EVOLUTIONARY THEORY

The evolutionary theory is based on Einstein's General theory of Relativity and was developed by Abbe Georges Lemaitre. According to this theory, the universe was evolved from a single atom which contained the entire matter of the present universe. Evidently the matter must have been in a highly crushed and unstable state. Consequently the atom should have exploded with great violence resulting in a uniform distribution of the matter in space in the form of hydrogen gas. As time passed, this gas might have settled down and condensations might have started due to gravitational forces forming the different galaxies.

8.9.3. THE STEADY STATE THEORY

The steady state theory was proposed by the British scientists Fred Hoyle. H.Bondi, T.Gold and W. H. McCrea. This is based on the assumption that the universe presents the same appearance at every instant. However, due to the receding away of the galaxies, the universe must be thinning out In order to overcome this difficulty it is proposed that new matter is being continually created at the rate of one hydrogen atom per $4 \cdot 5 \times 10^7$c.c in 100 years.

Exercise 8. 9

1. *Explain the (i) Big Bang theory and (ii) Steady state theory of the origin of the universe*

8.10 THE BIRTH AND DEATH OF A STAR

8.10.1. THE BIRTH OF A STAR

It has already been mentioned that the seemingly empty space of the galaxies are really filled with gas and dust particles. This gas and the dust particles form the raw material in the evolution of a star. Under suitable conditions a gas cloud may become unstable and begin to contract due to the gravitational forces between them. Such a contracting cloud, called a *protostar* releases gravitational energy. Part of the gravitational energy is radiated while the rest goes to increase the temperature and hence the pressure inside the cloud. The contraction continues till a delicate balance is reached between the increasing pressure trying to expand the cloud and the gravitational force contracting it. At this stage the temperature is sufficiently high (about 10 million degrees) to start a thermo-nuclear reaction (nuclear fusion) which then takes up the work of creating energy for the star to radiate. During the thermo-nuclear reaction, the more abundant constituent of the star viz. hydrogen is burnt producing helium. Our sun is considered to be in this stage.

8.10.2 THE DEATH OF A STAR

When all the hydrogen is used up, the internal pressure cannot be maintained and another gravitational contraction starts again till the temperature rises to about 100 million degrees. At this stage helium fusion starts, producing carbon-oxygen ash. When all the helium is used up, another gravitational contraction sets in taking the star to a highly crushed state. After this stage depending upon its initial mass, the star may continue to radiate its stored-up energy till it cools down sufficiently and disappears from sight or may explode resulting in a nova. According to the world-famous astrophysicist **Prof. S. Chandrasekar**, this explosion will take place when the initial mass is greater than 1·64 times the mass of the sun.

Exercise 8.10

1. *Explain the birth and death of a star.*

8.11. THE SOLAR SYSTEM

8.11.1. ITS ORIGIN

So far we have been travelling millions and millions of light years through space in an attempt to learn about the distant stars and galaxies. Now let us turn our attention towards our sun which is also a star very near to the earth at a distance of about 8 light minutes

Our sun is not alone but possesses a family of comparatively small celestial bodies called planets, satellities, asteroids, comets and meteors. The family is called the **Solar system.** In our solar system, there are nine major planets, thirty-one satellites belonging to six planets and more than 1500 minor planets called asteroids The planets revolve round the sun in elleptical orbits at the same time rotating about their own axes. They differ very much in their sizes, rotation periods, temperatures etc. The names of the different planets together with their distances from the sun, sizes etc., can be seen from Table 8-2.

8.11.2. McCREA'S THEORY

Let us now see how the solar system was formed. Any successful theory about the origin of the solar system must account for the following observations regarding the solar systems:

(i) All the planets and satellites with the exception of outer satellites of Jupiter and Saturn have a common direction of orbital motion and their orbital planes are practically coincident.

