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Notes

ELECTRIC CHARGE AND ELECTRIC FIELD

So far you have learnt about mechanical, thermal and optical systems and various phenomena exhibited by them. The importance of electricity in our daily life is too evident. The physical comforts we enjoy and the various devices used in daily life depend on the availability of electrical energy. An electrical power failure demonstrates directly our dependence on electric and magnetic phenomena; the lights go off, the fans, coolers and air-conditioners in summer and heaters and gysers in winter stop working. Similarly, radio, TV, computers, microwaves can not be operated. Water pumps stop running and fields cannot be irrigated. Even train services are affected by power failure. Machines in industrial units can not be operated. In short, life almost comes to a stand still, sometimes even evoking public anger. It is, therefore, extremely important to study electric and magnetic phenomena.

In this lesson, you will learn about two kinds of electric charges, their behaviour in different circumstances, the forces that act between them, the behaviour of the surrounding space etc. Broadly speaking, we wish to study that branch of physics which deals with electrical charges at rest. This branch is called **electrostatics**.



OBJECTIVES

After studying this lesson, you should be able to :

- *state the basic properties of electric charges;*
- *explain the concepts of quantisation and conservation of charge;*
- *explain Coulomb's law of force between electric charges;*
- *define electric field due to a charge at rest and draw electric lines of force;*
- *define electric dipole, dipole moment and the electric field due to a dipole;*

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- state Gauss' theorem and derive expressions for the electric field due to a point charge, a long charged wire, a uniformly charged spherical shell and a plane sheet of charge; and
- describe how a van de Graaff generator functions.

15.1 FRICTIONAL ELECTRICITY

The ancient Greeks observed electric and magnetic phenomena as early as 600 B.C. They found that a piece of amber, when rubbed, becomes electrified and attracts small pieces of feathers. The word **electric** comes from Greek word for amber meaning **electron**.

You can perform simple activities to demonstrate the existence of charges and forces between them. If you run a comb through your dry hair, you will note that the comb begins to attract small pieces of paper. Do you know how does it happen? Let us perform two simple experiments to understand the reason.



ACTIVITY 15.1

Take a hard rubber rod and rub it with fur or wool. Next you take a glass rod and rub it with silk. Suspend them (rubber rod and a glass rod) separately with the help of non-metallic threads, as shown in Fig. 15.1.

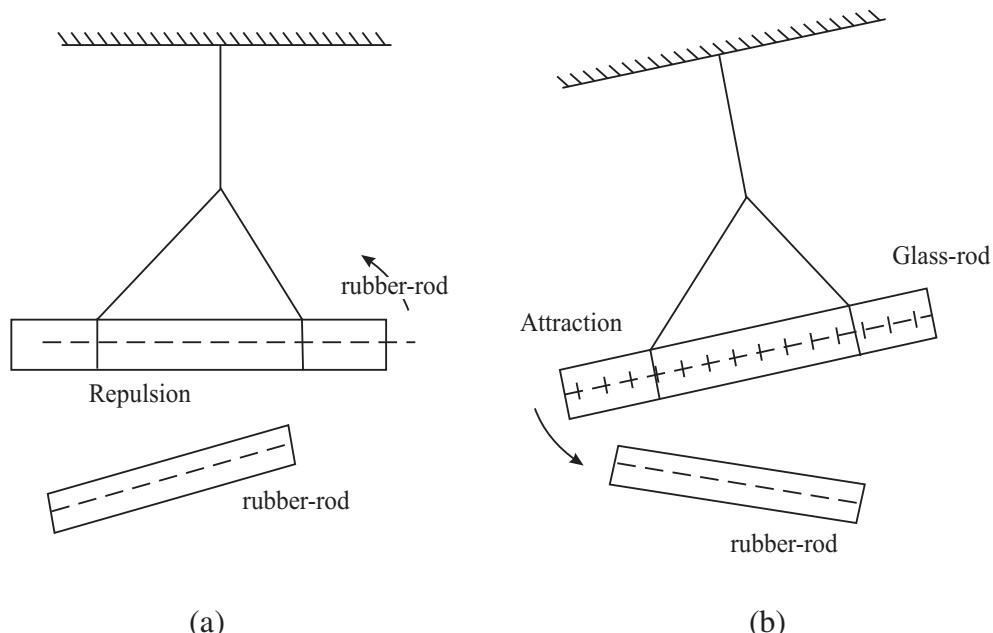


Fig. 15.1: Force of attraction/repulsion between charges: a) a charged rubber rod repels another charged rubber rod : like charges repel each other; and b) a charged glass rod attracts a charged rubber rod : unlike charges attract each other.



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Now bring rubber rod rubbed with wool near these rods one by one. What do you observe? You will observe that

- when a charged rubber rod is brought near the charged (suspended) rubber rod, they show repulsion [Fig. 15.1(a)]; and
- when the charged rubber rod is brought near the (suspended) charged glass rod, they show attraction [Fig 15.1(b)].

Similar results will be obtained by bringing a charged glass rod.

On the basis of these observations, we can say that

- A charged rubber rod attracts a charged glass rod but repels a charged rubber rod.
- A charged glass rod repels a charged glass rod but attracts a charged rubber rod.

From these activities we can infer that the rubber rod has acquired one kind of electricity and the glass rod has acquired another kind of electricity. Moreover, **like charges repel and unlike charges attract each other.**

Franklin (Benjamin Franklin, 1706 -1790) suggested that the charge on glass rod is to be called **positive** and that on the rubber rod is to be called **negative**. We follow this convention since then.

Once a body is charged by friction, it can be used to charge other conducting bodies by

conduction, i.e., by touching the charged body with an uncharged body; and

induction, i.e., by bringing the charged body close to an uncharged conductor and earthing it. Subsequently, the charged body and the earthing are removed simultaneously.

15.1.1 Conservation of Charge

In Activity 15.1, you have seen that when a glass rod is rubbed with silk, the rod acquires positive charge and silk acquires negative charge. Since both materials in the normal state are neutral (no charge), the positive charge on the glass rod should be equal in magnitude to the negative charge on silk. This means that the total charge of the system (glass + silk) is conserved. It is neither created nor destroyed. It is only transferred from one body of the system to the other. The transfer of charges takes place due to increase in the thermal energy of the system when the glass rod is rubbed; the less tightly bound electrons from the glass rod are transferred to silk. The glass rod (deficient in electrons) becomes positively charged and silk, which now has excess electrons, becomes negatively charged. When rubber is rubbed with fur, electrons from the fur are transferred to rubber.

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That is, rubber gains negative charge and fur gains an equal amount of positive charge. Any other kind of charge (other than positive and negative) has not been found till today.

15.1.2 Quantisation of Charge

In 1909, Millikan (Robert Millikan, 1861-1953) experimentally proved that charge always occurs as some integral multiple of a fundamental unit of charge, which is taken as the charge on an electron. This means that if Q is the charge on an object, it can be written as $Q = Ne$, where N is an integer and e is charge on an electron. Then we say that charge is quantised. It means that a charged body cannot have $2.5e$ or $6.4e$ amount of charge. In units 24-26, you will learn that an electron has charge $-e$ and a proton has charge $+e$. Neutron has no charge. Every atom has equal number of electrons and protons and that is why it is neutral. From this discussion, we can draw the following conclusions :

- There are only two kinds of charges in nature; positive and negative.
- Charge is conserved.
- Charge is quantised.



INTEXT QUESTIONS 15.1

1. A glass rod when rubbed with silk cloth acquires a charge $q = +3.2 \times 10^{-17} \text{ C}$.
 - i) Is silk cloth also charged?
 - ii) What is the nature and magnitude of the charge on silk cloth?
2. There are two identical metallic spheres A and B . A is given a charge $+Q$. Both spheres are then brought in contact and then separated.
 - (i) Will there be any charge on B ?
 - (ii) What will the magnitude of charge on B , if it gets charged when in contact with A .
3. A charged object has $q = 4.8 \times 10^{-16} \text{ C}$. How many units of fundamental charge are there on the object? (Take $e = 1.6 \times 10^{-19} \text{ C}$).

15.2 COULOMB'S LAW

You have learnt that two stationary charges either attract or repel each other. The force of attraction or repulsion between them depends on their nature. Coulomb studied the nature of this force and in 1785 established a fundamental law governing



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it. From experimental observations, he showed that the electrical force between two static point charges q_1 and q_2 placed some distance apart is

- directly proportional to their product ;
- inversely proportional to the square of the distance r between them;
- directed along the line joining the two charged particles ; and
- repulsive for same kind of charges and attractive for opposite charges.

The magnitude of force F can then be expressed as

$$F = k \frac{q_1 \times q_2}{r^2} \quad (15.1)$$

For free space, we write $F = \frac{1}{4\pi\epsilon_0} \times \frac{q_1 \times q_2}{r^2}$ (15.2)

where constant of proportionality $k = \frac{1}{4\pi\epsilon_0}$ for free space (vacuum) and $k = \frac{1}{4\pi\epsilon}$

for a material medium. ϵ_0 is called **permittivity** of free space and ϵ is the permittivity of the medium. It means that if the same system of charges is kept in a material medium, the magnitude of Coulomb force will be different from that in free space.

The constant k has a value which depends on the units of the quantities involved. The unit of charge in SI system is coulomb (C). The coulomb is defined in terms of the unit of current, called **ampere**. (You will learn about it later.) In SI system of units, the value of k is

$$k = \frac{1}{4\pi\epsilon_0} = 9 \times 10^9 \frac{\text{N m}^2}{\text{C}^2} \quad (15.3)$$

since $\epsilon_0 = 8.85 \times 10^{-12} \text{ C}^2 \text{ N}^{-1} \text{ m}^{-2}$.

Thus in terms of force, one coulomb charge can be defined as : *If two equal charges separated by one metre experience a force of $9 \times 10^9 \text{ N}$, each charge has a magnitude of one coulomb.* The value of electronic charge e is $1.60 \times 10^{-19} \text{ C}$.

Note that

- Coulomb's law is also an inverse square law just like Newton's law of Gravitation, which you studied in lesson 6.
- Coulomb's law holds good for point charges only.
- Coulomb's force acts at a distance, unlike mechanical force.



How Big is One Coulomb?

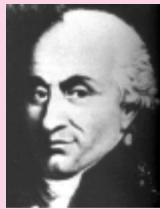
The unit of electrical charge is coulomb. Have you ever thought : How big a coulomb is? To know this, let us calculate the magnitude of force between two charges, each of one coulomb, placed at a distance of one metre from one another:

$$\begin{aligned} |\mathbf{F}| &= k \times \frac{q_1 \times q_2}{r^2} \\ &= 9.0 \times 10^9 \times \frac{1 \times 1}{1^2} \\ &= 9.0 \times 10^9 \approx 10^{10} \text{ N} \end{aligned}$$

If the mass of a loaded passenger bus is 5000 kg, its weight $mg = (5000 \times 10) \text{ N}$ (assume $g \approx 10 \text{ m s}^{-2}$) = $5 \times 10^4 \text{ N}$.

Let us assume that there are 10,000 such loaded buses in Delhi. The total weight of all these buses will be $5 \times 10^4 \times 10,000 = 5 \times 10^8 \text{ N}$. If there are 10 cities having same number of buses as those in Delhi, the total weight of all these loaded buses will be $5 \times 10^9 \text{ N}$. It means that the force between two charges, each of 1C and separated by one metre is equivalent to the weight of about two hundred thousand buses, each of mass 5000 kg.

Charles Augustin de Coulomb (1736–1806)



A French physicist, Coulomb started his career as military engineer in West Indies. He invented a torsional balance and used it to perform experiments to determine the nature of interaction forces between charges and magnets. He presented the results of these experiments in the form of Coulomb's law of electrostatics and Coulomb's law of magnetostatics. The SI unit of charge has been named in his honour.

You now know that the ratio of forces between two point charges q_1 and q_2 separated by a distance r ; when kept in free space (vacuum) and material medium, is equal to ϵ/ϵ_0 :

$$\frac{F_0 \text{ (in vacuum)}}{F \text{ (in medium)}} = \frac{\epsilon}{\epsilon_0} = \epsilon_r$$

where ϵ_r is known as relative permittivity or **dielectric constant**. Its value is always greater than one. We will define dielectric constant in another form later.



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15.2.1 Vector Form of Coulomb's Law

You know that force is a vector quantity. It means that force between two charges should also be represented as a vector. That is, Eqn. (15.1) should be expressed in vector form. Let us learn to do so now.

Let there be two point charges q_1 and q_2 separated by a distance r (Fig. 15.3). Suppose that \mathbf{F}_{12} denotes the force experienced by q_1 due to the charge q_2 and \mathbf{F}_{21} denotes the force on q_2 due to charge q_1 . We denote the unit vector pointing from q_1 to q_2 by $\hat{\mathbf{r}}_{12}$. Then from Fig. 15.3 (a), it follows that

$$\mathbf{F}_{12} = k \frac{q_1 q_2}{|r_{12}^2|} \hat{\mathbf{r}}_{12} \quad (15.4)$$

Similarly, for charges shown in Fig. 15.3 (b), we can write

$$\mathbf{F}_{21} = -k \frac{q_1 q_2}{|r_{12}^2|} \hat{\mathbf{r}}_{12} \quad (15.5)$$

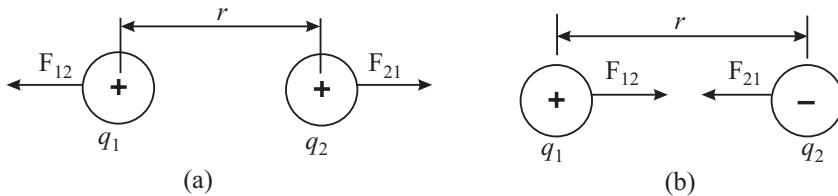


Fig. 15.3 : Two point charges q_1 and q_2 separated by a distance r : a) the direction of forces of repulsion between two positive charges, and b) the direction of forces of attraction between a positive and a negative charge.

The positive sign in Eqn. (15.4) indicates that the force is repulsive and the negative sign in Eqn. (15.5) indicates that the force is attractive.

The Coulomb's law obeys the principle of action and reaction between two charges q_1 and q_2 . Therefore,

$$\mathbf{F}_{12} = -\mathbf{F}_{21} \quad (15.6)$$

In general, we can write the expression for force between two charges as

$$\mathbf{F}_{12} = k \times \frac{q_1 q_2}{r^2} \hat{\mathbf{r}}_{12} \quad (15.7)$$

15.2.2 Principle of Superposition

If there are more than two charges, we can calculate the force between any two charges using Eqn. (15.7). Suppose now that there are several charges q_1, q_2, q_3, q_4 , etc. The force exerted on q_1 due to all other charges is given by Eqn. (15.7):

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$$\begin{aligned}\mathbf{F}_{12} &= k \frac{q_1 q_2}{|r_{12}|^2} \hat{\mathbf{r}}_{12} \\ \mathbf{F}_{13} &= k \frac{q_1 q_3}{|r_{13}|^2} \hat{\mathbf{r}}_{13} \\ \text{and } \mathbf{F}_{14} &= k \frac{q_1 q_4}{|r_{14}|^2} \hat{\mathbf{r}}_{14} \quad (15.8)\end{aligned}$$

The resultant of all these forces, i.e., the total force \mathbf{F} experienced by q_1 is their vector sum:

$$\mathbf{F} = \mathbf{F}_{12} + \mathbf{F}_{13} + \mathbf{F}_{14} + \dots \quad (15.9)$$

This is known as principle of superposition.

Example 15.1 : A charge $+q_1 = 12\text{C}$ is placed at a distance of 4.0 m from another charge $+q_2 = 6\text{C}$, as shown in the Fig. 15.5. Where should a negative charge q_3 be placed on the line joining q_1 and q_2 so that the charge q_3 does not experience any force?

Solution : Let q_3 be placed between q_1 and q_2 at a distance of x metre from q_1 . (It can be easily seen that on placing q_3 on the left of q_1 or on the right of q_2 or at any position other than the one between the line joining q_1 and q_2 , the resultant force can not be zero.) The force exerted on q_3 by q_1 will be

$$\mathbf{F}_{31} = k \frac{q_1 q_3}{r_{31}^2} \hat{\mathbf{r}}_{31} \text{ towards } q_1$$

$$\therefore |\mathbf{F}_{31}| = k \frac{q_3 q_1}{x^2}$$

The magnitude of force on q_3 due to q_2 is given by

$$|\mathbf{F}_{32}| = k \frac{q_3 q_2}{(4-x)^2} \text{ towards } q_2$$

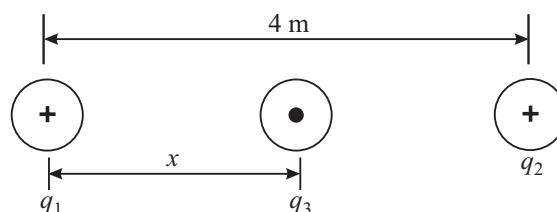


Fig. 15.5 : Three point charges q_1 , q_2 and q_3 placed in a straight line

The resultant force on q_3 will be zero when $\mathbf{F}_{31} = \mathbf{F}_{32}$. Therefore, on substituting the numerical values, we get



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Note that $6q_3k$ is common on both sides and cancels out. Therefore, on simplification, we get

$$\frac{2}{x^2} = \frac{1}{(4-x)^2}$$

or $2(4-x)^2 = x^2$

$\Rightarrow x^2 - 16x + 32 = 0$

On solving this, we get two values of x : 2.35 m and 13.65 m. The latter value is inadmissible because it goes beyond q_2 . Therefore, the charge q_3 should be placed at a distance of 2.35 m from q_1 .

It is a reasonable solution qualitatively also. The charge q_1 is stronger than q_2 . Hence the distance between q_1 and q_3 should be greater than that between q_2 and q_3 .

The roots of a quadratic equation of the form

$$ax^2 + bx + c = 0$$

are given by

$$x = \frac{-b \pm \sqrt{b^2 - 4ac}}{2a}$$

In this case, $a = 1$, $b = -16$ and $c = 32$.

$$\therefore x = \frac{16 \pm \sqrt{256 - 4 \times 32}}{2} \\ = 2.35, 13.65$$

Example 15.2 : Two charges, each of 6.0×10^{-10} C, are separated by a distance of 2.0 m. Calculate the magnitude of Coulomb force between them.

Solution : We know that the magnitude of Coulomb force between two charges is given by Eqn. (15.2) :

$$F = k \frac{q_1 \cdot q_2}{r^2}$$

Given, $q_1 = q_2 = 6.0 \times 10^{-10}$ C and $r = 2.0$ m, Therefore on putting these values, we get

$$F = \frac{(9 \times 10^9 \text{ N m}^2 \text{ C}^{-2}) \times (6.0 \times 10^{-10} \text{ C})^2}{2^2 \text{ m}^2} \\ = \frac{9 \times 10^9 \times 36.0 \times 10^{-20}}{4} \text{ N} \\ = 81 \times 10^{-11} \text{ N}$$



INTEXT QUESTIONS 15.2

- Two charges $q_1 = 16\mu\text{C}$ and $q_2 = 9 \mu\text{C}$ are separated by a distance 12m. Determine the magnitude of the force experienced by q_1 due to q_2 and also the direction of this force. What is the direction of the force experienced by q_2 due to q_1 ?

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2. There are three point charges of equal magnitude q placed at the three corners of a right angle triangle, as shown in Fig. 15.2. $AB = AC$. What is the magnitude and direction of the force exerted on $-q$?

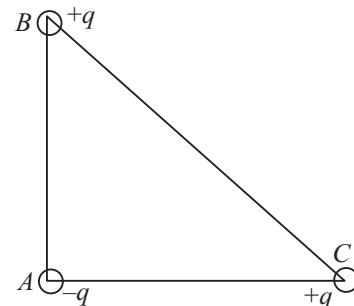


Fig. 15.2 : Three charges placed at the three corners of a right angle triangle.

15.3 ELECTRIC FIELD

To explain the interaction between two charges placed at a distance, Faraday introduced the concept of electric field. The electric field \mathbf{E} at a point is defined as the electric force \mathbf{F} experienced by a positive test charge q_0 placed at that point divided by the magnitude of the test charge. Mathematically, we write

$$\mathbf{E} = \frac{\mathbf{F}}{q_0} \quad (15.10)$$

This is analogous to the definition of acceleration due to gravity, $\mathbf{g} = \mathbf{F}/m_0$, experienced by mass m_0 in the gravitational field \mathbf{F} .

The electric field \mathbf{E} is a vector quantity and has the same direction as the electric force \mathbf{F} . Note that the electric field is due to an external charge and not due to the test charge. The test charge q_0 should, therefore, be so small in magnitude that it does not disturb the field due to external charge. (In practice, however, even the smallest test charge will disturb the external field.) Strictly speaking, mathematical definition given below is more accurate :

$$\mathbf{E} = \lim_{q_0 \rightarrow 0} \frac{\mathbf{F}}{q_0} \quad (15.11)$$

In SI system, the force is in newton and the charge is in coulomb. Therefore, according to Eqn.(15.10), the electric field has the unit newton per coulomb. The direction of \mathbf{E} is same as that of \mathbf{F} . Note that *the action of electric force is mediated through electric field*.

Let us now examine why the test charge q_0 should be infinitesimally small.

Refer to Fig. 15.6. It shows a uniformly charged metallic sphere with charge q and a test charge $q_0 (< < q)$. It means that charge density per unit area is same around points A, B, C and D. The test charge q_0 must measure the force \mathbf{F} without disturbing the charge distribution on the sphere. Fig. 15.6 (b) shows the situation when $q \approx q_0$. In this case, the presence of the test charge modifies the surface

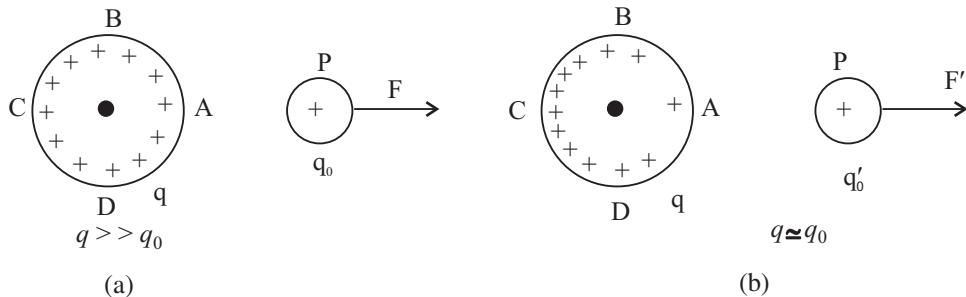


Fig. 15.6 : a) uniformly charged metallic sphere and a test charge, and b) redistribution of charge on the sphere when another charge is brought near it.

charge density. As a result, the electrical force experienced by the test charge q_0 will also change, say from \mathbf{F} to \mathbf{F}' . That is, the force in the presence of test charge is different from that in its absence. But without q_0 , the force cannot be measured. If q_0 is infinitesimally small in comparison to q , the charge distribution on the sphere will be minimally affected and the results of measurement will have a value very close to the true value. That is, \mathbf{F}' will be very nearly equal to \mathbf{F} . We hope you now appreciate the point as to why the test charge should be infinitesimally small.

Let there be a point charge q . A test charge q_0 is placed at a distance r from q . The force experienced by the test charge is given by

$$\mathbf{F} = k \frac{qq_0}{r^2} \hat{\mathbf{r}} \quad (15.12)$$

The electric field is defined as the force per unit charge. Hence

$$\mathbf{E} = k \times \frac{q}{r^2} \hat{\mathbf{r}} \quad (15.13)$$

If q is positive, the field \mathbf{E} will be directed away from it. If q is negative, the field \mathbf{E} will be directed towards it. This is shown in Fig. 15.7.

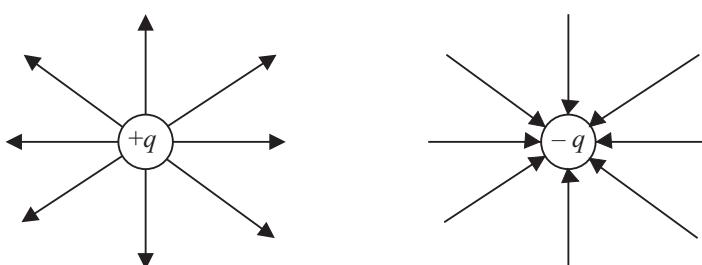


Fig. 15.7 : Direction of electric field due to positive and negative charges

The principle of superposition applies to electric field also. If there are a number of charges q_1, q_2, q_3, \dots , the corresponding fields at a point P according to Eqn. (15.13) are

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$$\mathbf{E}_1 = k \times \frac{q_1}{r_1^2} \hat{\mathbf{r}}_1, \quad \mathbf{E}_2 = k \times \frac{q_2}{r_2^2} \hat{\mathbf{r}}_2 \text{ and } \mathbf{E}_3 = k \times \frac{q_3}{r_3^2} \hat{\mathbf{r}}_3$$

The total field at point P due to all charges is the vector sum of all fields. Thus,

$$\mathbf{E} = \mathbf{E}_1 + \mathbf{E}_2 + \mathbf{E}_3 + \dots$$

or

$$\mathbf{E} = k \sum_{i=1}^N \frac{q_i \hat{\mathbf{r}}_i}{r_i^2} \quad (15.15)$$

where r_i is the distance between P and charge q_i and $\hat{\mathbf{r}}_i$ is the unit vector directed from q_i to P . The force on a charge q in an electric field \mathbf{E} is

$$\mathbf{F} = q \mathbf{E} \quad (15.16)$$

Example 15.3 : The electric force at some point due to a point charge $q = 3.5\mu\text{C}$ is $8.5 \times 10^{-4} \text{ N}$. Calculate the strength of electric field at that point.

Solution : From Eq. (15.16) we can write

$$E = \frac{F}{q} = \frac{8.5 \times 10^{-4} \text{ N}}{3.5 \times 10^{-6} \text{ C}}$$

$$= 2.43 \times 10^2 \text{ NC}^{-1}$$

Example 15.4 : Three equal positive point charges are placed at the three corners of an equilateral triangle, as shown in Fig. 15.8. Calculate the electric field at the centroid P of the triangle.

Solution : Suppose that a test charge q_0 has been placed at the centroid P of the triangle. The test charge will experience force in three directions making same angle between any two of them. The resultant of these forces at P will be zero. Hence the field at P is zero.

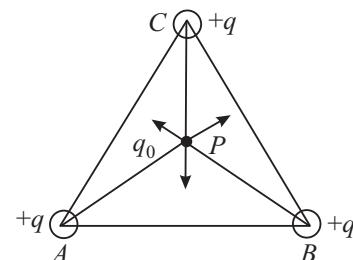


Fig. 15.8 : Electric field at the centroid of an equilateral triangle due to equal charges at its three corners is zero.



INTEXT QUESTIONS 15.3

- A charge $+Q$ is placed at the origin of co-ordinate system. Determine the direction of the field at a point P located on
 - $+x$ -axis
 - $+y$ -axis
 - $x = 4$ units and $y = 4$ units



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2. The ΔABC is defined by $AB = AC = 40\text{ cm}$. And angle at A is 30° . Two charges, each of magnitude $2 \times 10^{-6}\text{ C}$ but opposite in sign, are placed at B and C , as shown in Fig. 15.9. Calculate the magnitude and direction of the field at A .
3. A negative charge is located in space and the electric field is directed towards the earth. What is the direction of the force on this charge?
4. Two identical charges are placed on a plane surface separated by a distance d between them. Where will the resultant field be zero?

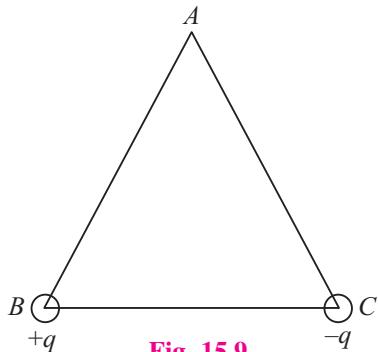


Fig. 15.9

15.3.1 Electric Field due to a Dipole

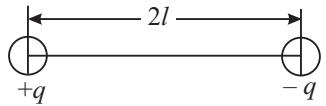


Fig. 15.10: Two unlike charges of equal magnitude separated by a small distance form a dipole.

If two equal and opposite charges are separated by a small distance, the system is said to form a dipole. The most familiar example is H_2O . Fig 15.10 shows charges $+q$ and $-q$ separated by a small distance $2l$. The product of the magnitude of charge and separation between the charges is called **dipole moment**, p :

$$p = q \times 2l \quad (15.17)$$

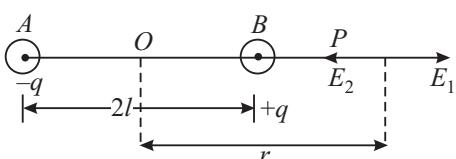
Its SI unit is coulomb-metre.

The dipole moment is a vector quantity. Eqn. (15.17) gives its magnitude and its direction is from negative charge to positive charge along the line joining the two charges (axis of the dipole). Having defined a dipole and dipole moment, we are now in a position to calculate the **electric field due to a dipole**. The calculations are particularly simple in the following cases.

CASE I : Electric field due to a dipole at an axial point : End-on position

To derive an expression for the electric field of a dipole at a point P which lies on the axis of the dipole, refer to Fig. 15.11. This is known as **end-on position**. The point charges $-q$ and $+q$ at points A and B are separated by a distance $2l$. The point O is at the middle of AB . Suppose that point P is at a distance r from the mid point O . Then electric field at P due to $+q$ at B is given by

$$\mathbf{E}_1 = k \times \frac{q}{(r-l)^2} \text{ in the direction } AP$$

Fig. 15.11 : Field at point P on the dipole axis

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Similarly, the electric field \mathbf{E}_2 at P due to $-q$ is given by

$$\mathbf{E}_2 = k \times \frac{q}{(r+l)^2} \text{ in the direction } PA$$

The resultant field \mathbf{E} at P will be in the direction of \mathbf{E}_1 , since \mathbf{E}_1 is greater than \mathbf{E}_2 [as $(r-l)$ is less than $(r+l)$]. Hence

$$\mathbf{E} = \frac{kq}{(r-l)^2} - \frac{kq}{(r+l)^2}$$

$$= kq \left[\frac{1}{(r-l)^2} - \frac{1}{(r+l)^2} \right]$$

$$= kq \left[\frac{(r+l)^2 - (r-l)^2}{(r^2 - l^2)^2} \right]$$

$$= kq \times \frac{4lr}{(r^2 - l^2)^2}$$

$$(a+b)^2 - (a-b)^2 = 4ab$$

$$(a+b)(a-b) = a^2 - b^2$$

$$= k \frac{(2lq) 2r}{(r^2 - l^2)^2}$$

$$= k \frac{2\mathbf{p}r}{(r^2 - l^2)^2}$$

where dipole moment $\mathbf{p} = 2lq$. Since $k = 1/4\pi\epsilon_0$, we can rewrite it as

$$\mathbf{E} = \frac{2\mathbf{p}}{4\pi\epsilon_0} \times \frac{r}{r^4 (1 - l^2/r^2)^2}$$

If $r \gg l$, l^2/r^2 will be very small compared to 1. It can even be neglected and the expression for electric field then simplifies to

$$\mathbf{E} = \frac{2\mathbf{p}}{4\pi\epsilon_0 r^3} \quad (15.18)$$

It shows that electric field is in the direction of \mathbf{p} and its magnitude is inversely proportional to the third power of distance of the observation point from the centre of the dipole.

CASE II : Electric field due to a dipole at a point on the perpendicular bisector : Broad-on position

Suppose that point P lies on the perpendicular bisector of the line joining the charges shown in Fig. 15.12. Note that $AB = 2l$, $OP = r$, and $AO = OB = l$.



Notes

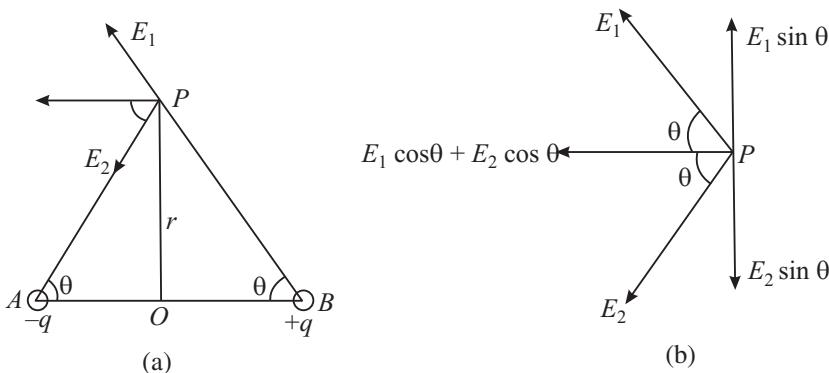


Fig. 15.12 : a) Field at point P on the perpendicular bisector of the line joining the charges, and b) resolution of field in rectangular components.

The angle θ is shown in Fig. 15.12(a). From right angled Δ s PAO and PBO , we can write

$$AP = BP = \sqrt{l^2 + r^2}$$

The field at P due to charge $+q$ at B in the direction of BP can be written as

$$\mathbf{E}_1 = k \frac{q}{l^2 + r^2}$$

Similarly, the field at P due to charge at A in the direction of PA is given as

$$\mathbf{E}_2 = k \frac{q}{l^2 + r^2}$$

Note that the magnitudes of \mathbf{E}_1 and \mathbf{E}_2 are equal.

Let us resolve the fields \mathbf{E}_1 and \mathbf{E}_2 parallel and perpendicular to AB . The components parallel to AB are $\mathbf{E}_1 \cos \theta$ and $\mathbf{E}_2 \cos \theta$, and both point in the same direction.

The components normal to AB are $\mathbf{E}_1 \sin \theta$ and $\mathbf{E}_2 \sin \theta$ and point in opposite directions. (Fig. 15.12b) Since these component are equal in magnitude but opposite in direction, they cancel each other. Hence, the magnitude of resultant electric field at P is given by

$$\begin{aligned} E &= E_1 \cos \theta + E_2 \cos \theta \\ &= k \frac{q}{l^2 + r^2} \cos \theta + k \frac{q}{l^2 + r^2} \cos \theta \end{aligned}$$

But $\cos \theta = \frac{l}{\sqrt{l^2 + r^2}}$. Using this expression in the above result, the electric field at P is given by

$$E = \frac{kq}{(l^2 + r^2)} \times \frac{2l}{\sqrt{l^2 + r^2}}$$

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$$= k \frac{2lq}{(l^2 + r^2)^{3/2}}$$

$$= k \frac{2lq}{r^3 (1 + l^2/r^2)^{3/2}}$$

But $p = 2lq$. If $r^2 \gg l^2$, the factor l^2/r^2 can be neglected in comparison to unity. Hence

$$E = \frac{p}{4\pi\epsilon_0 r^3} \quad (15.19)$$

Note that electric field due to a dipole at a point in broad-on position is inversely proportional to the third power of the perpendicular distance between P and the line joining the charges.

If we compare Eqns. (15.18) and (15.19), we note that the electric field in both cases is proportional to $1/r^3$. But there are differences in details:

- The magnitude of electric field in end-on-position is twice the field in the broad-on position.
- The direction of the field in the end-on position is along the direction of dipole moment, whereas in the broad-on position, they are oppositely directed.

15.3.2 Electric Dipole in a Uniform Field

A uniform electric field has constant magnitude and fixed direction. Such a field is produced between the plates of a charged parallel plate capacitor. Pictorially, it is represented by equidistant parallel lines. Let us now examine the behaviour of an electric dipole when it is placed in a uniform electric field (Fig 15.13).

Let us choose x -axis such that the electric field points along it. Suppose that the dipole axis makes an angle θ with the field direction. A force qE acts on charge $+q$ along the $+x$ direction and an equal force acts on charge $-q$ in the $-x$ direction. Two equal, unlike and parallel forces form a couple and tend to rotate the dipole in clockwise direction. This couple tends to align the dipole in the direction of the external electric field E . The magnitude of torque τ is given by

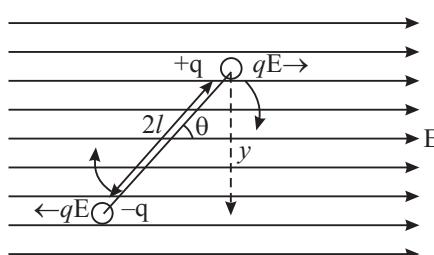


Fig. 15.13 : A dipole in a uniform electric field. The forces on the dipole form a couple and tend to rotate it.

$$\begin{aligned}\tau &= \text{Force} \times \text{arm of the couple} \\ &= qE \times y \\ &= qE \times 2l \sin \theta \\ &= pE \sin \theta\end{aligned}$$



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In vector form, we can express this result to

$$\tau = \mathbf{p} \times \mathbf{E} \quad (15.20)$$

We note that

- when $\theta = 0$, the torque is zero, and
- for $\theta = 90^\circ$, the torque on the dipole is maximum, equal to pE . So we may conclude that the electric field tends to rotate the dipole and align it along its own direction.

Example 15.5 : Two charges $+q$ and $-q$, each of magnitude 6.0×10^{-6} C, form a dipole. The separation between the charges is 4×10^{-10} m. Calculate the dipole moment. If this dipole is placed in a uniform electric field $E = 3.0 \times 10^2$ NC $^{-1}$ at an angle 30° with the field, calculate the value of torque on the dipole.

Solution : The dipole moment $p = qd$

$$\begin{aligned} &= (6.0 \times 10^{-6}\text{C}) \times (4.0 \times 10^{-10}\text{m}) \\ &= 24 \times 10^{-16}\text{Cm}. \end{aligned}$$

Since torque $\tau = pE \sin \theta$, we can write

$$\begin{aligned} \tau &= (24 \times 10^{-16}\text{cm}) \times 3.0 \times 10^2 \text{ NC}^{-1} \sin 30^\circ \\ &= \frac{72}{2} \times 10^{-14} \text{ Nm} \\ &= 36 \times 10^{-14} \text{ Nm} \end{aligned}$$

If a dipole is placed in a non-uniform electric field, the forces on the charges $-q$ and $+q$ will be unequal. Such as electric field will not only tend to rotate but also displace the dipole in the direction of the field.

15.3.3 Electric Lines of Force (Field Lines)

A very convenient method for depicting the electric field (or force) is to draw lines of force pointing in the direction of the field. The sketch of the electric field lines gives us an idea of the magnitude and direction of the electric field. **The number of field lines passing through a unit area of a plane placed perpendicular the direction of the field is proportional to the strength of the field.** A tangent at any point on the field lines gives the direction of the field at that point.

Note that the *electric field lines are only fictitious construction to depict the field. No such lines really exist.* But the behaviour of charges in the field and the interaction between charges can be effectively explained in terms of field lines. Some illustrative examples of electric field lines due to point charges are shown in Fig 15.14. The field lines of a stationary positive charge point radially in outward

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direction. But for stationary negative charge, the lines start from infinity and terminate at the point charge in radially inward direction (towards the point charge). You must understand that the electric field lines in both cases are in all directions in the space. Only those which are in the plane containing the charge are shown here.

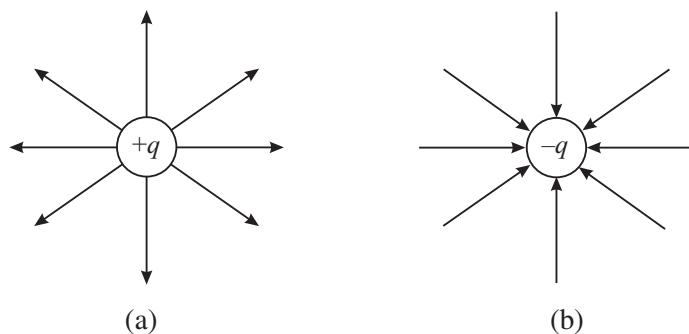


Fig. 15.14 : Electrical field lines of single point charges : a) The field lines of positive charge, and b) the field lines of negative charge.

Fig 15.15(a) shows a sketch of electric field lines of two equal and similar positive charges placed close to each other. The lines are almost radial at points very close to the positive charges and repel each other, bending outwards. There is a point P midway between the charges where no lines are present. The fields of the two charges at this point cancel each other and the resultant field at this point is zero.

Fig. 15.15(b) depicts the field lines due to a dipole. The number of lines leaving the positive charge is equal to the number of lines terminating on the negative charge.

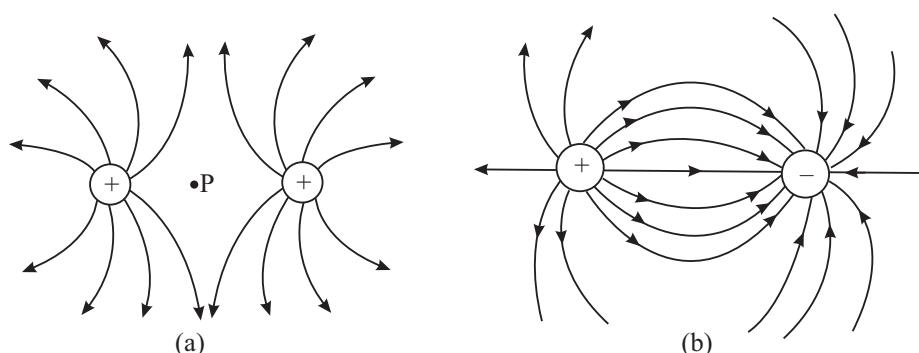


Fig. 15.15 : Electric field lines due to a system of two point charges : a) Two positive charges at rest, and b) The field lines due to a dipole start from the positive charge and terminate on the negative charge.

You must remember the following properties of the electric field lines :

- The field lines start from a positive charge radially outward in all directions and terminate at infinity.
- The field lines start from infinity and terminate radially on a negative charge.



Notes

- For a dipole, field lines start from the positive charge and terminate on the negative charge.
- A tangent at any point on field line gives the direction of electric field at that point.
- The number of field lines passing through unit area of a surface drawn perpendicular to the field lines is proportional to the field strength on this surface.
- Two field lines never cross each other.

15.4 ELECTRIC FLUX AND GAUSS' LAW

Let us consider a sphere of radius r having charge $+q$ located at its center. The magnitude of electric field at every point on the surface of this sphere is given by

$$E = k \times \frac{q}{r^2}$$

The direction of the electric field is normal to the surface and points outward. Let us consider a small element of area Δs on the spherical surface. Δs is a vector whose magnitude is equal to the element of area Δs and its direction is perpendicular to this element (Fig.15.16). The electric flux $\Delta\phi$ is defined as the scalar product of Δs and E :

$$\Delta\phi = \mathbf{E} \cdot \Delta\mathbf{s}$$

The total flux over the entire spherical surface is obtained by summing all such contributions:

$$\phi_E = \sum_{\Delta s_i \rightarrow 0} \mathbf{E}_i \cdot \Delta\mathbf{s}_i \quad (15.21)$$

Since the angle between \mathbf{E} and $\Delta\mathbf{s}$ is zero, the total flux through the spherical surface is given by

$$\phi_E = k \times \frac{q}{r^2} \sum \Delta s$$

The sum of all elements of area over the spherical surface is $4\pi r^2$. Hence the net flux through the spherical surface is

$$\phi_E = k \times \frac{q}{r^2} \times 4\pi r^2$$

$$= 4\pi k \times q$$

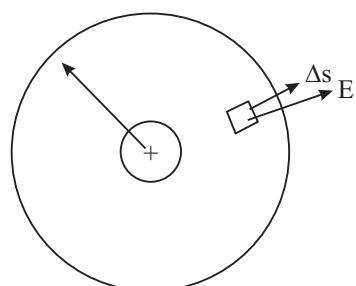


Fig. 15.16

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On substituting for $k = 1/4\pi\epsilon_0$, we get

$$\begin{aligned}\phi_E &= \frac{1}{4\pi\epsilon_0} \times 4\pi q \\ &= q/\epsilon_0\end{aligned}\quad (15.22)$$

The spherical surface of the sphere is referred to as *Gaussian surface*. Eqn. (15.22) is known as Gauss' law. It states that *the net electric flux through a closed gaussian surface is equal to the total charge q inside the surface divided by ε₀*.