Table 8.2 Solar system

Planet (1)	*Mean distance from the sun 10^6 km* (2)	*Mean radius km* (3)	*Mass relative to earth** (4)	*Mean density $kg\ m^3$* (5)	*Period of rotation round the sun* (6)	*Axial Rotation period* (7)	*Temperature K* (8)	*Number of Satellites* (9)
Mercury	57·91	2420	0·054	5300	87·97 days	59 days	690	0
Venus	108·21	6100	0·815	5000	224·70 days	243 days	290	0
Earth	149·60	6378	1·000	5520	365·256 days	1 day	287	1
Mars	227·94	3380	0·108	3900	687·0 days	24·6 hrs.	235	2
Jupiter	778·30	71350	317·8	1330	11·86 yrs.	9·9 hrs	135	12
Saturn	1427·00	60400	95·2	710	29·46 yrs.	10·2 hrs.	120	9
Uranus	2869·00	23800	14·5	1600	84·02 yrs.	10·7 hrs.	Below 90	5
Neptune	4498·00	22200	17·2	2200	164·8 yrs	15·8 hrs.	,,	2
Pluto	5900·00	3000	0·8	5000	248·4 yrs.	6·3 days	,,	0
Sun		6,96,000	3,29·390	1,410		25·38 days	6000	

* Mass of earth $6 \cdot 6 \times 10^{21}$ tons

(ii) The inner planets *viz.* Mercury, Venus, Earth and Mars have small masses, high specific densities, low rotational velocities and a few satellites while the outer planets viz. Jupiter, Saturn, Uranus and Neptune have large masses, low specific densities, relatively higher rotational velocities and a larger number of satellites. Pluto is an exception.

(iii) The sun, although it has more than 98 percent of the total mass of the solar system possesses only 2 per cent of the total angular momentum of the solar system.

Though a number of theories have been proposed so far none of them is able to explain fully all the observed facts about the solar system. However, the theory put forward by McCrea in 1960 is worth mentioning here.

McCrea's theory is based on two well established results about star formation. They are: (i) stars are formed in clusters of several hundred members and (ii) the interstellar matter which forms the raw material is in a state of random motion with velocities larger than the velocity of sound.

McCrea starts with the situation when the interstellar matter is in the form of *cloudlets* moving in random directions. These cloudlets grow bigger and bigger just as the drops of mercury collide with each other to form larger drops. When a sufficiently large aggregate is formed, it will contract under gravitation attracting other *cloudlets* moving near it. Such a condensation is taken to be the main condensation forming the stars.

A stellar condensation described above captures only the matter that falls on it at any stage. A cloudlet that has too much of angular momentum about the condensation will not be captured by it. Such a cloudlet will be captured by another condensation about which its angular momentum is small. Also some of the angular momentum goes into the relative motions of the condensations.

Our sun being a star is the result of main condensation. Now the sun captures some of the cloudlets which are in its vicinity and are revolving in the same plane and in the same way in the gravitational field of the sun. However, the cloudlets on their way are subjected to secondary condensations. When such condensations reach sufficiently large size, they contract due to gravitational force forming the planets. However, this theory does not deal with the formation of satellites.

Exercise 8. 11

1. *Explian the McCreas theory about the origin of the solar system.*
2. *Name the planets of the solar system.*

8.12. RADIO ASTRONOMY

8.12.1. INTRODUCTION

Radio astronomy is that branch of science wherein the universe is explored with the help of radio waves emitted by the celestial bodies. Until about three decades ago the only means of exploring the universe was the optical telescope using radiations in the wavelength range from 390 nm to 750 nm (nm = 10^{-9}m). This range includes ultraviolet, visible, and infra red radiations. Radio astronomy provides an additional range of wavelengths from a few centimetres to about 10 metres.

The birth of radio astronomy is quite an accidental one like many other scientific discoveries. The first discovery about the radiowaves arriving at the earth from outer space was accidentally made by Karl. G. Jansky in 1932 while he was measuring the general noise level for the construction of short wave receivers for long distance communication. He discovered that the noise of the greatest intensity always came from a fixed direction which was later found to point towards the centre of milky way galaxy. This was followed by another accidental discovery by British military authorities during World War II about the radio waves from the sun.

8.12.2. RADIO TELESCOPE

The instrument used for the reception of the radio waves emitted by the celestial bodies is called the radio telescope. A radio telescope is a very powerfultool of observation. However the optical telescope is still made use of to recognise the object which emits radio waves. During the last two decades, some very large and spectacular radio telescopes have been built and are engaged in the task of exploring the universe. There is one very large radio telescope at Muthorai, a village near Ootacamnnd in Tamil Nadu built by TIFR. It was entirely designed and constructed by the Indian scientists using mostly indigenous material.