Gauss' law is a useful tool for determining the electric field. You must also note that gaussian surface is an imaginary mathematical surface. It may not necessarily coincide with any real surface.

Carl Friedrich Gauss (1777 – 1855)



German genius in the field of physics and mathematics, Gauss has been one of the most influential mathematicians. He contributed in such diverse fields as optics, electricity and magnetism, astronomy, number theory, differential geometry, and mathematical analysis.

As child prodigy, Gauss corrected an error in his father's accounts when he was only three year old. In primary school, he stunned his teacher by adding the integers 1 to 100 within a second.

Though he shun interactions with scientific community and disliked teaching, many of his students rose to become top class mathematicians – Richard Dedekind, Berhard Riemann, Friedrich Bessel and Sophie Germain are a few among them. Germany issued three postal stamps and a 10 mark bank note in his honour. A crater on moon called Gauss crater, and asteroid 100 called Gaussia have been named after him.

15.4.1 Electric Field due to a Point Charge

Let us apply Gauss' law to calculate electric field due to a point charge. Draw a spherical surface of radius r with a point charge at the centre of the sphere, as shown in Fig. 15.17.

The electric field \mathbf{E} is along the radial direction pointing away from the centre and normal to the surface of the

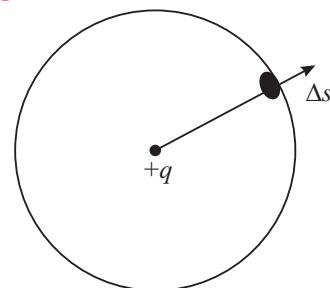


Fig. 15.17 : Electric field on a spherical surface due to a charge $+q$ at its centre



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sphere at every point. The normal to the element of area Δs is parallel to E . According to Gauss' law, we can write

$$\phi_E = \sum_i E_i \cdot \Delta s_i = q/\epsilon_0$$

Since $\cos \theta = 1$ and E is same on all points on the surface, we can write

$$\phi_E = E \times 4\pi r^2$$

or

$$q/\epsilon_0 = E \times 4\pi r^2$$

$$\Rightarrow E = \frac{q}{4\pi\epsilon_0 r^2} \quad (15.23)$$

If there is a second charge q_0 placed at a point on the surface of the sphere, the magnitude of force on this charge would be

$$F = q_0 \times E$$

so that

$$F = \frac{qq_0}{4\pi\epsilon_0 r^2} \quad (15.24)$$

Do you recognise this result? It is expression for Coulomb's force between two static point charges.

15.4.2 Electric Field due to a Long Line Charge

A line charge is in the form of a **thin charged wire of infinite length** with a uniform linear charge density σ_l (charge per unit length). Let there be a charge $+q$ on the wire. We have to calculate the electric field at a point P at a distance r . Draw a right circular cylinder of radius r with the long wire as the axis of the cylinder. The cylinder is closed at both ends. The surface of this cylinder is the gaussian surface and shown in Fig. 15.18. The magnitude of the electric field E is same at every point on the curved surface of the cylinder because all points are at the same distance from the charged wire. The electric field direction and the normal to area element Δs are parallel.

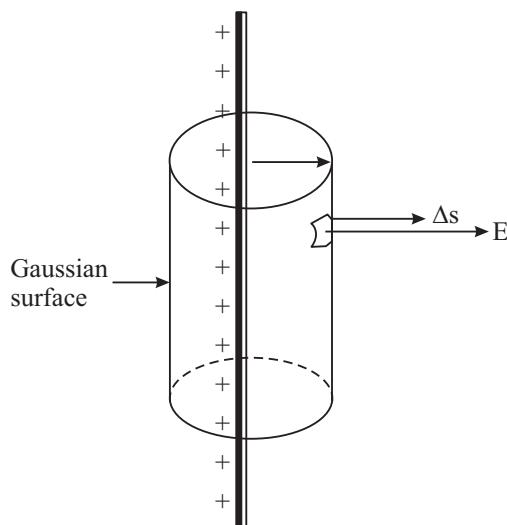


Fig. 15.18 : Electric field due to an infinite line of charges having uniform linear charge density. The gaussian surface is a right circular cylinder.

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Let the length of the gaussian cylinder be l . The total charge enclosed in the cylinder is $q = \sigma_l l$. The area of the curved surface of the cylinder is $2\pi r l$.

For the flat surfaces at the top and bottom of the cylinder, the normals to these areas are perpendicular to the electric field ($\cos 90^\circ = 0$). These surfaces, therefore, do not contribute to the total flux. Hence

$$\begin{aligned}\phi_E &= \sum \mathbf{E} \cdot \Delta \mathbf{s} \\ &= E \times 2\pi r l\end{aligned}$$

According to Gauss' law, $\phi_E = q/\epsilon_0$. Hence

$$E \times 2\pi r l = q/\epsilon_0 = \sigma_l l / \epsilon_0$$

or $E = \frac{\sigma_l}{2\pi\epsilon_0 r}$ (15.25)

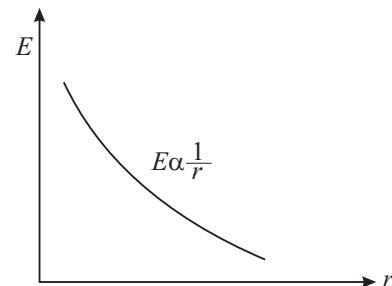


Fig. 15.19 : Variation of E with r for a line charge

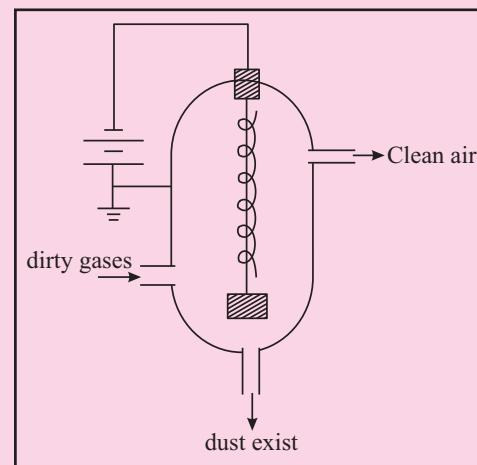
This shows that electric field varies inversely with distance. This is illustrated in Fig. 15.19.

Electrostatic Filter

You must have seen black smoke and dirt particles coming out of a chimney of a thermal power station or brick klin. The smoke consists of not only gases but large quantities of small dust (coal) particles. The smoke along with the dirt is discharged into the atmosphere. The dust particles settle down on earth and pollute the soil. The gases contribute to global warming. These are extremely injurious to living systems (health). It is therefore essential that the dirt is removed from smoke before it is discharged into the atmosphere.

A very important application of electrical discharge in gases by application of high electric field is the construction of a device called *Electrostatic Filter or Precipitator*.

The basic diagram of the device is shown here. The central wire inside a metallic container is maintained at a very high negative potential (about 100 kV). The wall of the container is connected to the positive terminal of a high volt battery and is **earthed**. A weight W keeps the wire straight in the central part. The electric field thus created is from the wall towards the wire. The dirt and gases are passed





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through the container. An electrical discharge takes place because of the high field near the wire. Positive and negative ions and electrons are generated. These negatively charged particles are accelerated towards the wall. They collide with dust particles and charge them. Most of the dust particles become negatively charged because they capture electrons or negative ions. They are attracted towards the wall of the container. The container is periodically shaken so that the particles leave the surface and fall down at the bottom of the container. These are taken out through the exit pipe.

The undesirable dust particles are thus removed from the gases and the clean air goes out in the atmosphere. Most efficient systems of this kind are able to remove about 98% of the ash and dust from the smoke.

15.4.3 Electric Field due to a Uniformly Charged Spherical Shell

A spherical shell, by definition, is a hollow sphere having infinitesimal small thickness. Consider a spherical shell of radius R carrying a total charge Q which is uniformly distributed on its surface. We shall calculate the electric field due to the spherical charge distribution at points external as well as internal to the shell.

(a) Field at an external point

Let P be an external point distant r from the center O of the shell. Draw a spherical surface (called Gaussian surface) passing through P and concentric with the charge distribution. By symmetry, the electric field is radial, being directed outward as shown in Fig. 15.20.

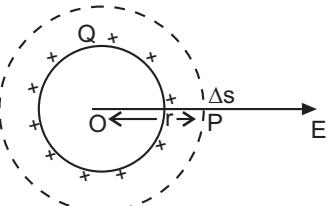


Fig. 15.20

The electric field \vec{E} is normal to the surface element everywhere. Its magnitude at all points on the Gaussian surface has the same value E .

According to Gauss' law,

$$\Sigma E \Delta s \cos 0^\circ = \frac{Q}{\epsilon_0}$$

or
$$\Delta E \cdot 4\pi r^2 = \frac{Q}{\epsilon_0}$$

or
$$E = \frac{Q}{4\pi\epsilon_0 r^2}$$

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From the result we can conclude that for a point external to the spherical shell, the entire charge on the shell can be treated as though located at its centre. The electric field decreases with distance.

Instead of a spherical shell if we had taken a charged solid conducting sphere, we would have obtained the same result. This is because the charge of a conductor always resides on its outer surface.

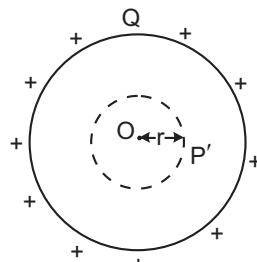


Fig. 15.21

(b) Field at an Internal Point

Let P' be an internal point distant r from the centre of the shell. Draw a concentric sphere passing through the point P' .

Applying Gauss' Law,

$$\Sigma E \Delta s \cos 0^\circ = \frac{Q}{\epsilon_0}$$

or $E \cdot 4\pi r^2 = \frac{Q}{\epsilon_0}$

$\Rightarrow E = 0$ as $Q = 0$

the electric field at an internal point of the shell is zero. The same result is applicable to a charged solid conducting sphere.

The variation of the electric field with the radial distance r has been shown in Fig 15.22.

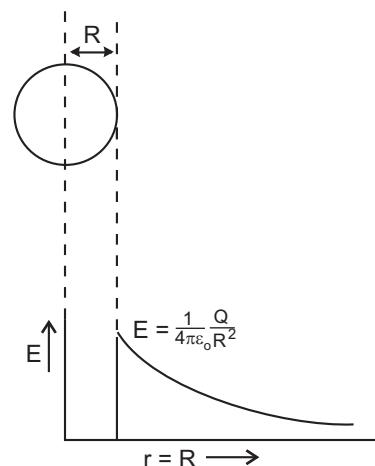


Fig. 15.22

15.4.4 Electric Field due to a Plane Sheet of Charge

Consider an infinite plane sheet of charge $ABCD$, charged uniformly with surface charge density σ .

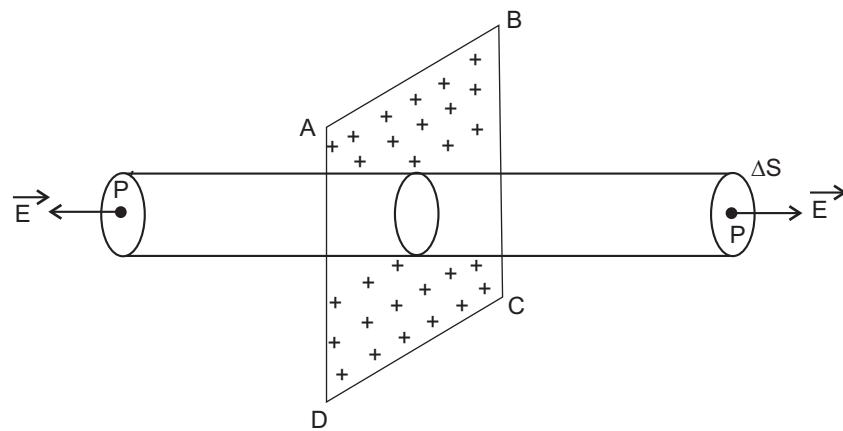


Fig. 15.23



Notes

For symmetry reasons, the electric field will be perpendicular to the sheet, directed away from it, if $\sigma > 0$. Let P be the point in front of the sheet where we want to find the electric field. Draw a Gaussian surface in the form of a cylinder with its axis parallel to the field and one of its circular caps passing through P . The other circular cap of the cylinder lies symmetrically opposite at P' , on the other side of the sheet, being situated at the same distance as P .

The electric flux through both the circular caps is

$$\begin{aligned}\vec{E} \cdot \Delta \vec{s} + \vec{E} \cdot \Delta \vec{s} &= E\Delta s + E\Delta s \\ &= 2E\Delta s\end{aligned}$$

The electric flux through the curved surface of the Gaussian surface is $\vec{E} \cdot \Delta \vec{s} = E\Delta s \cos 90^\circ = 0$. Hence, the total electric flux through the Gaussian cylinder is

$$\begin{aligned}\phi_E &= \sum \vec{E} \cdot \vec{\Delta s} \\ &= 2E\Delta s\end{aligned}$$

As the charge enclosed by the Gaussian cylinder is $\sigma\Delta s$, using Gauss' Law we have

$$2E\Delta s = \frac{1}{\epsilon_0} \sigma \Delta s$$

or

$$E = \frac{\sigma}{2\epsilon_0}$$

Please note that the electric field is independent of the distance from the sheet.

15.5 VAN DE GRAAFF GENERATOR

Van de Graaff Generator is an electrostatic device that can produce potential differences of the order of a few million volts. It was named after its designer Robert J. van de Graaff.

It consists of a large hollow metallic sphere S mounted on an insulating stand. A long narrow belt, made of an insulating material, like rubber or silk, is wound around two pulleys P_1 and P_2 as shown in Fig. 15.5. The pulley P_2 is mounted at the centre of the sphere S while the pulley P_1 is mounted near the bottom. The belt is made to rotate continuously by driving the pulley P_1 by an electric motor M . Two comb-shaped conductors C_1 and C_2 , having a number of sharp points in the shape of metallic needles, are mounted near the pulleys.

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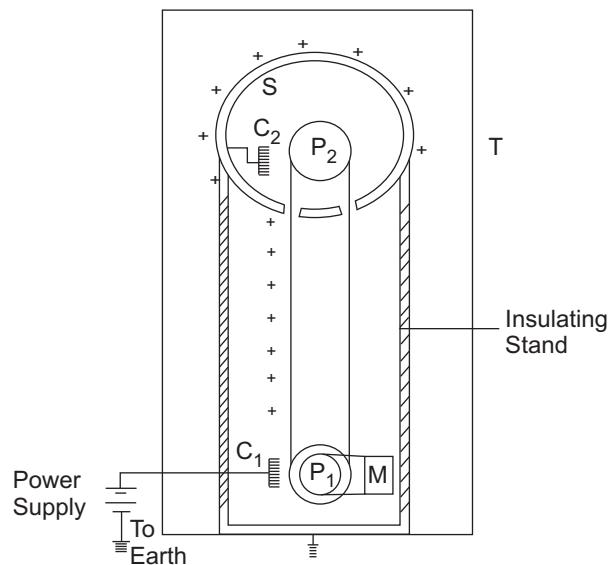


Fig. 15.24

The needles point towards the belt. The comb-shaped conductor C_1 is maintained at a high positive potential ($\sim 10^4$ V) relative to the ground with the help of a power supply. The upper comb C_2 is connected to the inner surface of the metallic sphere S .

Near the sharp points of the comb-shaped conductor C_1 , the charge density and electrostatic field are very high. Large electrostatic field near their pointed ends causes dielectric breakdown of the air, producing ions (both positive and negative) in the process. This phenomenon is known as corona discharge. The negative charges from the air move towards the needles and the positive charge towards the belt. The negative charges neutralize some of the positive charges on the comb C_1 . However, by supplying more positive charges to C_1 , the power supply maintains its positive potential. As the belt carrying the positive charges moves towards C_2 , the air near it becomes conducting due to corona discharge. The negative charges of the air move towards the belt neutralizing its positive charges while the positive charges of the air move towards the needles of the comb C_2 . These positive charges are then transferred to the conducting sphere S which quickly moves them to its outer surface.

The process continues and positive charges keep on accumulating on the sphere S and it acquires a high potential.

As the surrounding air is at ordinary pressure, the leakage of charge from the sphere takes place. In order to prevent the leakage, the machine is surrounded by an earthed metallic chamber T whose inner space is filled with air at high pressure.



Notes



INTEXT QUESTIONS 15.4

- If the electric flux through a gaussian surface is zero, does it necessarily mean that
 - the total charge inside the surface is zero?
 - the electric field is zero at every point on the surface?
 - the electric field lines entering into the surface is equal to the number going out of the surface?
- If the electric field exceeds the value $3.0 \times 10^6 \text{ NC}^{-1}$, there will be sparking in air. What is the highest value of charge that a metallic sphere can hold without sparking in the surrounding air, if the radius of the sphere is 5.0 cm?
- What is the magnitude and direction of the net force and net torque on a dipole placed along a a) uniform electric field, and b) non-uniform field.



WHAT YOU HAVE LEARNT

- Electric charge is produced when glass rod is rubbed with silk or rubber is rubbed with fur.
- By convention, the charge on glass rod is taken **positive** and that on rubber is taken **negative**.
- Like charges repel and unlike charges attract each other.
- Coulomb's law gives the magnitude and direction of force between two point charges :

$$\mathbf{F} = k \frac{q_1 \times q_2}{r^2} \hat{\mathbf{r}}$$

where $k = \frac{1}{4\pi\epsilon_0} = 9.0 \times 10^9 \text{ Nm}^2\text{C}^{-2}$.

- The smallest unit of charge in nature is the charge on an electron :

$$e = 1.60 \times 10^{-19} \text{ C}$$

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Electric Charge and Electric Field

- Charge is conserved and quantised in terms of electronic charge.
- The electric field \mathbf{E} due to a charge q at a point in space is defined as the force experienced by a unit test charge q_0 :

$$\mathbf{E} = \mathbf{F}/q_0 = k \times \frac{q}{r^2} \hat{\mathbf{r}}$$

- Superposition principle can be used to obtain the force experienced by a charge due to a group of charges. It is also applicable to electric field at a point due to a group of charges.
- Electric dipole is a system of two equal and unlike charges separated by a small distance. It has a dipole moment $|\mathbf{p}| = qr$; the direction of \mathbf{p} is from negative charge to positive charge along the line joining the two charges.
- The electric field due to a dipole in end-on position and broad-on position is respectively given by

$$\mathbf{E} = \frac{1}{4\pi\epsilon_0} \frac{2\mathbf{p}}{r^3}$$

and

$$\mathbf{E} = -\frac{1}{4\pi\epsilon_0} \frac{\mathbf{p}}{r^3}.$$

- Electric field lines (line of force) are only a pictorial way of depicting field.
- Electric flux is the total number of electric lines of force passing through an area and is defined as $\phi_E = \mathbf{E} \cdot \mathbf{A}$.
- Gauss's law states that the total flux passing through a closed area is $\frac{1}{\epsilon_0}$ times the total charge enclosed by it.
- The electric field due to a line charge is given by $E = \frac{\sigma_l}{2\pi\epsilon_0 r}$.



TERMINAL EXERCISE

1. A $+12\mu\text{C}$ charge is at $x=20\text{ cm}$ and a $-18\mu\text{C}$ ($-q$) charge is at $x=29\text{ cm}$ on the x -axis. Calculate the magnitude and direction of the force on a charge of $18\mu\text{C}$. What is the direction of force on $12\mu\text{C}$ charge?
2. Two point Charges q_1 and q_2 separated by a distance of 3.0 m experience a mutual force of $16 \times 10^{-15}\text{ N}$. Calculate the magnitude of force when $q_1 = q_2 = q$. What will be the magnitude of force if separation distance is changed to 6.0 m ?



Notes

3. There are two points A and B separated by a distance x . If two point charges $+q$ each are on the points A and B , the force between them is F . The point charges are now replaced by two identical metallic spheres having the same charge $+q$ on each. The distance between their centers is again x only. Will the force between them change? Give reasons to support your answer.
4. The force of repulsion between two point chargers placed 16 cm apart in vaccum is 7.5×10^{-10} N. What will be force between them, if they are placed in a medium of dielectric constant $k = 2.5$?
5. Compare the electrical force with the gravitational force between two protons separated by a distance x . Take charge on proton as 1.60×10^{-19} C, mass of proton as 1.67×10^{-27} kg and Gravitational constant $G = 6.67 \times 10^{-11}$ Nm 2 kg $^{-2}$.
6. Four identical point charges $+q$ each are placed at the four corners (one q at one corner) of a square of side 1. Find the force experienced by a test charge q_0 placed at the center of the square.
7. When are the electric field lines parallel to each other?
8. How many electrons should be removed from a metallic sphere to give it a positive charge $= 6.4 \times 10^{-7}$ C.
9. Consider an electric dipole of $q = 3.0 \times 10^{-6}$ C and $2l = 4 \times 10^{-10}$ m. Calculate the magnitude of dipole moment. Calculate electric field at a point $r = 6 \times 10^{-6}$ m on the equatorial plane.
10. A Charge $-q = 15 \times 10^{-6}$ C is placed on a metallic sphere of radius $R = 3.0$ mm. Calculate the magnitude and direction of the electric field at a point $r = 15$ cm from the center of the sphere. What will be the magnitude and direction of the field at the same point if 3.0 mm sphere is replaced by 9.0 mm sphere having the same Charge.
11. A charge of $+15\mu\text{C}$ is located at the center of a sphere of radius 20 cm. Calculate the electric flux through the surface of the sphere.
12. A proton is placed in a uniform electric field $E = 8.0 \times 10^4$ NC $^{-1}$. Calculate the acceleration of the proton.
13. Two point charges q_1 and q_2 are 3.0 cm apart and $(q_1 + q_2) = 20\mu\text{C}$. If the force of repulsion between them is 750N, calculate q_1 and q_2 .



ANSWERS TO INTEXT QUESTIONS

15.1

1. (i) Yes (ii) Charge $= 3.2 \times 10^{-17}$ C.

MODULE - 5

Electricity and
Magnetism



Notes

Electric Charge and Electric Field

2. A has charge $+Q$. When A and B are brought in contact, charge will get distributed equally.
(i) Yes., (ii) $+Q/2$
3. $q = 4.8 \times 10^{-16}$

Since $Ne = q$, we get

$$N = \frac{4.8 \times 10^{-16}}{1.6 \times 10^{-19}} = 3.0 \times 10^3 \text{ charges}$$

15.2

1. $Q_1 = 16\mu\text{C}$, $Q_2 = q\mu\text{C}$ and $r = 12\text{m}$

Since

$$\begin{aligned} F &= \frac{1}{4\pi\epsilon_0} \frac{q_1 q_2}{r^2} \\ &= \frac{(9 \times 10^9 \text{ Nm}^2\text{C}^{-2})(16 \times 10^{-6}\text{C}) \times (12 \times 10^{-6}\text{C})}{144\text{m}^2} \\ &= 9 \times 10^{-3} \text{ N} \end{aligned}$$

- (i) direction from q_2 to q_1
- (ii) direction from q_1 to q_2

2. The force at A due to charge at B, $F_1 = k \frac{q^2}{a^2}$ where $AB = a$

Since $AB = AC$, the force at A due to charge at B is

$$\begin{aligned} F_2 &= k \frac{q^2}{a^2} \\ R^2 &= F_1^2 + F_2^2 = 2F^2 \\ R &= F\sqrt{2} \text{ at } 45^\circ \end{aligned}$$

15.3

1. (a) E along the $+x$ axis.
(b) along the $+y$ axis.
(c) at 45° with the x axis
2. $AB = AC = 40\text{ cm}$

$$|\mathbf{E}_1| = \frac{kq}{r^2} = |\mathbf{E}_2| = \frac{9 \times 10^9 \text{ Nm}^2\text{C}^{-2} \times (2 \times 10^{-6}\text{C})}{(0.40\text{m})^2} = 1.125 \times 10^5 \text{ NC}^{-1}$$

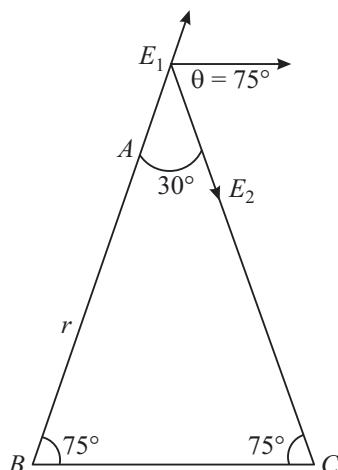


Notes

The resultant of \mathbf{E}_1 and \mathbf{E}_2 will be parallel to BC . Hence

$$\begin{aligned} R^2 &= E_1^2 + E_2^2 + 2E_1 E_2 \cos 150^\circ \\ &= 2E^2 + 2E^2 \cos(180^\circ - 30^\circ) \\ &= 2E^2 - 2E^2 \cos 30^\circ = 2E^2 \left(1 + \frac{\sqrt{3}}{2}\right) = 4.723 \times 10^{10} \text{ N}^2\text{C}^{-2}. \end{aligned}$$

Direction will be parallel to BC in the direction $B \rightarrow C$.



3. \mathbf{E} is directed towards the earth. The force on $-ve$ charge will be vertically upwards.
4. The field will be zero at the mid point between the charge.

15.4

1. (i) Yes (ii) not necessarily (iii) Yes.

$$2. E = \frac{Q}{4\pi\epsilon_0 r^2}$$

$$\therefore Q = 4\pi\epsilon_0 r^2 E$$

$$= (3 \times 10^6 \text{ NC}^{-1}) \times \frac{1}{(9 \times 10^9 \text{ Nm}^2\text{C}^{-2})} \times (25 \times 10^{-4} \text{ m}^2)$$

$$= 8.3 \times 10^{-7} \text{ C}$$

3. (a) $\mathbf{F} = 0, \tau = 0$
(b) $\mathbf{F} \neq 0, \tau = 0$

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Electricity and
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Notes

Electric Charge and Electric Field

Answers to Problems in Terminal Exercise

1. 240 N towards negative x -direction force on $+12 \mu\text{C}$ charge is towards $+x$ direction.
2. $q = 4 \times 10^{-3}\text{C}$
3. $3 \times 10^{-10}\text{N}$
4. Electric force is 10^{36} times the gravitational force.
5. zero.
6. 4×10^{12} electrons
7. $12 \times 10^{-16} \text{ Cm. } 0.5 \times 10^{15}$ or Nc^{-1}
8. $6 \times 10^6 \text{ NC}^{-1}$ towards the centre, same field.
9. $1.7 \times 10^6 \mu\text{m}$
10. $7.6 \times 10^{12} \text{ ms}^{-2}$
11. $15 \mu\text{C}$ and $5 \mu\text{C}$.

□

16

Notes

ELECTRIC POTENTIAL AND CAPACITORS

In modules 2 and 3, you learnt about the direction of flow of fluids and thermal energy. You may recall that the level of water in a container determines the direction in which it flows. If the level of water in one container is higher than that in the other, water will flow from higher level to lower level, irrespective of the quantity of water in the containers. Temperature plays a similar role in case of flow of thermal energy from one object to another. Thermal energy always flows from a body at higher temperature to the one at lower temperature. Here also, the direction of flow does not depend on the quantity of thermal energy possessed by an object.

Electric potential plays a similar role in the flow of charges from one point to another. The positive charge always moves from a point at higher potential to a point at lower potential. A positive test charge, when left free in an electric field, moves in the direction of the electric field. From this behaviour of a positive test charge, you may be tempted to say that the **electric field (E)** and **electric potential (V)** are closely related. In this lesson, you will learn to establish a relation between these physical quantities. You will also learn about a device called capacitor, which is used to store charge, filter alternating current and finds wide applications in electronic circuitry as well as power transmission.



OBJECTIVES

After studying this lesson, you should be able to :

- *explain the meaning of electric potential at a point and potential difference;*
- *derive expressions for electric potential due to a point charge and a dipole;*
- *explain the principle of capacitors and state their applications;*
- *derive an expression for the capacitance of a parallel plate capacitor;*
- *obtain equivalent capacitance in grouping of capacitors;*

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Notes

Electric Potential and Capacitors

- calculate the energy stored in a capacitor; and
- explain polarization of dielectric materials in an electric field.

16.1 ELECTRIC POTENTIAL AND POTENTIAL DIFFERENCE

When a charged particle is made to move in an electrostatic field in a direction opposite to the direction of the field, work is done by an external agency. This work is stored as potential energy of charge in accordance with the law of conservation of energy. So, we can say that an electric charge placed at a point in an electric field has potential energy, which is a function of its position. We can visualize the potential energy of charge in the field as a scalar function of position and for a unit charge call it potential. It means that different points in an electric field would be at different potentials. And if a positively charged particle is placed in an electric field, it will tend to move from higher to lower potential to minimize its potential energy. In the next lesson, you will learn how the concept of potential difference leads to flow of current in electric circuits.

The **electric potential** at any point in an electric field is equal to the work done against the electric force in moving a unit positive charge from outside the electric field to that point. Electric potential is a scalar quantity, as it is related to work done.

Alessandro, Conte Volta (1745-1827)



Born at Como, Italy, Volta was a professor at Pavia for more than 20 years. A well travelled man, he was known to many famous men of his times. He decisively proved that animal electricity observed by Luigi Galvani in frog muscles was a general phenomenon taking place between two dissimilar metals separated by acidic or salt solutions. On the basis of this observation, he invented first electro-chemical cell, called voltaic cell. The unit of potential difference is named volt in his honour.

The potential at a point is taken positive when work is done against the field by a positive charge but negative when work is done by the electric field in moving the unit positive charge from infinity to the point in the field.

Consider two points *A* and *B* in an electric field (Fig. 16.1). If a test charge q_0 is moved from point *A* to point *B* along any path by an external force, the amount of work done by the external force is given by

$$W_{AB} = q_0 (V_B - V_A) \quad (16.1)$$

Thus, potential difference between points *A* and *B* will be

$$V_{AB} = V_B - V_A = \frac{W_{AB}}{q_0} \quad (16.2)$$

Here V_A and V_B are potentials at points *A* and *B*, respectively.



Notes

A potential difference is said to exist between two points in an electric field, if work is done against the electric force in moving a positive test charge from one point to the other. Note that this work is independent of the path. (*For this reason, the electric field is said to be a conservative field*). The SI unit of potential and potential difference is **volt** :

$$1 \text{ volt} = 1 \text{ joule}/1 \text{ coulomb}$$

If one joule of work is done in taking a test charge of one coulomb from one point to the other in an electric field, the potential difference between these points is said to be one volt. If one joule of work is done in bringing a test charge of one coulomb from infinity to a point in the field, the potential at that point is one volt.

Note that potential at a point is not a unique quantity as its value depends on our choice of zero potential energy (infinity). However, the potential difference between two points in a stationary field will have a unique value. Let us now learn to calculate potential at a point due to a single charge.

16.1.1 Potential at a point due to a Point Charge

Suppose we have to calculate electric potential at point P due to a single point charge $+q$ situated at O (Fig. 16.2), where $OP = r$. The magnitude of electric field at P due to the point charge is given by

$$E_p = \frac{1}{4\pi\epsilon_0} \times \frac{q}{r^2} \quad (16.3)$$

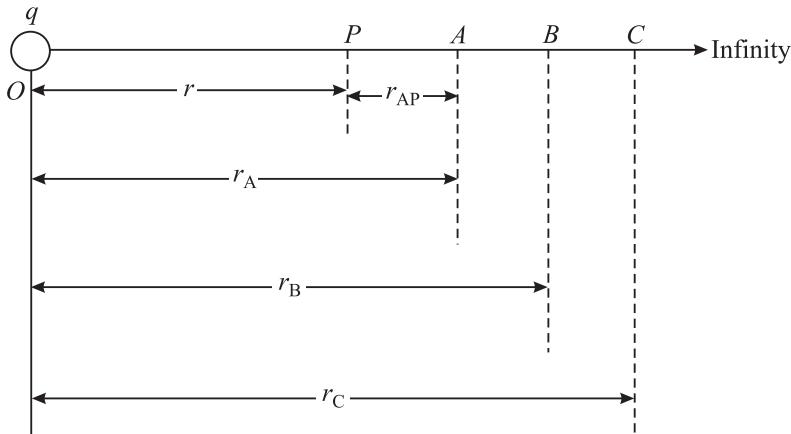


Fig. 16.2 : Work done per unit charge in moving a charge q_0 from infinity to a point P in an electric field E is the potential at that point.

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Electric Potential and Capacitors

Similarly, the electric field at point A will be

$$E_A = \frac{1}{4\pi\epsilon_0} \frac{q}{r_A^2} \quad (16.4)$$

If points P and A are very close, the average field E_{AP} between these points can be taken as the geometric mean of E_p and E_A :

$$\begin{aligned} E_{AP} &= \sqrt{E_A \times E_p} \\ &= \sqrt{\frac{1}{4\pi\epsilon_0} \frac{q}{r_A^2} \times \frac{1}{4\pi\epsilon_0} \frac{q}{r^2}} \\ &= \frac{1}{4\pi\epsilon_0} \frac{q}{r_A r} \end{aligned} \quad (16.5)$$

Therefore, the magnitude of force experienced by a test charge q_0 over this region will be

$$F_{AP} = q_0 E_{AP} = \frac{1}{4\pi\epsilon_0} \frac{q q_0}{r_A r} \quad (16.6)$$

and the work done in moving charge q_0 from A to P is given by

$$\begin{aligned} W_{AP} &= F_{AP} \times r_{AP} \\ &= \frac{1}{4\pi\epsilon_0} \frac{q q_0}{r_A r} \times (r_A - r) \\ &= \frac{q q_0}{4\pi\epsilon_0} \times \left(\frac{1}{r} - \frac{1}{r_A} \right) \end{aligned} \quad (16.7)$$

where r_{AP} is the distance between points A and P.

Similarly, work done in moving this charge from B to A will be given by

$$W_{BA} = \frac{q q_0}{4\pi\epsilon_0} \times \left(\frac{1}{r_A} - \frac{1}{r_B} \right) \quad (16.8a)$$

And work done in moving the test charge from C to B will be

$$W_{CB} = \frac{q q_0}{4\pi\epsilon_0} \times \left(\frac{1}{r_B} - \frac{1}{r_C} \right) \quad (16.8b)$$

and so on. The total work done in moving the charge from infinity to the point P will be



Notes

$$\begin{aligned}
 W &= \frac{q q_0}{4\pi\epsilon_0} \times \left(\frac{1}{r} - \frac{1}{r_A} + \frac{1}{r_A} - \frac{1}{r_B} + \frac{1}{r_B} - \frac{1}{r_C} \dots + \dots - \frac{1}{\infty} \right) \\
 &= \frac{q q_0}{4\pi\epsilon_0} \times \left(\frac{1}{r} - \frac{1}{\infty} \right) \\
 &= \frac{q q_0}{4\pi\epsilon_0 r}
 \end{aligned} \tag{16.9}$$

By definition, potential at a point is given by

$$\begin{aligned}
 V_p &= \frac{W}{q_0} \\
 &= \frac{1}{4\pi\epsilon_0} \frac{q}{r}
 \end{aligned} \tag{16.10}$$

Note that potential is inversely proportional to distance. It is positive or negative depending on whether q is positive or negative.

If there are several charges of magnitudes q_1, q_2, q_3, \dots , the electric potential at a point is the scalar sum of the potentials due to individual charges (Fig.16.3) :

$$\begin{aligned}
 V &= V_1 + V_2 + V_3 + \dots \\
 &= \sum_{i=1}^{\infty} \frac{q_i}{4\pi\epsilon_0 r_i}
 \end{aligned} \tag{16.11}$$

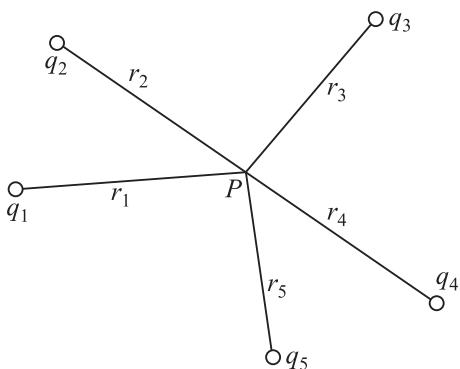


Fig. 16.3 : Potential at a point P due to a system of charges

16.1.2 Potential at a Point due to an Electric Dipole

Let us consider an electric dipole consisting of two equal and opposite point charges $-q$ at A and $+q$ at B , separated by a distance $2l$ with centre at O . We wish to calculate potential at a point P , whose polar co-ordinates are (r, θ) ; i.e. $OP = r$ and $\angle BOP = \theta$, as shown in Fig. 16.4. Here $AP = r_1$ and $BP = r_2$. We can easily calculate potential as P due to point charges at A and B using Eqn.(16.10) :

$$V_1 = \frac{1}{4\pi\epsilon_0} \times \frac{(-q)}{r_1}$$

and

$$V_2 = \frac{1}{4\pi\epsilon_0} \times \frac{q}{r_2}$$

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Electric Potential and Capacitors

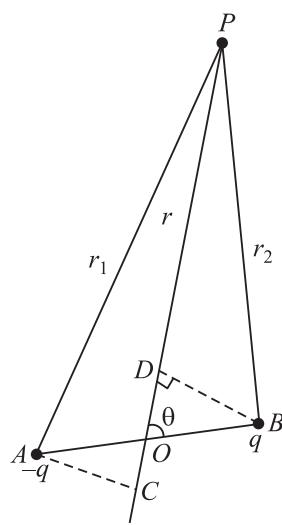
Total potential at P due to both the charges of the dipole is given by

$$V = V_1 + V_2$$

That is,

$$V = \frac{q}{4\pi\epsilon_0} \left[\frac{1}{r_2} - \frac{1}{r_1} \right] \quad (16.12)$$

To put this result in a more convenient form, we draw normals from A and B on the line joining O and P . From ΔBOD , we note that $OD = l \cos \theta$ and from ΔOAC we can write $OC = l \cos \theta$. For a small dipole ($AB \ll OP$), from Fig. 16.4, we can take $PB = PD$ and $PA = PC$. Hence



$$r_1 = r + l \cos \theta$$

$$r_2 = r - l \cos \theta$$

Using these results in Eqn (16.12), we get

$$\begin{aligned} V &= \frac{q}{4\pi\epsilon_0} \left[\frac{1}{(r - l \cos \theta)} - \frac{1}{(r + l \cos \theta)} \right] \\ &= \frac{q}{4\pi\epsilon_0} \left[\frac{(2l \cos \theta)}{(r^2 - l^2 \cos^2 \theta)} \right] \\ &= \frac{q \times 2l \cos \theta}{4\pi\epsilon_0 r^2} \end{aligned}$$

Fig. 16.4 : Electric potential at a point P due to an electric dipole.

where we have neglected the term containing second power of l since $l \ll r$.

In terms of dipole moment ($p = q \times 2l$), we can express this result as

$$V = \frac{p \cos \theta}{4\pi\epsilon_0 r^2} \quad (16.13)$$

This result shows that unlike the potential due to a point charge, the potential due to a dipole is inversely proportional to the square of the distance.

Let us now consider its special cases.

Special Cases

Case I : When point P lies on the axial line of the dipole on the side of positive charge, $\theta = 0$ and $\cos \theta = 1$. Then Eqn. (16.13) reduces to

$$V_{\text{AXIS}} = \frac{p}{4\pi\epsilon_0 r^2} \quad (16.14)$$



Notes

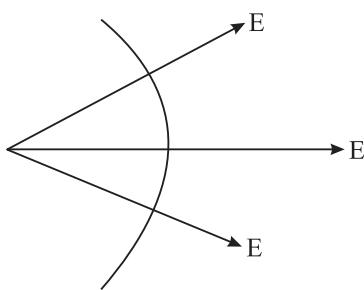
Case II : When point P lies on the axial line of the dipole but on the side of negative charge, $\theta = 180^\circ$ and $\cos \theta = -1$. Hence

$$V_{\text{AXIS}} = -\frac{p}{4\pi\epsilon_0 r^2} \quad (16.15)$$

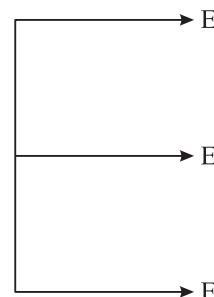
Case III : When point P lies on the equatorial line of the dipole (perpendicular bisector of AB), $\theta = 90^\circ$ and $\cos \theta = 0$. Then

$$V_{\text{equatorial}} = 0 \quad (16.16)$$

That is, electric potential due to a dipole is zero at every point on the equatorial line of the dipole. When a dipole is kept in 3D space, the equatorial line will lie in the plane of the paper. The potential at all points in this plane will be same, i.e. zero. Such a surface is referred to as *equipotential surface*. The electric field is always perpendicular to an equipotential surface. No work is done in moving a charge from one point to another on the equipotential surface.



(a) Spherical equipotential surface



(b) Plane equipotential surface

Fig. 16.5 : Equipotential surfaces and electric field directions

16.1.3 Potential Energy of a System of Point Charges

The electric potential energy is the energy possessed by a system of point charges by virtue of their being in an electric field. When charges are infinite distance apart, they do not interact and their potential energy is zero. If we want to assemble a charge system, i.e. bring charges near each other, work will have to be done. This work is stored in the form of potential energy in the system of these charges. This is called the electric potential energy of the charge system. Hence, we can define *potential energy of a system of point charges as the total amount of work done in bringing various point charges of the system to their respective positions from infinitely large mutual separations.*

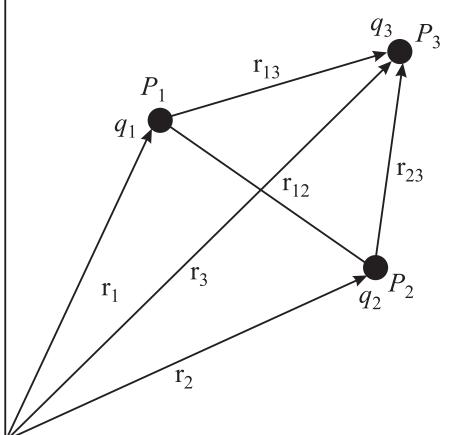


Fig. 16.6 : Potential energy of a system of point charges separated by a distance

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Notes

Electric Potential and Capacitors

Suppose that a point charge q_1 is located at a point P_1 with position vector \mathbf{r}_1 in space. Assume that point charge q_2 is at infinity. This is to be brought to the point P_2 having position vector \mathbf{r}_2 where $P_1P_2 = \mathbf{r}_{12}$, as shown in Fig. 16.6. We know that electric potential at P_2 due to charge q_1 at P_1 is

$$V_2 = \frac{1}{4\pi\epsilon_0} \frac{q_1}{|\mathbf{r}_{12}|} \quad (16.17)$$

From the definition of potential, work done in bringing charge q_2 from infinity to point P_2 is

$$W = (\text{Potential at } P_2) \times \text{value of charge}$$

This work is stored in the system of charges q_1 and q_2 in the form of electric potential energy U . Thus,

$$U = \frac{q_1 \times q_2}{4\pi\epsilon_0 |\mathbf{r}_{12}|} \quad (16.18)$$

In case the two charges have same sign, work is done against the repulsive force to bring them closer and hence, electric potential energy of the system increases. On the other hand, in separating them from one another, work is done by the field. As a result, potential energy of the system decreases. If charges are of opposite sign, i.e. one is positive and the other is negative, the potential energy of the charge system decreases in bringing the charges closer and increases in separating them from one another.

For a three point charge system (Fig. 16.6), Eqn. (16.18) can be written as

$$U = \frac{1}{4\pi\epsilon_0} \left(\frac{q_1 q_2}{\mathbf{r}_{12}} + \frac{q_1 q_3}{\mathbf{r}_{13}} + \frac{q_2 q_3}{\mathbf{r}_{23}} \right) \quad (16.19)$$

Proceeding in the same way, we can calculate the potential energy of a system of any number of charges.

By combining Eqns. (16.3) and (16.13), the potential energy of a dipole in a uniform electric field can be written as

$$U_0 = -pE \cos\theta = -\mathbf{p} \cdot \mathbf{E} \quad (16.20)$$

where \mathbf{p} is the dipole moment in electric field \mathbf{E} and θ is the angle between \mathbf{p} and \mathbf{E} .

16.2 RELATION BETWEEN ELECTRIC FIELD AND POTENTIAL

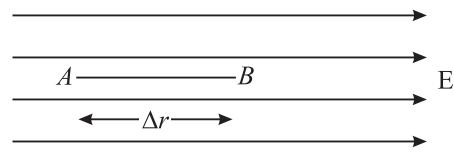
Consider two points A and B in a uniform electric field \mathbf{E} , separated by a small distance Δr . By definition, potential difference ΔV between A and B is equal to the work done in moving a unit positive test charge from A to B :

Electric Potential and Capacitors

$$\begin{aligned}\Delta V &= (\text{Force on unit positive charge}) \times (AB) \\ &= \mathbf{E} \cdot \Delta \mathbf{r} = E(\Delta r) \cos 180^\circ \\ &= -\mathbf{E} \Delta \mathbf{r}\end{aligned}$$

or

$$\mathbf{E} = -\frac{\Delta V}{\Delta r}$$



(16.21)

The negative sign indicates that work is done against the electric field.

Hence, at any point, the electric field is equal to negative rate of change of potential with distance (called **potential gradient**) at that point in the direction of field. Remember that electric potential is a scalar quantity but electric potential gradient is a vector as it is numerically equal to electric field.

From the above relation, for a uniform electric field, we can write

$$E = \frac{V_A - V_B}{d} \quad (16.22)$$

Here V_A and V_B are potentials at points A and B , respectively separated by a distance d .

Example 16.1 : In a 10 volt battery, how much work is done when a positively charged particle having charge $1.6 \times 10^{-19} \text{ C}$ is moved from its negative terminal to the positive terminal?

Solution : According to Eqn. (16.2)

$$V_{AB} = W_{AB} / q_0$$

Since $V_{AB} = 10 \text{ V}$ and $q_0 = 1.6 \times 10^{-19} \text{ C}$, we get

$$\begin{aligned}W_{AB} &= (10 \text{ V}) \times (1.6 \times 10^{-19} \text{ C}) \\ &= 1.6 \times 10^{-18} \text{ J}\end{aligned}$$

Example 16.2 : A point charge q is at the origin of Cartesian co-ordinate system. The electric potential is 400 V and the magnitude of electric field is 150 N C^{-1} at a point x . Calculate x and q .

Solution : The electric field

$$E = \frac{V}{x}$$

On inserting the numerical values, we get

$$150 = \frac{400}{x}$$

or

$$x = 2.67 \text{ m}$$

Recall that electric field is given by the expression

$$E = \frac{1}{4\pi\epsilon_0} \frac{q}{x^2}$$

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We substitute $\frac{1}{4\pi\epsilon_0} = 9 \times 10^9 \text{ N C}^{-2} \text{ m}^2$, $E = 150 \text{ N C}^{-1}$ and $x = 2.67 \text{ m}$ and obtain

$$q = \frac{(150 \text{ N C}^{-1}) \times (2.67 \text{ m})^2}{9 \times 10^9 \text{ N C}^{-2}}$$
$$= 11.9 \times 10^{-8} \text{ C}$$



INTEXT QUESTIONS 16.1

1. A metallic sphere of radius R has a charge $+q$ uniformly distributed on its surface. What is the potential at a point r ($> R$) from the centre of the sphere?
2. Calculate the work done when a point charge is moved in a circle of radius r around a point charge q .
3. The electric potential V is constant in a region. What can you say about the electric field \mathbf{E} in this region?
4. If electric field is zero at a point, will the electric potential be necessarily zero at that point?
5. Can two equipotential surfaces intersect?

On the basis of charge conduction, substances are broadly classified as **conductors** and **insulators**. In solids, conduction of electricity usually takes place due to free electrons, whereas in fluids, it is due to ions. Conductors have free charge carriers through which electric currents can be established on applying an electric field. Metals are good conductors. Substances having no free charge carriers are called **insulators**. The common insulators are wood, ebonite, glass, quartz, mica etc. Substances which have electrical conductivity in between those of conductors and insulators are called **semiconductors**. The ratio of electrical conductivities of good conductors and good insulators is of the order of 10^{20} . Let us now learn how conductors behave in an electric field.

16.2.1 Behaviour of Conductors in an Electric Field

Conductors have electrons which are not bound tightly in their atoms. These are free to move within the conductor. However, there is no net transfer of electrons (charges) from one part of the conductor to the other in the absence of any applied electric field. The conductor is said to be in electrostatic equilibrium.

Refer to Fig. 16.7(a) which shows a conductor placed in an external electric field \mathbf{E} . The free electrons are accelerated in a direction opposite to that of the electric field. This results in build up of electrons on the surface $ABCD$ of the conductor. The surface $FGHK$ becomes positively charged because of removal of electrons. These charges (-ve on surface $ABCD$ and +ve on surface $FGHK$) create their



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own fields, which are in a direction opposite to \mathbf{E} . The transfer of electrons from $FGHK$ to $ABCD$ continues till \mathbf{E} becomes equal to \mathbf{E}_1 . Such a state of electrostatic equilibrium is reached usually in 10^{-16} s. We then say that equilibrium is reached almost instantaneously. If there is a cavity inside a conductor, the electric field inside the cavity is zero (Fig. 16.7(b)).

These results are true for a charged conductor or when charges are induced on a neutral conductor by an external electric field.

This property of a conductor is used in **Electrostatic Shielding** — a phenomenon of protecting a certain region of space from external electric fields. To protect delicate instruments from external electric fields, they are enclosed in hollow conductors. That is why in a thunder storm accompanied by lightning, it is safer to be inside a car or a bus than outside. The metallic body of the car or bus provides electrostatic shielding from lightning.

Conductors in electrostatic equilibrium exhibit the following properties :

- There is no electric field inside a conductor.
- The electric field outside a charged conductor is perpendicular to the surface of the conductor, irrespective of the shape of the conductor.
- Any charge on the conductor resides on the surface of the conductor.

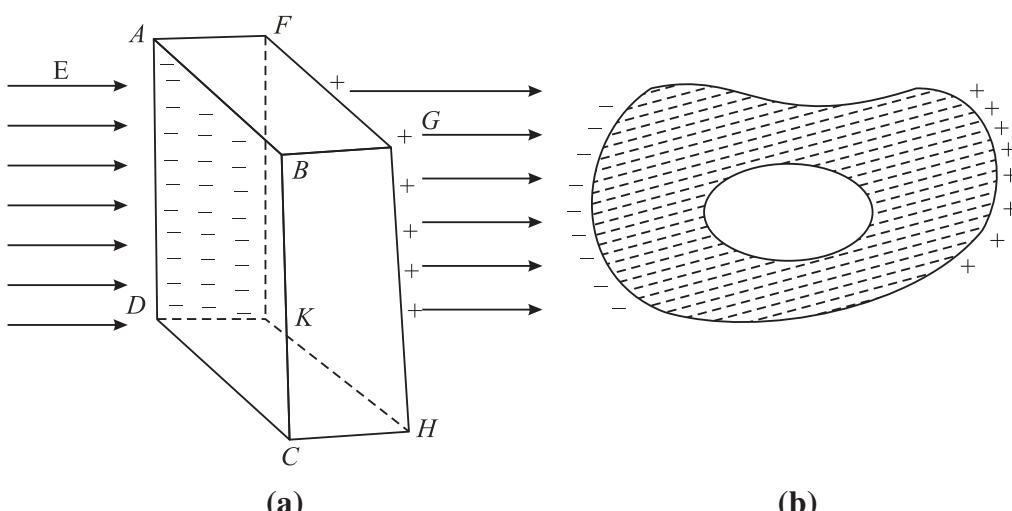


Fig.16.7 : Electrostatic shielding: (a) External electric field \mathbf{E} pulls free electrons on the surface $ABCD$. The surface $FGHK$, which is deficient in electrons, becomes positively charged; the net field inside the conductor is zero. (b) If there is a cavity inside a conductor, the field inside the cavity is zero.

16.3 CAPACITANCE

Let us consider two conductors having equal but opposite charges $+Q$ and $-Q$ on them. There is a potential difference V between them. Such a system of conductors is called a **capacitor**. Experimentally it is found that the potential difference is directly proportional to charge on a conductor. As charge increases,

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the potential difference between them also increases but their ratio remains constant. This ratio is termed as capacitance of the capacitor:

$$C = Q / V \quad (16.23)$$

The capacitance is defined as the ratio between the charge on either of the conductors and the potential difference between them. It is a measure of the capability of a capacitor to store charge.

In SI system of units, capacitance is measured in farad (F). The capacitance is one farad, if a charge of one coulomb creates a potential difference of one volt :

$$1 \text{ farad} = \frac{1 \text{ coulomb}}{1 \text{ volt}} \quad (16.24)$$

You may recall from the previous unit that coulomb is a very large unit of charge. It means that farad is also a very large unit of capacitance. Usually we use capacitors of values in microfarad or picofarad:

$$1 \text{ microfarad} = 10^{-6} \text{ farad}, \text{ written as } \mu\text{F}$$

$$1 \text{ picofarad} = 10^{-12} \text{ farad}, \text{ written as pF}$$

In an electrical circuit, a capacitor is represented by two parallel lines.

16.3.1 Capacitance of a Spherical Conductor

Suppose that a sphere of radius r is given charge q . Let the potential of the sphere be V . Then

$$V = \frac{1}{4\pi\epsilon_0} \frac{q}{r}$$

Since $C = q/V$, we find that

$$C = \frac{q}{q/4\pi\epsilon_0 r} = 4\pi\epsilon_0 r = \frac{r}{9 \times 10^9} \quad (16.25)$$

This shows that capacitance of a spherical conductor is directly proportional to its radius. In fact, it is numerically equal to its radius divided by 9×10^9 , where radius is taken in metre. For example, the capacitance of a sphere of radius 0.18 m is

$$C = \frac{0.18}{9} \times 10^{-9} \text{ F} = 20 \text{ pF}$$

16.3.2 Types of Capacitors

You will come across many types of capacitors in your physics laboratory. The power supply system of your city also uses capacitors. These also form important components of devices such as radio, T.V., amplifiers and oscillators. A capacitor essentially consists of two conductors, one charged and the other usually earthed. To understand the principle of a capacitor, let us consider an insulated metal plate A and give it positive charge (q) till its potential (V) becomes maximum. (Any



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further charge given to it would leak out.) The capacitance of this plate is equal to q/V .

Now bring another insulated metal plate *B* near plate *A*. By induction, negative charge is produced on the nearer face of *B* and equal positive charge develops on its farther face (Fig. 16.8a). The induced negative charge tends to decrease whereas induced positive charge tends to increase the potential of *A*. If plate *B* is earthed (Fig. 16.8b), the induced positive charge on it, being free, flows to earth. (In reality, it is the negative charge that flows from the earth to the plate. Positive charges in the plate are immobile.) But negative charge will stay as it is bound to positive charge on *A*. Due to this induced negative charge on *B*, the potential of *A* decreases and its capacitance increases.

Hence, we can say that *capacitance of an insulated conductor can be increased by bringing near it an uncharged earthed conductor*. This is the basic principle of a capacitor. Capacitors are used for storing large amounts of electric charge and hence electrical energy in a small space for a small interval of time.

A Parallel Plate Capacitor

A parallel plate capacitor is one of the simplest capacitors in which two parallel metallic plates, each of area *A*, are separated from one another by a small distance *d*. An insulating medium like air, paper, mica, glass etc separates the plates. The plates are connected to the terminals of a battery, as shown in Fig. 16.9. Suppose that these plates acquire $+q$ and $-q$ charge when the capacitor is fully charged. These charges set up a uniform electric field **E** between the plates. When the separation *d* is small compared to the size of the plates, distortion of electric field at the boundaries of the plates can be neglected.

If σ is surface charge density on either plate, the magnitude of electric field between the plates is given by

$$E = \frac{\sigma}{\epsilon_0} = \frac{q}{\epsilon_0 A}$$

and the potential difference between the plates is given by

$$V = Ed$$

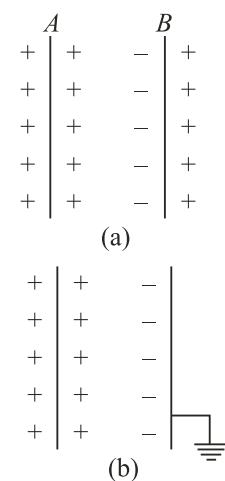


Fig.16.8 : Working principle of a capacitor

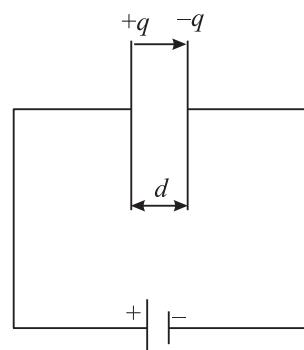


Fig. 16.9 : Working principle of a capacitor

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Hence, capacitance of a parallel plate capacitor, whose plates are separated by d and have air in-between them is given by

$$\begin{aligned} C_0 &= \frac{q}{V} = \frac{q}{qd/\epsilon_0 A} \\ &= \frac{\epsilon_0 A}{d} \end{aligned} \quad (16.26)$$

It shows that capacitance of a parallel plate capacitor is directly proportional to the area of the plates and inversely proportional to their separation. It means that to obtain high capacitance, area of the plates should be large and separation between them should be small.

If the plates of a capacitor are separated by a dielectric material other than air or vacuum, the capacitance of a parallel plate capacitor is given by

$$C = \frac{\epsilon A}{d} = \frac{k \epsilon_0 A}{d}$$

where ϵ is called permittivity of the medium. Therefore, we find that capacitance of a dielectric filled parallel plate capacitor becomes K times the capacitance with air or vacuum as dielectric :

$$C = KC_0 \quad (16.27)$$

16.3.3 Relative Permittivity or Dielectric Constant

We can also define dielectric constant by calculating the force between the charges. According to Coulomb's law, the magnitude of force of interaction between two charges q_1 and q_2 separated by a distance r in vacuum is :

$$F_v = \frac{1}{4\pi\epsilon_0} \frac{q_1 q_2}{r^2} \quad (16.28)$$

where ϵ_0 is the permittivity of free space.

If these charges are held at the same distance in a material medium, the force of interaction between them will be given by

$$F_m = \frac{1}{4\pi\epsilon} \frac{q_1 q_2}{r^2} \quad (16.29)$$

On combining Eqns. (16.28) and (16.29), we get

$$\frac{F_v}{F_m} = \frac{\epsilon}{\epsilon_0} = \epsilon_r \quad (16.30)$$



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where ϵ_r (or K) is relative permittivity. It is also termed as dielectric constant of the medium. Note that it is the ratio of permittivity of the material medium to the permittivity of free space. We can also define the dielectric constant of a medium as the ratio of the electrostatic force of interaction between two point charges held at certain distance apart in air or vacuum to the force of interaction between them held at the same distance apart in the material medium.

The dielectric constant can also be expressed as

$$K = \frac{\text{Capacitance with dielectric between the plates}}{\text{Capacitance with vacuum between the plates}}$$

$$= \frac{C_m}{C_0}$$

Thus

$$C_m = KC_0 \quad (16.31)$$

For metals, $K = \infty$, for mica $K \approx 6$, and for paper $K = 3.6$.

16.4. GROUPING OF CAPACITORS

Capacitors are very important elements of electrical and electronic circuits. We need capacitors of a variety of capacitances for different purposes. Sometimes a capacitance of a proper value may not be available. In such situations, grouping of capacitors helps us to obtain desired (smaller or larger) value of capacitance with available capacitors. Two most common capacitor groupings are :

- Series grouping, and
- Parallel grouping

Let us learn about these now.

16.4.1 Parallel Grouping of Capacitors

In parallel grouping, one plate of each capacitor is connected to one terminal and the other plate is connected to another terminal of a battery, as shown in Fig. 16.10. Let V be the potential difference applied to the combination between points A and B . Note that *in parallel combination, potential difference across each capacitor is the same*. Therefore, charge on these will be different, say q_1 , q_2 and q_3 such that

$$\begin{aligned} q_1 &= C_1 V \\ q_2 &= C_2 V \\ q_3 &= C_3 V \end{aligned} \quad (16.32)$$

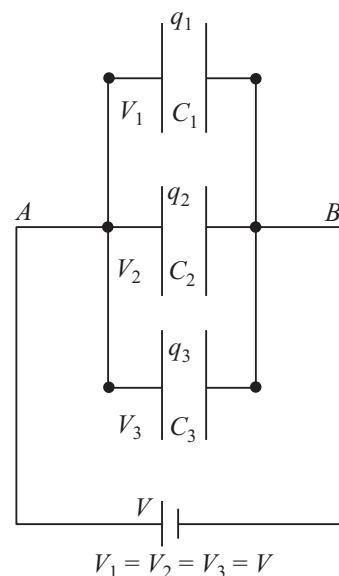


Fig. 16.10 : Capacitors joined in parallel



Total charge on all the capacitors of the combination is :

$$q = q_1 + q_2 + q_3$$

$$q = (C_1 + C_2 + C_3 + \dots)V \quad (16.33)$$

Let C_p be the equivalent capacitance in parallel combination. Then

$$q = C_p V$$

From these relations, we get

$$q = C_p V = (C_1 + C_2 + C_3)V$$

In general, we can write

$$C_p = C_1 + C_2 + C_3 = \sum_{i=1}^n C_i \quad (16.34)$$

Thus, we see that *equivalent capacitance of a number of capacitors joined in parallel is equal to the sum of the individual capacitances.*

Remember that in parallel combination, all the capacitors have the same potential difference between their plates but charge is distributed in proportion to their capacitances. Such a combination is used for charge accumulation.

16.4.2 Series Grouping of Capacitors

In the series combination of capacitors, the first plate of the first capacitor is connected to

the electrical source. The second plate of the first capacitor is connected to the first plate of the second capacitor. The second plate of second capacitor is connected to first plate of the next capacitor of the combination and so on. The second plate of last capacitor of the combination is connected to the electrical source, as shown in Fig.16.11. Let $+q$ unit of charge be given to the first plate of capacitor C_1 from the source. Due to electrical induction, as explained in the principle of capacitor, $-q$ charge appears on the inner side of right plate of C_1 and $+q$ charge develops on the outer side of the second plate of C_1 . The $+q$ unit of charge flows to the first plate of C_2 and so on. Thus, each capacitor receives the same charge of magnitude q . As their capacitances are different, potential difference across these capacitors will be

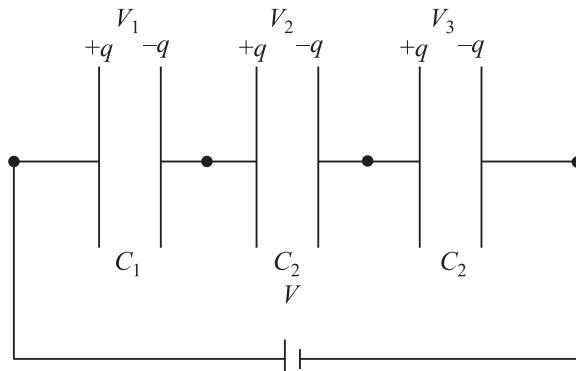


Fig.16.11 : Capacitors in series grouping. The amount of charge on each capacitor plate is same.

$$V_1 = \frac{q}{C_1}, V_2 = \frac{q}{C_2}, V_3 = \frac{q}{C_3} \quad (16.35)$$

If C_s is the total capacitance of the series grouping, then

$$V = \frac{q}{C_s}$$

and

$$V = V_1 + V_2 + V_3 \quad (16.36)$$

Hence

$$\frac{q}{C_s} = \frac{q}{C_1} + \frac{q}{C_2} + \frac{q}{C_3}$$

or

$$\frac{1}{C_s} = \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3} \quad (16.37)$$

For n capacitors joined in series, we can write

$$\frac{1}{C_s} = \sum_{i=1}^n \frac{1}{C_i}$$

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Types of Capacitors

There are three common varieties of capacitors in commercial use. Their schematic diagrams are shown in Fig.16.12.

- 1. Paper capacitor:** Several large thin sheets of paraffin impregnated paper or mylar are cut in proper size (rectangular). Several sheets of metallic foils are also cut to the same size. These are spread one over the other alternately. The outer sheet is mylar, then over it a sheet of metal foil, again over it a sheet of mylar and then a sheet of metal foil and so on. The entire system is then rolled in the form of a cylinder to form a small device.
- 2. Metal plate capacitors:** A large number of metals are alternately joined to two metal rods as shown in Fig.16.12 (b). The entire plate system is immersed in silicon oil which works as dielectric material between the plates. High voltage capacitors are usually of this type. Variable capacitors of micro farad capacitance are usually of this type and use air as dielectric. One set of plates is fixed and the other set is movable. The movable plates, when rotated, change their effective area, thereby changing the capacitance of the system. You might see such capacitors in a radio receiver. Variable capacitance helps in tuning to different radio stations.

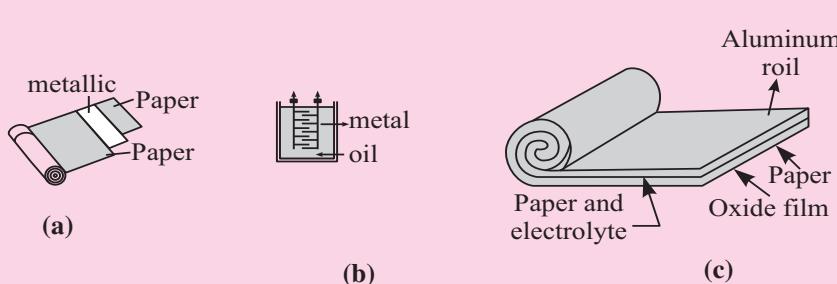


Fig.16.12 : Different types of capacitors : a) paper capacitor, b) variable capacitor, and c) electrolytic capacitor

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3. **Electrolytic capacitor:** An electrolytic capacitor is shown Fig. 16.12(c). A metal foil is rolled in the shape of a cylinder with increasing diameter so that there is always a space between one surface and the other. The system is immersed in an electrolyte in the form of a solution. This solution is conducting because of ions in the solution. A voltage is applied between the electrolyte and the metallic foil. Because of the conducting nature of the electrolyte, a thin layer of metal oxide, which is an insulator, is formed on the foil. The oxide layer works as dielectric material. Since the dielectric layer is extremely thin, the system provides a very high value of capacitance. It is important in this type of capacitor to mark the **positive** and **negative** terminals. A wrong connection of positive and negative terminals removes the oxide layer. (The capacitor then starts conducting.) This type of capacitor is used in storing large amount of charge at low voltage.

Thus, *the reciprocal of equivalent capacitance of any number of capacitors connected in series is equal to the sum of the reciprocals of individual capacitances.* From the above relation, you will agree that C_s is less than the least of C_1 , C_2 , and C_3 .

Note that *all the capacitors in series grouping have the same amount of charge but the potential difference between their plates are inversely proportional to their capacitances.* It means that the capacitor with minimum capacitance of the combination will have maximum potential difference between its plates.

Example 16.3 : The capacitance of a parallel plate air capacitor is $22.0 \mu\text{F}$. The separation between the plates is d . A dielectric slab of thickness $d/2$ is put in-between the plates. Calculate the effective capacitance, if the dielectric constant $K = 5$.

Solution: The Capacitance of the air capacitor is given by

$$C_0 = \frac{\epsilon_0 A}{d} = 22.0 \mu\text{F}$$

The new system can be considered as a series combination of two capacitors:

$$C_1 = \frac{K \epsilon_0 A}{d/2} = \frac{2 K \epsilon_0 A}{d} = 2 K C_0$$

and

$$C_2 = \frac{\epsilon_0 A}{d/2} = \frac{2 \epsilon_0 A}{d} = 2 C_0$$

The effective capacitance C is given by

$$\frac{1}{C} = \frac{1}{C_1} + \frac{1}{C_2}$$

or

$$C = \frac{C_1 C_2}{C_1 + C_2}$$

$$= \frac{2KC_0 \times 2C_0}{2KC_0 + 2C_0}$$

$$= \frac{2KC_0}{K + 1}$$

$$= \frac{10 \times 22 \times 10^{-6} \text{F}}{6}$$

$$= 36.7 \mu\text{F}$$



Notes



INTEXT QUESTIONS 16.2

1. Write the dimensions of capacitance.
2. What is the potential difference between two points separated by a distance d in a uniform electric field \mathbf{E} ?
3. The usual quantities related with an air capacitor are C_0 , E_0 and V_0 . How are these related with C , \mathbf{E} and V of the same capacitor filled with dielectric constant K ?
4. Calculate the area of air filled capacitor plate when the separation between the plates is 50 cm and capacitance is 1.0 μF .

16.4.3 Energy Stored in a Capacitor

The charging of a capacitor can be visualized as if some external agent, say a battery, pulls electrons from the positive plate of a capacitor and transfers them to the negative plate. Some work is done in transferring this charge, which is stored in the capacitor in the form of electrostatic potential energy. This energy is obtained from the battery (stored as chemical energy). When this capacitor is discharged through a resistor, this energy is released in the form of heat.

Let us assume that an uncharged capacitor, when connected to a battery, develops a maximum charge q . The charging takes place slowly. The initial potential difference between the capacitor plates is zero and the final potential difference is V . The average potential difference during the entire process of charging is

$$\frac{0 + V}{2} = \frac{V}{2} = \frac{q}{2C}$$

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The work done during charging is given by

$$W = \text{Charge} \times \text{potential difference}$$

$$= q \frac{q}{2C} = \frac{1}{2} \frac{q^2}{C}$$

Hence potential energy

$$U = \frac{1}{2} qV = \frac{1}{2} \frac{q^2}{C} = \frac{1}{2} CV^2 \quad (16.38)$$

This energy is stored in the electric field between the plates. The stored energy is directly proportional to the capacitance. It also increases as potential difference increases. However, every capacitor can store only a limited amount of energy. An automatic discharge will take place when the potential difference becomes more than its threshold value.

It is dangerous to touch the plates of a charged capacitor. The capacitor may get discharged through your body resulting in an electric shock. Such a shock could be fatal for high value capacitors when fully charged.

16.5 DIELECTRICS AND DIELECTRIC POLARIZATION

We know that dielectrics are insulating materials, which transmit electric effects without conducting. Dielectrics are of two types : **non-polar** and **polar**. We now learn about these.

(a) Non-polar dielectrics

In the molecules of non-polar dielectrics, the centre of positive charge coincides with the centre of negative charge. Each molecule has zero dipole moment in its normal state. These molecules are mostly symmetrical such as nitrogen, oxygen, benzene, methane, CO_2 , etc.

(b) Polar dielectrics

Polar dielectrics have asymmetric shape of the molecules such as water, NH_3 , HCl etc. In such molecules, the centres of positive and negative charges are separated through a definite distance and have finite permanent dipole moment.

When a non-polar dielectric is held in an external electric field, the centre of positive charge in each molecule is pushed in the direction of \mathbf{E} and the centre of negative charge is displaced in the direction opposite to \mathbf{E} . Because of external electric field, centres of positive and negative charges in the non-polar dielectric molecules are separated. Dielectric is then said to be polarized and a tiny dipole moment develops in each molecule. In fact, the force due to external electric field pulling the charge centres apart balances the force of mutual attraction between the centres (i.e. equilibrium is set) and the molecule is said to be polarised. Induced dipole moment \mathbf{p} acquired by the molecule may be written as

$$\mathbf{p} = \alpha \epsilon_0 \mathbf{E}$$



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where α is constant of proportionality and is called atomic/molecular polarizability.

Let us now consider a non-polar slab $ABCD$ placed in an electric field \mathbf{E} maintained between the plates of a capacitor. As shown in Fig.16.13, the dielectric slab gets polarised. The nuclei of dielectric molecules are displaced towards the negative plate and electrons towards the positive plate. Because of polarisation, an electric field \mathbf{E}_p is produced within the dielectric, which is opposite to \mathbf{E} . Hence, due to the presence of a non-polar dielectric, the field between the plates is reduced, i.e. effective electric field in a polarised dielectric is given by

$$\mathbf{E}(\text{effective}) = \mathbf{E} - \mathbf{E}_p \quad (16.39)$$

Thus, the potential difference between the capacitor plates is correspondingly reduced (as $V = Ed$), increasing the value of capacitance of the capacitor (as $C = q/V$).

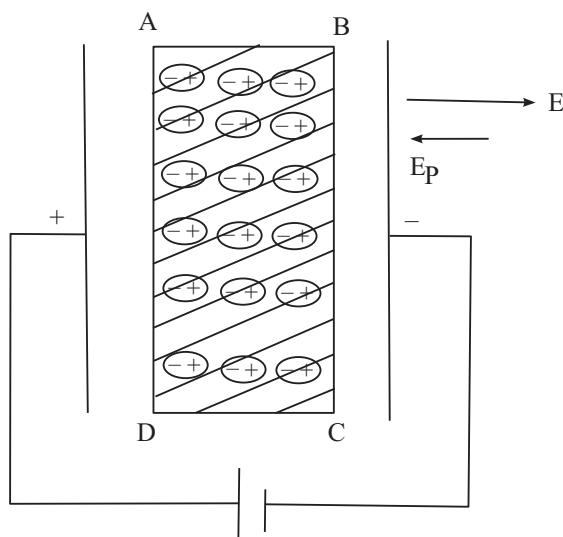


Fig.16.13 : A dielectric slab between the charged capacitor plates.

Applications of Electrostatics

Electrostatics provides basis for the theory of electromagnetics, apart from useful assistance in many fields of science and technology.

- Capacitors are essential parts of most electronic and electrical circuitry. These play a very crucial role in power transmission.
- Gold leaf electroscope – the simple device used for detecting charge, paved the way for cosmic ray research.
- Lightning conductor devised by Benjamin Franklin is still used to protect sky-scrappers from the strokes of lightning and thunder.
- The working of photocopiers, so common these days, is based on the principle of electrostatics.



INTEXT QUESTIONS 16.3

- Two capacitors $C_1 = 12 \text{ mF}$ and $C_2 = 4 \text{ mF}$ are in group connections. Calculate the effective capacitance of the system when they are connected (a) in series (b) in parallel.
- Four capacitors are connected together as shown in Fig.16.14. Calculate the equivalent capacitance of the system.
- An air capacitor $C = 8 \text{ mF}$ is connected to a 12V battery. Calculate
 - the value of Q when it is fully charged?
 - the charge on the plates, when slab of dielectric constant $K = 5$ fills the gap between the plates completely.
 - potential difference between the plates; and
 - capacitance of the new capacitor
- A parallel plate capacitor of capacitance C_0 is connected to a battery and charged to a potential difference V_0 . After disconnecting the battery, the gap between the plates is completely filled with a slab of dielectric constant K . How much energy is stored in the capacitor (a) in the first state? (b) in the second state? and (c) which one is larger and why?

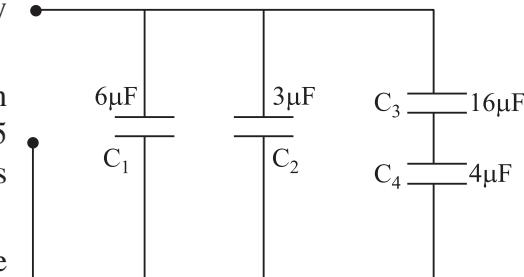


Fig.16.14 : Grouping of capacitors



WHAT YOU HAVE LEARNT

- The potential at any point in an electric field is equal to the work done against the electric field in moving a unit charge from infinity to that point.
- Work done in transferring a charge from one point to another in an electrostatic field is path independent.
- If one joule of work is done in bringing a test charge of one coulomb from infinity to a point in the field, we say that potential at that point is one volt.
- Electric potential due to a dipole is zero at every point on the equatorial line of the dipole.
- In an equipotential surface, every point has same electric potential.
- At any point in an electric field, the negative rate of change of potential with distance (called potential gradient) gives the field.
- Electrostatic shielding is the phenomenon of protecting a region of space from electric field.



Notes

- Capacitance of a conductor depends on its shape, size and nature of medium, rather than its material.
- The capacitance of a dielectric filled parallel plate capacitor becomes K times the capacitance with air or vacuum as dielectric.
- Relative permittivity is the ratio of capacitance with dielectric between the plates to the capacitance with air or vacuum between the plates.
- In series combination of capacitors, the equivalent capacitance is less than the least of any of the individual capacitances.
- In parallel combination of capacitors, the equivalent capacitance is equal to the sum of individual capacitances.
- Due to the presence of a non-polar dielectric, the field between the plates of a capacitor is reduced.



TERMINAL EXERCISES

1. Calculate the potential at a point P at a distance of 30 cm from a point charge $q = 20 \mu C$
2. Three point charges q_1 , q_2 and q_3 , each of magnitude $200 \mu C$, are placed at the corners A , B and C respectively of an equilateral triangle. The length of the side is 10cm. Calculate the potential energy of the system.
3. The potential difference between the plates of a capacitor separated by 3mm is 12.0 V. Calculate the magnitude of E between the plates?
4. Two ions having charges $+e$ and $-e$ are 4.0×10^{-10} m apart. Calculate the potential energy of the system.
5. The plates A and B of a parallel plate capacitor have a potential difference of 15 V. A proton ($m = 1.67 \times 10^{-27}$ kg) is moved from the positive plate A to B . Calculate the speed of the proton near plate B .
6. Show that dimensionally the quantities Vq and $(\frac{1}{2})mv^2$ are equivalent. The symbols carry the usual meaning.
7. Under what condition, the electric field between the plates of a parallel plate capacitor is uniform?
8. A metallic sphere of radius r has a charge $+q$. Calculate the work done in moving a test charge q_0 from one end of a diameter to its other end.
9. A parallel plate air capacitor of value C_0 is charged to a potential V_0 between the plates and $+q_0$ is charge on one plate. Separation between plates is d . A dielectric of dielectric constant $K = 3$ fills the space between the plates. Which of these quantities will change and why. (i) capacitance (ii) charge (iii) potential difference and (iv) field density?

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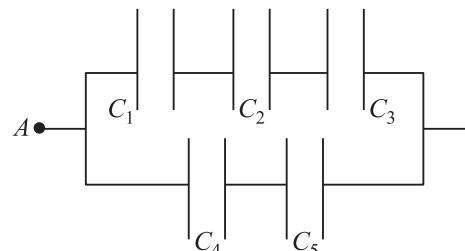
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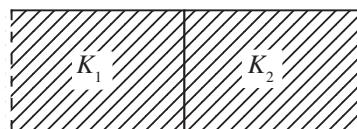
10. Examine the following network of capacitors. The potential difference between A and B is 16V :



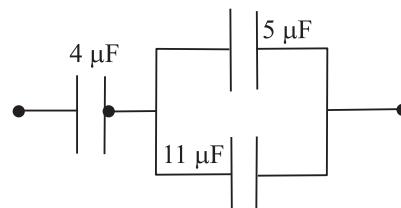
$$C_1 = 2 \mu\text{F}, C_2 = 4 \mu\text{F}, C_3 = 8 \mu\text{F}, C_4 = 3 \mu\text{F}, C_5 = 3 \mu\text{F}$$

Calculate (a) the effective capacitance between A and B, (b) the charge on each capacitor, and (c) the potential difference across each capacitor.

11. The value of capacitance of an air capacitor is $8\mu\text{F}$. Two dielectrics of identical size fill the space between the plates as shown. Dielectric constants are $K_1 = 3.0$ and $K_2 = 6.0$. Calculate the value of the new capacitance.



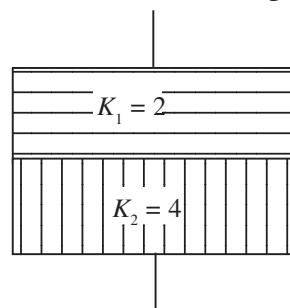
12. Calculate the equivalent capacitance of the following system.



13. A $3.0 \mu\text{F}$ air capacitor is charged to a potential 12.0 V. A slab of dielectric constant $K = 7$ is made to fill the space. Calculate the ratio of the energies stored in the two systems.

14. A dipole of dipole moment $P = 3.5 \times 10^{-15} \text{ Cm}$ is placed in a uniform electric field $E = 2.0 \times 10^4 \text{ NC}^{-1}$. The dipole makes an angle of 60° with the field. Calculate the (a) Potential energy of the dipole and (b) the torque on the dipole.

15. The capacitance of a parallel plate air capacitor is $12\mu\text{F}$. The separation between the plates is 8mm. Two dielectric slabs of the same size fill the air space. Calculate the new value of capacitance.





ANSWERS TO INTEXT QUESTIONS

**Notes****16.1**

1. The potential at r ($r > R$)

$$V = \frac{1}{4\pi\epsilon_0} \frac{q}{r}$$

2. The field around a point charge possesses spherical symmetry. Thus every point on the surface of the sphere is equipotential. And no work is done when a charge moves on an equipotential surface
3. $E = -\frac{dV}{dr}$ Since V is constant, E is zero.

We can obtain the same result using Eqn. (16.22) :

$$E = \frac{V_A - V_B}{d}. \text{ Since } V_A = V_B, E \text{ is zero}$$

4. No. Not necessarily. When $E = 0$, the potential is either constant or zero.
5. Two equipotential surfaces never intersect. If they do so, at the point of intersection we can draw two normals giving directions of electric field.

16.2

$$\begin{aligned} 1. \quad C &= \frac{Q}{V} = \frac{\frac{Q}{\text{Work done}}}{\frac{\text{Charge}}{\text{Work done}}} = \frac{Q \times Q}{\text{Work done}} \\ &= \frac{Q^2}{\text{N.m.}} \end{aligned}$$

The basic unit is

$$A = \frac{C}{s}$$

$$\therefore C^2 = A^2 s^2 \text{ and newton} = \text{mass} \times \text{acc} = \text{kg} \frac{\text{m}}{\text{s}^2}$$

$$\begin{aligned} \text{Capacitance} &= \frac{C^2}{\text{Nm}} = \frac{\frac{\text{A}^2 \text{s}^2}{\text{kg m}^2}}{\frac{\text{m}}{\text{s}^2}} = \frac{\text{A}^2 \text{s}^4}{\text{kg m}^2} \\ &= \text{A}^2 \text{s}^4 (\text{kg m}^2)^{-1} \end{aligned}$$

2. In a capacitor, E is uniform between the plates. Potential difference between the plates

$$V_A - V_B = E \times d.$$

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Notes

Electric Potential and Capacitors

3. C_0, E_0, V_0 for air capacitor and
 C, E, V for dielectric capacitor. Then

$$k = \frac{C}{C_0}, k = \frac{V_0}{V} \quad k = \frac{E_0}{E}.$$

4. $C = 1.0 \mu F = 1.0 \times 10^{-6} F$.
 $d = 50 \text{ cm} = 0.5 \text{ m}$.

$$C = \frac{\epsilon_0 A}{d}$$

$$\therefore A = \frac{Cd}{\epsilon_0}. \text{ Since } \epsilon_0 = 8.85 \times 10^{-12},$$

$$A = \frac{1.0 \times 10^{-6} \times 0.5}{8.85 \times 10^{-12}}$$

$$= \frac{5 \times 10^{-7}}{8.85 \times 10^{-12}}$$

$$= 0.56 \times 10^5 \text{ m}^2$$

16.3

1. (a) 3 mF (b) 16 mF 2. $12.2 \mu F$
3. (a) 96 mC (b) 0.480 C (c) 12 v (d) 40 mF
4. (a) $\frac{1}{2} C_0 V_0^2$ (b) $\frac{1}{2} \frac{(C_0 V_0)^2}{C_0 R} = \frac{1}{2k} C_0 V_0^2$
(c) The energy in the first case is more, because same energy is used up for sucking in the dielectric slab.

Answers to Problems in Terminal Exercises

1. $6 \times 10^5 \text{ V}$. 2. $1.08 \times 10^4 \text{ J}$.
3. $4 \times 10^3 \text{ V m}^{-1}$. 4. $-5.76 \times 10^{-19} \text{ J}$
5. $1.4 \times 10^9 \text{ ms}^{-2}$
10. (a) $\frac{37}{14} \mu F$, (b) $\frac{128}{7} \mu C, \frac{128}{7} \mu C, \frac{128}{7} \mu C, 24 \mu C$,
(c) $\frac{64}{7} \text{ V}, \frac{32}{7} \text{ V}, \frac{16}{7} \text{ V}, 8 \text{ V}, 8 \text{ V}$
11. $36 \mu F$. 12. $\frac{16}{5} \mu F$.
13. 1 : 7 14. (a) $3.5 \times 10^{-11} \text{ J}$ (b) $6 \times 10^{-11} \text{ Nm}$.
15. $32 \mu F$

17

Notes

ELECTRIC CURRENT

In our daily life we use electricity for various activities. The electric lamps and tubes light our houses, we listen music on a tape recorder or radio, see different programmes on television, enjoy cool breeze from electric fan or cooler, and use electric pump to irrigate fields. In fact, electricity is a unique gift of science to mankind. We can not imagine life without electricity in the modern world. At home you might have observed that as soon as you switch on an electric lamp, it starts glowing. Why does it happen? What is the function of a switch?