A radio telescope is basically the same as a domestic radio receiver but with the loud speaker replaced by an automatic recorder. Thus a radio telescope esesentially consists of three parts viz. (i) *an antenna* or aerial to collect the radio waves, (ii) *a radio receiver* to amplify and suitably detect or rectify the received signal and (iii) *a recorder* to record the detected signals. We shall first briefly explain the receiver and then the antenna.

As already mentioned, the receiver of a radio telescope performs the functions similar to that of a domestic radio receiver. However, these receivers must be extraordinarily sensitive as the radio signals received are very weak. The antenna receives not only the radio signal from the object under observation but from other objects in the form of noise. The frequencies of the signals are spread over a broad band. The receiver selects the required signal and amplifies it employing a superhetrodyne system. The amplified signal which is of oscillatory nature is then rectified to produce unidirectional current. This is then passed onto a pen recorder which records it in the form of curves on a uniformly moving strip of paper or to magnetic tapes which record the varying current in the form of varying current in the form of varying magnetization. The tapes are then fed onto computers for further analysis.

Considering the antenna, its function is to collect the radio waves that fall on it. The antenna in general consists of a parabolic reflector which collects and focusses the radio waves onto what are known as dipoles. The dipoles are nothing but two metallic rods placed along a horizontal line with a gap between them. The radio waves focussed on to the dipoles induce weak currents of oscillatory nature and of the same frequencies as the waves. These weak currents are given an initial amplification and then passed onto the amplifier by coaxial cables.

As the radio waves are very weak in intensity, the antenna should be very large and sensitive. High resolving power can also be obtained by increasing the size of the antenna. This is similar to the fact that, in the case of optical telescopes the light gathering power as well as the resolving power are increased by increasing the aperture (diameter) of the objective. However, there are other factors which set an upper limit to the size of the antenna. For example, the telescope at Cambridge, England has got an antenna of about 1000 m long; the one at Jodrell Bank observatory, Manchester, England, has a parabolic reflector of about 76 m diameter. The antenna near Cotacamund, which is in the form of a cylindrieal paraboloid, has got a length of 529 m and a width of 30 m.

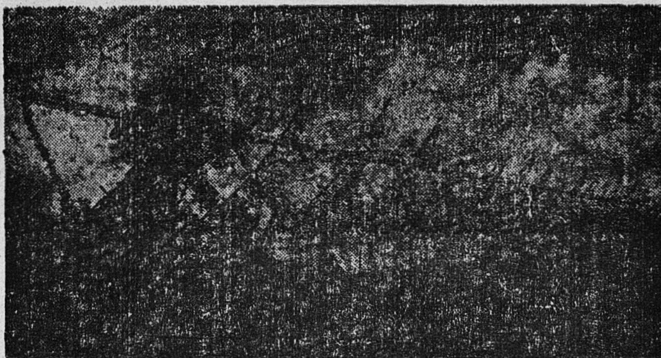


Fig. 8-4

A view of the antenna at Mathorai, Ootacamund

Courtesy : TIFR, Bombay

8.12.3. THE ACHIEVEMENTS OF RADIO ASTRONOMY

Though radio astronomy is almost a new-born child in the family of science, it has enabled the astronomers to learn a lot more about the universe than with the optical astronomy. New celestial objects like *quasars* and *pulsars* are some of its very interesting discoveries. The pulsars are celestial objects which give out radio waves in the form of short pulses at regular intervals. They are now identified to be neutron stars which were predicted as early as 1930. Measurements on the radio waves from the planets give information about the thermal and electrical properties of the surfaces of the planets. Since radio waves unlike light waves can penetrate even clouds, radio astronomy has enabled astronomers to examine the physical properties of planets like Venus which are permanently surrounded by clouds. It is only with the help of radio astronomy that we are able to know about the shape of our galaxy.

Exercise 8.12

1. *What is radio astronomy ?*
2. *Explain briefly the working of a radio telescope.*
3. *What are the achievements of radio astronomy ?*

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