In the preceding lessons of this module, you have studied about static electric charges and forces between them. In this lesson, you will learn about electric charges in motion. You will also learn that the rate of flow of charge through a conductor depends on the potential difference across it. You will also study the distribution of current in circuits and Kirchhoff's laws which govern it. Elementary idea of primary and secondary cells will also be discussed in this lesson.

Physics is an experimental science and the progress it has made to unfold laws of nature became possible due to our ability to verify theoretical predictions or reproduce experimental results. This has led to continuous improvement in equipment and techniques. In this lesson you will learn about potentiometer, which is a very versatile instrument. It can be used to measure resistance as well as electro-motive force using null method.



OBJECTIVES

After studying this lesson, you should be able to :

- *state Ohm's law and distinguish between ohmic and non-ohmic resistances;*
- *obtain equivalent resistance for a series and parallel combination of resistors;*
- *distinguish between primary and secondary cells;*

MODULE - 5

Electricity and
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Notes

Electric Current

- apply Kirchhoff's rules to closed electrical circuits;
- apply Wheatstone bridge equation to determine an unknown resistance; and
- explain the principle of potentiometer and apply it to measure the e.m.f. and internal resistance of a cell.

Free and Bound Electrons

An atom is electrically neutral, i.e. as many negatively charged electrons revolve around the nucleus in closed orbits as there are positively charged protons inside it. The electrons are bound with the nucleus through Coulomb (attractive) forces.

Farther the electrons from the nucleus, weaker is the Coulomb force. The electrons in the outermost orbit are, therefore, most loosely bound with the nucleus. These are called *valence electrons*. In metallic solids, the valence electrons become free to move when a small potential difference is applied.

17.1 ELECTRIC CURRENT

You have studied in the previous lesson that when a potential difference is applied across a conductor, an electric field is set up within it. The free electrons move in a direction opposite to the field through the conductor. This constitutes an electric current. Conventionally, *the direction of current is taken as the direction in which a positive charge moves*. The electrons move in the opposite direction. To define current precisely, let us assume that the charges are moving perpendicular to a surface of area A , as shown in Fig. 17.1. The current is the rate of flow of charge through a surface area placed perpendicular to the direction of flow. If charge Δq flows in time Δt , the average current is defined as :

$$I_{av} = \frac{\Delta q}{\Delta t} \quad (17.1)$$

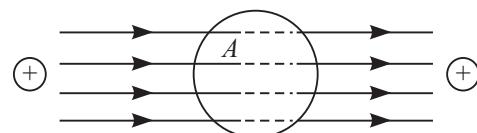


Fig. 17.1 : Motion of charges inside a conductor of surface area A

If the rate of flow of charge varies with time, the current also varies with time. The instantaneous current is expressed as :

$$I = \frac{dq}{dt} \quad (17.2)$$

The electric current through a conductor is the rate of transfer of charge across a surface placed normal to the direction of flow.

**Notes**

The SI unit of current is ampere. Its symbol is A :

$$1 \text{ ampere} = \frac{1 \text{ coulomb}}{1 \text{ second}} \quad (17.3)$$

The smaller units of current are milliampere, $1 \text{ mA} = 10^{-3} \text{ A}$, and microampere, $1 \mu\text{A} = 10^{-6} \text{ A}$. The current can arise due to flow of negative charges (electrons), as in metals. In a semiconductor, flow of electrons (negative charge) and holes constitutes current. Holes are vacancies in a crystal. These are taken as positively charged particles having the same amount of charge as that on an electron. You will study about these particles in more detail in lesson 28.

Let us consider a conductor of cross sectional area A shown in Fig. 17.2. The volume element for a length Δx is $A \Delta x$. If n is the number of electrons per unit volume, the number of electrons in this volume element will be $nA\Delta x$. The total charge in this volume

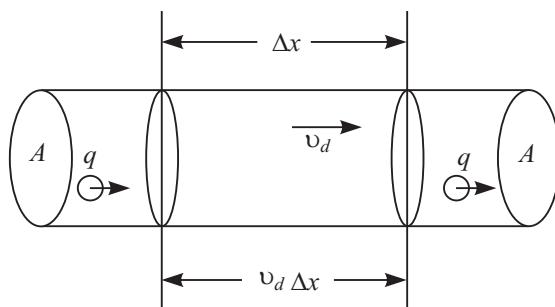


Fig. 17.2 : The charges move with a speed v_d through a surface of area A . The number of charges in a length Δx is $nA v_d \Delta t$.

element is $\Delta q = nA\Delta x e$, where e is charge on the electron. If electrons drift with a speed v_d due to thermal energy, the distance travelled in time Δt is $\Delta x = v_d \Delta t$. On substituting this value of Δx in the expression for Δq , we find that total charge in the volume element under consideration is given by

$$\Delta q = nAe v_d \Delta t$$

so that

$$\frac{\Delta q}{\Delta t} = I = nAev_d \quad (17.4)$$

You will learn more about the drift velocity in sec. 17.9.

17.2 OHM'S LAW

In 1828, Ohm studied the relation between current in a conductor and potential difference applied across it. He expressed this relation in the form of a law, known as **Ohm's law**.



George Simon Ohm (1787-1854)

German physicist, George Simon Ohm is famous for the law named after him. He arrived at the law by considering an analogy between thermal and electrical conduction. He also contributed to theory of sirens, interference of polarised light in crystals etc. Ohm, the practical unit of resistance, is named in his honour.



According to Ohm's law, **the electric current through a conductor is directly proportional to the potential difference across it, provided the physical conditions such as temperature and pressure remain unchanged.**

Let V be the potential difference applied across a conductor and I be the current flowing through it. According to Ohm's law,

$$V \propto I$$

or

$$V = RI$$

$$\Rightarrow \frac{V}{I} = R \quad (17.5)$$

where constant of proportionality R signifies the electrical resistance offered by a conductor to the flow of electric current. *Resistance is the property of a conductor by virtue of which it opposes the flow of current through it.* The $I-V$ graph for a metallic conductor is a straight line (Fig. 17.3(a)).

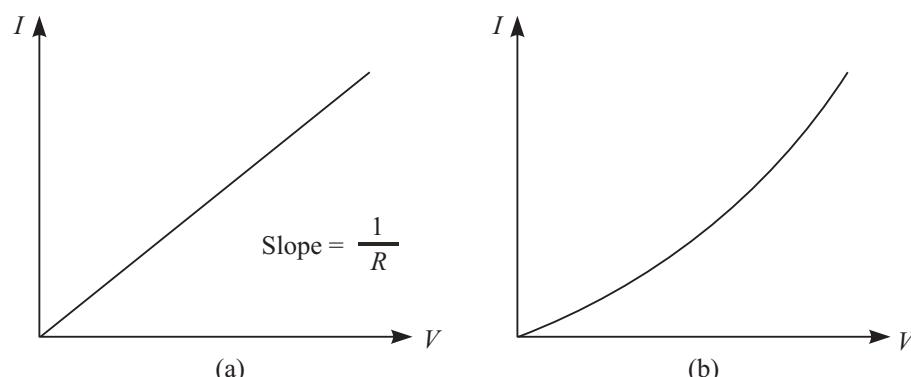


Fig. 17.3 : Current-voltage graph for a) an ohmic device, and b) a semiconductor diode

The SI unit of resistance is ohm. It is expressed by symbol Ω (read as omega)

$$1 \text{ ohm} = 1 \text{ volt}/1 \text{ ampere}$$

Most of the metals obey Ohm's law and the relation between voltage and current is linear. Such resistors are called *ohmic*. Resistors which do not obey Ohm's law are called *non-ohmic*. Devices such as vacuum diode, semiconductor diode, transistors



ACTIVITY 17.1

Aim : To study conduction of electricity through an electrolyte.

Material Required Ammeter, Voltmeter, a jar containing copper sulphate solution, two copper plates, a battery, plug key, connecting wires and a rheostat.

How to Proceed :

1. Set up the apparatus as shown in Fig. 17.4.
2. Plug in the key and note ammeter and voltmeter readings.
3. Change the value of ammeter reading by moving the sliding contact of rheostat and note voltmeter reading again.
4. Repeat step –3 at least five times and record ammeter and voltmeter readings each time.
5. Repeat the experiment by changing (a) separation between P_1 and P_2 , (b) plate area immersed in electrolyte, and (c) concentration of electrolyte.
6. Plot $I-V$ graph in each case.

What do you conclude?

- If $I-V$ graph is a straight line passing through the origin, as shown in Fig. 17.5, we say that ionic solution behaves as an ohmic resistor.
- The slope of the graph changes steeply with change in volume of electrolyte between the plates. It means that resistivity of an electrolyte depends not only on its nature but also on the area of the electrodes and the separation between them.

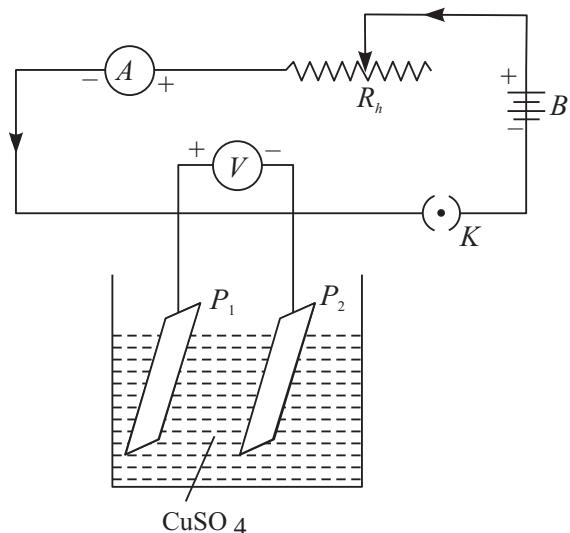


Fig. 17.4 : Electrical conduction through an electrolyte

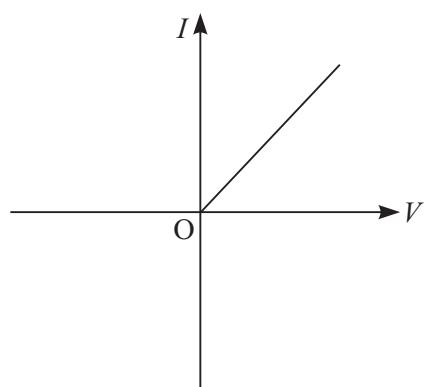


Fig. 17.5 : $I-V$ graph for an ionic solution

Notes





17.2.1 Resistance and Resistivity

Let us now study the factors which affect the resistance of a conductor. You can perform two simple experiments. To do so, set up a circuit as shown in Fig. 17.6.



ACTIVITY 17.2

Take a long conducting wire of uniform cross section. Cut out pieces of different lengths, say l_1, l_2, l_3 , etc from it. This makes sure that wires have same area of cross-section. Connect l_1 between A and B and note down the current through this wire. Let this current be I . Perform the same experiment with wires of lengths l_2 and l_3 , one by one. Let the currents in the wires be I_2 and I_3 respectively.

Plot a graph between l^{-1} and I . You will find that the graph is a straight line and longer wires allow smaller currents to flow. That is, longer wires offer greater resistance [Fig. 17.7(a)]. Mathematically, we express this fact as

$$R \propto l \quad (17.6)$$

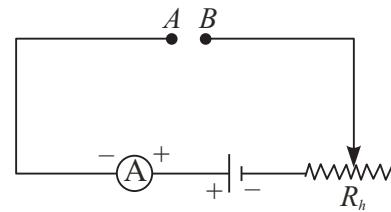


Fig. 17.6 : Electrical circuit to study factors affecting resistance of conductors

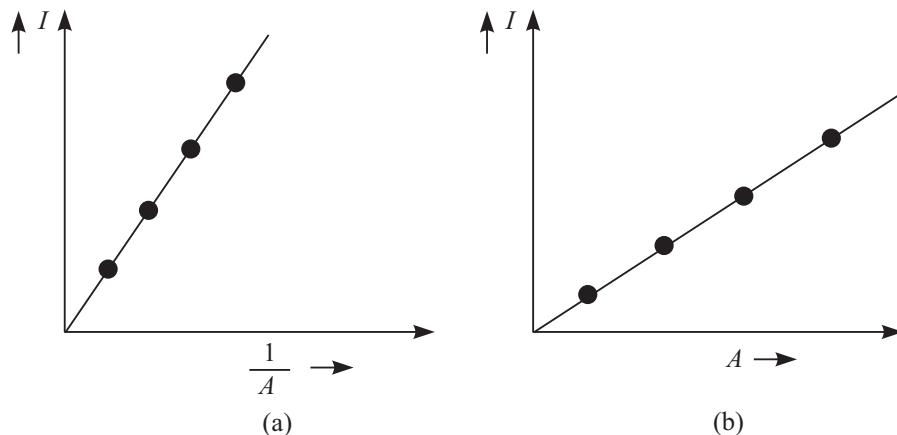


Fig. 17.7 : a) The graph between I and $1/l$ for wires of uniform cross-section and b) the graph between current and area of cross section for wires of same length



ACTIVITY 17.3

Take wires of the same length of a given material but having different areas of cross section, say A_1, A_2, A_3 etc. Connect the wires between A and B one by one and note down the currents I_1, I_2, I_3 etc. in each case. A plot of I and A will give



Notes

a straight line. Wires of greater cross sectional area allow greater currents to flow. You may say that wires of larger area of cross-section offer smaller resistance [Fig. 17.7 (b)]. Mathematically, we can write

$$R \propto \frac{1}{A} \quad (17.7)$$

On combining Eqns.(17.6) and (17.7), we can write

$$R \propto \frac{\ell}{A}$$

or $R = \rho \frac{\ell}{A}$ (17.8)

where ρ is a constant for the material at constant temperature. It is called the **specific resistance** or **resistivity** of the material. By rearranging terms, we can write

$$\rho = \frac{RA}{\ell} \quad (17.9)$$

If $\ell = 1\text{m}$ and $A = 1\text{m}^2$, then $\rho = R$ ohm-metre. Thus *resistivity of a material is the resistance offered by a wire of length one metre and area of cross section one m*².

The unit of resistivity is ohm metre (Ωm)

Reciprocal of resistivity is called conductivity (specific conductance) and is denoted by σ :

$$\sigma = \frac{1}{\rho} \quad (17.10)$$

Unit of **conductivity** is Ohm⁻¹ metre⁻¹ or mho-metre⁻¹ or Sm⁻¹.

Resistivity depends on the nature of the material rather than its dimensions, whereas the resistance of a conductor depends on its dimensions as well as on the nature of its material.

You should now study the following examples carefully.

Example 17.1 : In our homes, the electricity is supplied at 220V. Calculate the resistance of the bulb if the current drawn by it is 0.2A.

Solution :

$$R = \frac{V}{I} = \frac{220 \text{ volt}}{0.2 \text{ amp.}} = 1100 \Omega$$

Example 17.2 : A total of 6.0×10^{16} electrons pass through any cross section of a conducting wire per second. Determine the value of current in the wire.

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Electricity and
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Notes

Electric Current

Solution : Total charge passing through the cross-section in one second is

$$\Delta Q = ne = 6.0 \times 10^{16} \times 1.6 \times 10^{-19} \text{ C} = 9.6 \times 10^{-3} \text{ C}$$

$$\therefore I = \frac{\Delta Q}{\Delta t} = \frac{9.6 \times 10^{-3} \text{ C}}{1 \text{ s}} \\ = 9.6 \times 10^{-3} \text{ A} \\ = 9.6 \text{ mA}$$

Example 17.3 : Two copper wires *A* and *B* have the same length. The diameter of *A* is twice that of *B*. Compare their resistances.

Solution : From Eqn. (17.8) we know that

$$R_A = \rho \frac{\ell}{\pi r_A^2} \text{ and } R_B = \rho \frac{\ell}{\pi r_B^2}$$

$$\therefore \frac{R_A}{R_B} = \frac{r_B^2}{r_A^2}$$

Since diameter of *A* = 2 × diameter of *B*, we have $r_A = 2r_B$. Hence

Resistance of *B* will be four times the resistance of *A*.

Example 17.4 : The length of a conducting wire is 60.0 m and its radius is 0.5 cm. A potential difference of 5.0 V produces a current of 2.5 A in the wire. Calculate the resistivity of the material of the wire.

$$\text{Solution : } R = \frac{V}{I} = \frac{5.0 \text{ V}}{2.5 \text{ A}} = 2.0 \Omega$$

Radius of the wire = 0.5 cm = 5.0×10^{-3} m

Area of cross section $A = \pi R^2 = 3.14 \times (5.0 \times 10^{-3})^2 \text{ m}^2 = 78.5 \times 10^{-6} \text{ m}^2$

$$\therefore \rho = \frac{2.0 \times 78.5 \times 10^{-6} \Omega \text{m}^2}{60.0 \text{ m}} = 2.6 \times 10^{-6} \Omega \text{m}$$



INTEXT QUESTIONS 17.1

1. (a) A current I is established in a copper wire of length ℓ . If the length of the wire is doubled, calculate the current due to the same cell.
(b) What happens to current in an identical copper wire if the area of cross section is decreased to half of the original value?
2. The resistivity of a wire of length l and area of cross section A is $2 \times 10^{-8} \Omega \text{m}$. What will be the resistivity of the same metallic wire of length $2l$ and area of cross section $2A$?



17.3 GROUPING OF RESISTORS

An electrical circuit consists of several components and devices connected together. Some of these are batteries, resistors, capacitors, inductors, diodes, transistors etc. (They are known as circuit elements.) These are classified as resistive and reactive. The most common resistive components are resistors, keys, rheostats, resistance coils, resistance boxes and connecting wires. The reactive components include capacitors, inductors and transformers. In addition to many other functions performed by these elements individually or collectively, they control the current in the circuit. In the preceding lesson you learnt how grouping of capacitors can be used for controlling charge and voltage. Let us now discuss the role of combination of resistors in controlling current and voltage.

Two types of groupings of resistors are in common use. These are : **series grouping and parallel grouping.** We define equivalent resistance of the combination as a single resistance which allows the same current to flow as the given combination when the same potential difference is applied across it.

17.3.1 Series Combination

You may connect many resistors in series by joining them end-to-end such that the same current passes through all the resistors. In Fig. 17.8, two resistors of resistances R_1 and R_2 are connected in series. The combination is connected to a battery at the ends A and D . Suppose that current I flows through the series combination when it is connected to a battery of voltage V . Potential differences V_1 and V_2 develop across R_1 and R_2 , respectively. Then $V_1 = IR_1$ and $V_2 = IR_2$. But sum of V_1 and V_2 is equal to V , i.e.

$$\Rightarrow V = V_1 + V_2 = IR_1 + IR_2$$

If equivalent resistance of this series combination is R , then

$$V = IR = I(R_1 + R_2)$$

so that

$$R = R_1 + R_2$$

This arrangement may be extended for any number of resistors to obtain

$$R = R_1 + R_2 + R_3 + R_4 + \dots \quad (17.11)$$

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Notes

Electric Current

That is, the equivalent resistance of a series combination of resistors is equal to the sum of individual resistances. If we wish to apply a voltage across a resistor (say electric lamp) less than that provided by the constant voltage supply source, we should connect another resistor (lamp) in series with it.

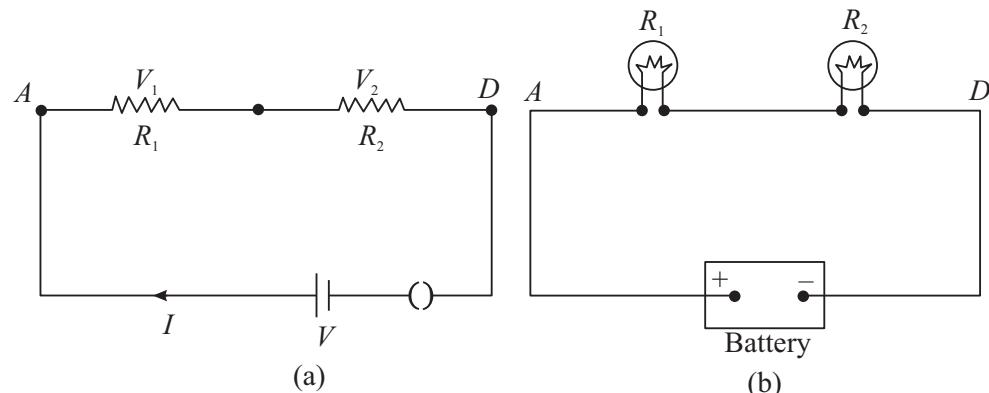


Fig. 17.8 : a) Two resistors connected in series to a battery, and b) two lamps joined in series connected to a dc source.

17.3.2 Parallel Combination

You may connect the resistors in parallel by joining their one end at one point and the other ends at another point. In parallel combination, **same potential difference exists across all resistors**. Fig. 17.9 shows a parallel combination of two resistors R_1 and R_2 . Let the combination be connected to a battery of voltage V and draw a current I from the source.

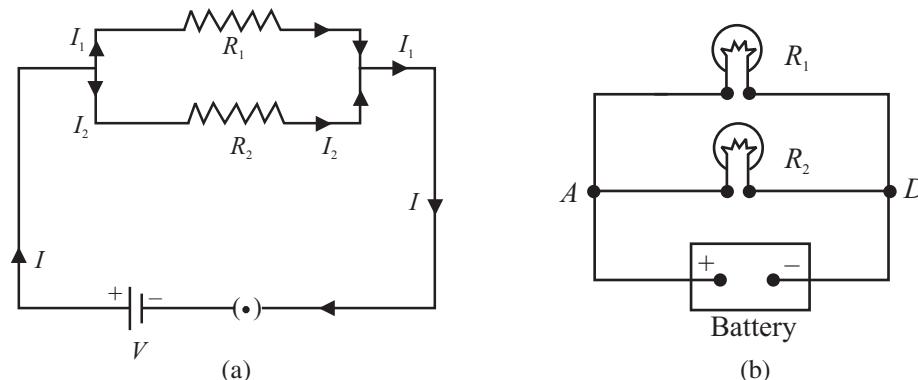


Fig. 17.9 : a) Two resistors connected in parallel. The battery supplies the same voltage to both resistors, and b) lamps connected in parallel to a battery.

The main current divides into two parts. Let I_1 and I_2 be the currents flowing through resistors R_1 and R_2 , respectively. Then $I_1 = V/R_1$ and $I_2 = V/R_2$.

The main current is the sum of I_1 and I_2 . Therefore, we can write

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



Notes

If the equivalent resistance of combination is R , we write $V = IR$ or $I = V/R$:

$$I = \frac{V}{R} = \frac{V}{R_1} + \frac{V}{R_2}$$

$$\therefore \frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} \quad (17.12a)$$

or $R = \frac{R_1 R_2}{R_1 + R_2} \quad (17.12b)$

From Eqn. (17.12a) we note that **reciprocal of equivalent resistance of parallel combination is equal to the sum of the reciprocals of individual resistances.** The process may be extended for any number of resistors, so that

$$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} + \frac{1}{R_4} + \dots \quad (17.13)$$

Note that the equivalent resistance of parallel combination is smaller than the smallest individual resistance. You may easily see this fact by a simple electrical circuit having a resistor of $2\ \Omega$ connected across a $2V$ battery. It will draw a current of one ampere. When another resistor of $2\ \Omega$ is connected in parallel, it will also draw the same current. That is, total current drawn from the battery is $2A$. Hence, resistance of the circuit is halved. As we increase the number of resistors in parallel, the resistance of the circuit decreases and the current drawn from the battery goes on increasing.

In our homes, electrical appliances such as lamps, fans, heaters etc. are connected in parallel and each has a separate switch. Potential difference across each remains the same and their working is not influenced by others. As we switch on bulbs and fans, the resistance of the electrical circuit of the house decreases and the current drawn from the mains goes on increasing (Fig.17.10).

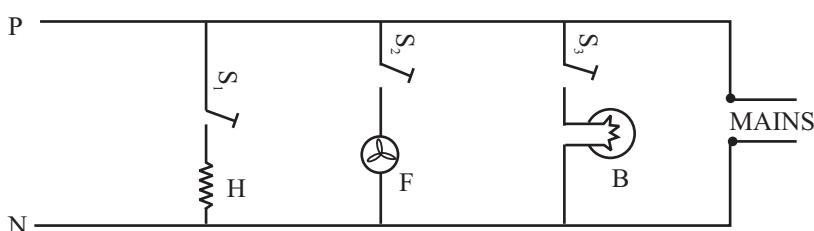


Fig. 17.10 : Arrangement of appliances in our homes. These are connected in parallel so that every appliance is connected to 220 V main supply. The total current drawn from the mains is the sum of the currents drawn by each appliance.

Example 17.5: For the circuit shown in Fig. 17.11, calculate the value of resistance R_2 , and current I_2 flowing through it.

Solution: If the equivalent resistance of parallel combination of R_1 and R_2 is R , then

$$R = \frac{R_1 R_2}{R_1 + R_2} = \frac{10 R_2}{10 + R_2}$$

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Notes

Electric Current

According to Ohm's law,

$$R = \frac{50}{10} = 5\Omega$$

$$\therefore \frac{10 R_2}{10 + R_2} = 5$$

$$\Rightarrow 10 R_2 = 50 + 5 R_2 \text{ or } R_2 = 10 \Omega$$

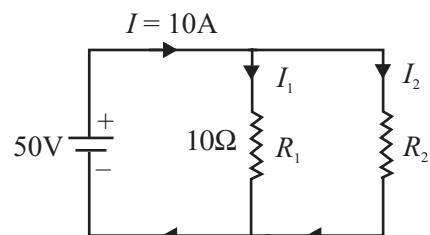


Fig. 17.11 : Two resistors in parallel

Since R_1 and R_2 are equal, current will be equally divided between them. Hence, $I_2 = 5A$

Example 17.6 : For the circuit shown in Fig. 17.12, calculate the equivalent resistance between points a and d .

Solution : 15Ω and 3Ω resistors are connected in parallel. The equivalent resistance of this combination is

$$R_1 = \frac{15 \times 3}{15 + 3} \Omega = \frac{45}{18} = \frac{5}{2} = 2.5\Omega$$

Now we can regard the resistances 5Ω , $R_1 = 2.5\Omega$ and 7Ω as connected in series. Hence, equivalent resistance between points a and d is

$$R = (5 + 2.5 + 7) = 14.5 \Omega$$

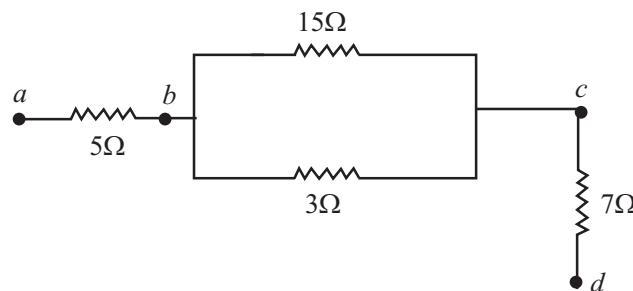


Fig. 17.12: A combination of series and parallel groupings

Example 17.7 : Refer to the network shown in Fig. 17.13. Calculate the equivalent resistance between the points (i) b and c (ii) c and d , and (iii) a and e .

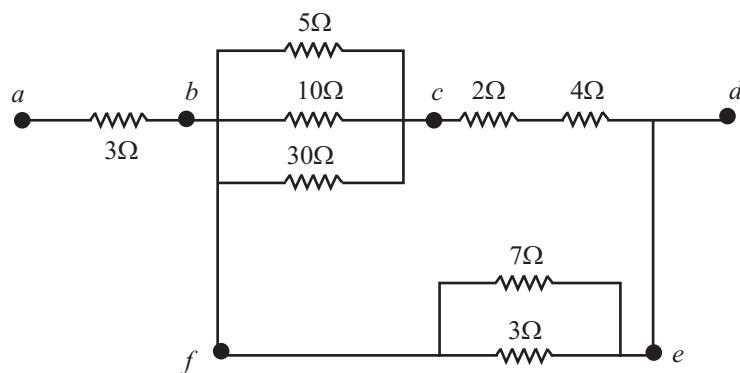


Fig. 17.13 : A combination of series and parallel groupings of resistors



Notes

Solution :

(i) Three resistors (5Ω , 10Ω and 30Ω) are connected in parallel. Therefore, equivalent resistance is given by

$$\frac{1}{R_1} = \frac{1}{5} + \frac{1}{10} + \frac{1}{30} = \frac{6+3+1}{30} = \frac{10}{30} \Omega$$

or

$$R_1 = 3 \Omega$$

(ii) The resistors with resistances 2Ω and 4Ω are in series. The equivalent resistance

$$R_2 = (2 + 4) = 6\Omega$$

(iii) The resistances 7Ω and 3Ω are in parallel. So equivalent resistance

$$\frac{1}{R_3} = \left(\frac{1}{7} + \frac{1}{3} \right) = \frac{3+7}{21} = \frac{10}{21}$$

or,

$$R_3 = \frac{21}{10} \Omega = 2.1 \Omega$$

Now we can treat equivalent resistance R_1 and R_2 to be in series. Therefore

$$R_4 = R_1 + R_2 = (3 + 6) = 9\Omega$$

Now R_4 and R_3 are in parallel. Therefore equivalent resistance

$$\begin{aligned}\frac{1}{R_5} &= \frac{1}{R_4} + \frac{1}{R_3} \\ &= \frac{1}{9} + \frac{1}{2.1} \\ &= \frac{1}{9} + \frac{10}{21} = \frac{37}{63} \\ R_5 &= \frac{63}{37} \Omega = 1.70 \Omega\end{aligned}$$

(iv) Finally R_5 and 3Ω (between a and b) are in series. Hence

$$R = (1.70 + 3) = 4.79 \Omega$$

Note : For ease and convenience, you should draw a new equivalent circuit after every calculation.

**INTEXT QUESTIONS 17.2**

- There are two bulbs and a fan in your bed room. Are these connected in series or in parallel? Why?

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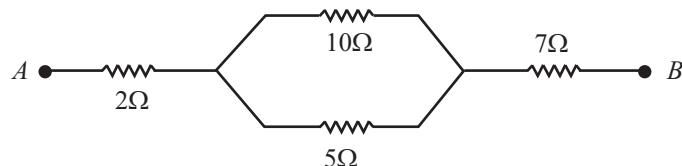
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2. The electric supply in a town is usually at 220 V. Sometimes the voltage shoots upto 300 V and may harm your T V set and other gadgets. What simple precaution can be taken to save your appliances?
3. Calculate the equivalent resistance between points A and B for the following circuit :



17.4 TYPES OF RESISTORS

We use resistors in all electrical and electronic circuits to control the magnitude of current. Resistors usually are of two types :

- carbon resistors
- wire wound resistors

In a wire wound resistor, a resistance wire (of manganin, constantan or nichrome) of definite length, which depends on the required value of resistance, is wound two-fold over an insulating cylinder to make it non-inductive. In carbon resistors, carbon with a suitable binding agent is molded into a cylinder. Wire leads are attached to the cylinder for making connections to electrical circuits. Resistors are colour coded to give their values :

$$R = AB \times 10^C \Omega, D$$

where A, B and C are coloured stripes. The values of different colours are given in Table 17.1. As may be noted,

- first two colours indicate the first two digits of the resistance value;
- third colour gives the power of ten for the multiplier of the value of the resistance; and
- fourth colour (the last one) gives the tolerance of the resistance, which is 5% for golden colour, 10% for silver colour and 20% for body colour.

Table 17.1 : Colour codes of resistors

| Colour | Number | Multiplier |
|--------|--------|------------|
| Black | 0 | 1 |
| Brown | 1 | 10^1 |
| Red | 2 | 10^2 |
| Orange | 3 | 10^3 |



Suppose that four colours on a resistor are Blue, Grey, Green and Silver. Then

The first digit will be 6 (blue)

The second digit will be 8 (Grey)

The third colour signifies multiplier 10^5 (Green)

The fourth colour defines tolerance = 10% (Silver)

Hence value of the resistance is

$$\begin{aligned} & 68 \times 10^5 \pm 10\% \\ & = 68 \times 10^5 \pm (68 \times 10^5 \times 10/100) \\ & = 68 \times 10^5 \pm 68 \times 10^4 \\ & = (6.8 \pm 0.68) \text{ M}\Omega \end{aligned}$$

17.5 TEMPERATURE DEPENDENCE OF RESISTANCE

The resistivity of a conductor depends on temperature. For most metals, the resistivity increases with temperature and the change is linear over a limited range of temperature :

$$\rho = \rho_0 [1 + \alpha (T - T_0)] \quad (17.14)$$

where ρ and ρ_0 are the resistivities at temperatures T and T_0 , respectively. The temperatures are taken in $^{\circ}\text{C}$ and T_0 is the reference temperature. α is called the temperature co-efficient of resistivity. Its unit is per degree celcius.

Superconductors

Temperature dependence of resistivity led scientists to study the behaviour of materials at very low temperatures. They observed that certain metals and their alloys lost their resistivity completely below a certain temperature, called **transition temperature**, which is specific to the material. In such materials, current, once set up, remained, unchanged for ever without the use of an external source to maintain it. Such materials were termed as **superconductors**.

It was soon realised that superconductors, if they may exist near room temperature, will bring in revolutionary changes in technology. (These have

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been termed as *high temperature superconductors*.) For example, energy efficient powerful electromagnets made of superconducting coils may levitate vehicles above a magnetic track and make a high speed transportation system possible.

Efforts are being made to develop high temperature superconductors. The work done so far suggests that oxides of copper, barium and ytterium are showing good possibilities. A superconductor ($T_2\text{Ba}_2\text{Ca}_2\text{Cu}_3\text{O}_{10}$) which can exist at -153°C has been developed. India is a front runner in this area of research.

Eqn. (17.14) can be rearranged to obtain an expression for temperature coefficient of resistivity :

$$\rho = \rho_0 + \rho_0 \alpha (T - T_0)$$

or

$$\alpha = \frac{(\rho - \rho_0)}{\rho_0 (T - T_0)} = \frac{1}{\rho_0} \frac{\Delta \rho}{\Delta T}$$

where $\Delta \rho = (\rho - \rho_0)$ and $\Delta T = T - T_0$.

The resistivity versus temperature graph for a metal like copper is shown in Fig. 17.14(a). The curve is linear over a wide range of temperatures.

You may recall that resistance of a conductor is proportional to its resistivity. Therefore, temperature variation of resistance can written as :

$$R = R_0 [1 + \alpha (T - T_0)] \quad (17.15)$$

The resistances corresponding to two different temperatures T_1 and T_2 are given by

$$R_1 = R_0 [1 + \alpha (T_1 - T_0)] \quad (17.16)$$

and

$$R_2 = R_0 [1 + \alpha (T_2 - T_0)] \quad (17.17)$$

On combining these equations, we can write an expression for temperature coefficient of resistivity :

$$\alpha = \frac{(R_2 - R_1)}{R_0(T_2 - T_1)} = \frac{1}{R_0} \frac{\Delta R}{\Delta T} \quad (17.18)$$

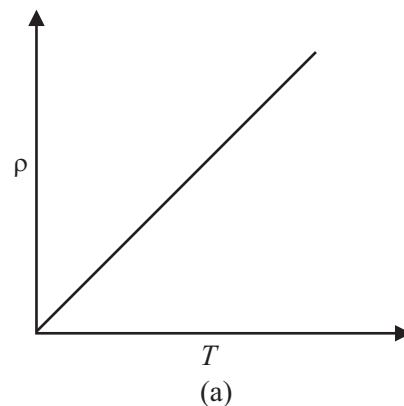


Fig. 17.14 : Typical resistivity–temperature graph for a metal



Notes

If $R_0 = 1\Omega$ and $(T_2 - T_1) = 1^\circ\text{C}$, then $\alpha = (R_2 - R_1)$. Thus **temperature coefficient of resistance is numerically equal to the change in resistance of a wire of resistance 1Ω at 0°C when the temperature changes by 1°C .** This property of metals is used in making resistance thermometers.

The resistivity of alloys also increases with increase in temperature. But the increase is very small compared to that for metals. For alloys such as **manganin**, **constantan** and **nichrome**, the temperature coefficient of resistivity is vanishingly small ($\sim 10^{-6} \text{ }^\circ\text{C}^{-1}$) and resistivity is high. That is why these materials are used for making resistance wires or standard resistances.

Semiconductors such as germanium and silicon have resistivities which lie between those of metals and insulators.

The resistivity of semiconductors usually decreases with increase in temperature [Fig.17.14(b)]. This gives a negative temperature coefficient of resistance. This will be discussed in detail in the lesson on semiconductors.

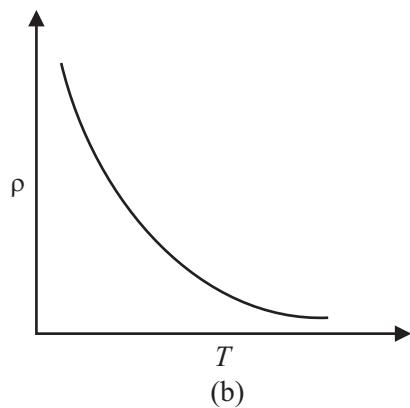


Fig. 17.14(b) : Resistivity of semiconductors decreases with temperature

17.6 ELECTROMOTIVE FORCE (EMF) AND POTENTIAL DIFFERENCE

EMF is the short form of electromotive force. EMF of a cell or battery equals the potential difference between its terminals when these are not connected (open circuit) externally. You may easily understand the difference between e.m.f. and potential difference of a cell by performing the following activity.



ACTIVITY 17.4

Connect a cell in a circuit having a resistor R and key K . A voltmeter of very high resistance is connected in parallel to the cell, as shown in Fig.17.15. When key K is closed, voltmeter reading will decrease. Can you give reasons for this decrease in the voltmeter reading? Actually when key K is open, no current flows through the loop having cell and voltmeter: (The resistance in the circuit is infinite.) Hence the voltmeter reading gives e.m.f. E of the cell, which is the potential difference between the terminals of the cell when no current is drawn from it. When key K is closed, current flows outside and inside the cell. The cell

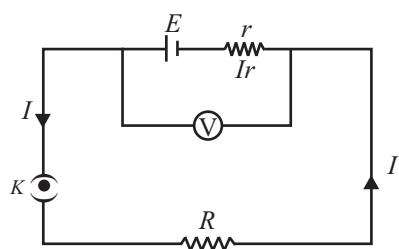


Fig. 17. 15

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introduces a resistance r , called **internal resistance** of the cell. Let current I be flowing in the circuit. Potential drop Ir across internal resistance r due to current flow acts opposite to the e.m.f. of the cell. Hence, the voltmeter reading will be

$$E - Ir = V$$

or

$$E = V + Ir \quad \{17.19\}$$

Thus while drawing current from a cell, e.m.f. of the cell is always greater than the potential difference across external resistance, unless internal resistance is zero.

E.M.F. of a cell depends on :

- the electrolyte used in the cell;
- the material of the electrodes; and
- the temperature of the cell.

Note that the e.m.f. of a cell does not depend on the size of the cell, i.e. on the area of plates and distance between them. This means that if you have two cells of different sizes, one big and one small, the e.m.f.s can be the same if the material of electrodes and electrolyte are the same. However, cells of larger size will offer higher resistance to the passage of current through it but can be used for a longer time.

Example 17.8 : When the current drawn from a battery is 0.5A, potential difference at the terminals is 20V. And when current drawn from it is 2.0A, its voltage reduces to 16V. Calculate the e.m.f. and internal resistance of the battery.

Solution : Let E and r be the e.m.f. and internal resistance of battery. When current I is drawn from it, the potential drop across internal resistance of the cell is Ir . Then we can write

$$V = E - Ir$$

For $I = 0.5\text{A}$ and $V = 20$ volt, we have

$$20 = E - 0.5 r \quad (i)$$

For $I = 2.0\text{A}$ and $V = 16$ volt, we can write

$$16 = E - 2r \quad (ii)$$

We can rewrite Eqns. (i) and (ii) as

$$2E - r = 40$$

and $E - 2r = 16$

Solving these, we get

$$E = 21.3 \text{ V} \text{ and } r = 2.67\Omega$$



17.6.1 Elementary Idea of Primary and Secondary cells

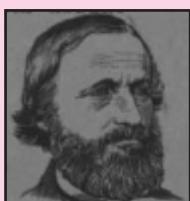
We have seen that to pass electric current through a conductor continuously we have to maintain a potential difference between its ends. For the purpose, generally, we use a device called chemical cell.

Chemical cells are of two types :

- (i) **Primary Cells** : In these cells, the chemical energy is directly converted into electrical energy. The material of a primary cell is consumed as we use the cell and, therefore, it cannot be recharged and reused. Dry cell, Daniel Cell, Voltaic Cell etc are examples of primary cells.
- (ii) **Secondary Cells** : These are chemical cells in which electrical energy is stored as a reversible chemical reaction. When current is drawn from the cells the chemical reaction runs in the reverse direction and the original substances are obtained. These cells, therefore, can be charged again and again. Acid-accumulator, the type of battery we use in our inverter or car, is a set of secondary cells.

17.7 KIRCHHOFF'S RULES

You now know that Ohm's law gives current–voltage relation for resistive circuits. But when the circuit is complicated, it is difficult to know current distribution by Ohm's law. In 1842, Kirchhoff formulated two rules which enable us to know the distribution of current in complicated electrical circuits or electrical networks.



Gustav Robert Kirchhoff
(1824-1887)

The fundamental contributions of German physicist Kirchhoff were in the fields of black body radiation and spectroscopy. But he also contributed in many other fields. His rules that you will study in this lesson enable us to analyse complex electric networks. With the help of Bunsen spectrum analysis, he discovered elements Rubidium and Cesium.

- (i) **Kirchhoff's First Rule (Junction Rule)** : It states that *the sum of all currents directed towards a junction (point) in an electrical network is equal to the sum of all the currents directed away from the junction.*

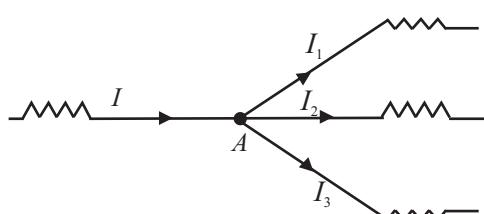


Fig. 17.16 : Kirchhoff's first rule : Sum of currents coming to a junction is equal to the sum of currents going away from it.

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Refer to Fig. 17.16. If we take currents approaching point A as positive and those leaving it as negative, then we can write

$$I = I_1 + I_2 + I_3 \quad \text{or} \quad I - (I_1 + I_2 + I_3) = 0 \quad (17.20)$$

In other words, the algebraic sum of all currents at a junction is zero.

Kirchhoff's first rule tells us that there is no accumulation of charge at any point if steady current flows in it. The net charge coming towards a point should be equal to that going away from it in the same time. In a way, it is an extension of continuity theorem in electrical circuits.

(ii) Kirchhoff's Second Rule (Loop Rule) : This rule is an application of law of conservation of energy for electrical circuits. It tells us that *the algebraic sum of the products of the currents and resistances in any closed loop of an electrical network is equal to the algebraic sum of electromotive forces acting in the loop.*

While using this rule, we start from a point on the loop and go along the loop either clockwise or anticlockwise to reach the same point again. The product of current and resistance is taken as positive when we traverse in the direction of current. The e.m.f is taken positive when we traverse from negative to positive electrode through the cell. Mathematically, we can write

$$\sum IR = \sum E \quad (17.21)$$

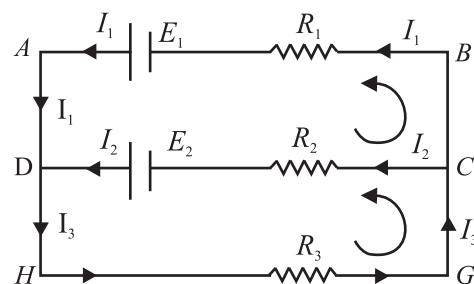


Fig. 17.17 : A network to illustrate Kirchhoff's second rule

Let us consider the electrical network shown in Fig. 17.17. For closed mesh ADCBA, we can write

$$I_1 R_1 - I_2 R_2 = E_1 - E_2$$

Similarly, for the mesh DHGCD

$$I_2 R_2 + (I_1 + I_2) R_3 = E_2$$

And for mesh AHGBA

$$I_1 R_1 + I_3 R_3 = E_1$$

At point D

$$I_1 + I_2 = I_3$$



Notes

In more general form, Kirchhoff's second rule is stated as : *The algebraic sum of all the potential differences along a closed loop in a circuit is zero.*

Example 17.9 : Consider the network shown in Fig. 17.18. Current is supplied to the network by two batteries. Calculate the values of currents I_1 , I_2 and I_3 . The directions of the currents are as indicated by the arrows.

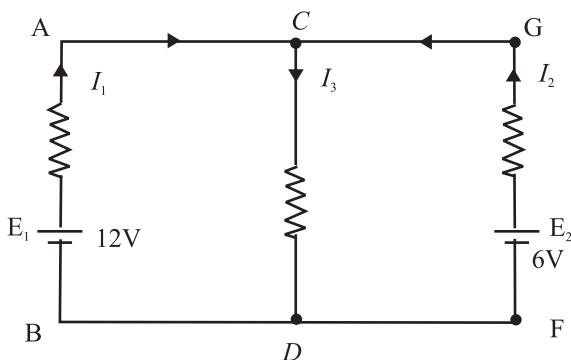


Fig. 17.18 : Calculation of currents in a network of resistors and batteries.

Solution : Applying Kirchhoff's first rule to junction C, we get

$$I_1 + I_2 - I_3 = 0 \quad (\text{i})$$

Applying Kirchhoff's second rule to the closed loops $ACDBA$ and $GCDFG$, we get

$$5I_1 + 2I_3 = 12 \quad (\text{ii})$$

and $3I_2 + 2I_3 = 6 \quad (\text{iii})$

On combining these equations, we get

$$5I_1 - 3I_2 = 6 \quad (\text{iv})$$

Multiply (i) by 2 and add to (ii) to obtain

$$7I_1 + 2I_2 = 12 \quad (\text{v})$$

On multiplying Eqn. (iv) by 2 and Eqn. (v) by 3 and adding them, we get

$$31I_1 = 48$$

or $I_1 = 1.548\text{A}$

Putting this value of I_1 in eqn. (v), we get

$$I_2 = 0.582\text{A}$$

And from (i), we get

$$I_3 = I_1 + I_2 = 2.13\text{A}$$



17.7.1 Wheatstone Bridge

You have learnt that a resistance can be measured by Ohm's law using a voltmeter and an ammeter in an electrical circuit. But this measurement may not be accurate for low resistances. To overcome this difficulty, we use a wheatstone bridge. It is an arrangement of four resistances which can be used to measure one of them in terms of the other three.

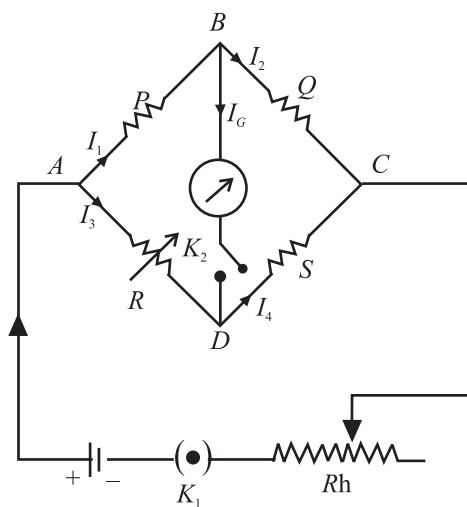


Fig. 17.19 : A wheatstone bridge.

Consider the circuit shown in Fig. 17.19 where

- P and Q are two adjustable resistances connected in arms AB and BC .
- R is an adjustable known resistance.
- S is an unknown resistance to be measured.
- A sensitive galvanometer G along with a key K_2 is connected in the arm BD .
- A battery E along with a key K_1 is connected in the arm AC .

On closing the keys, in general, some current will flow through the galvanometer and you will see a deflection in the galvanometer. It indicates that there is some potential difference between points B and D . We now consider the following three possibilities:

- Point B is at a higher potential than point D :** Current will flow from B towards D and the galvanometer will show a deflection in one direction, say right
- Point B is at a lower potential than point D :** Current will flow from point D towards B and the galvanometer will show a deflection in the opposite direction.



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(iii) **Both points B and D are at the same potential:** In this case, no current will flow through the galvanometer and it will show no deflection, i.e. the galvanometer is in null condition. In this condition, the Wheatstone bridge is said to be in the *state of balance*.

The points B and D will be at the same potential only when the potential drop across P is equal to that across R . Thus

$$I_1 P = I_3 R \quad (17.22)$$

But

$$I_1 = I_2 + I_G$$

and

$$I_4 = I_3 + I_G \quad (17.23)$$

Applying Kirchhoff's first rule at junctions B and D in the null condition ($I_G = 0$), we get

$$I_1 = I_2$$

and

$$I_3 = I_4 \quad (17.24)$$

Also potential drop across Q will be equal to that across S . Hence

$$I_2 Q = I_4 S \quad (17.25)$$

Dividing Eqn. (17.22) by Eqn. (17.25), we obtain

$$\frac{I_1 P}{I_2 Q} = \frac{I_3 R}{I_4 S} \quad (17.26)$$

Using Eqn. (17.24), we get

$$\frac{P}{Q} = \frac{R}{S} \quad (17.27)$$

This is the condition for which a Wheatstone bridge will be balanced. From Eqn. (17.27), we find that the unknown resistance S is given by

$$S = \frac{QR}{P}$$

You can easily see that measurement of resistance by Wheatstone bridge method has the following merits.

(i) *The balance condition given by Eqn. (17.27) at null position is independent of the applied voltage V . In other words, even if you change the e.m.f of the cell, the balance condition will not change.*

(ii) *The measurement of resistance does not depend on the accuracy of calibration of the galvanometer. Galvanometer is used only as a null indicator (current detector).*



The main factor affecting the accuracy of measurement by Wheatstone bridge is its sensitivity with which the changes in the null condition can be detected. It has been found that the bridge has the greatest sensitivity when the resistances in all the arms are nearly equal.

Example 17.9: Calculate the value of R shown in Fig. 17.20. when there is no current in 50Ω resistor.

Solution: This is Wheatstone bridge where galvanometer has been replaced by 50Ω resistor. The bridge is balanced because there is no current in 50Ω resistor. Hence,

$$\frac{20}{10} = \frac{40}{R}$$

or $R = \frac{40 \times 10}{20} = 20\Omega$

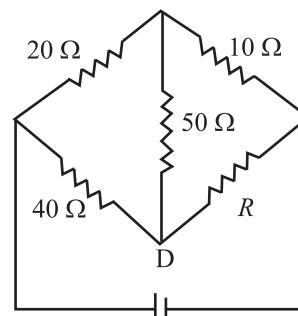
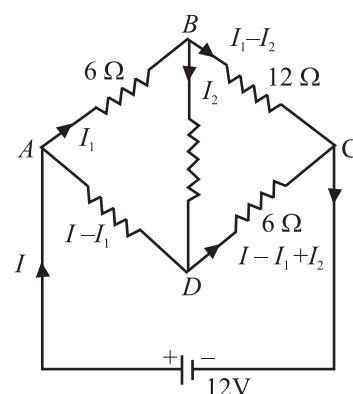


Fig. 17.20 : When there is no current through 50Ω resistor, the bridge is balanced.

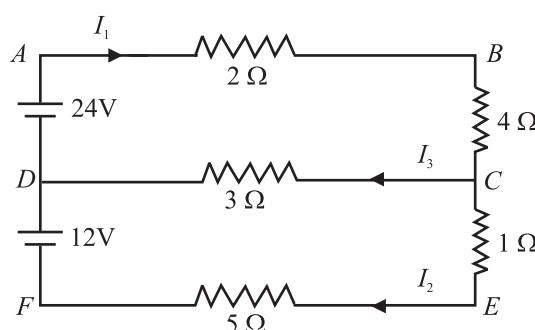


INTEXT QUESTIONS 17.3

1. Refere to figure below. Calculate the value of currents in the arms AB , AD and BD .



2. Examine the following circuit containing resistors and batteries. Calculate the current I_1 , I_2 and I_3 ,





Notes

17.8 POTENTIOMETER

You now know how to measure e.m.f. of a source or potential difference across a circuit element using a voltmeter. (An ideal voltmeter should have infinite resistance so that it does not draw any current when connected across a source of e.m.f.) Practically it is not possible to manufacture a voltmeter which will not draw any current. To overcome this difficulty, we use a potentiometer, which draws no current from it. It employs a null method. The potentiometer can also be used for measurement of internal resistance of a cell, the current flowing in a circuit and comparison of resistances.

17.8.1 Description of a Potentiometer

A potentiometer consists of a wooden board on which a number of resistance wires (usually ten) of uniform cross-sectional area are stretched parallel to each other. The wire is of manganin or nichrome. These wires are joined in series by thick copper strips. In this way, these wires together act as a single wire of length equal to the sum of the lengths of all the wires. The end terminals of the wires are provided with connecting screws.

A metre scale is fixed on the wooden board parallel to wires. A jockey (a sliding contact maker) is provided with the arrangement. It makes a knife edge contact at any desired point on a wire. Jockey has a pointer which moves over the scale. It determines the position of the knife edge contact. In Fig. 17.21 a ten wire potentiometer is shown. A and B are ends of the wire. K is a jockey and S is a scale. Jockey slides over a rod CD.

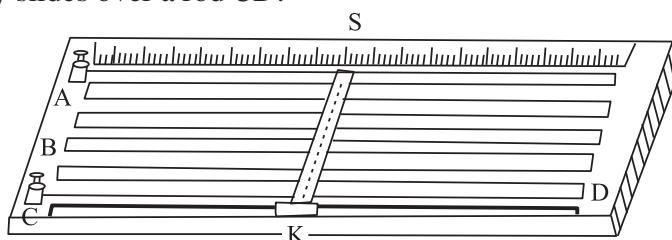


Fig. 17.21 : An illustrative diagram of a potentiometer

17.8.2 Measurements with a Potentiometer

Let us suppose that a steady source of e.m.f. E (say an accumulator) is connected across a uniform wire AB of length l . Positive terminal of the accumulator is connected at end A (Fig. 17.22). A steady current I flows through the wire. The potential difference across AB is given by

$$V_{AB} = RI$$

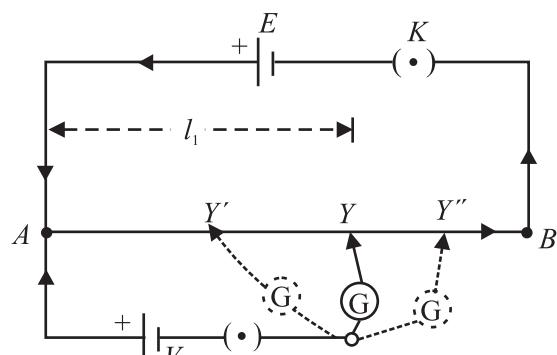


Fig. 17.22 : Potentiometer circuit to measure potential difference between the terminals of a cell.

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If r is the resistance per unit length of the wire, and k is the potential drop across unit length of the wire, then

$$R = r\ell$$

and

$$E = k \ell$$

or

$$k = \frac{E}{\ell}$$

For length ℓ_1 of wire, potential drop is given by

$$V_1 = k\ell_1 = \frac{E}{\ell} \ell_1 \quad (17.28)$$

Thus potential falls linearly with distance along the wire from the positive to the negative end.

We wish to measure an unknown voltage V . The positive terminal of the cell is connected to end A of the wire and negative terminal through a galvanometer to the jockey having variable contact Y . Note that for $V > E$, it will not be possible to obtain a null point. So we use a standard cell of emf $E (> V)$, as shown in Fig.17.22. To check this, insert keys K and K_1 and tap at ends A and B . The galvanometer should show deflection in opposite directions. If so, all is well with the circuit.

Insert key K_1 and start moving jockey from A towards B . Suppose that at position Y' potential drop across the length AY of the wire is less than voltage V . The current in the loop $AY'XA$ due to voltage V exceeds the current due to potential difference across AY' . Hence galvanometer shows some deflection in one direction. Then jockey is moved away, say to Y'' such that potential drop across AY'' is greater than the voltage V . If galvanometer shows deflection in the other direction, the voltage drop across AY'' is greater than that across AY' . Therefore, the jockey is moved slowly between Y' and Y'' . A stage is reached, say at point Y , where potential drop across AY is equal to voltage V . Then points X and Y will be at the same potential and hence the galvanometer will not show any deflection, i.e. null point is achieved. If l_1 is the length between A and Y , then

$$V = k\ell_1 = \frac{E\ell_1}{\ell} \quad (17.29)$$

Thus, the unknown voltage V is measured when no current is drawn

The measurements with potentiometer have following advantages :

- When the potentiometer is balanced, no current is drawn from the circuit on which the measurement is being made.
- It produces no change in conditions in a circuit to which it is connected.
- It makes use of null method for the measurement and the galvanometer used need not be calibrated.



Notes

17.8.3 Comparison of E.M.Fs of two Cells

You have learnt to measure the e.m.f. of a cell using a potentiometer. We shall now extend the same technique for comparison of e.m.fs of two cells. Let us take, for example, a Daniel cell and a Leclanche cell and let E_1 and E_2 be their respective e.m.fs.

Refer to circuit diagram shown in Fig.17.23. The cell of e.m.f. E_1 is connected in the circuit through terminals 1 and 3 of key K_1 . The balance point is obtained by moving the jockey on the potentiometer wire as explained earlier. Note that e.m.f of cell E should be greater than the emfs of E_1 and E_2 separately. (Otherwise, balance point will not be obtained.) Let the balance point on potentiometer be at point Y_1 and length $AY_1 = l_1$. The cell of e.m.f. E_2 is connected in the circuit through terminals 2 and 3 of the key K_2 . Suppose balance is obtained at point Y_2 and length $AY_2 = l_2$.

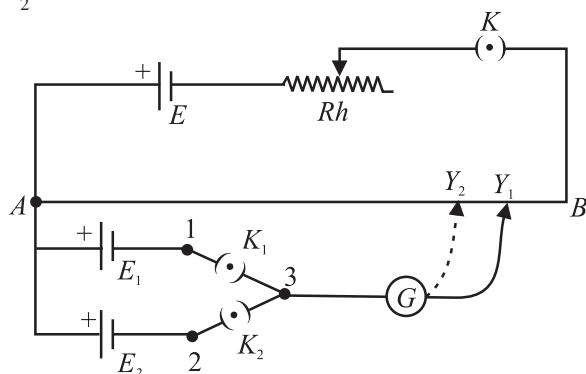


Fig. 17.23 : Circuit diagram for comparison of e.m.fs of two cells E_1 and E_2 .

Applying potentiometer principle, we can write

$$E_1 = kl_1 \text{ and } E_2 = kl_2$$

where k is the potential gradient along the wire AB . Hence

$$\frac{E_1}{E_2} = \frac{l_1}{l_2} \quad (17.30)$$

17.8.4 Determination of Internal Resistance of a Cell

You have learnt that cells always offer resistance to the flow of current through them, which is often very small. This resistance is called the internal resistance of the cell and depends on the size of the cell, i.e. the area of the plates immersed in the liquid, the distance between the plates and strength of electrolyte used in the cell.

Let us now learn how to measure internal resistance of a cell using a potentiometer. Refer to Fig. 17.24, which shows the circuit diagram for measuring internal resistance ' r ' of a cell of emf E_1 . A resistance box R with a key K_1 is connected in parallel with the cell. The primary circuit has a standard cell, a rheostat and a one way key K . As soon as key K is closed, a current I begins to flow through the wire

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AB . The key K_1 is kept open and on moving the jockey, a balance is obtained with the cell E_1 at point, say Y_1 . Let $AY_1 = l_1$. Then we can write

$$E_1 = kl_1 \quad (17.31a)$$

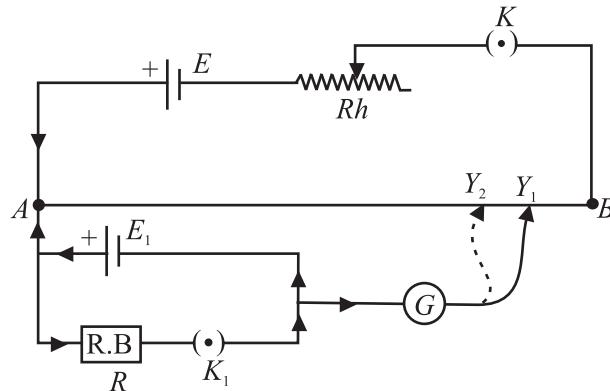


Fig. 17.24 : Measurement of the internal resistance r of a cell

Now key K_1 is closed. This introduces a resistance across the cell. A current, say I_1 , flows in the loop $E_1RK_1E_1$ due to cell E_1 . Using Ohm's law, we can write

$$I_1 = \frac{E_1}{R+r}$$

where r is internal resistance of the cell. It means that terminal potential difference V_1 of the cell will be less than E_1 by an amount I_1r . The value of V_1 is

$$V_1 = I_1R = \frac{E_1}{R+r} R$$

Then, potential difference V_1 is balanced on the potentiometer wire without change in current I . Let the balance point be at point Y_2 such that $AY_2 = l_2$. Then

$$V_1 = kl_2 \quad (17.31b)$$

Using Eqns. (17.31a,b) we get

$$\frac{E_1}{V_1} = \frac{l_1}{l_2} = \frac{R+r}{R}$$

$$\text{or } r = R \left(\frac{l_1}{l_2} - 1 \right) \quad (17.32)$$

Thus by knowing R_1 , l_1 and l_2 , the value of r can be easily calculated.

Example 17.10 : Length of a potentiometer wire is 5 m. It is connected with a battery of fixed e.m.f. Null point is obtained for the Daniel cell at 100 cm. If the length of the wire is kept 7 m, what will be the position of null point?



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Solution: Let e.m.f. of battery be E volt. The potential gradient for 5 m length is

$$k_1 = \frac{E}{5} \text{ Vm}^{-1}$$

When the length of potentiometer wire is 7 m, potential gradient is

$$k_2 = \frac{E}{7} \text{ Vm}^{-1}$$

Now, if null point is obtained at length l_2 , then

$$E_1 = k_2 l_2 = \frac{E}{7} l_2$$

Here same cell is used in two arrangements. Hence

$$\frac{E}{5} = \frac{E}{7} l_2$$

$$\Rightarrow l_2 = 7 / 5 = 1.4 \text{ m}$$

17.9 DRIFT VELOCITY OF ELECTRONS

Let us now understand the microscopic picture of electrical conduction in a metal. The model presented here is simple but its strength lies in the fact that it conforms to Ohm's law.

We assume that a metallic solid consists of atoms arranged in a regular fashion. Each atom usually contributes free electrons, also called conduction electrons. These electrons are free to move in the metal in a random manner, almost the same way as atoms or molecules of a gas move about freely in the a container. It is for this reason that sometimes conduction electrons are referred to as **electron gas**. The average speed of conduction electrons is about 10^6 ms^{-1} .

We know that no current flows through a conductor in the absence of an electric field, because the **average velocity** of free electrons is zero. On an average, the number of electrons moving in $+x$ direction is same as number of electrons moving in $-x$ direction. There is no net flow of charge in any direction.

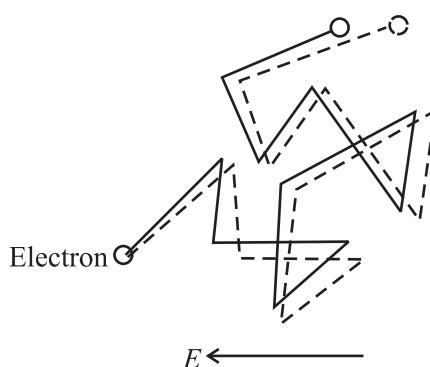


Fig. 17.25 : Motion of electrons in a conductor placed in an electric field.

The conduction electrons frequently collide with the atoms in the solid. The free electrons drift slowly in a direction opposite to the direction of the applied electric field. The average drift velocity is of the order of 10^{-4} ms^{-1} . This is very small compared to the average speed of free electrons between two successive collisions (10^6 ms^{-1}). On applying an electric field, the conduction electrons get accelerated.

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The excess energy gained by the electrons is lost during collisions with the atoms. The atoms gain energy and vibrate more vigorously. The conductor gets heated up. Fig. 17.25 shows how the motion of electrons is modified when an electric field is applied is applied.

Let us now obtain an expression for the drift velocity of conduction electrons. Let e and m be the charge and mass respectively of an electron. If E is the electric field, the force on the electron is eE . Hence acceleration experienced by the electron is given by

$$\mathbf{a} = \frac{e\mathbf{E}}{m}$$

If τ is the average time between collisions, we can write the expression for velocity of drifting electrons in terms of electric field as

$$\mathbf{v}_d = \frac{e\mathbf{E}}{m}\tau$$

On combining this result with Eqn. (17.4), we obtain the expression for current :

$$\begin{aligned} I &= -neAv_d \\ &= -neA \frac{eE}{m}\tau \\ &= -\frac{Ane^2E}{m}\tau \end{aligned}$$

Since electric field is negative spatial gradient of potential $\left(E = -\frac{\partial V}{\partial r} \right)$ we can rewrite the expression for current as

$$I = +\frac{ne^2 A}{m} \frac{V}{\ell} \tau \quad (17.33)$$

$$\Rightarrow \frac{V}{I} = \frac{m}{ne^2 \tau} \frac{\ell}{A} = R \quad (17.34)$$

Eqn. (17.34) implies that conduction current obeys Ohm's law.

On combining this result with Eqn. (17.9), we get

$$\rho = \frac{1}{\sigma} = \frac{m}{ne^2 \tau} \quad (17.35)$$

17.10 POWER CONSUMED IN AN ELECTRICAL CIRCUIT

Let us examine the circuit in Fig. 17.26 where a battery is connected to an external resistor R . The positive charges (so to say) flow in the direction of the current in the resistor and from negative to positive terminal inside the battery. The potential



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difference between two points gives kinetic energy to the charges. These moving charges collide with the atoms (ions) in the resistor and thus lose a part of their kinetic energy. This energy increases with the temperature of the resistor. The loss of energy by moving charges is made up at the expense of chemical energy of the battery.

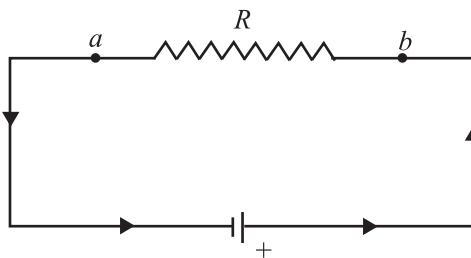


Fig. 17.26 : A circuit containing a battery and a resistor. The power consumed depends on the potential difference between the points *a* and *b*, the current through the resistor.

The rate of loss of potential energy by moving charge ΔQ in going through the resistor is

$$\frac{\Delta U}{\Delta t} = V \frac{\Delta Q}{\Delta t} = VI \quad (17.36)$$

where I is the current in the circuit and V is potential difference between the ends of the resistor.

It is assumed that the resistance of the connecting wires is negligible. The total loss is in the resistor R only. Rate of loss of energy is defined as power :

$$P = VI$$

Since $V = IR$, we can write

$$P = I^2 R = V^2 / R \quad (17.37)$$

The SI unit of power is watt (W).

The electrical power lost in a conductor as heat is called *joule heat*. The heat produced is proportional to : (i) square of current (I), (ii) resistance of conductor (R), and (iii) time for which current is passed (t).

The statement $Q = I^2 Rt$, is called Joule's law for heating effect of current.

Example: 17.11 : A 60W lamp is connected to 220V electricity supply in your home. Calculate the power consumed by it, the resistance of its filament and the current through it.

Solution : We know that $I = P/V$

$$\therefore I = \frac{60W}{220V} = \frac{3}{11} A = 0.27A$$



Resistance of the lamp

$$\begin{aligned}
 R &= \frac{V}{I} \\
 &= \frac{220V}{3/11A} \\
 &= \frac{220 \times 11}{3} \Omega = 807\Omega
 \end{aligned}$$

The lamp consumes 60J of energy per second. It will consume 60 Wh energy in one hour and $60 \times 24 = 1440$ Wh energy in one day.

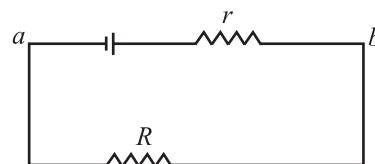
Energy consumed per day = 1.440 kWh

In common man's language, it is known as 1.4 unit of energy.



INTEXT QUESTIONS 17.4

- When current drawn from a cell increases, the potential difference between the cell electrodes decreases. Why?
- A metallic wire has a resistance of 30Ω at $20^\circ C$ and 30.16Ω at $40^\circ C$. Calculate the temperature coefficient of resistance.
- The e.m.f of a cell is 5.0 V and R in the circuit is 4.5Ω . If the potential difference between the points a and b is 3.0 V, calculate the internal resistance r of the cell.



- In a potentiometer circuit, balance point is obtained at 45 cm from end A when an unknown e.m.f is measured. The balance point shifts to 30 cm from this end when a cell of 1.02 V is put in the circuit. Standard cell E always supplies a constant current. Calculate the value of unknown e.m.f.
- A potentiometer circuit is used to compare the e.m.f. of two cells E_1 and E_2 . The balance point is obtained at lengths 30 cm and 45 cm, respectively for E_1 and E_2 . What is the e.m.f. of E_1 , if E_2 is 3.0 V?
- A current of 0.30 A flows through a resistance of 500Ω . How much power is lost in the resistor?
- You have two electric lamps. The printed specifications on them are 40W, 220V and 100W, 220 V. Calculate the current and resistance of each lamp when put in a circuit of 220 V supply line.



WHAT YOU HAVE LEARNT

- Drift velocity is the average velocity with which electrons move opposite to the field when an electric field exists in a conductor.
- Electric current through any cross-sectional area is the rate of transfer of charge from one side to other side of the area. Unit of current is ampere and is denoted by A.
- Ohm's law states that the current flowing through a conductor is proportional to the potential difference when physical conditions like pressure and temperature remain unchanged.
- Ratio V/I is called resistance and is denoted by R . Unit of resistance is ohm (denoted by Ω)
- Resistivity (or specific resistance) of a material equals the resistance of a wire of the material of one metre length and one m^2 area of cross section. Unit of resistivity is ohm metre.
- For a series combination of resistors, the equivalent resistance is sum of resistances of all resistors.
- For a parallel combination of resistors, inverse of equivalent resistance is equal to the sum of inverses of all the resistances.
- Primary cells cannot be recharged and reused, whereas, secondary cells can be charged again and again.
- Kirchhoff's rules help us to study systematically the complicated electrical circuits. The first rule states that the sum of all the currents directed towards a point in an electrical network is equal to the sum of all currents directed away from the point. Rule II : The algebraic sum of all potential differences along a closed loop in an electrical network is zero.
- The Wheatstone bridge circuit is used to measure accurately an unknown resistance (S) by comparing it with known resistances (P, Q and R). In the balance condition,
$$P/Q = R/S.$$
- The e.m.f. of a cell is equal to the potential difference between its terminals when a circuit is not connected to it.
- A potentiometer measures voltages without drawing current. Therefore, it can be used to measure e.m.f. of a source that has appreciable internal resistance.
- Drift velocity of electrons in a conductor is given by $v_d = -\frac{eE}{m}\tau$.
- Power consumed in an electrical circuit through Joule heating is given by

$$P = VI = I^2R = \frac{V^2}{R}.$$

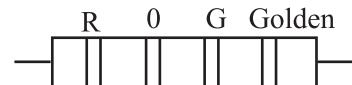


Notes



TERMINAL EXERCISES

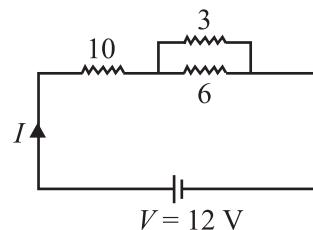
1. Explain the drift of free electrons in a metallic conductor under external electric field. Derive an expression for drift velocity.
2. Define electric current and discuss Ohm's law.
3. Define resistivity of a conductor. How does the resistance of a wire depend on the resistivity of its material, its length and area of cross-section?
4. Define electrical conductivity. Write its unit. How does electrical conductivity depend on free electron concentration of the conductor?
5. Explain the difference between ohmic and non-ohmic resistances. Give some examples of non-ohmic resistances.
6. What is the effect of temperature on the resistivity of a material? Why does electrical conductivity of a metal decrease with increase in temperature?
7. The colours on the resistor shown here are red, orange, green and gold as read from left to right. How much is the resistance according to colour code?
8. Three resistors of resistances R_1 , R_2 and R_3 are connected (i) in series, and (ii) in parallel. Calculate the equivalent resistance of combination in each case.
9. What is the difference between emf and potential difference between the electrodes of a cell. Derive relation between the two.
10. Explain the difference between primary cells and secondary cells.
11. State Kirchhoff's rules governing the currents and electromotive forces in an electrical network?
12. Give theory of Wheatstone's bridge method for measuring resistances.
13. Discuss the theory of potentiometer.
14. How will you measure unknown potential difference with the help of a potentiometer?
15. Describe potentiometer method of comparing e.m.f. of two cells.
16. How will you determine internal resistance of a cell with the help of a potentiometer? What factors are responsible for internal resistance of a cell?
17. A wire of length 1 m and radius 0.1 mm has a resistance of 100Ω . Calculate the resistivity of the material.
18. Consider a wire of length 4m and cross-sectional area 1mm^2 carrying a current of 2A. If each cubic meter of the material contains 10^{29} free electrons, calculate the average time taken by an electron to cross the length of the wire.



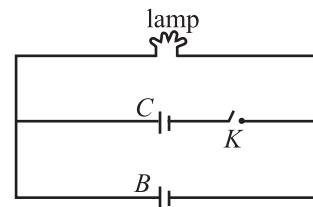
19. Suppose you have three resistors, each of value 30Ω . List all the different resistances that you can obtain by combining them.

20. The potential difference between the terminals of a battery of e.m.f. 6.0V and internal resistance 1Ω drops to 5.8V when connected across an external resistor. Find the resistance of the external resistor.

21. For the circuit shown here, calculate the value of current I and equivalent resistance R .

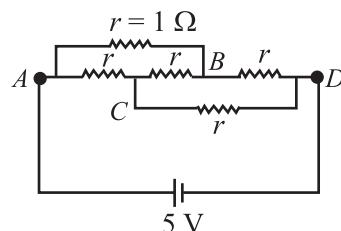


22. Examine the following network containing a lamp, a capacitor and a battery. The lamp is lighted when connected directly to the battery. What happens to it in this circuit when the switch is closed.



23. The following Wheatstone's bridge is balanced. Calculate

- (a) the value of equivalent resistance R in the circuit, and
 (b) the current in the arms AB and DC .



ANSWERS TO INTEXT QUESTIONS

17.1

- (a) The current reduces to half as resistance of the wire is doubled.
 (b) The current is doubled as resistance is halved.
- Resistivity is a property of the material of wire. It will not change with change in length and area of cross-section.

$$\rho = 2 \times 10^{-8} \Omega\text{m}$$

$$3. R = \frac{V}{I} = \frac{8}{0.15} = \frac{800}{15} = 53.3 \Omega$$

$$R = \frac{P\ell}{A} \Rightarrow \frac{800}{15} = \rho \frac{3}{2 \times 10^{-4}} \Rightarrow \rho = \frac{800 \times 2 \times 10^{-4}}{15 \times 3} = 35.5 \times 10^{-4} \Omega\text{m.}$$

- No. Only metallic conductor obey Ohm's law upto a certain limit. Semiconductors and electrolytes do not obey Ohm's law.



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$$5. I = \frac{q}{\tau} = \frac{n|e|}{t} = \frac{5 \times 10^{17} \times 1.6 \times 10^{-19}}{1} A = 0.8 \times 10^{-3} A = 0.8 \text{ mA}$$

The direction of current is opposite to the direction of flow of electrons, i.e., from right to left.

17.2

1. In parallel. They may draw different currents needed for their operation and are operated separately using different switches.
2. We use a voltage stabilizer

$$3. R = R_1 + \frac{R_2 R_3}{R_2 + R_3} + R_4$$

$$= 2 + \frac{10}{3} + 7$$

$$= 12.3 \Omega$$

17.3

1. Applying Kirchhoff's second rule on loop ABCDA, we get

$$2I_1 + 4I_1 + 3I_3 = 24$$

$$6I_1 + 3I_3 = 24 \quad \dots(1) \quad \Rightarrow \quad 2I_1 + I_3 = 8 \quad \dots(1)$$

Similarly, for loop DCBFD, we can write

$$-3I_3 + 6I_2 = 12 \quad \Rightarrow \quad 2I_2 - I_3 = 4 \quad \dots(2)$$

Also applying Kirchoff's first rule at junction D we get

$$I_2 + I_3 = I_1$$

Substituting in (1) we get

$$2I_2 + 3I_3 = 8$$

$$2I_2 - I_3 = 4$$

$$4I_3 = 4$$

$$I_3 = 1A$$

Substituting in (2)

$$2I_2 = 5 \Rightarrow I_2 = 2.5 A$$

$$2. \frac{P}{Q} = \frac{6}{12} = \frac{1}{2} \quad \text{and} \quad \frac{R}{S} = \frac{3}{6} = \frac{1}{2}$$



Notes

$$\frac{P}{Q} = \frac{R}{S} \quad \therefore \text{bridge is balanced}$$

Hence $V_B = V_D$ and $I_2 = 0$

$$I_1 = \frac{V}{I} = \frac{12}{18} = \frac{2}{3} \text{ A}$$

and

$$I - I_1 = \frac{12}{9} = \frac{4}{3} \text{ A}$$

17.4

1. $V = E - Ir$ as I increases V decreases.

$$2. R_{20} = R_0 (1 + 20\alpha)$$

$$R_{40} = R_0 (1 + 40\alpha)$$

$$\frac{R_{40}}{R_{20}} = \frac{1 + 40\alpha}{1 + 20\alpha}$$

$$\frac{1 + 40\alpha}{1 + 20\alpha} = \frac{30.16}{30} = 1 + \frac{0.16}{30}$$

$$1 + \frac{20\alpha}{1 + 20\alpha} = 1 + \frac{0.16}{30}$$

$$\frac{20\alpha}{1 + 20\alpha} = \frac{0.16}{30}$$

On cross-multiplication, we get $600\alpha = 0.16 + 3.2\alpha$

$$\Rightarrow \alpha \approx \frac{0.16}{600} = 2.67 \times 10^{-4} \text{ K}^{-1}$$

$$3. I = \frac{V}{R} = \frac{3}{4.5} = \frac{30}{45} = \frac{2}{3} \text{ A}$$

$$V = \sum -Ir \Rightarrow 3 = 5 - \frac{2}{3}r$$

$$\therefore r = \frac{2 \times 3}{2} = 3 \Omega$$

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$$4. \frac{E_2}{E_1} = \frac{\ell_2}{\ell_1} \Rightarrow \frac{1.02}{E_1} = \frac{30}{45} \Rightarrow E_1 = 0.51 \times \frac{3}{2} = 1.53V$$

$$5. \frac{E_2}{E_1} = \frac{\ell_2}{\ell_1}$$

$$\frac{E_1}{3} = \frac{2}{3}$$

$$E_1 = 2 V$$

$$6. P = IV \\ = 3 \times 0.3 \times 500 \\ = 45 \text{ WaH.}$$

$$7. I = \frac{P}{V} \Rightarrow I_1 = \frac{40}{220} = \frac{2}{11} A \quad \text{and} \quad I_2 = \frac{100}{220} = \frac{5}{11} A$$

$$R = \frac{V^2}{P} \Rightarrow I_1 = \frac{40}{220} = \frac{2}{11} A \quad \frac{V^2}{P} \Rightarrow R_1 = \frac{220 \times 220}{40} = 1210 \Omega$$

$$\text{and} \quad R_2 = \frac{220 \times 220}{100} = 484 \Omega$$

Answers to Problems in Terminal Exercises

17. $3.14 \times 10^{-6} \Omega m$. 18. 32 ms.
19. (i) All resistance in series; equivalent resistance 90Ω
(ii) All resistances in parallel; equivalent resistance 10Ω
(iii) One resistance in series with two others which are connected in parallel; equivalent resistance 45Ω
(iv) Two resistances in series and one resistance in parallel to them; equivalent resistance 20Ω .
20. 29Ω 21. $I = 1A, R = 12\Omega$
23. (a) $R = r = 1\Omega$ (b) $I = 2.5A$



MAGNETISM AND MAGNETIC EFFECT OF ELECTRIC CURRENT

In lesson 15, you learnt how charged rods attract each other or small bits of paper. You might have also played with magnets – the substances having the property of attracting small bits of iron. But did you ever think of some relation between electricity and magnetism? Such a relationship was discovered by Oersted in 1820. Now we know, for sure, how intimately magnetism and electricity are related.

In this lesson, you will learn the behaviour of magnets and their uses as also the magnetic effects of electric current. The behaviour of current carrying conductors and moving charges in a magnetic field are also discussed. On the basis of these principles, we will discuss the working of electric devices like motors and measuring devices like an ammeter, a voltmeter and a galvanometer.



OBJECTIVES

After studying this lesson, you should be able to :

- *define magnetic field and state its SI unit;*
- *list the elements of earth's magnetic field and write the relation between them;*
- *describe the magnetic effect of electric current : Oersted's experiment;*
- *state Biot-Savart's law and explain its applications;*
- *explain Ampere's circuital law and its application;*
- *describe the motion of a charged particle in uniform electric field and magnetic field;*
- *explain the construction and working of a cyclotron;*

MODULE - 5

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Notes

Magnetism and Magnetic Effect of Electric Current

- derive an expression for the force experienced by a current carrying conductor placed in a uniform magnetic field;
- derive an expression for the force between two infinitely long current carrying conductors placed parallel to each other; and
- explain the working principle of a galvanometer, an ammeter and a voltmeter.

18.1 MAGNETS AND THEIR PROPERTIES

The phenomenon of magnetism was known to Greeks as early as 600 B.C. They observed that some stones called magnetite (Fe_3O_4) attracted iron pieces. The pieces of naturally occurring magnetite are called **natural magnets**. Natural magnets are weak, but materials like iron, nickel, cobalt may be converted into strong permanent magnets. All magnets—natural or artificial—have same properties. You must be familiar with basic properties of magnets. However, for completeness, we recapitulate these.

- (i) **Directive Property :** A small bar magnet, when suspended freely on its center of mass so as to rotate about a vertical axis, always stays in approximately geographical north-south direction.
- (ii) **Attractive Property :** A magnet attracts small pieces of magnetic materials like iron, nickel and cobalt. The force of attraction is maximum at points near the ends of the magnet. These points are called **poles** of the magnet. In a freely suspended magnet, the pole which points towards the geographical north is called **north pole** and the one which points towards the geographical south is called **south pole**. Do directive and attractive properties suggest that our earth also acts like a magnet? Yes, it does.
- (iii) Unlike poles of two magnets attract each other and like poles repel (Fig.18.1).
- (iv) The poles of a magnet are inseparable, i.e. the simplest specimen providing magnetic field is a magnetic dipole.
- (v) When a magnet is brought close to a piece of iron, the nearer end of the piece of iron acquires opposite polarity and the farther end acquires same polarity. This phenomenon is called **magnetic induction**.

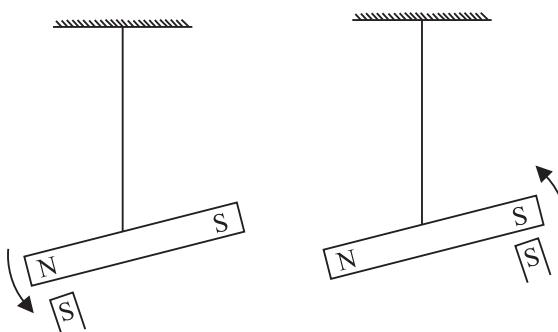


Fig. 18.1 : Unlike poles of two magnets attract each other and like poles repel.



Notes

18.1.1 Magnetic Field Lines

Interactions between magnets or a magnet and a piece of iron essentially represent action at a distance. This can be understood in terms of magnetic field. A very convenient method to visualize the direction and magnitude of a field is to draw the field lines :

- The direction of magnetic field vector \mathbf{B} at any point is given by the tangent to the field line at that point.
- The number of field lines that pass through unit area of a surface held perpendicular to the lines is proportional to the strength of magnetic field in that region. Thus, the magnetic field \mathbf{B} is large where the field lines are closer together and smaller where they are far apart.

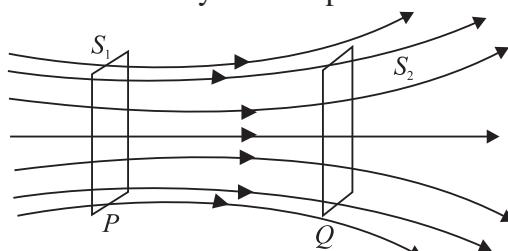


Fig 18.2: Magnetic field lines passing through two parallel surfaces

Fig 18.2 shows a certain number of field lines passing through parallel surfaces S_1 and S_2 . The surface area of S_1 is same as that of S_2 but the number of field lines passing through S_1 is greater than those passing through S_2 . Hence, the number of lines per unit area passing through S_1 is greater than that through S_2 . We can, therefore, say that the magnetic field in the region around P is stronger than that around Q .

- Outside the magnet, the field lines run from north pole to south pole and inside it, these run from south pole to north pole forming closed curves (Fig. 18.3).
- Two magnetic field lines can never cross each other.

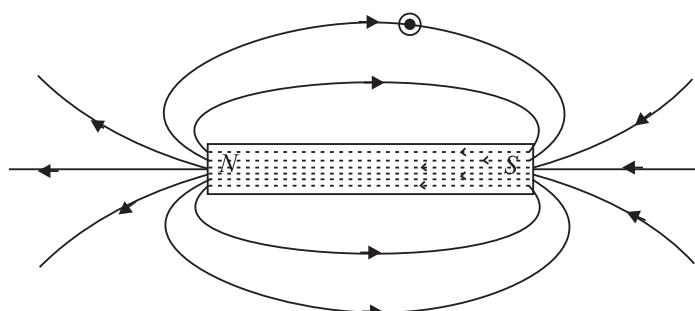


Fig. 18.3 : Magnetic field lines of a bar magnet



INTEXT QUESTIONS 18.1

1. You are given a magnet. How will you locate its north pole?

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Magnetism



Notes

Magnetism and Magnetic Effect of Electric Current

2. You are provided two identical looking iron bars. One of these is a magnet. Using just these two, how will you identify which of the two is a magnet.
3. You are given a thread and two bar magnets. Describe a method by which you can identify the polarities of the two magnets.

Magnetic field of the Earth

The directive property of magnets could be explained by considering that the earth acts as a magnet, i.e., as if a large bar magnet is placed inside the earth. The south pole of this magnet is considered near the geographical north pole and the magnetic north pole near the geographical south pole. RR_1 is the rotation axis of earth and MM_1 is the magnetic axis of the earth.

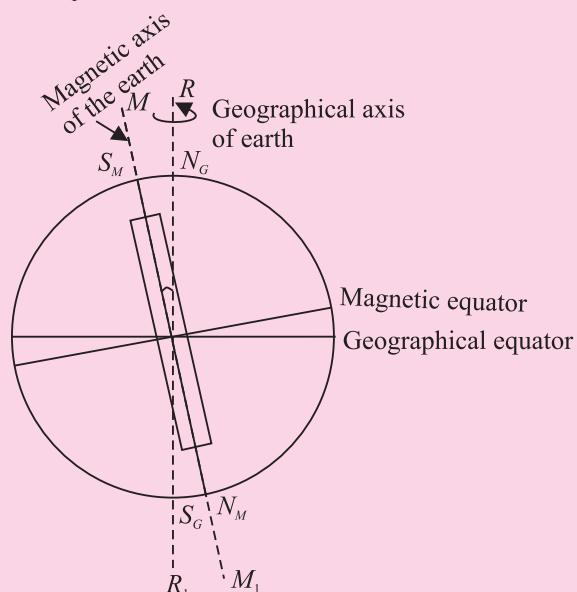


Fig. 18.4 : Magnetic field of the earth



ACTIVITY 18.1

Let us perform an experiment with a magnetic needle. (You can actually perform the experiment with a globe containing a bar magnet along its axis of rotation with north pole of the magnet pointing south.) Suspend the needle freely in such a manner that it can rotate in horizontal as well as vertical planes. If the needle is near the equator on earth's surface, it rests in horizontal plane. Suppose this needle is taken to places in the northern hemisphere. The needle rotates in the vertical plane and the north pole dips towards the earth, as we move towards geographical north pole. Finally at a point very near to Hudson bay in Canada, the north pole of the needle will point vertically downward. This place, located at 6° east of north, is considered to be the south pole of the earth's magnet.



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This place is about 650 km away from the earth's geographical north pole. If we take the same magnetic needle to places in the southern hemisphere, the south pole of the needle will dip downward and point vertically downward at a point 650 km west of the geographical south. This point could be considered as the *N pole of the earth's magnet*. **From this we conclude that the magnetic axis of the earth does not coincide with the geographical axis.**

An important aspect of earth's magnetic field is that it does not remain constant; its magnitude and direction change with time.

Elements of the Earth's Magnetic Field

Three measurable quantities are used to describe the magnetic field of earth. These are called elements of earth's magnetic field :

- Inclination or dip (δ);
- Declination (θ); and
- Horizontal component of the earth's field (\mathbf{B}_M).

(a) Inclination or Dip

If you suspend a magnetic needle freely at a place, you will observe that the needle does not rest in the horizontal plane. It will point in the direction of the resultant intensity of earth's field.

Fig. 18.5 shows the plane $PCDE$, which is the magnetic meridian at the point P (i.e. the vertical plane passing through the north and south poles of the earth's magnet) on the surface of the earth and $PABC$ is the geographic meridian (i.e. the vertical plane passing through the geographical north and south poles of the earth). Suppose that PR represents the magnitude and direction of the earth's magnetic field at the point P . Note that PR makes an angle δ with the horizontal direction. This angle is known as inclination or dip at P on the surface of the earth.

The angle which the earth's magnetic field makes with the horizontal direction in the magnetic meridian is called the **dip or inclination**.

(b) Declination

Refer to Fig 18.5 again. The plane $PCDE$ contains the magnetic field vector (\mathbf{PR}) of the earth. The angle between the planes $PCDE$ and $PABC$ is called the declination at the point P . It is shown as angle θ .

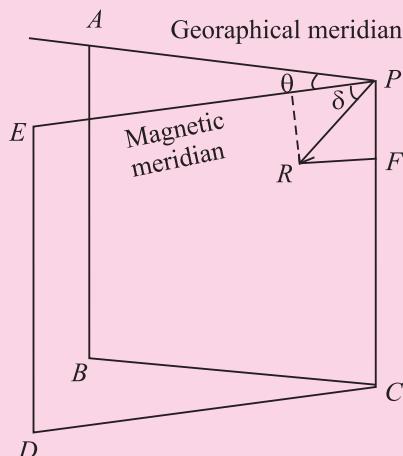


Fig. 18.5: Elements of earth's magnetic field

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The angle which the magnetic meridian at a place makes with the geographical meridian is called the declination at that place.

(c) Horizontal component

Fig. 18.5 shows that **PR** is the resultant magnetic field at the point **P**. **PH** represents the horizontal component and **PF** the vertical component of the earth's magnetic field in magnitude and direction. Let the magnetic field at the point **P** be **B**. The horizontal component

$$B_H = B \cos \delta \quad (18.1)$$

and the vertical component

$$B_V = B \sin \delta \quad (18.2)$$

By squaring and adding Eqns. (18.1) and (18.2), we get

$$B_H^2 + B_V^2 = B^2 \cos^2 \delta + B^2 \sin^2 \delta = B^2 \quad (18.3)$$

On dividing Eqn. (18.2) by Eqn. (18.1), we have

$$\frac{B_V}{B_H} = \tan \delta \quad (18.4)$$

18.2 ELECTRICITY AND MAGNETISM : BASIC CONCEPTS

You now know that flow of electrons in a conductor due to a potential difference across it constitutes electric current. The current flowing in a conductor is seen to exert a force on a free magnetic needle placed in a region around it. A magnetic needle is also affected by a magnet and hence we say that a current carrying conductor has a magnetic field around it. The magnetic field **B** is visualized by magnetic field lines. You will learn about these and some more terms such as magnetic permeability later in this lesson.

18.2.1 Magnetic Field around an Electric Current

Let us do a simple experiment.



ACTIVITY 18.2

Take a 1.5 volt battery, a wire about 1 m in length, a compass needle and a match box. Wind 10-15 turns of the electric wire on its base. Under the windings, place a compass needle, as shown in Fig. 18.6. Place the match box on the table so as to have the wires running along the north – south direction. Connect the free ends of the wire to the battery. What happens to the needle? You will observe that



Notes

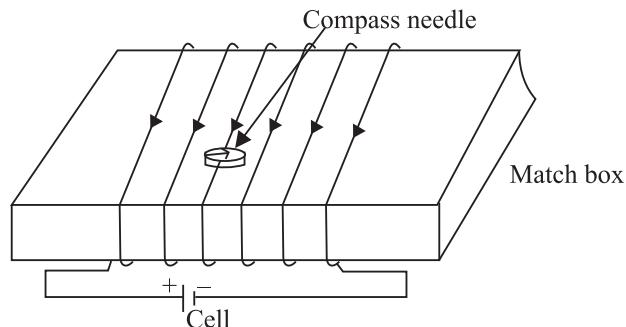
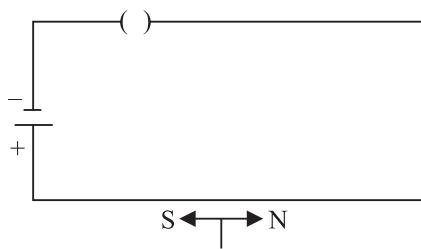
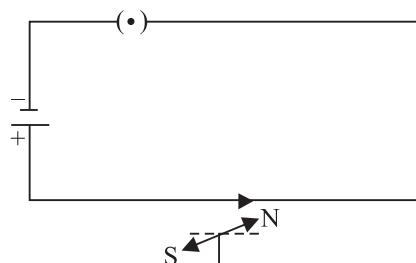


Fig. 18.6 : Demonstration of magnetic field due to electric current

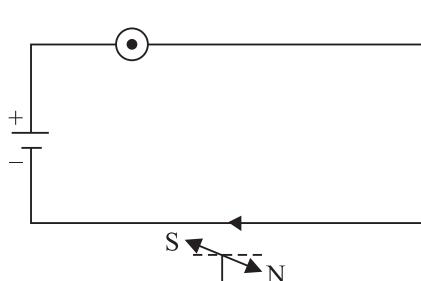
current in the wire, the compass needle points in the north – south direction (Fig. 18.7 a, b & c). When a magnetic needle is brought close to a vertical current carrying wire, the magnetic field lines are concentric circles around the wire, as shown in Fig 18.7 (d).



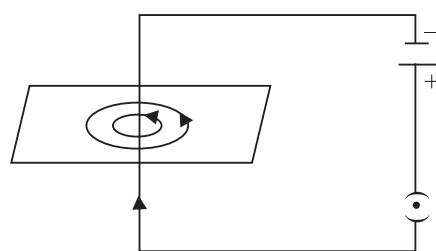
(a) No current, No deflection
north



(b) Current towards north deflection of
pole towards west



(c) When direction of current is reversed,
direction of deflection is reversed



(d) Circular field lines around a straight
current carrying conductor

Fig. 18.7 : Magnetic field around a current carrying conductor

In 1820 Hans Christian Oersted, Professor of Physics at Copenhagen in Denmark performed similar experiments and established that there is a magnetic field around a current carrying conductor.



18.3 BIOT-SAVART'S LAW

Biot-Savart's law gives a quantitative relationship between current in conductor and the resulting magnetic field at a point in the space around it. Each part of a current carrying conductor contributes to magnetic field around it. The net value of \mathbf{B} at a point is thus the combined effect of all the individual parts of the conductor. As shown in Fig. 18.8, the net magnetic field due to any current carrying conductor is the vector sum of the contributions due to the current in each infinitesimal element of length Δl .

Experiments show that the field B due to an element Δl depends on

- current flowing through the conductor, I ;
- length of the element Δl ;
- inversely proportional to the square of the distance of observation point P from the element Δl ; and
- the angle between the element and the line joining the element to the observation point.

Thus, we can write

$$\begin{aligned} |\Delta \mathbf{B}_0| &\propto \frac{I \Delta l \sin \theta}{r^2} \\ &= \frac{\mu_0}{4\pi} \frac{I d\ell \sin \theta}{r^2} \end{aligned} \quad (18.5)$$

where μ_0 is permeability of vacuum. Its value is $4\pi \times 10^{-7} \text{ W A}^{-1} \text{ m}^{-1}$. The value of permeability of air is also nearly equal to μ_0 .

If the conductor is placed in a medium other than air, the value of the field is altered and is given by $|\mathbf{B}| = \mu |\mathbf{B}_0|$. Here μ represents the permeability of the medium.

Direction of \mathbf{B} : Magnetic field at a point is a vector quantity. The direction of \mathbf{B} may be determined by applying the right hand grip rule. To apply this rule, let us consider the direction of the field produced in some simple cases. As shown in the Fig. 18.9 (a), grasp the wire in your right hand so that the thumb points in the direction of the current. Then the curled fingers of the hand will point in the direction of the magnetic field. To represent the magnetic field on paper, let us consider that current is flowing into the plane of the paper. Then according to the right hand rule, the field lines shall be in the plane of the paper (Fig. 18.9 b).

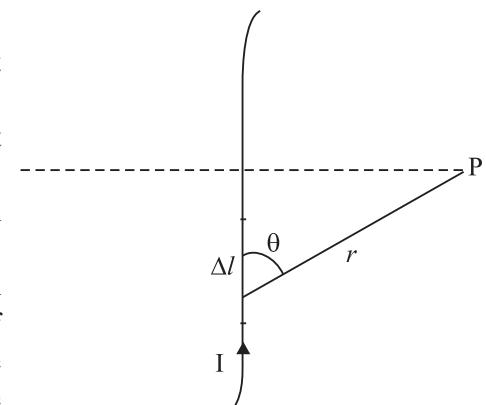


Fig. 18.8 : Magnetic field at P due to a current element Δl

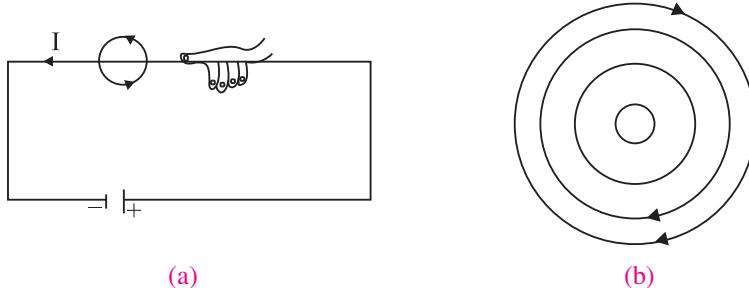


Fig. 18.9 : Direction of magnetic field : **a)** Right hand rule : thumb in the direction of current, field lines in the direction of curling fingers, and **b)** when current is in the plane of paper, the field lines shall be in the plane of paper, according to the right hand rule.

Notes**18.3.1 Applications of Biot-Savart's Law**

You now know that Biot-Savart's law gives the magnitude of the magnetic field. Let us now apply it to find the field around conductors of different shapes. Note that to calculate the net field due to different segments of the conductor, we have to add up the field contributions due to each one of them. We first consider a circular coil carrying current and calculate magnetic field at its centre.

(a) Magnetic field at the centre of a circular coil carrying current : Refer to Fig.18.10. It shows a circular coil of radius r carrying current I . To calculate magnetic field at its centre O , we first consider a small current element $\Delta\ell$ of the circular coil. Note that the angle between current element $\Delta\ell$ and r is 90° . From Eqn. (18.5) we know that the field at the centre O due to $\Delta\ell$ is

$$\begin{aligned} |\Delta\mathbf{B}| &= \frac{\mu_0}{4\pi} I \frac{\Delta\ell}{r^2} \sin 90^\circ \\ &= \frac{\mu_0}{4\pi} I \frac{\Delta\ell}{r^2} \quad (\text{as } \sin 90^\circ = 1) \end{aligned}$$

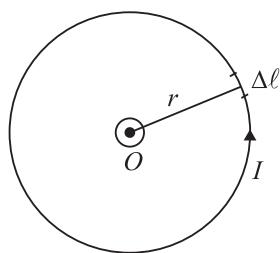


Fig. 18.10: Circular coil carrying current

The direction of $\Delta\mathbf{B}$ is normal to the plane of the coil. Since the field due to every element of the circular coil will be in the same direction, the resultant is obtained by adding all the contributions at the centre of the loop. Therefore

$$|\mathbf{B}| = \sum |\Delta\mathbf{B}| = \frac{\mu_0 I}{4\pi r^2} \sum \Delta\ell = \frac{\mu_0 I}{4\pi r^2} \cdot 2\pi r$$

Hence, magnetic field at the centre of a coil of radius r carrying current I is given by

$$|\mathbf{B}| = \frac{\mu_0 I}{2r} \quad (18.6)$$

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In case there is more than one loop of wire (say there are n turns), the field is given by

$$|\mathbf{B}| = \frac{\mu_0 nI}{2r}$$

You can check the direction of the net field using the rule given in Fig. 18.7. You can use right hand rule in any segment of the coil and will obtain the same result. (Another simple quick rule to identify the direction of magnetic field due to a current carrying coil is the so called End-rule, illustrated in Fig. 18.11 (a, b)).

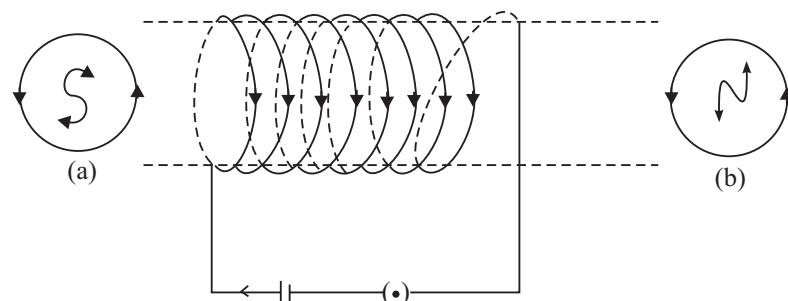


Fig 18.11: Direction of magnetic field : End-rule

When an observer looking at the circular coil at its either end finds the current to be flowing in the clockwise sense, the face of the coil behaves like the south pole of the equivalent magnet, i.e., B is directed inwards. On the other hand, if the current is seen to flow in the anticlockwise sense, the face of the coil behaves like the north pole of the equivalent magnet or the field is directed out of that end.



INTEXT QUESTIONS 18.2

1. What can you say about the field developed by
 - (i) a stationary electron ?
 - (ii) a moving electron ?
2. Electrons in a conductor are in constant motion due to thermal energy. Why do they not show magnetism till such time that a potential difference is applied across it ?
3. A current is flowing in a long wire. It is first shaped as a circular coil of one turn, and then into a coil of two turns of smaller radius. Will the magnetic field at the centre coil change? If so, how much ?

18.4 AMPERE'S CIRCUITAL LAW

Ampere's circuital law provides another way of calculating magnetic field around a current carrying conductor in some simple situations.

Ampere's circuital law states that the line integral of the magnetic field \mathbf{B} around a closed loop is μ_0 times the total current, I . Mathematically, we write

$$\oint \mathbf{B} \cdot d\mathbf{l} = \mu_0 I \quad (18.7)$$

Note that this is independent of the size or shape of the closed loop.

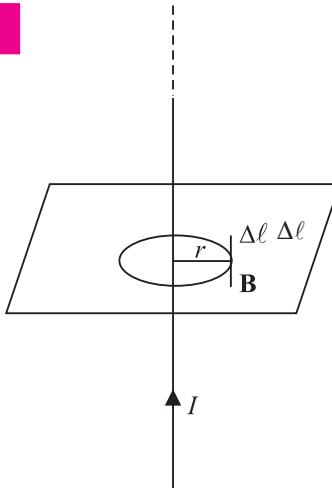


Fig. 18.12 : Ampere's circuital law

Notes



Andre Marie Ampere
(1775 – 1836)

French Physicist, mathematician and chemist, Ampere was a child prodigy. He mastered advanced mathematics at the age of 12. A mix of experimental skills and theoretical acumen, Ampere performed rigorous experiments and presented his results in the form of a theory of electrodynamics, which provides mathematical formulation of electricity and its magnetic effects. Unit of current is named in his honour. Lost in his work and ideas, he seldom cared for honours and awards. Once he forgot an invitation by emperor Napoleon to dine with him. His gravestone bears the epitaph : Tendun felix (Happy at last), which suggests that he had to face a very hard and unhappy life. But it never lowered his spirit of creativity.

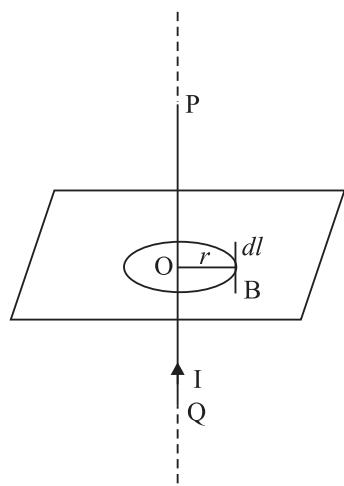


Fig. 18.13: Infinitely long current carrying conductor

18.4.1 Applications of Ampere's Circuital Law

We now apply Ampere's circuital law to obtain magnetic field in two simple situations.

(a) Magnetic field due to an infinitely long current carrying conductor

Refer to Fig. 18.13. It shows an infinitely long current carrying conductor POQ carrying current I . Consider a circular loop of radius r around it in the plane as shown. Then

$$\oint \mathbf{B} \cdot d\mathbf{l} = B \cdot 2\pi r$$

By applying Ampere's circuital law, we can write

$$|B| \cdot 2\pi r = \mu_0 I$$

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or

$$|\mathbf{B}| = \frac{\mu_0 I}{2\pi r} \quad (18.8)$$

This gives the magnetic field around an infinitely long straight current carrying conductor.

Solenoids and toroids are widely used in motors, generators, toys, fan-windings, transformers, electromagnets etc. They are used to provide uniform magnetic field. When we need large fields, soft iron is placed inside the coil.

(b) Magnetic field due to a solenoid

A solenoid is a straight coil having a large number of loops set in a straight line with a common axis, as shown in Fig. 18.14. We know that a current I flowing through a wire, sets up a magnetic field around it. Suppose that the length of the solenoid is ℓ and it has N number of turns. To calculate the magnetic field inside the solenoid along its axis (Fig 18.14), we can treat it to be a section of a toroidal solenoid of a very large radius. Thus :

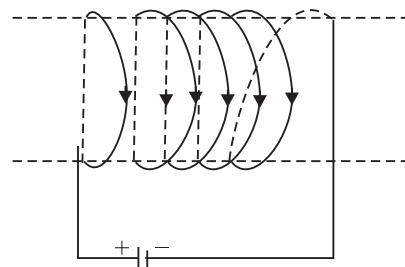


Fig. 18.14 : A solenoid

$$|\mathbf{B}| = \mu_0 nI$$

The direction of the field is along the axis of the solenoid. A straight solenoid is finite. Therefore, $|\mathbf{B}| = \mu_0 nI$ should be correct well inside the solenoid, near its centre.

For solenoids of small radius, the magnitude of B at the ends is given by

$$|\mathbf{B}| = \frac{\mu_0 nI}{2} \quad (18.9)$$

The solenoid behaves like a bar magnet and the magnetic field is as shown in Fig. 18.15.

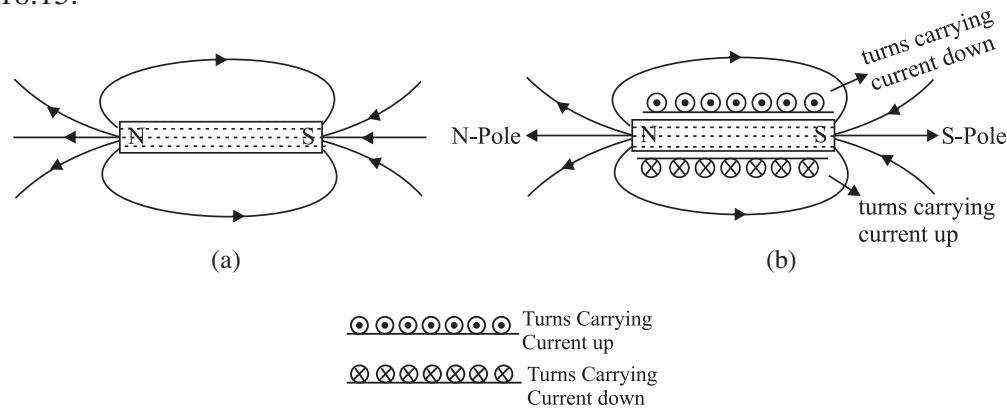


Fig. 18.15: Solenoid behaves like a bar magnet : a) Magnetic field due to a bar magnet, and b) magnetic field due to a current carrying solenoid

18.4.2 Application of Ampere's Circuital Law

(b) Magnetic Field due to a Straight Solenoid

A solenoid is a straight coil having a large number of loops set in a straight line with a common axis, as shown in Fig. 18.4.2. We know that a current I flowing through a wire, sets up a magnetic field around it. Suppose that the length of the solenoid is ℓ and it has N number of turns.

The magnetic field inside the solenoid, in its middle, is uniform and parallel to its axis. Outside the solenoid, however, the field is negligibly weak. These statements hold true, strictly speaking, if the length of the solenoid is very large as compared to its diameter. For a long solenoid, whose windings are very tightly and uniformly wound, the magnetic field inside it is fairly uniform everywhere and is zero outside it.

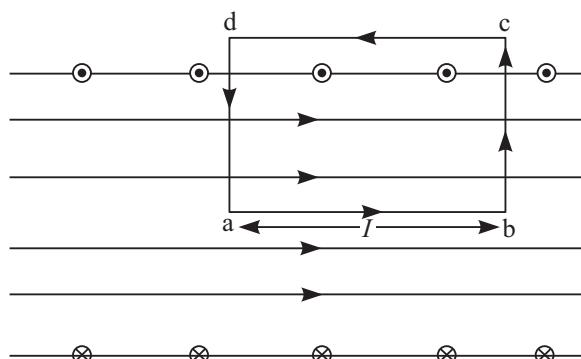


Fig. 18.4.2

Let us take a rectangular loop abcd as shown in Fig 18.4.2. Along the path ab, the magnetic field is uniform. Hence, for this path $\mathbf{B} \cdot d\ell = B\ell$. Along the paths cd, as the magnetic field is weak it may be taken as zero. Hence, for this path $\mathbf{B} \cdot d\ell = 0$. The two short sides bc and da also do not contribute anything to $\mathbf{B} \cdot d\ell$ as \mathbf{B} is either zero (outside the solenoid), or perpendicular to $d\ell$ (inside the solenoid).

If n be the number of turns per unit length along the length of the solenoid, then the number of turns enclosed by the rectangular loop of length ℓ is $n\ell$. If each turn of the solenoid carries a current i , then the total current threading the loop is nli . Hence, from Ampere's circuital law,

$$\sum \mathbf{B} \cdot d\ell = \mu_0(nli)$$

or

$$B\ell = \mu_0 nli$$

or

$$B = \mu_0 ni$$



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(c) Magnetic field due to a toroid

A toroid is basically an endless solenoid which may be formed by bending a straight solenoid so as to give it a circular shape.

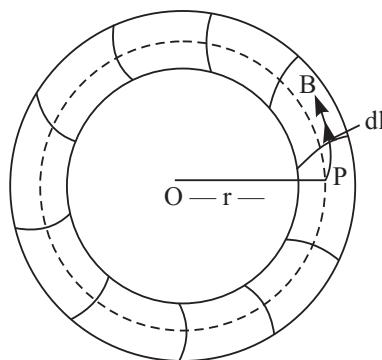


Fig. 18.4.3

Suppose, we want to find the magnetic field at a point P , inside the toroid, whose distance from the centre O is r . Draw a circle passing through the point P and concentric with the toroid. The magnetic field will everywhere be tangential to the circle, its magnitude being the same at all points of it. So, we can write:

$$\sum \mathbf{B} \cdot d\ell = \sum B d\ell = B \sum d\ell$$

But $\sum d\ell = 2\pi r$, the circumference of the circular path.

Therefore,

$$\sum \mathbf{B} \cdot d\ell = 2\pi r B$$

If N be the total number of turns and i the current flowing through the windings of the toroid, then the total current threaded by the circular path of radius r is Ni . Hence, from Ampere's circuital law,

$$\sum \mathbf{B} \cdot d\ell = \mu_0 Ni$$

or

$$2\pi r B = \mu_0 Ni$$

or

$$B = \frac{\mu_0 Ni}{2\pi r}$$

**Notes****18.4.3 Electromagnets and Factors Affecting their Strength**

We have seen that a current-carrying solenoid behaves as a bar magnet, with one end behaving as north pole and the other as south pole depending on the direction of flow of current. The polarity of such magnets is determined by the end rule and the strength of the magnetic field is given by

$$|\mathbf{B}| = \mu_0 n I$$

where μ_0 is the permeability of free space, n is the number of turns per unit length and I is the current flowing through the solenoid.

It is clear that the solenoid remains a magnet as long as the current is flowing through it. Thus, a current-carrying solenoid is called an electromagnet.

Its strength depends on :

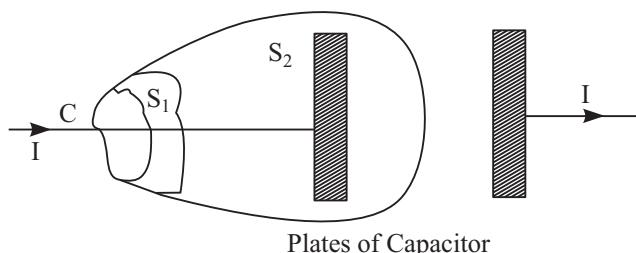
- (i) Number of turns per unit length of the solenoid, and
- (ii) The current flowing through it.

It may also be noted that the strength of the magnetic field of an electromagnet increases when a soft iron core is introduced inside it.

18.4.4 Concept of Displacement Current

The concept of displacement current was introduced by Maxwell. As we know, magnetic field is produced due to the conduction current. However, according to Maxwell in empty space (where no conduction current exists), the magnetic field is produced due to the displacement current which, unlike conduction current, is not associated with the motion of charges.

Consider a simple circuit consisting of a small parallel-plate capacitor being charged by a current I .

**Fig. 18.4.4**

Applying Ampere's circuital law to the contour C and the surface S_1 , we find

$$\oint \mathbf{B} \cdot d\ell = \mu_0 I$$

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However, applying Ampere's circuital law to the contour C and the surface S_2 , as there is no current through this surface, we get

$$\oint \mathbf{B} \cdot d\ell = 0$$

The above two equations are mutually contradictory. To avoid this contradiction, Maxwell assumed that a current exists between the capacitor plates. He called this current displacement current and showed that this current arises due to the variation of electric field with time.

A simple expression for the displacement current can be derived as follows. Consider a parallel plate capacitor. Let q be the charge on the capacitor plates at any instant t .

The electric field inside the capacitor is given by

$$E = \frac{q}{A\epsilon_0}$$

When A is the surface area of the plates. Therefore, the electric flux through the capacitor is

$$\phi_E = EA = \frac{q}{\epsilon_0}$$

The rate of change of the instantaneous flux can be written as

$$\frac{\Delta\phi_E}{\Delta t} = \frac{1}{\epsilon_0} \frac{\Delta q}{\Delta t} = \frac{I}{\epsilon_0}$$

So, we can write

$$\epsilon_0 \frac{\Delta\phi_E}{\Delta t} = I$$

The expression on the left hand side is equivalent to a current, which though equal to the conduction current I is actually different from it as it is not associated with the motion of free charges. It is called displacement current. Unlike the conduction current I , the displacement current arises whenever the electric field and hence the electric flux changes with time.

Adding displacement current to the conduction currents I , Maxwell modified the Ampere's circuital law in the form,

$$\oint \mathbf{B} \cdot d\ell = \mu_0 \left(I + \epsilon_0 \frac{\Delta\phi_E}{\Delta t} \right)$$

Maxwell's modification of Ampere's law tells us that, in addition to conduction current, a time-varying electric field can also produce magnetic field.

Example 18.1 : A 50 cm long solenoid has 3 layers of windings of 250 turns each. The radius of the lowest layer is 2cm. If the current through it is 4.0 A,



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calculate the magnitude of \mathbf{B} (a) near the centre of the solenoid on and about the axis; (b) near the ends on its axis; and (c) outside the solenoid near the middle.

Solution :

a) At the centre or near it

$$\begin{aligned} B &= \mu_0 n I \\ &= 4\pi \times 10^{-7} \times \frac{3 \times 250}{0.5} \times 4 \\ &= 16\pi \times 1500 \times 10^{-7} \text{ T} \\ &= 24\pi \times 10^{-4} \text{ T} \end{aligned}$$

b) At the ends

$$B_{\text{ends}} = \frac{1}{2} B_{\text{centre}} = 12\pi \times 10^{-4} \text{ T}$$

c) Outside the solenoid the field is zero.

Example 18.2: Calculate the distance from a long straight wire carrying a current of 12 A at which the magnetic field will be equal to 3×10^{-5} T.

Solution :

$$B = \frac{\mu_0 I}{2\pi r} \Rightarrow r = \frac{\mu_0 I}{2\pi B}$$

$$\therefore r = \frac{2 \times 10^{-7} \times 12}{3 \times 10^{-5}} = 0.25 \text{ m}$$

**INTEXT QUESTIONS 18.3**

1. A drawing of the lines of force of a magnetic field provides information on
 - direction of field only
 - magnitude of field only
 - both the direction and magnitude of the field
 - the force of the field
2. What is common between Biot-Savart's law and Ampere's circuital law?
3. In the following drawing of lines of force of a non-uniform magnetic field, at which point is the field (i) uniform, (ii) weakest, (iii) strongest?

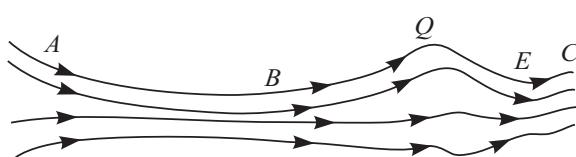


Fig. 18.16 : A typical magnetic field

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4. A 10 cm long solenoid is meant to have a magnetic field 0.002T inside it, when a current of 3A flows through it. Calculate the required no. of turns.
5. Derive an expression for the field due to a toroid using Ampere's circuital law.

18.5 FORCE ON A MOVING CHARGE IN A MAGNETIC FIELD

When a charged body moves in a magnetic field, it experiences a force. Such a force experienced by a moving charge is called the *Lorentz force*. The Lorentz force on a particle with a charge $+q$ moving with a velocity \mathbf{v} in a magnetic field \mathbf{B} is given by

$$\mathbf{F} = q(\mathbf{v} \times \mathbf{B})$$

$$\text{or } |\mathbf{F}| = q v B \sin \theta \quad (18.10)$$

where θ is the angle between the directions of \mathbf{v} and \mathbf{B} . The direction of \mathbf{F} is given by Fleming's left hand rule.

Fleming's left hand rule states that if we stretch the fore finger, the central finger and the thumb of our left hand at right angles to each other and hold them in such a way that the fore finger points in the direction of magnetic field and the central finger points in the direction of motion of positively charged particle, then the thumb will point in the direction of the Lorentz force (Fig. 18.17).

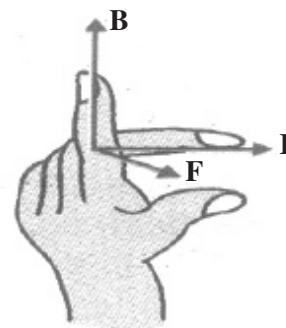


Fig. 18.17 : Fleming's left hand rule

Some important points to note

- \mathbf{F} is a mechanical force resulting in a pull or a push.
- The direction of force is given by Fleming's left hand rule.
- In case of negative charges, the central finger should point opposite to the direction of its motion.
- If the charge stops, the force becomes zero instantly.
- Force is zero when charges move along the field \mathbf{B} .
- Force is maximum when charges move perpendicular to the field : $F = qvB$

18.5.1 Force on a Current Carrying Conductor in a Uniform Magnetic Field

The concept of Lorentz force can be easily extended to current carrying conductors placed in uniform magnetic field \mathbf{B} . Suppose that the magnetic field is parallel to the plane of paper and a conductor of length Δl carrying current I is placed normal to the field. Suppose further that the current is flowing downward with a drift velocity \mathbf{v}_d and hence each free electron constituting the current experiences a Lorentz force $\mathbf{F} = e \mathbf{v}_d \cdot \mathbf{B}$



Notes

If there are N free electrons in the conductor, the net force on it is given by

$$F = N e v_d B = nA \Delta\ell e v_d B \quad (18.11)$$

where n denotes the number of free electrons per unit volume. But $neAv_d = I$. Hence

$$\therefore F = I \Delta\ell B \quad (18.12)$$

If conductor makes an angle θ with \mathbf{B} , then $|F| = I \Delta\ell B \sin\theta$.

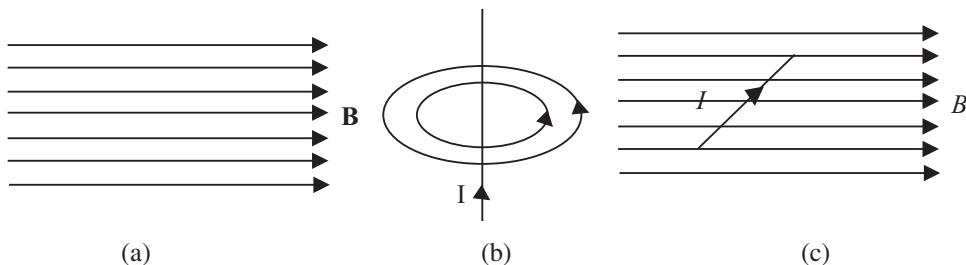


Fig. 18.18: a) Uniform magnetic field, b) field due to current carrying inductor, and c) force on a current carrying conductor

The direction of the force is again given by Fleming's left hand rule.

Eqn. (18.12) can be used to define the unit of magnetic field in terms of the force experienced by a current carrying conductor. By rearranging terms, we can write

$$B = \frac{F}{I\Delta\ell}$$

Since F is taken in newton, I in ampere and $\Delta\ell$ in metre, the unit of B will be $\text{NA}^{-1} \text{m}^{-1}$. It is called tesla (T).

18.5.2 Force Between two Parallel Wires Carrying Current

You now know that every current carrying conductor is surrounded by a magnetic field. It means that it will exert force on a nearby current carrying conductor. The force between two current carrying conductors placed parallel to each other is mutual and magnetic in origin. A current carrying wire has no net electric charge, and hence cannot interact electrically with another such wire.

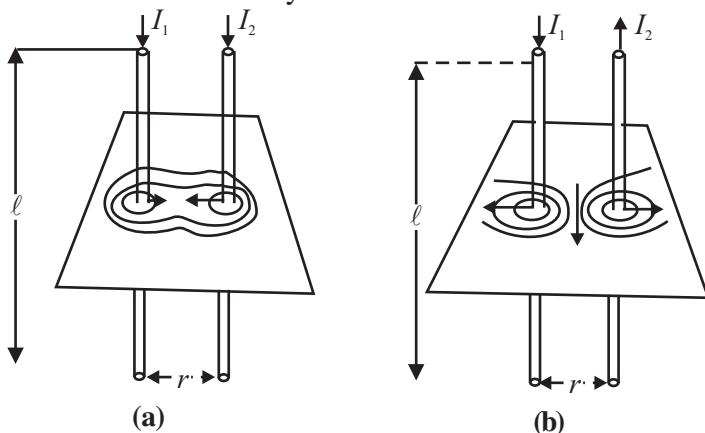


Fig. 18.19: Experimental demonstration of force between two parallel wires carrying current

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Fig 18.19 shows two parallel wires separated by distance r and carrying currents I_1 and I_2 , respectively. The magnetic field due to one wire at a distance r from it is $B_1 = \frac{\mu_0 I_1}{2\pi r}$.

Similarly, the field due to second wire at a distance r from it will be $B_2 = \frac{\mu_0 I_2}{2\pi r}$.

These fields are perpendicular to the length of the wires and therefore the force on a length l , of the other current carrying conductor is given by

$$F = B I \ell = \frac{\mu_0 I_1}{2\pi r} I_2 \ell$$

or force per unit length

$$\frac{F}{\ell} = \frac{\mu_0 I_1 I_2}{2\pi r} \quad (18.13)$$

The forces are attractive when the currents are in the same direction and repulsive when they are in opposite directions.

Eqn (18.13) can be used to define the unit of current. If $I_1 = I_2 = 1\text{A}$, $\ell = 1\text{m}$ and $r = 1\text{m}$, then

$$F = \frac{\mu_0}{2\pi} = 2 \times 10^{-7} \text{ N}$$

Thus, if two parallel wires carrying equal currents and placed 1 m apart in vacuum or air experience a mutual force of $2 \times 10^{-7} \text{ N m}^{-1}$, the current in each wire is said to be one ampere.

18.5.3 Motion of a Charged Particle in a Uniform Field

We can now think of various situations in which a moving charged particle or a current carrying conductor in a magnetic field experiences Lorentz force. The work done by a force on a body depends on its component in the direction of motion of the body. When the force on a charged particle in a magnetic field is perpendicular to its direction of motion, no work is said to be done. Hence the particle keeps the same speed and kinetic energy which it had while moving in the field, even though it is deflected. On the other hand, the speed and energy of a charged particle in an electrical field is always affected due to the force by the field on the particle. A charged particle moving perpendicular to a magnetic field follows a circular path (Fig. 18.20) because it experiences a force at right angles to the direction of motion at every position.

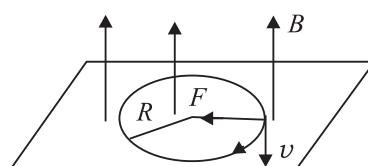


Fig. 18.20: Path of a charged particle in a uniform magnetic field



Notes

To know the radius R of the circular path of the charged particle, we note that the magnetic force $q v B$ provides the particle with the centripetal force ($m v^2/R$) that keeps it moving in a circle. So we can write

$$q v B = \frac{mv^2}{R}$$

On rearrangement, we get

$$R = \frac{mv}{qB} \quad (18.14)$$

The radius of the path traced by a charged particle in a uniform magnetic field is directly proportional to its momentum (mv) and inversely proportional to its charge and the magnetic field. It means that greater the momentum, larger the circle, and stronger the field, the smaller the circle. The time period of rotation of the particle in a circular path is given by

$$T = \frac{2\pi R}{v} = \frac{2\pi m}{qB} \quad (18.14 \text{ a})$$

Note that the time period is independent of velocity of the particle and radius of the orbit. It which means that once the particle is in the magnetic field, it would go round and round in a circle of the same radius. If m , B , q , remain constant, the time period does not change even if v and R are changed.

Now think, what happens to R and T if a) field B is made stronger; b) field B is made weaker; c) field B ceases to exist; d) direction of B is changed; e) the particle is made to enter the magnetic field at a higher speed; f) the particle enters at an angle to B ; and g) the charged particle loses its charge.

18.5.4 Motion of a Charged Particle in uniform Electric Field and Magnetic Field

(a) Motion in Electric Field

When a charged particle q is placed in a uniform electric field E , it experiences a force,

$$\mathbf{F} = q\mathbf{E}$$

Thus, the charged particle will be accelerated under the influence of this force. The acceleration is given by

$$\mathbf{a} = \frac{\mathbf{F}}{m} = \frac{q\mathbf{E}}{m}$$

The acceleration will be in the direction of the force. If it is a positive charge, it will accelerate in the direction of the field and if it is a negative charge it will

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accelerate in a direction opposite that of the field. The velocity and displacement of charged particle can also be calculated by using the equations of motion:

$$v = u + \left(\frac{qE}{m} \right) t$$

$$s = ut + \frac{1}{2} \left(\frac{qE}{m} \right) t^2$$

where t denotes time.

(b) Motion in magnetic field

In article 18.5 (Page 114, Book 2), it has been discussed that the force experienced by a charged particle in a magnetic field is given by

$$F = qBv\sin\theta$$

Where θ is angle between the velocity and magnetic field.

If $\theta = 0^\circ$, $F = 0$ and charged particle will move along a straight line with constant speed.

If $\theta = 90^\circ$, F will be maximum and its direction, according to Fleming's left hand rule, will be perpendicular to the plane of v and \mathbf{B} and the charged particle will move along a circular path with a constant speed and frequency.

If $\theta \neq 0^\circ \neq 90^\circ$, then the velocity of the charged particle will be $v\sin\theta$ perpendicular to the field and $v\cos\theta$ parallel to the field. The particle, therefore, moves along a helical path.

What we note from the above discussion is that a magnetic field does not change the speed of a moving charge, it only changes its direction of motion.

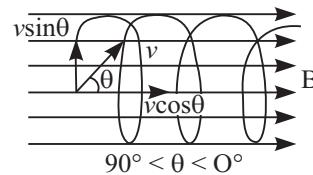
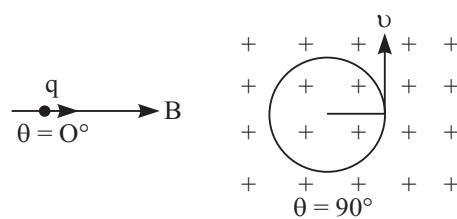


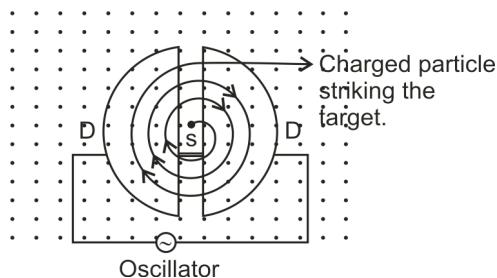
Fig. 18.5.4



Notes

18.5.5 Cyclotron

The cyclotron is a device invented by E.O. Lawrence in 1929, that is used for accelerating charged particles (such as protons, deuteron or α -particles) to high velocities. It consists of two semi-circular hollow metallic disks DD , called dees, on account of their shape resembling the letter D of English alphabet. They are insulated from one another with a small gap between them. The dees are placed in an evacuated chamber.



18.5.5

A magnetic field perpendicular to the plane of the dees (out of the paper in Fig) is maintained with the help of an electromagnet having flat pole-pieces. A rapidly oscillating potential difference is applied between the dees with the help of an oscillator. This produces an oscillating electric field in the gap between the dees.

Consider a charged particle of mass m and charge q in the gap between the dees. The particle is accelerated by the electric field towards one of them. Inside the dees, it moves with constant speed in a semicircle in a clockwise direction. If the frequency of the oscillator is equal to the frequency of revolution of the charged particle, then it reaches the gap at the instant when the opposite dee becomes negative because of the reversal of the direction of electric field.

The frequency of revolution of the charged particle is given by (see Eq. 18. 14a):

$$\nu = \frac{1}{T} = \frac{\nu}{2\pi R} = \frac{Bq}{2\pi m}$$

where B is the magnetic field.

It is also called cyclotron frequency and denoted by ν_c . When $\nu_c = \nu_o$, the frequency of the oscillator, the particle reaches the gap when the electric potential at the opposite 'D' has just reversed its sign. This condition is also known as cyclotron resonance condition. On account of this, the particle gains energy and, therefore, it moves in a circle of larger radius. This energy gain can be repeated many times.

Thus, the energy and the radius of the path of the particle keep on increasing progressively. However, the maximum radius which the path can have is limited

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by the radius R of the dees. The high energy charged particle finally comes out through an opening in the dee.

Example 18.3 : Refer to Fig. 18.21 and calculate the force between wires carrying current 10A and 15A, if their length is 5m. What is the nature of this force ?

Solution : When currents flow in two long parallel wires in the same direction, the wires attract each other and the force of attraction is given by

$$\frac{F}{\ell} = \frac{\mu_0 I_1 I_2}{2\pi r} = \frac{2 \times 10^{-7} \times 10 \times 15}{3} = 10^{-4} \text{ N m}^{-1}$$
$$\therefore F = 5 \times 10^{-4} \text{ N}$$

Fig. 18.21 The force is attractive in nature.

Example 18.4 : An electron with velocity $3 \times 10^7 \text{ ms}^{-1}$ describes a circular path in a uniform magnetic field of 0.2T, perpendicular to it. Calculate the radius of the path.

Solution :

We know that

$$R = \frac{mv}{Bq}$$

Here, $m_e = 9 \times 10^{-31} \text{ kg}$, $e = 1.6 \times 10^{-19} \text{ C}$, $v = 3 \times 10^7 \text{ m s}^{-1}$ and $B = 0.2 \text{ T}$. Hence

$$R = \frac{9 \times 10^{-31} \times 3 \times 10^7}{0.2 \times 1.6 \times 10^{-19}}$$
$$= 0.85 \times 10^{-3} \text{ m}$$
$$= 8.5 \times 10^{-4} \text{ m}$$



INTEXT QUESTION 18.4

1. A stream of protons is moving parallel to a stream of electrons but in the opposite direction. What is the nature of force between them ?
2. Both electrical and magnetic fields can deflect an electron. What is the difference between them?
3. A body is suspended from a vertical spring. What shall be the effect on the position of the body when a current is made to pass through the spring.
4. How does a cyclotron accelerate charged particles?

18.6 CURRENT LOOP AS A DIPOLE

From Eqn. (18.6) you will recall that the field at the centre of a coil is given by

$$B = \frac{\mu_0 I}{2r}$$



Notes

On multiplying the numerator and denominator by $2\pi r^2$, we can rewrite it as

$$B = \frac{\mu_0 2I \cdot \pi r^2}{4\pi r^3} = \frac{\mu_0 2IA}{4\pi r^3} = \frac{\mu_0}{4\pi} \frac{2M}{r^3}$$

where A is area of coil and M is magnetic moment. This shows that a current carrying coil behaves like a magnetic dipole having north and south poles. One face of the loop behaves as north pole while the other behaves as south pole.

Let us now undertake a simple activity.



ACTIVITY 18.3

Suspend a bar magnet by a thread between pole pieces of a horse shoe magnate, as shown in Fig 18.22.

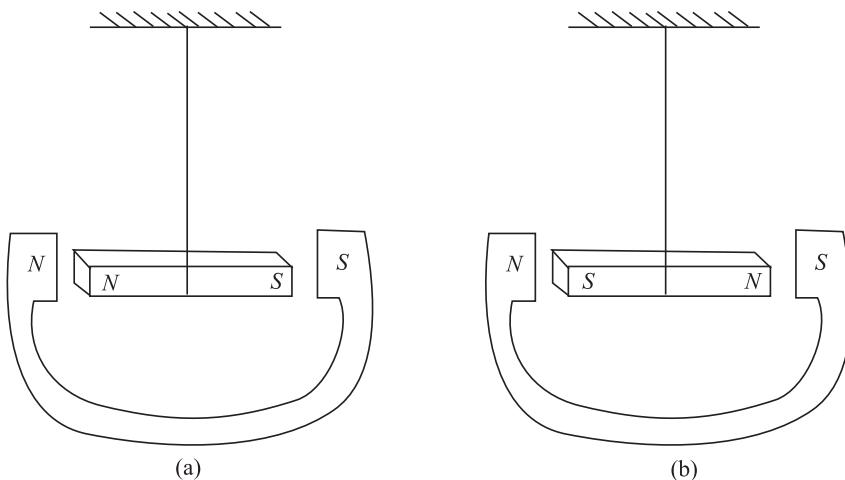


Fig. 18.22 : A bar magnet suspended between a horse shoe magnet

What will happen when the bar magnet shown in Fig. 18.24(a) is displaced slightly sideways? Since like poles repel, the bar magnet experiences a torque and tends to turn through 180° and get aligned, as shown in Fig. 18.22 (b). Since a current loop behaves as a magnet, it will align in an external field in the same way.

You have already studied the following equations in the lesson on electrostatics. The electric field of a dipole at a far point on its axis is given by

$$\mathbf{E} = \frac{1}{4\pi\mu_0} \frac{2\mathbf{P}}{x^3} \quad (18.15 \text{ b})$$

The magnetic field due to a current carrying coil is given by

$$\mathbf{B} = \frac{\mu_0}{4\pi} \frac{2\mathbf{NI}\mathbf{A}}{x^3} = \frac{\mu_0}{4\pi} \frac{2\mathbf{M}}{x^3} \quad (18.15 \text{ c})$$

where \mathbf{M} is the magnetic dipole moment.

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A comparison between these expressions leads us to the following analogies :

- A current loop behaves as a magnetic dipole with magnetic moment
- $$\mathbf{M} = NIA \quad (18.15\text{ d})$$
- Like the poles of a magnetic dipole, the two faces of a current loop are inseparable.
 - A magnetic dipole in a uniform magnetic field behaves the same way as an electric dipole in a uniform electric field.
 - A magnetic dipole also has a magnetic field around it similar to the electric field around an electric dipole.

Thus magnetic field due to a magnetic dipole at an axial point is given by

$$\mathbf{B} = \frac{\mu_0}{4\pi} \frac{2\mathbf{M}}{x^3} \quad (18.16)$$

whereas the field at an equatorial point is given by

$$\mathbf{B} = - \frac{\mu_0}{4\pi} \frac{\mathbf{M}}{x^3} \quad (18.17)$$

Magnetism in Matter

Based on the behaviour of materials in magnetic field, we can divide them broadly into three categories : (i) **Diamagnetic** materials are feebly repelled by a magnet. (ii) **Paramagnetic** materials are feebly attracted by a magnet. (iii) **Ferromagnetic** materials are very strongly attracted by a magnet. Substances like iron, nickel and cobalt are ferromagnetic. Let us study ferromagnetic behaviour of materials in some details.

Ferromagnetic materials, when placed even in a weak magnetic field, become magnets, because their atoms act as permanent magnetic dipoles. The atomic dipoles tend to align parallel to each other in an external field. These dipoles are not independent of each other. Any dipole strongly feels the presence of a neighboring dipole. A correct explanation of this interaction can be given only on the basis of quantum mechanics. However, we can qualitatively understand the ferromagnetic character along the following lines.

A ferromagnetic substance contains small regions called **domains**. All magnetic dipoles in a domain are fully aligned. The magnetization of domains is maximum. But the domains are randomly oriented. As a result, the total magnetic moment of the sample is zero. When we apply an external magnetic field, the domains slightly rotate and align themselves in the direction of the field giving rise to resultant magnetic moment. The process can be easily understood with the help of a simple diagram shown in Fig.18.23.



Notes

Table 18.1: Ferromagnetic substances and their curie temperatures

| Substances | Curie temperature T_c (K) |
|-------------------------|-----------------------------|
| Iron | 1043 |
| Nickel | 631 |
| Cobalt | 1394 |
| Gadolinium | 317 |
| Fe_2O_3 | 893 |

Fig. 18.23 (a) shows ten domains. For simplicity we take a two dimensional example. All the domains are so directed that the total magnetization of the sample is zero. Fig. 18.23 (b) shows the state after the application of an external magnetic field. The boundaries of the domains (Domain Walls) reorganise in such a way that the size of the domain having magnetic moment in the direction of the field becomes larger at the cost of others. On increasing the strength of external field, the size of favorable domains increases, and the orientation of the domain changes slightly resulting in greater magnetization (Fig. 18.23 (c)). Under the action of very strong applied field, almost the entire volume behaves like a single domain giving rise to saturated magnetization. When the external field is removed, the sample retains net magnetization. The domain in ferromagnetic samples can be easily seen with the help of high power microscope.

When the temperature of a ferromagnetic substance is raised beyond a certain critical value, the substance becomes paramagnetic. This critical temperature is known as Curie temperature T_c .

Example 18.5 : The smallest value of magnetic moment is called the Bohr

Magneton $\mu_B = \frac{e\hbar}{4\pi m}$. It is a fundamental constant. Calculate its value.

Solution :

$$\mu_B = \frac{e\hbar}{4\pi m} = \frac{(1.6 \times 10^{-19} \text{ C}) \times (6.6 \times 10^{-34} \text{ Js})}{4 \times 3.14 \times (9 \times 10^{-31} \text{ kg})}$$

$$= 9.34 \times 10^{-24} \text{ J T}^{-1}$$

18.6.1 Torque on a Current Loop

A loop of current carrying wire placed in a uniform magnetic field (\mathbf{B}) experiences no net force but a torque acts on it. This torque tends to rotate the loop to bring its plane perpendicular to the field direction. This is the principle that underlines the operation of all electric motors, meters etc.

Let us examine the force on each side of a rectangular current carrying loop where plane is parallel to a uniform magnetic field \mathbf{B} . (Fig. 18.24 (a).)

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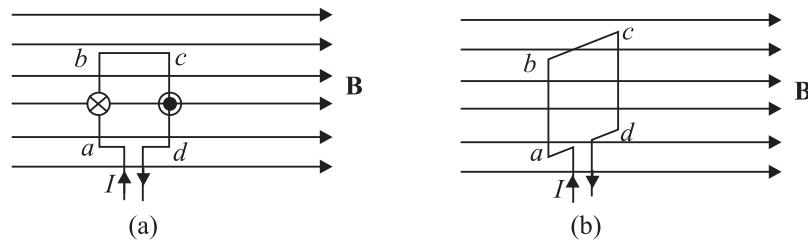


Fig. 18.24: Force on the sides of a rectangular loop when (a) the loop is parallel to the field, and (b) the coil is perpendicular to the field.

The sides ad and bc of the loop are parallel to \mathbf{B} . So no force will act on them. Sides ab and cd are however, perpendicular to \mathbf{B} , and these experience maximum force. We can easily find the direction of the force on ab and cd .

In fact, $|\mathbf{F}_{ab}| = |\mathbf{F}_{cd}|$ and these act in opposite directions. Therefore, there is no net force on the loop. Since \mathbf{F}_{ab} and \mathbf{F}_{cd} do not act along the same line, they exert a torque on the loop that tends to turn it. This holds good for a current loop of any shape in a magnetic field.

In case the plane of the loop were perpendicular to the magnetic field, there would neither be a net force nor a net torque on it (see Fig 18.26 (b)).

$$\begin{aligned}\text{Torque} &= \text{force} \times \text{perpendicular distance between the force} \\ &= B IL \cdot b \sin \theta\end{aligned}$$

Refer to Fig. 18.25 which shows a loop $PQRS$ carrying current I . θ is the angle between the magnetic field \mathbf{B} and the normal to the plane of the coil n . The torque is then

$$\tau = NBIL b \sin \theta$$

where N is the number of turns of the coil. We can rewrite it as

$$|\tau| = NBI A \sin \theta \quad (18.18)$$

where A is area of the coil $= L \times b$

$$|\tau| = |\mathbf{B}| |\mathbf{M}| \sin \theta \quad (18.19)$$

where $\mathbf{M} = NI \mathbf{A}$ is known as the magnetic moment of the current carrying coil.

Thus, we see that the torque depends on B, A, I, N and θ

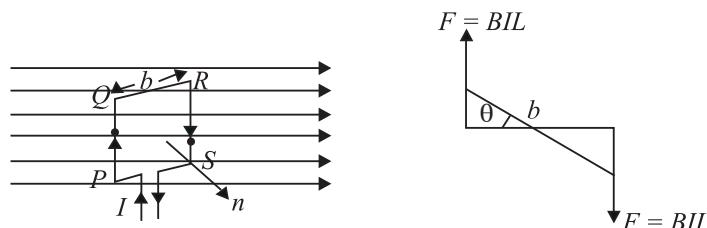


Fig. 18.25 : Torque on the current carrying loop

If a uniform rotation of the loop is desired in a magnetic field, we need to have a constant torque. The couple would be approximately constant if the plane of the



Notes

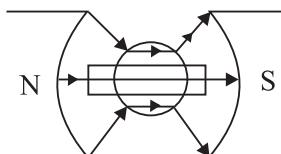


Fig. 18.26: Constant torque on a coil in a radial field

18.6.1 (a) Magnetic Dipole

The term magnetic dipole includes

- (i) a current-carrying circular coil of wire, and
- (ii) a small bar magnet

The magnetic field due to a magnetic dipole at a point

- (i) situated at a distance r on the axis of the dipole is given by :

$$|\mathbf{B}| = \frac{\mu_0}{4\pi} \frac{2\mathbf{M}}{r^3}$$

- (ii) situated at a distance r on the equatorial line is given by :

$$|\mathbf{B}| = \frac{\mu_0}{4\pi} \frac{\mathbf{M}}{r^3}$$

This implies that the field has a cylindrical symmetry about the dipole axis.

18.6.1(b) The Torque on a Magnetic Dipole Placed in a uniform magnetic field

We have seen in section 18.6 that a current loop behaves as a magnetic dipole.

In 18.6.1 we have also seen that a current loop placed in a uniform magnetic field experiences a torque

$$\tau = \mathbf{M} \times \mathbf{B}$$

$$\Rightarrow |\tau| = |\mathbf{M}| \times |\mathbf{B}| \sin \theta$$

The direction of τ is normal to the plane containing \mathbf{M} and \mathbf{B} and is determined by the right hand cork screw rule.

Note that in all these expressions $\mathbf{M} = NIA$.

where the direction of \mathbf{A} is determined by the right hand rule.

18.6.2 Galvanometer

From what you have learnt so far, you can think of an instrument to detect current in any circuit. A device doing precisely this is called a galvanometer, which works on the principle that a current carrying coil, when placed in a magnetic field, experiences a torque.

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A galvanometer consists of a coil wound on a non-magnetic frame. A soft iron cylinder is placed inside the coil. The assembly is supported on two pivots attached to springs with a pointer. This is placed between the pole pieces of a horse shoe magnet providing radial field (see Fig. 18.27).

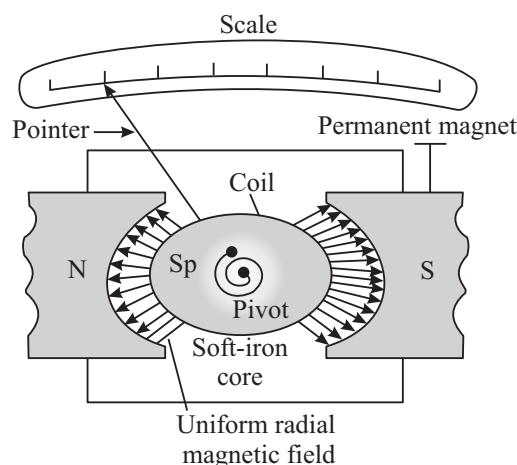


Fig. 18.27 : A moving coil galvanometer

To understand the working of a moving coil galvanometer, we recall that when a current is passed through the coil, it will rotate due to the torque acting on it. The spring sets up a restoring force and hence, a restoring torque. If α is the angle of twist and k is the restoring torque per unit twist or torsional constant, we can write $NBA \sin\theta = k\alpha$. For $\theta = 90^\circ$, $\sin\theta = 1$. So, in the instant case, we can write

$$\therefore NBA = k\alpha$$

$$\text{or } \frac{INBA}{k} = \alpha$$

$$\text{That is, } I = \frac{k\alpha}{NBA} \quad (18.20)$$

where $\frac{k}{NBA}$ is called galvanometer constant. From this we conclude that

$$\alpha \propto I$$

That is, deflection produced in a galvanometer is proportional to the current flowing through it provided N, BA and k are constant. The ratio α/I is known as current sensitivity of the galvanometer. It is defined as the deflection of the coil per unit current. The more the current stronger the torque and the coil turns more. Galvanometer can be constructed to respond to very small currents (of the order of $0.1\mu\text{A}$).



Notes

Sensitivity of a galvanometer : In order to have a more sensitive galvanometer,

- N should be large;
- B should be large, uniform and radial;
- area of the coil should be large; and
- torsional constant should be small.

The values of N and A cannot be increased beyond a certain limit. Large values of N and A will increase the electrical and inertial resistance and the size of the galvanometer. \mathbf{B} can be increased using a strong horse shoe magnet and by mounting the coil on a soft iron core. The value of k can be decreased by the use of materials such as quartz or phospher bronze.

18.6.3 An Ammeter and a Voltmeter

(a) Ammeter : An Ammeter is a suitably shunted galvanometer. Its scale is calibrated to give the value of current in the circuit. To convert a galvanometer into an ammeter, a low resistance wire is connected in parallel with the galvanometer. The resistance of the shunt depends on the range of the ammeter and can be calculated as follows :

Let G be resistance of the galvanometer and N be the number of scale divisions in the galvanometer. Let k denote figure of merit or current for one scale deflection in the galvanometer. Then current which produces full scale deflection in the galvanometer is $I_g = Nk$

Let I be the maximum current to be measured by the galvanometer.

Refer to Fig. 18.28. The voltage between points A and B is given by

$$V_{AB} = I_g G = (I - I_g) S$$

so that

$$S = \frac{I_g G}{I - I_g} \quad (18.21)$$

where S is the shunt resistance.

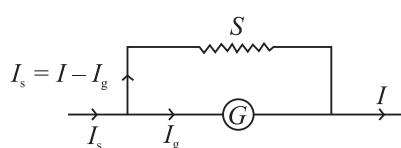


Fig. 18.28 : A shunted galvanometer acts as an ammeter

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As G and S are in parallel, the effective resistance R of the ammeter is given by

$$R = \frac{GS}{G + S}.$$

As the shunt resistance is small, the combined resistance of the galvanometer and the shunt is very low and hence, ammeter resistance is lower than that of the galvanometer. An ideal ammeter has almost negligible resistance. That is why when it is connected in series in a circuit, all the current passes through it without any observable drop.

(b) Voltmeter : A voltmeter is used to measure the potential difference between two points in a circuit. We can convert a galvanometer into a voltmeter by connecting a high resistance in series with the galvanometer coil, as shown in Fig 18.29. The value of the resistance depends on the range of voltmeter and can be calculated as follows :

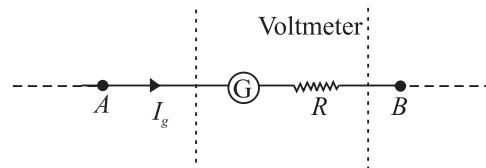


Fig. 18.29 : Galvanometer as a voltmeter

A high resistance, say R is connected in series with the galvanometer coil. If the potential difference across AB is V volt, then total resistance of the voltmeter will be $G + R$. From Ohm's law, we can write

$$\begin{aligned} I_g (G + R) &= V \\ \text{or } G + R &= \frac{V}{I_g} \\ \Rightarrow R &= \frac{V}{I_g} - G \end{aligned} \quad (18.22)$$

This means that if a resistance R is connected in series with the coil of the galvanometer, it works as a voltmeter of range $0-V$ volts.

Now the same scale of the galvanometer which was recording the maximum potential $I_g \times G$ before conversion will record the potential V after conversion into voltmeter. The scale can be calibrated accordingly. The resistance of the voltmeter is higher than the resistance of galvanometer. Effective resistance of the voltmeter, is given by

$$R_v = R + G$$

The resistance of an ideal voltmeter is infinite. It is connected in parallel to the points across which potential drop is to be measured in a circuit. It will not draw



Notes

any current. But the galvanometer coil deflects. Seems impossible! Think about it.

Example 18.6 : A circular coil of 30 turns and radius 8.0 cm, carrying a current of 6.0 A is suspended vertically in a uniform horizontal magnetic field of magnitude 1.0 T. The field lines make an angle of 90° with the normal to the coil. Calculate the magnitude of the counter torque that must be applied to prevent the coil from turning.

Solution : Here, $N = 30$, $I = 6.0 \text{ A}$, $B = 1.0 \text{ T}$, $\theta = 90^\circ$, $r = 8.0 \text{ cm} = 8 \times 10^{-2} \text{ m}$.

$$\text{Area (A) of the coil} = \pi r^2 = \frac{22}{7} \times (8 \times 10^{-2})^2 = 2.01 \times 10^{-2} \text{ m}^2$$

∴

$$\begin{aligned}\text{Torque} &= NI BA \sin\theta \\ &= 30 \times 6 \times 1.0 \times (2.01 \times 10^{-2}) \times \sin 90^\circ \\ &= 30 \times 6 \times (2.01 \times 10^{-2}) \\ &= 3.61 \text{ Nm}\end{aligned}$$

Example 18.7 : A galvanometer with a coil of resistance 12.0Ω shows a full scale deflection for a current of 2.5 mA. How will you convert it into (a) an ammeter of range 0 – 2A, and (b) voltmeter of range 0 – 10 volt?

Solution : (a) Here, $G = 12.0 \Omega$, $I_g = 2.5 \text{ mA} = 2.5 \times 10^{-3} \text{ A}$, and $I = 2 \text{ A}$. From Eqn. (18.21), we have

$$\begin{aligned}S &= \frac{I_g G}{I - I_g} \\ &= \frac{2.5 \times 10^{-3} \times 12}{2 - 2.5 \times 10^{-3}} \\ &= 15 \times 10^{-3} \Omega\end{aligned}$$

So, for converting the galvanometer into an ammeter for reading 0 – 2V, a shunt of $15 \times 10^{-3} \Omega$ resistance should be connected parallel to the coil.

(b) For conversion into voltmeter, let R be the resistance to be connected in series.

$$\begin{aligned}R &= \frac{V}{I_g} - G \\ &= \frac{10}{2.5 \times 10^{-3}} - 12 = 4000 - 12 \\ &= 3988 \Omega\end{aligned}$$

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Thus, a resistance of 3988Ω should be connected in series to convert the galvanometer into voltmeter.



INTEXT QUESTIONS 18.5

1. What is radial magnetic field ?
2. What is the main function of a soft iron core in a moving coil galvanometer ?
3. Which one has the lowest resistance - ammeter, voltmeter or galvanometer ? Explain.
4. A galvanometer having a coil of resistance 20Ω needs 20 mA current for full scale deflection. In order to pass a maximum current of 3A through the galvanometer, what resistance should be added and how ?



WHAT YOU HAVE LEARNT

- Every magnet has two poles. These are inseparable.
- The term magnetic dipole may imply (i) a magnet with dipole moment $\mathbf{M} = m\ell$ (ii) a current carrying coil with dipole moment $\mathbf{M} = NIA$
- Magnetic field at the axis of a magnetic dipole is given by $\mathbf{B} = \frac{\mu_0}{4\pi} \frac{2\mathbf{M}}{x^3}$ and on the equatorial line by $\mathbf{B} = \frac{\mu_0}{4\pi} \frac{2\mathbf{M}}{x^3}$.
- A magnetic dipole behaves the same way in a uniform magnetic field as an electric dipole does in a uniform electric field, i.e., it experience no net force but a torque
$$\tau = \mathbf{M} \times \mathbf{B}$$
.
- Earth has a magnetic field which can be completely described in terms of three basic quantities called elements of earth's magnetic field :
 - angle of inclination,
 - angle of declination, and
 - horizontal component of earth's field.
- Every current carrying conductor develops a magnetic field around it. The magnetic field is given by Biot-Savart's Law :

$$|dB| = \frac{\mu_0 I}{4\pi} \frac{dl \sin\theta}{r^2}$$



Notes

- Unit of magnetic field is tesla.
 - Field at the centre of a flat coil carrying current is given by $|\mathbf{B}| = \frac{\mu_0 I}{2r}$.
- Ampere's circuital law gives the magnitude of the magnetic field around a conductor $\oint \mathbf{B} \cdot d\mathbf{l} = \mu_0 I$
- The Lorentz force on a moving charge q is $\mathbf{F} = q(\mathbf{v} \times \mathbf{B})$ and its direction is given by Fleming's left hand rule.
 - The mechanical force on a wire of length L and carrying a current of I in a magnetic field \mathbf{B} is $\mathbf{F} = \mathbf{B} I L$.
 - Mutual force per unit length between parallel straight conductors carrying currents I_1 and I_2 is given by $\frac{F}{L} = \frac{\mu_0 I_1 I_2}{2\pi r}$.
 - Magnetic field due to a toroid, $B = \frac{\mu_0 Ni}{2\pi r}$
 - A charged particle traces a circular path of radius $R = \frac{mv}{Bq}$.
 - Cyclotron is a device used to accelerate charged particles to high velocities.
 - Cyclotron frequency $v_c = \frac{Bq}{2m\pi}$
 - A current loop behaves like a magnetic dipole.
 - A current carrying coil placed in a magnetic field experiences a torque given by
- $$\tau = N B I A \sin\theta$$
- $$= N B I A, (\text{if } \theta = 90^\circ)$$
- Galvanometer is used to detect electric current in a circuit.
 - An ammeter is a shunted galvanometer and voltmeter is a galvanometer with a high resistance in series. Current is measured by an ammeter and potential difference by a voltmeter.



TERMINAL EXERCISES

1. A small piece of the material is brought near a magnet. Complete the following by filling up the blanks by writing Yes or No.

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Electricity and Magnetism



Notes

Magnetism and Magnetic Effect of Electric Current

| Material | Repulsion | | Attraction | |
|---------------|-----------|--------|------------|--------|
| | weak | strong | weak | strong |
| Diamagnetic | | | | |
| Paramagnetic | | | | |
| Ferromagnetic | | | | |

- You have to keep two identical bar magnets packed together in a box. How will you pack and why?
N S OR N S
N S S N
- The magnetic force between two poles is 80 units. The separation between the poles is doubled. What is the force between them?
- The length of a bar magnet is 10 cm and the area of cross-section is 1.0 cm^2 . The magnetization $I = 10^2 \text{ A/m}$. Calculate the pole strength.
- Two identical bar magnets are placed on the same line end to end with north pole facing north pole. Draw the lines of force, if no other field is present.
- The points, where the magnetic field of a magnet is equal and opposite to the horizontal component of magnetic field of the earth, are called neutral points
 - Locate the neutral points when the bar magnet is placed in magnetic meridian with north pole pointing north.
 - Locate the neutral points when a bar magnet is placed in magnetic meridian with north pole pointing south.
- If a bar magnet of length 10 cm is cut into two equal pieces each of length 5 cm then what is the pole strength of the new bar magnet compare to that of the old one.
- A 10 cm long bar magnet has a pole strength 10 A.m. Calculate the magnetic field at a point on the axis at a distance of 30 cm from the centre of the bar magnet.
- How will you show that a current carrying conductor has a magnetic field around it? How will you find its magnitude and direction at a particular place ?
- A force acts upon a charged particle moving in a magnetic field, but this force does not change the speed of the particle, Why ?
- At any instant a charged particle is moving parallel to a long, straight current carrying wire. Does it experience any force ?



Notes

12. A current of 10 ampere is flowing through a wire. It is kept perpendicular to a magnetic field of 5T. Calculate the force on its $1/10$ m length.
13. A long straight wire carries a current of 12 amperes. Calculate the intensity of the magnetic field at a distance of 48 cm from it.
14. Two parallel wires, each 3m long, are situated at a distance of 0.05 m from each other. A current of 5A flows in each of the wires in the same direction. Calculate the force acting on the wires. Comment on its nature ?
15. The magnetic field at the centre of a 50cm long solenoid is $4.0 \times 10^{-2} \text{ NA}^{-1}$ m⁻¹ when a current of 8.0A flows through it, calculate the number of turns in the solenoid.
16. Of the two identical galvanometer one is to be converted into an ammeter and the other into a milliammeter. Which of the shunts will be of a larger resistance ?
17. The resistance of a galvanometer is 20 ohms and gives a full scale deflection for 0.005A. Calculate the value of shunt required to change it into an ammeter to measure 1A. What is the resistance of the ammeter ?
18. An electron is moving in a circular orbit of radius 5×10^{-11} m at the rate of 7.0×10^{15} revolutions per second. Calculate the magnetic field **B** at the centre of the orbit.
19. Calculate the magnetic field at the centre of a flat circular coil containing 200 turns, of radius 0.16m and carrying a current of 4.8 ampere.
20. Refer to Fig. 18.30 and calculate the magnetic field at A, B and C.

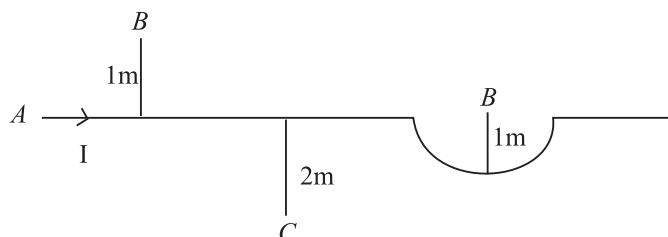


Fig. 18.30



ANSWERS TO INTEXT QUESTIONS

18.1

1. Suspend the magnet with a thread at its centre of mass. Let it come to equilibrium. The end of the magnet which points towards geographical north is its north pole.

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Magnetism



Notes

Magnetism and Magnetic Effect of Electric Current

2. Bring the ends of any two bars closer together. If there is attraction between them, one of the bars is a magnet and the other is an iron bar. Now lay down one of these bars on the table and strock along its length with the other. If uniform force is experienced, the bar in hand is a magnet and that on the table is iron piece. If non-uniform force is experienced, reverse is the case.
3. Suspending one of the bar magnets with thread, we can find its south pole. Then the end of the second magnet, which is repelled by the first, is its south pole.

18.2

1. (i)electrical (ii) magnetic as well as electrical.
2. A conductor in equilibrium is neutral i.e. it has no net electrical current. Due to their random motion, thermal electrons cancel the magnetic fields produced by them.
3. In first case length of wire $l_1 = 2\pi r$ In second case length of wire $l_2 = (2\pi r_2)2$.
But $l_1 = l_2$

$$\therefore 2\pi r = 4\pi r_2 \Rightarrow r_2 = \frac{r}{2}$$

Using $|\mathbf{B}| = \frac{\mu_0 nI}{2r}$

$$|\mathbf{B}_1| = \frac{\mu_0 I}{2r}, \quad |\mathbf{B}_2| = \frac{\mu_0 \cdot 2 \cdot I}{2 \times \frac{r}{2}} = \frac{2\mu_0 I}{r} = 4 \mathbf{B}$$

That is, the magnetic \mathbf{B} at the centre of a coil with two turns is four times stronger than the field in first case.

18.3

1. c
2. Both laws specify magnetic field due to current carrying conductors.
3. (i) B, (ii) A, (iii) C.
4. $B = \mu_0 \frac{n}{\ell} I \Rightarrow 4\pi \times \frac{10^{-7} \times n}{0.1m} \times 3A = 0.002 \text{ or } n = \frac{0.0002 \times 10^7}{12\pi} = 50 \text{ turns}$

**Notes****18.4**

1. The nature of the force will be attractive because the stream of protons is equivalent to electrons in the opposite direction.
2. The force exerted by a magnetic field on a moving charge is perpendicular to the motion of the charge and the work done by the force on the charge is zero. So the KE of the charge does not change. In an electric field, the deflection is in the direction of the field. Hence the field accelerates it in the direction of field lines.
3. The direction of current in each turn of the spring is the same. Since parallel currents in the same direction exert force of attraction, the turns will come closer and the body shall be lifted upward, whatever be the direction of the current in the spring.

18.5

1. Radial magnetic field is one in which plane of the coil remains parallel to it.
2. This increases the strength of magnetic field due to the crowding of magnetic lines of force through the soft iron core, which in turn increases the sensitivity of the galvanometer.
3. Ammeter has the lowest resistance whereas voltmeter has the highest resistance. In an ammeter a low resistance is connected in parallel to the galvanometer coil whereas in a voltmeter, a high resistance is connected in series with it.
4. A low resistance R_s should be connected in parallel to the coil :

$$R_s = \frac{G I_g}{I - I_g} = \frac{20 \times 20 \times 10^{-3}}{3 - 20 \times 10^{-3}} = 0.13 \Omega$$

Answers To Problems in Terminal Exercise

- | | |
|---|---|
| 1. 10^{-2} T m $^{-1}$ | 7. same. |
| 8. 2.3×10^{-6} T | 12. 5 N |
| 13. 5 μ N | 14. attractive force of 10^{-4} N m $^{-1}$ |
| 15. $\frac{625}{\pi}$ turns. | 17. 0.1 Ω . |
| 18. 4.48π T | 19. 1.2π mT |
| 20. $B_A = 2 \times 10^{-7}$ T, $B_B = \pi \times 10^{-7}$ T and $B_C = 10^{-7}$ T. | |

MODULE - 5

Electricity and
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Notes

19

ELECTROMAGNETIC INDUCTION AND ALTERNATING CURRENT

Electricity is the most convenient form of energy available to us. It lights our houses, runs trains, operates communication devices and makes our lives comfortable. The list of electrical appliances that we use in our homes is very long. Have you ever thought as to how is electricity produced?

Hydro-electricity is produced by a generator which is run by a turbine using the energy of water. In a coal, gas or nuclear fuel power station, the turbine uses steam to run the generator. Electricity reaches our homes through cables from the town substation. Have you ever visited an electric sub-station? What are the big machines installed there? These machines are called transformers. Generators and transformers are the devices, which basically make electricity easily available to us. These devices are based on the principle of electromagnetic induction.

In this lesson you will study electromagnetic induction, laws governing it and the devices based on it. You will also study the construction and working of electric generators, transformers and their role in providing electric power to us. A brief idea of eddy current and its application will also be undertaken in this chapter.



OBJECTIVES

After studying this lesson, you should be able to :

- *explain the phenomenon of electromagnetic induction with simple experiments;*
- *explain Faraday's and Lenz's laws of electromagnetic induction;*
- *explain eddy currents and its applications;*



Notes

19.1 ELECTROMAGNETIC INDUCTION

In the previous lesson you have learnt that a steady current in a wire produces a steady magnetic field. Faraday initially (and mistakenly) thought that a steady magnetic field could produce electric current. Some of his investigations on magnetically induced currents used an arrangement similar to the one shown in Fig.19.1. A current in the coil on the left produces a magnetic field concentrated in the iron ring. The coil on the right is connected to a galvanometer G , which can indicate the presence of an induced current in that circuit. It is observed that there is no deflection in G for a steady current flow but when the switch S in the left circuit is closed, the galvanometer shows deflection for a moment. Similarly, when switch S is opened, momentary deflection is recorded but in opposite direction. It means that current is induced only when the magnetic field due to the current in the circuit on the left **changes**.

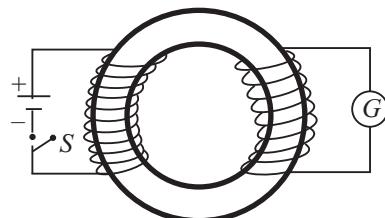


Fig. 19.1: Two coils are wrapped around an iron ring. The galvanometer G deflects for a moment when the switch is opened or closed.

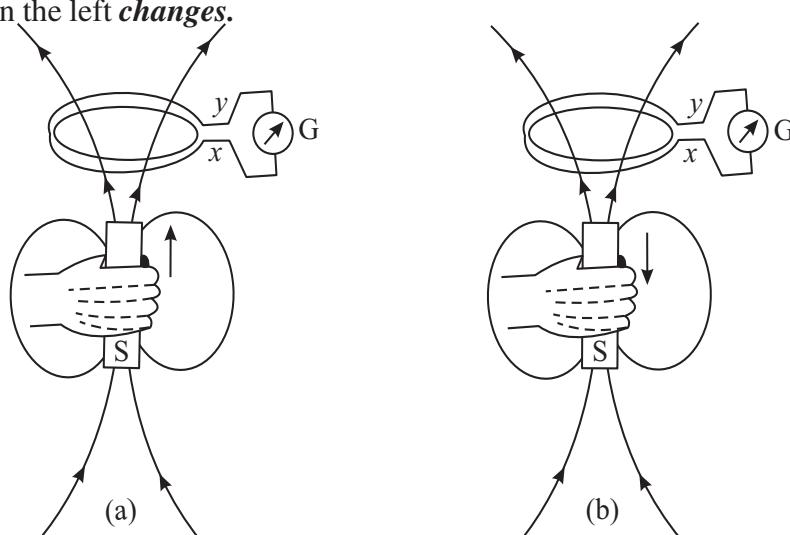


Fig. 19.2 : a) A current is induced in the coil if the magnet moves towards the coil, and
b) the induced current has opposite direction if the magnet moves away from the coil.

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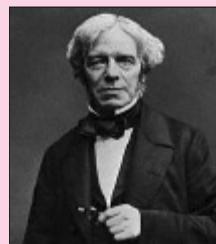


Notes

Electromagnetic Induction and Alternating Current

The importance of a change can also be demonstrated by the arrangement shown in Fig. 19.2. If the magnet is at rest relative to the coil, no current is induced in the coil. But when the magnet is moved towards the coil, current is induced in the direction indicated in Fig. 19.2a. Similarly, if the magnet is moved away from the coil, the current is induced in the opposite direction, as shown in Fig. 19.2b. Note that in both cases, the magnetic field changes in the neighbourhood of the coil. An induced current is also observed to flow through the coil, if this is moved relative to the magnet. The presence of such currents in a circuit implies the existence of an **induced electromotive force (emf)** across the free ends of the coil, i.e., x and y .

This phenomenon in which a magnetic field induces an emf is termed as **electromagnetic induction**. Faraday's genius recognised the significance of this work and followed it up. The quantitative description of this phenomenon is known as Faraday's law of electromagnetic induction. We will discuss it now.



Michael Faraday (1791-1867)

British experimental scientist Michael Faraday is a classical example of a person who became great by sheer hardwork, perseverance, love for science and humanity. He started his carrier as an apprentice with a book binder, but utilized the opportunity to read science books that he received for binding. He sent his notes to Sir Humphry Davy, who immediately recognised the talent in the young man and appointed him his permanent assistant in the Royal Institute.

Sir Humphry Davy once admitted that the greatest discovery of his life was Michael Faraday. And he was right because Faraday made basic discoveries which led to the electrical age. It is because of his discoveries that electrical generators, transformers, electrical motors, and electrolysis became possible.

19.1.1 Faraday's Law of Electromagnetic Induction

The relationship between the changing magnetic field and the induced emf is expressed in terms of magnetic flux ϕ_B linked with the surface of the coil. You will now ask: What is magnetic flux? To define **magnetic flux** ϕ_B refer to Fig. 19.3a, which shows a typical infinitesimal element of area ds , into which the given surface can be considered to be divided. The direction of ds is normal to the surface at that point. By analogy with electrostatics, we can define the magnetic flux $d\phi_B$ for the area element ds as

$$d\phi_B = \mathbf{B} \cdot \mathbf{ds} \quad (19.1a)$$



Notes

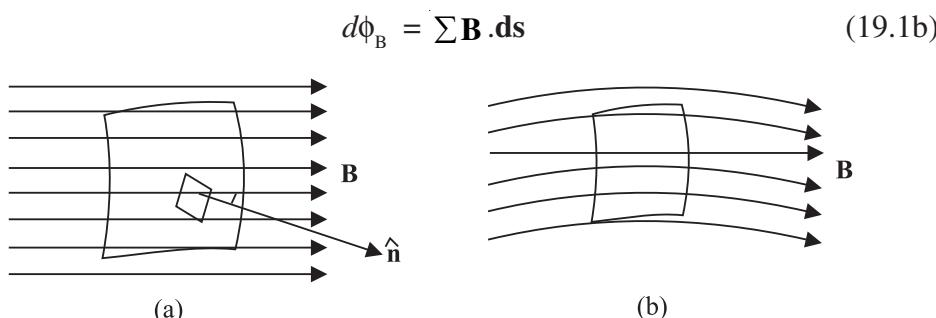


Fig. 19.3: a) The magnetic flux for an infinitesimal area ds is given by $d\phi_B = B.ds$, and b) The magnetic flux for a surface is proportional to the number of lines intersecting the surface.

The SI unit of magnetic flux is **weber** (Wb), where $1 \text{ Wb} = 1 \text{ Tm}^2$.

In analogy with electric lines and as shown in Fig. 19.3b, the number of magnetic lines intersecting a surface is proportional to the magnetic flux through the surface.

Faraday's law states that *an emf is induced across a loop of wire when the magnetic flux linked with the surface bound by the loop changes with time. The magnitude of induced emf is proportional to the rate of change of magnetic flux.* Mathematically, we can write

$$|\varepsilon| = \frac{d\phi_B}{dt} \quad (19.3)$$

From this we note that weber (Wb), the unit of magnetic flux and volt (V), the unit of emf are related as $1 \text{ V} = 1 \text{ Wb s}^{-1}$.

Now consider that an emf is induced in a closely wound coil. Each turn in such a coil behaves approximately as a single loop, and we can apply Faraday's law to determine the emf induced in each turn. Since the turns are in series, the total induced emf ε_r in a coil will be equal to the sum of the emfs induced in each turn. We suppose that the coil is so closely wound that the magnetic flux linking each turn of the coil has the same value at a given instant. Then the same emf ε is induced in each turn, and the total induced emf for a coil with N turns is given by

$$|\varepsilon_r| = N|\varepsilon| = N \left(\frac{d\phi_B}{dt} \right) \quad (19.4)$$

where ϕ_B is the magnetic flux linked with a single turn of the coil.

Let us now apply Faraday's law to some concrete situations.

Example 19.1 : The axis of a 75 turn circular coil of radius 35mm is parallel to a uniform magnetic field. The magnitude of the field changes at a constant rate

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Electricity and
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Electromagnetic Induction and Alternating Current

from 25mT to 50 mT in 250 millisecond. Determine the magnitude of induced emf in the coil in this time interval.

Solution : Since the magnetic field is uniform and parallel to the axis of the coil, the flux linking each turn is given by

$$\phi_B = B\pi R^2$$

where R is radius of a turn. Using Eq. (19.4), we note that the induced emf in the coil is given by

$$|\varepsilon_r| = N \frac{d\phi_B}{dt} = N \frac{d(B\pi R^2)}{dt} = N \pi R^2 \frac{dB}{dt} = N \pi R^2 \left(\frac{B_2 - B_1}{t} \right)$$

Hence, the magnitude of the emf induced in the coil is

$$|\varepsilon_r| = 75\pi (0.035\text{m})^2 (0.10\text{Ts}^{-1}) = 0.030\text{V} = 30\text{mV}$$

This example explains the concept of emf induced by a time changing magnetic field.

Example 19.2 : Consider a long solenoid with a cross-sectional area 8cm^2 (Fig. 19.4a and 19.4b). A time dependent current in its windings creates a magnetic field $B(t) = B_0 \sin 2\pi vt$. Here B_0 is constant, equal to 1.2 T. and v , the frequency of the magnetic field, is 50 Hz. If the ring resistance $R = 1.0\Omega$, calculate the emf and the current induced in a ring of radius r concentric with the axis of the solenoid.

Solution : We are told that magnetic flux

$$\phi_B = B_0 \sin 2\pi vt A$$

since normal to the cross sectional area of the solenoid is in the direction of magnetic field.

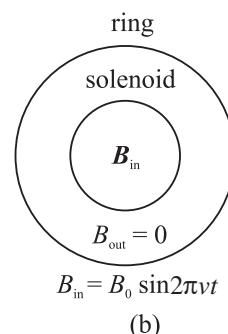
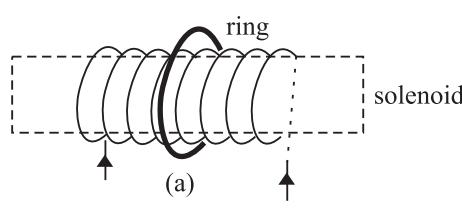


Fig.19.4 : a) A long solenoid and a concentric ring outside it, and b) cross-sectional view of the solenoid and concentric ring.

$$\begin{aligned} \text{Hence } |\varepsilon| &= \frac{d\phi_B}{dt} = 2\pi v A B_0 \cos 2\pi vt \\ &= 2\pi \cdot (50\text{s}^{-1}) \cdot (8 \times 10^{-4}\text{m}^2) \cdot (1.2 \text{ T}) \cos 2\pi vt \\ &= 0.3 \cos 2\pi vt \text{ volts} \\ &= 0.3 \cos 100\pi t \text{ V} \end{aligned}$$



INTEXT QUESTIONS 19.1

1. A 1000 turn coil has a radius of 5 cm. Calculate the emf developed across the coil if the magnetic field through the coil is reduced from 10 T to 0 in (a) 1s (b) 1ms.
2. The magnetic flux linking each loop of a 250-turn coil is given by $\phi_B(t) = A + Dt^2$, where $A = 3 \text{ Wb}$ and $D = 15 \text{ Wbs}^{-2}$ are constants. Show that a) the magnitude of the induced emf in the coil is given by $\epsilon = (2ND)t$, and b) evaluate the emf induced in the coil at $t = 0\text{s}$ and $t = 3.0\text{s}$.
3. The perpendicular to the plane of a conducting loop makes a fixed angle θ with a spatially uniform magnetic field. If the loop has area S and the magnitude of the field changes at a rate dB/dt , show that the magnitude of the induced emf in the loop is given by $\epsilon = (dB/dt) S \cos\theta$. For what orientation(s) of the loop will ϵ be a) maximum and b) minimum?

19.1.2 Lenz's Law

Consider a bar magnet approaching a conducting ring (Fig.19.5a). To apply Faraday's law to this system, we first choose a positive direction with respect to the ring. Let us take the direction from O to Z as positive. (Any other choice is fine, as long as we are consistent.) For this configuration, the positive normal for the area of the ring is in the z -direction and the magnetic flux is negative. As the distance between the conducting ring and the N-pole of the bar magnet decreases, more and more field lines go through the ring, making the flux more and more negative. Thus $d\phi_B/dt$ is negative. By Faraday's law, ϵ is positive relative to our chosen direction. The current I is directed as shown.

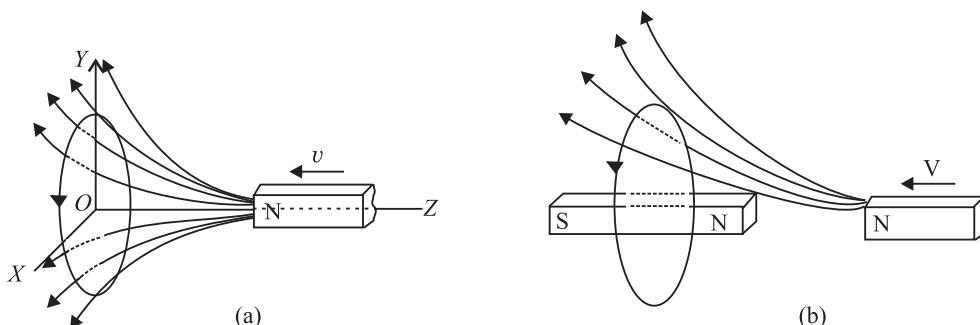


Fig.19.5: a) A bar magnet approaching a metal ring, and b) the magnetic field of the induced current opposes the approaching bar magnet.

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Notes

Electromagnetic Induction and Alternating Current

The current induced in the ring creates a secondary magnetic field in it. This induced magnetic field can be taken as produced by a bar magnet, as shown in Fig. 19.5 (b). Recall that induced magnetic field repels or opposes the original magnetic field. This opposition is a consequence of the law of conservation of energy, and is formalized as Lenz's law. **When a current is induced in a conductor, the direction of the current will be such that its magnetic effect opposes the change that induced it.**

The key word in the statement is 'oppose' - it tells us that we are not going to get something for nothing. When the bar magnet is pushed towards the ring, the current induced in the ring creates a magnetic field that opposes the change in flux. The magnetic field produced by the induced current repels the incoming magnet. If we wish to push the magnet towards the ring, we will have to do work on the magnet. This work shows up as electrical energy in the ring. Lenz's law thus follows from the law of conservation of energy. We can express the combined form of Faraday's and Lenz's laws as

$$\varepsilon = -\frac{d\phi}{dt} \quad (19.5)$$

The negative sign signifies opposition to the cause.

As an application of Lenz's law, let us reconsider the coil shown in Example 19.2. Suppose that its axis is chosen in vertical direction and the magnetic field is directed along it in upward direction. To an observer located directly above the coil, what would be the sense of the induced emf? It will be clockwise because only then the magnetic field due to it (directed downward by the right-hand rule) will oppose the changing magnetic flux. You should learn to apply Lenz's law before proceeding further. Try the following exercise.

19.1.3 Eddy currents

We know that the induced currents are produced in closed loops of conducting wires when the magnetic flux associated with them changes. However, induced currents are also produced when a solid conductor, usually in the form of a sheet or plate, is placed in a changing magnetic field. Actually, induced closed loops of currents are set up in the body of the conductor due to the change of flux linked with it. These currents flow in closed paths and in a direction perpendicular to the magnetic flux. These currents are called eddy currents as they look like eddies or whirlpools and also sometimes called Foucault currents as they were first discovered by Foucault.

The direction of these currents is given by Lenz's law according to which the direction will be such as to oppose the flux to which the induced currents are due. Fig. 19.1.3 shows some of the eddy currents in a metal sheet placed in



Notes

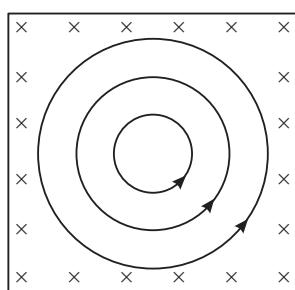


Fig. 19.1.3

The eddy currents produced in metallic bodies encounter little resistance and, therefore, have large magnitude. Obviously, eddy currents are considered undesirable in many electrical appliances and machines as they cause appreciable energy loss by way of heating. Hence, to reduce these currents, the metallic bodies are not taken in one solid piece but are rather made in parts or strips, called lamination, which are insulated from one another.

Eddy currents have also been put to some applications. For example, they are used in induction furnaces for making alloys of different metals in vacuum. They are also used in electric brakes for stopping electric trains.



INTEXT QUESTIONS 19.2

1. The bar magnet in Fig. 19.6 moves to the right. What is the sense of the induced current in the stationary loop A? In loop B?
2. A cross-section of an ideal solenoid is shown in Fig. 19.7. The magnitude of a uniform magnetic field is increasing inside the solenoid and $\mathbf{B} = 0$ outside the solenoid. Which conducting loops is there an induced current? What is the sense of the current in each case?
3. A bar magnet, with its axis aligned along the axis of a copper ring, is moved along its length toward the ring. Is there an induced current in the ring? Is there an induced electric field in the ring? Is there a magnetic force on the bar magnet? Explain.
4. Why do we use laminated iron core in a transformer.

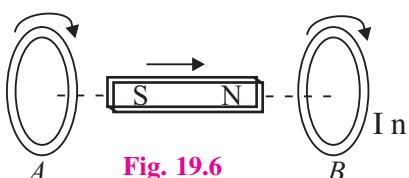


Fig. 19.6

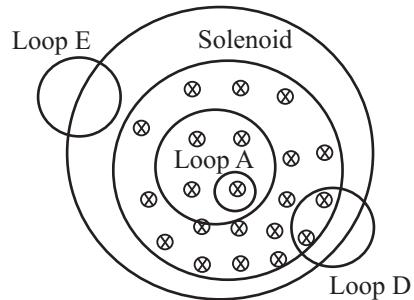


Fig. 19.7



19.2 INDUCTANCE

When current in a circuit changes, a changing magnetic field is produced around it. If a part of this field passes through the circuit itself, current is induced in it. Now suppose that another circuit is brought in the neighbourhood of this circuit. Then the magnetic field through that circuit also changes, inducing an emf across it. Thus, induced emfs can appear in these circuits in two ways:

- By changing current in a coil, the magnetic flux linked with each turn of the coil changes and hence an induced emf appears across that coil. This property is called ***self-induction***.
- for a pair of coils situated close to each other such that the flux associated with one coil is linked through the other, a changing current in one coil induces an emf in the other. In this case, we speak of ***mutual induction*** of the pair of coils.

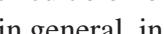
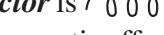
19.2.1 Self-Inductance

Let us consider a loop of a conducting material carrying electric current. The current produces a magnetic field **B**. The magnetic field gives rise to magnetic flux. The total magnetic flux linking the loop is

$$d\phi = \mathbf{B} \cdot d\mathbf{s}$$

In the absence of any external source of magnetic flux (for example, an adjacent coil carrying a current), the Biot-Savart's law tells us that the magnetic field and hence flux will be proportional to the current (*I*) in the loop, i.e.

$$\phi \propto I \quad \text{or} \quad \phi = LI \quad (19.6)$$

where *L* is called self-inductance of the coil. The circuit elements which oppose change in current are called **inductors**. These are in general, in the form of coils of varied shapes and sizes. The symbol for an **inductor** is . If the coil is wrapped around an iron core so as to enhance its magnetic effect, it is symbolised by putting two lines above it, as shown here . The inductance of an indicator depends on its geometry.

(a) Faraday's Law in terms of Self-Inductance: So far you have learnt that if current in a loop changes, the magnetic flux linked through it also changes and gives rise to self-induced emf between the ends. In accordance with Lenz's law, the self-induced emf opposes the change that produces it.

To express the combined form of Faraday's and Lenz's Laws of induction in terms of *L*, we combine Eqns. (19.5) and (19.6) to obtain

$$\varepsilon = -\frac{d\phi}{dt} = -L \frac{dI}{dt} \quad (19.7a)$$

$$= -L \left(\frac{I_2 - I_1}{t} \right) \quad (19.7b)$$



Notes

where I_1 and I_2 respectively denote the initial and final values of current at $t = 0$ and $t = \tau$. Using Eqn. (19.7b), we can define the unit of self-inductance:

$$\begin{aligned}\text{units of } L &= \frac{\text{unit of emf}}{\text{units of } dI/dt} \\ &= \frac{\text{volt}}{\text{ampere / second}} \\ &= \text{ohm-second}\end{aligned}$$

An ohm-second is called a *henry*, (abbreviated H). For most applications, henry is a rather large unit, and we often use millihenry, mH (10^{-3} H) and microhenry μH (10^{-6} H) as more convenient measures.

The self-induced emf is also called the *back emf*. Eqn.(19.7a) tells us that the *back emf in an inductor* depends on the rate of change of current in it and *opposes the change in current*. Moreover, since an infinite emf is not possible, from Eq.(19.7b) we can say that an instantaneous change in the inductor current cannot occur. Thus, we conclude that *current through an inductor cannot change instantaneously*.

The inductance of an inductor depends on its geometry. In principle, we can calculate the self-inductance of any circuit, but in practice it is difficult except for devices with simple geometry. A solenoid is one such device used widely in electrical circuits as inductor. Let us calculate the self-inductance of a solenoid.

(b) Self-inductance of a solenoid : Consider a long solenoid of cross-sectional area A and length ℓ , which consists of N turns of wire. To find its inductance, we must relate the current in the solenoid to the magnetic flux through it. In the preceding lesson, you used Ampere's law to determine magnetic field of a long solenoid:

$$|\mathbf{B}| = \mu_0 n I$$

where $n = N/\ell$ denotes is the number of turns per unit length and I is the current through the solenoid.

The total flux through N turns of the solenoid is

$$\phi = N |\mathbf{B}| A = \frac{\mu_0 N^2 A I}{\ell} \quad (19.8)$$

and self-inductance of the solenoid is

$$L = \frac{\phi}{I} = \frac{\mu_0 N^2 A}{\ell} \quad (19.9)$$

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Using this expression, you can calculate self-inductance and back emf for a typical solenoid to get an idea of their magnitudes.



INTEXT QUESTIONS 19.3

1. A solenoid 1m long and 20cm in diameter contains 10,000 turns of wire. A current of 2.5A flowing in it is reduced steadily to zero in 1.0ms. Calculate the magnitude of back emf of the inductor while the current is being reduced.
2. A certain length (ℓ) of wire, folded into two parallel, adjacent strands of length $\ell/2$, is wound on to a cylindrical insulator to form a type of wire-wound non-inductive resistor (Fig.19.8). Why is this configuration called non-inductive?
3. What rate of change of current in a 9.7 mH solenoid will produce a self-induced emf of 35mV?

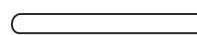


Fig.19.8: Wire wound on a cylindrical insulator

19.2.2 LR Circuits

Suppose that a solenoid is connected to a battery through a switch (Fig.19.9). Beginning at $t = 0$, when the switch is closed, the battery causes charges to move in the circuit. A solenoid has inductance (L) and resistance (R), and each of these influence the current in the circuit. The inductive and resistive effects of a solenoid are shown schematically in Fig.19.10. The inductance (L) is shown in series with the resistance (R). For simplicity, we assume that total resistance in the circuit, including the internal resistance of the battery, is represented by R . Similarly, L includes the self-inductance of the connecting wires. A circuit such as that shown in Fig.19.9, containing resistance and inductance in series, is called an *LR* circuit. The role of the inductance in any circuit can be understood qualitatively. As the current $i(t)$ in the circuit increases (from $i = 0$ at $t = 0$), a self-induced emf $\varepsilon = -L di/dt$ is produced in the inductance whose sense is opposite to the sense of the increasing current. This opposition to the increase in current prevents the current from rising abruptly.

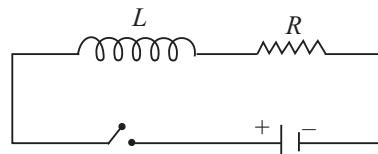
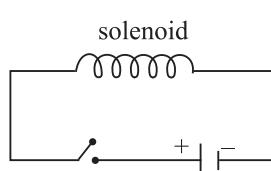


Fig. 19.9: LR Circuit



Notes

If there been no inductance in the circuit, the current would have jumped immediately to the maximum value defined by ϵ_0/R . But due to an inductance coil in the circuit, the current rises gradually and reaches a steady state value of ϵ_0/R as $t \rightarrow \tau$. The time taken by the current to reach about two-third of its steady state value is equal to by L/R , which is called the **inductive time constant** of the circuit. Significant changes in current in an LR circuit cannot occur on time scales much shorter than L/R . The plot of the current with time is shown in Fig. 19.10.

You can see that greater the value of L , the larger is the back emf, and longer it takes the current to build up. (This role of an inductance in an electrical circuit is somewhat similar to that of mass in mechanical systems.) That is why while switching off circuits containing large inductors, you should be mindful of back emf. The spark seen while turning off a switch connected to an electrical appliance such as a fan, computer, geyser or an iron, essentially arises due to back emf.

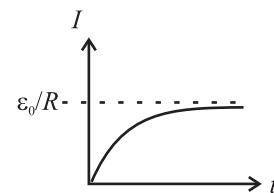


Fig.19.10 : Variation of current with time in a LR -circuit.



INTEXT QUESTIONS 19.4

1. A light bulb connected to a battery and a switch comes to full brightness almost instantaneously when the switch is closed. However, if a large inductance is in series with the bulb, several seconds may pass before the bulb achieves full brightness. Explain why.
2. In an LR circuit, the current reaches 48mA in 2.2 ms after the switch is closed. After sometime the current reaches its steady state value of 72mA. If the resistance in the circuit is 68Ω , calculate the value of the inductance.

19.2.3 Mutual Inductance

When current changes in a coil, a changing magnetic flux develops around it, which may induce emf across an adjoining coil. As we see in Fig. (19.11), the magnetic flux linking each turn of coil B is due to the magnetic field of the current in coil A .

Therefore, a changing current in each coil induces an emf in the other coil, i.e.

$$\text{i.e., } \phi_2 \propto \phi_1 \propto I_1 \Rightarrow \phi_2 = MI_1 \quad (19.10)$$

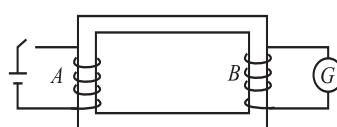


Fig. 19.11 : Mutual inductance of a pair of coils

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where M is called the mutual inductance of the pair of coils. Also back emf induced across the second coil

$$\begin{aligned} e_2 &= -\frac{d\phi}{dt} \\ &= -M \frac{dI}{dt} = -M \left(\frac{I_2 - I_1}{t} \right) \end{aligned} \quad (19.11)$$

where the current in coil A changes from I_1 to I_2 in t seconds.

The mutual inductance depends only on the geometry of the two coils, if no magnetic materials are nearby. The SI unit of mutual inductance is also henry (H), the same as the unit of self-inductance.

Example 19.3 : A coil in one circuit is close to another coil in a separate circuit. The mutual inductance of the combination is 340 mH. During a 15 ms time interval, the current in coil 1 changes steadily from 28mA to 57 mA and the current in coil 2 changes steadily from 36 mA to 16 mA. Determine the emf induced in each coil by the changing current in the other coil.

Solution : During the 15ms time interval, the currents in the coils change at the constant rates of

$$\frac{di_1}{dt} = \frac{57\text{mA} - 23\text{mA}}{15\text{ms}} = 2.3 \text{ As}^{-1}$$

$$\frac{di_2}{dt} = \frac{16\text{mA} - 36\text{mA}}{15\text{ms}} = -1.3 \text{ As}^{-1}$$

From Eq. (19.11), we note that the magnitudes of the induced emfs are

$$\varepsilon_1 = -(340\text{mH})(2.3\text{As}^{-1}) = -0.78 \text{ V}$$

$$\varepsilon_2 = (340\text{mH})(1.3\text{As}^{-1}) = 0.44 \text{ V}$$

Remember that the minus signs in Eq. (19.11) refer to the sense of each induced emf.

One of the most important applications based on the phenomenon of mutual inductance is transformer. You will learn about it later in this lesson. Some commonly used devices based on self-inductance are the choke coil and the ignition coil. We will discuss about these devices briefly. Later, you will also learn that a combination of inductor and capacitor acts as a basic oscillator. Once the capacitor is charged, the charge in this arrangement oscillates between its two plates through the inductor.



INTEXT QUESTIONS 19.5

- Consider the sense of the mutually induced emf's in Fig.19.11, according to an observer located to the right of the coils. (a) At an instant when the current i_1 is increasing, what is the sense of emf across the second coil? (b) At an instant when i_2 is decreasing, what is the sense of emf across the first coil?
- Suppose that one of the coils in Fig.19.11 is rotated so that the axes of the coils are perpendicular to each other. Would the mutual inductance remain the same, increase or decrease? Explain.



Notes

19.3 ALTERNATING CURRENTS AND VOLTAGES

When a battery is connected to a resistor, charge flows through the resistor in one direction only. If we want to reverse the direction of the current, we have to interchange the battery connections. However, the magnitude of the current will remain constant. Such a current is called *direct current*. But a current whose magnitude changes continuously and direction changes periodically, is said to be an *alternating current* (Fig. 19.12).

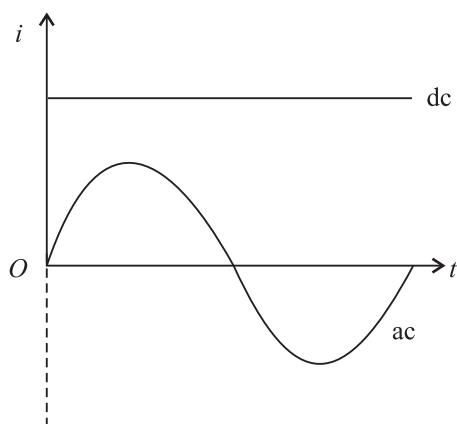


Fig. 19.12 : dc and ac current waveforms

In general, alternating voltage and currents are mathematically expressed as

$$V = V_m \cos \omega t \quad (19.12a)$$

and

$$I = I_m \cos \omega t \quad (19.12b)$$

V_m and I_m are known as the **peak values** of the alternating voltage and current respectively. In addition, we also define the root mean square (*rms*) values of V and I as

$$V_{\text{rms}} = \frac{V_m}{\sqrt{2}} = 0.707 V_m \quad (19.13a)$$

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$$I_{\text{rms}} = \frac{I_m}{\sqrt{2}} = 0.707 I_m \quad (19.13b)$$

The relation between V and I depends on the circuit elements present in the circuit. Let us now study a.c. circuits containing (i) a resistor (ii) a capacitor, and (iii) an inductor only



George Westinghouse

(1846-1914)

If ac prevails over dc all over the world today, it is due to the vision and efforts of George Westinghouse. He was an American inventor and entrepreneur having about 400 patents to his credit. His first invention was made when he was only fifteen year old. He invented air brakes and automatic railway signals, which made railway traffic safe.

When Yugoslav inventor Nicole Tesla (1856-1943) presented the idea of rotating magnetic field, George Westinghouse immediately grasped the importance of his discovery. He invited Tesla to join him on very lucrative terms and started his electric company. The company shot into fame when he used the energy of Niagra falls to produce electricity and used it to light up a town situated at a distance of 20km.

19.3.1 AC Source Connected to a Resistor

Refer to Fig. 19.13 which shows a resistor in an ac circuit. The instantaneous value of the current is given by the instantaneous value of the potential difference across the resistor divided by the resistance.

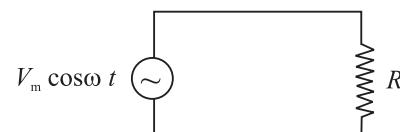


Fig. 19.13 : An ac circuit containing a resistor

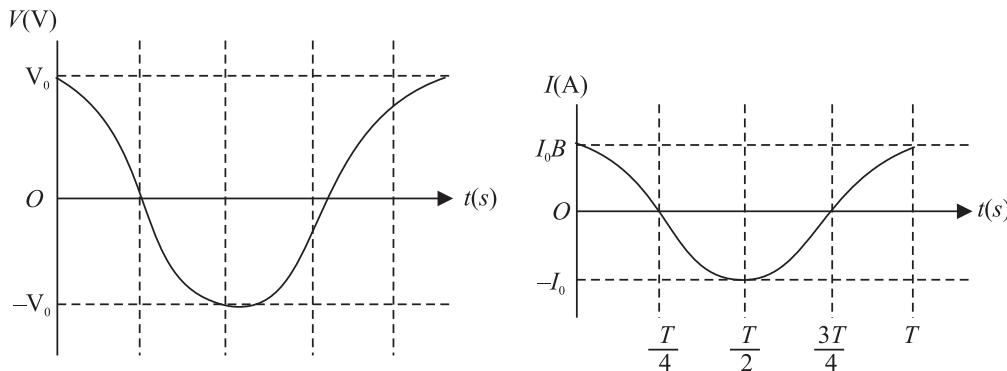
$$\begin{aligned} I &= \frac{V}{R} \\ &= \frac{V_m \cos \omega t}{R} \end{aligned} \quad (19.14a)$$

The quantity V_m/R has units of volt per ohm,(i.e., ampere). It represents the maximum value of the current in the circuit. The current changes direction with time, and so we use positive and negative values of the current to represent the two possible current directions. Substituting I_m , the maximum current in the circuit, for V_m/R in Eq. (19.14a), we get

$$I = I_m \cos \omega t \quad (19.14b)$$



Notes

**Fig. 19.14 :** Time variation of current and voltage in a purely resistive circuit

In India, we have $V_m = 310\text{V}$ and $v = 50 \text{ Hz}$. Therefore for $R = 10 \Omega$, we get

$$V = 310 \cos(2\pi 50t)$$

and

$$\begin{aligned} I &= \frac{310}{10} \cos(100\pi t) \\ &= 31 \cos(100\pi t)\text{A} \end{aligned}$$

Since V and I are proportional to $\cos(100\pi t)$, the average current is zero over an integral number of cycles.

The **average power** $P = I^2R$ developed in the resistor is not zero, because square of instantaneous value of current is always positive. As I^2 varies periodically between zero and I^2_m , we can determine the average power, P_{av} , for single cycle:

$$P_{av} = (I^2 R)_{av} = R(I^2)_{av} = R \left(\frac{I^2_m + 0}{2} \right)$$

$$P_{av} = R \left(\frac{I^2_m}{2} \right) = R I^2_{rms} \quad (19.15)$$

Note that the same power would be produced by a constant *dc* current of value $(I_m/\sqrt{2})$ in the resistor. It would also result if we were to connect the resistor to a potential difference having a constant value of $V_m/\sqrt{2}$ volt. The quantities $I_m/\sqrt{2}$ and $V_m/\sqrt{2}$ are called the rms values of the current and potential difference. The term rms is short for root-mean-square, which means “the square root of the mean value of the square of the quantity of interest.” For an electric outlet in an Indian home where $V_m = 310\text{V}$, the rms value of the potential difference is

$$V_{rms} = V_m/\sqrt{2} \approx 220\text{V}$$

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This is the value generally quoted for the potential difference. Note that when potential difference is 220 V, the peak value of a.c voltage is 310V and that is why it is so fatal.



INTEXT QUESTIONS 19.6

1. In a light bulb connected to an ac source the instantaneous current is zero two times in each cycle of the current. Why does the bulb not go off during these times of zero current?
2. An electric iron having a resistance 25Ω is connected to a 220V, 50 Hz household outlet. Determine the average current over the whole cycle, peak current, instantaneous current and the rms current in it.
3. Why is it necessary to calculate root mean square values of ac current and voltage.

19.3.2 AC Source Connected to a Capacitor

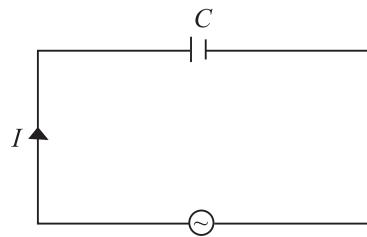


Fig.19.15 shows a capacitor connected to an ac source. From the definition of capacitance, it follows that the instantaneous charge on the capacitor equals the instantaneous potential difference across it multiplied by the capacitance ($q = CV$). Thus, we can write

Fig.19.15 : Capacitor in an ac circuit

$$q = CV_m \cos \omega t \quad (19.16)$$

Since $I = dq/dt$, we can write

$$I = -\omega CV_m \sin \omega t \quad (19.17)$$

Time variation of V and I in a capacitive circuit is shown in Fig.19.16.

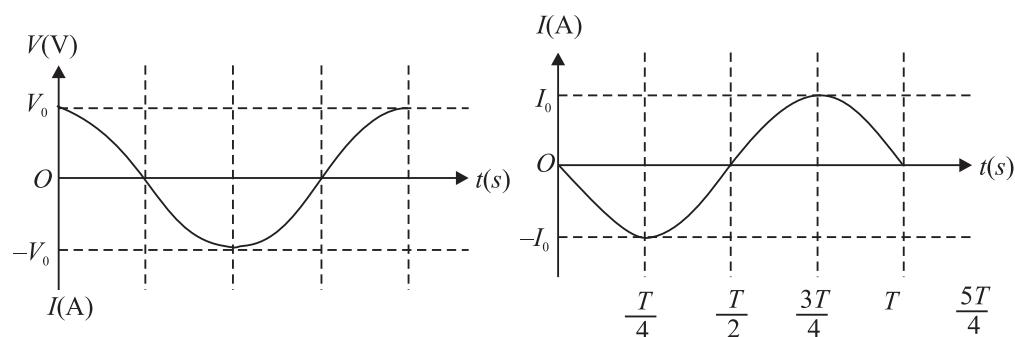


Fig.19.16: Variation in V and I with time in a capacitive circuit

Unlike a resistor, the current I and potential difference V for a capacitor are not in phase. The first peak of the current-time plot occurs one quarter of a cycle before



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the first peak in the potential difference-time plot. Hence we say that the capacitor current leads capacitor potential difference by one quarter of a period. One quarter of a period corresponds to a phase difference of $\pi/2$ or 90° . Accordingly, we also say that the potential difference lags the current by 90° .

Rewriting Eq. (19.17) as

$$I = -\frac{V_m}{1/(\omega C)} \sin \omega t \quad (19.18)$$

and comparing Eqs. (19.14a) and (19.18), we note that $(1/\omega C)$ must have units of resistance. The quantity $1/\omega C$ is called the capacitive reactance, and is denoted by the symbol X_C :

$$\begin{aligned} X_C &= \frac{1}{\omega C} \\ &= \frac{1}{2\pi\nu C} \end{aligned} \quad (19.19)$$

Capacitive reactance is a measure of the extent to which the capacitor limits the ac current in the circuit. It depends on capacitance and the frequency of the generator. The capacitive reactance decreases with increase in frequency and capacitance. Resistance and capacitive reactance are similar in the sense that both measure limitations to ac current. But unlike resistance, capacitive reactance depends on the frequency of the ac (Fig.19.17). The concept of capacitive reactance allows us to introduce an equation analogous to the equation $I = V/R$:

$$I_{\text{rms}} = \frac{V_{\text{rms}}}{X_C} \quad (19.20)$$

The instantaneous power delivered to the capacitor is the product of the instantaneous capacitor current and the potential difference :

$$\begin{aligned} P &= VI \\ &= -\omega CV^2 \sin \omega t \cos \omega t \\ &= -\frac{1}{2}\omega CV^2 \sin 2\omega t \end{aligned} \quad (19.21)$$

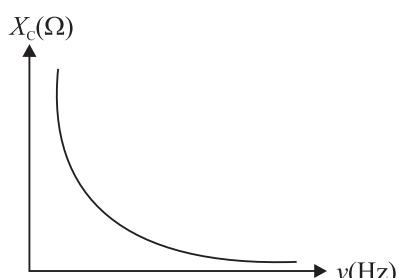


Fig.19.17 : Frequency variation of capacitive reactance

The sign of P determines the direction of energy flow with time. When P is positive, energy is stored in the capacitor. When P is negative, energy is released by the capacitor. Graphical representations of V , I , and P are shown in Fig.19.18. Note

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that whereas both the current and the potential difference vary with angular frequency ω , the power varies with angular frequency 2ω . The average power is zero. The electric energy stored in the capacitor during a charging cycle is completely recovered when the capacitor is discharged. On an average, there is no energy stored or lost in the capacitor in a cycle.

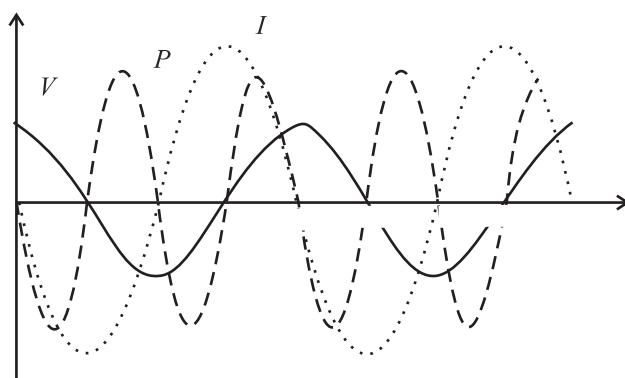


Fig.19.18 : Time variation of V , I and P

Example 19.5 : A $100 \mu\text{F}$ capacitor is connected to a 50Hz ac generator having a peak amplitude of 220V . Calculate the current that will be recorded by an rms ac ammeter connected in series with the capacitor.

Solution : The capacitive reactance of a capacitor is given by

$$X_C = \frac{1}{\omega C} = \frac{1}{2\pi(50\text{rads}^{-1})(100 \times 10^{-6}\text{F})} = 31.8\Omega$$

Assuming that ammeter does not influence the value of current because of its low resistance, the instantaneous current in the capacitor is given by

$$\begin{aligned} I &= \frac{V}{X_C} \cos \omega t = \frac{220}{31.8} \cos \omega t \\ &= (-6.92 \cos \omega t) \text{A} \end{aligned}$$

The rms value of current is

$$\begin{aligned} I_{\text{rms}} &= \frac{I_m}{\sqrt{2}} \\ &= \frac{6.92}{\sqrt{2}} \\ &= 4.91\text{A} \end{aligned}$$

Now answer the following questions.



INTEXT QUESTIONS 19.7

- Explain why current in a capacitor connected to an ac generator increases with capacitance.
- A capacitor is connected to an ac generator having a fixed peak value (V_m) but variable frequency. Will you expect the current to increase as the frequency decreases?
- Will average power delivered to a capacitor by an ac generator to be zero? Justify your answer.
- Why do capacitive reactances become small in high frequency circuits, such as those in a TV set?

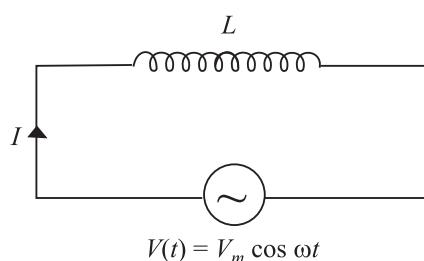
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19.3.3 AC Source Connected to an Inductor

Fig.19.19 : An ac generator connected to an inductor

We now consider an ideal (zero-resistance) inductor connected to an ac source. (Fig. 19.19). If V is the potential difference across the inductor, we can write

$$V(t) = L \frac{dI(t)}{dt} = V_m \cos \omega t \quad (19.22)$$

To integrate Eqn. (19.22) with time, we rewrite it as

$$\int dI = \frac{V_m}{L} \int \cos \omega t dt .$$

Since integral of $\cos x$ is $\sin x$, we get

$$I(t) = \frac{V_m}{\omega L} \sin \omega t + \text{constant} \quad (19.23a)$$

When $t = 0$, $I = 0$. Hence constant of integration becomes zero. Thus

$$I(t) = \frac{V_m}{\omega L} \sin \omega t \quad (19.23b)$$

To compare $V(t)$ and $I(t)$ let us take $V_m = 220V$, $\omega = 2\pi(50) \text{ rads}^{-1}$, and $L = 1H$. Then

$$V(t) = 220 \cos (2\pi 50t) \text{ volt}$$

$$I(t) = \frac{220}{2\pi \cdot 50} \sin (2\pi 50t) = 0.701 \sin (2\pi 50t) \text{ ampere}$$

Fig.19.20. Shows time variation of V and I The inductor current and potential difference across it are not in phase. In fact the potential difference peaks one-quarter cycle before the current. We say that in case of an inductor current lags

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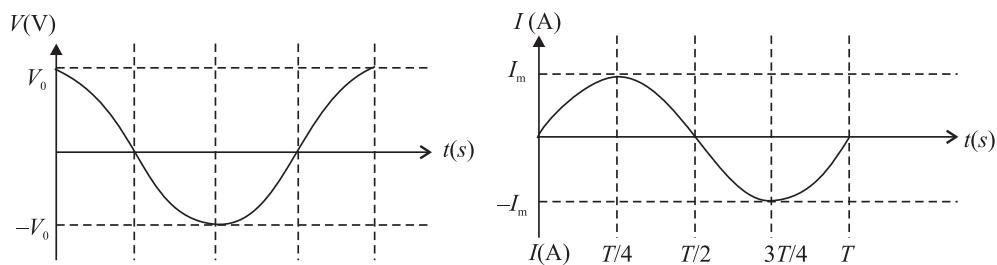


Fig. 19.20 : Time variation of the potential difference across an inductor and the current flowing through it. These are not in phase

the potential difference by $\pi/2$ rad (or 90°). This is what we would expect from Lenz's law. Another way of seeing this is to rewrite Eq. (19.23b) as

$$I = \frac{V_m}{\omega L} \cos\left(\omega t - \frac{\pi}{2}\right)$$

Because $V = V_m \cos \omega t$, the phase difference ($-\pi/2$) for I means that current lags voltage by $\pi/2$. This is in contrast to the current in a capacitor, which leads the potential difference. For an inductor, the current lags the potential difference.

The quantity ωL in Eq.(19.23b) has units of resistance and is called **inductive reactance**. It is denoted by symbol X_L :

$$X_L = \omega L = 2 \pi v L \quad (19.24)$$

Like capacitive reactance, the inductive reactance, X_L , is expressed in ohm. **Inductive reactance** is a measure of the extent to which the inductor limits ac current in the circuit. It depends on the inductance and the frequency of the generator. Inductive reactance increases, if either frequency or inductance increases. (This is just the opposite of capacitive reactance.) In the limit frequency goes to zero, the inductive reactance goes to zero. But recall that as $\omega \rightarrow 0$, capacitive reactance tends to infinity (see Table 19.1). Because inductive effects vanish for a dc source, such as a battery, zero inductive reactance for zero frequency is consistent with the behaviour of an inductor connected to a dc source. The frequency variation of X_L is shown in Fig. 19.21.

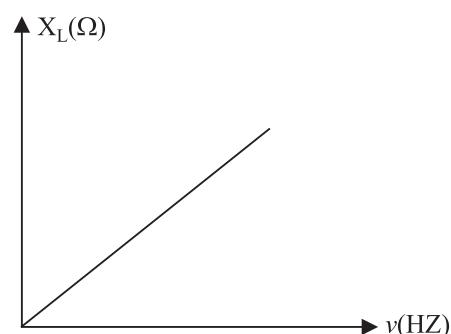


Fig.19.21 : The reactance of an inductor ($X_L = 2\pi v L$) as a function of frequency. The inductive reactance increases as the frequency increases.



Notes

Table 19.1: Frequency response of passive circuit elements

| Circuit element | Opposition to flow of current | Value at low-frequency | Value at high-frequency |
|-----------------|-------------------------------|------------------------|-------------------------|
| Resistor | R | R | R |
| Capacitor | $X_C = \frac{1}{\omega C}$ | ∞ | 0 |
| Inductor | $X_L = \omega L$ | 0 | ∞ |

The concept of inductive reactance allows us to introduce an inductor analog in the equation $I = V/R$ involving resistance R :

$$I_{\text{rms}} = \frac{V_{\text{rms}}}{X_L} \quad (19.25)$$

The instantaneous power delivered to the inductor is given by

$$P = VI$$

$$= \frac{V_m^2}{2\omega L} \sin \omega t \cos \omega t = \frac{V_m^2}{2\omega L} \sin 2\omega t \quad (19.26)$$

Graphical representations of V , I and P for an inductor are shown in Fig. 19.21. Although both the current and the potential difference vary with angular frequency, the power varies with twice the angular frequency. The average power delivered over a whole cycle is zero. Energy is alternately stored and released as the magnetic field alternately grows and windles, decays

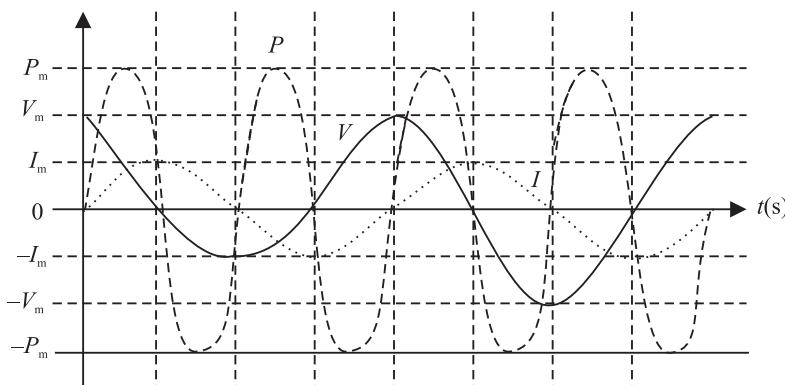


Fig. 19.21: Time variation of potential difference, current and power in an inductive circuit

Example 19.6 : An air cored solenoid has a length of 25cm and diameter of 2.5cm, and contains 1000 closely wound turns. The resistance of the coil is measured to be 1.00Ω . Compare the inductive reactance at 100Hz with the resistance of the coil.

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Solution : The inductance of a solenoid, whose length is large compared to its diameter, is given by

$$L = \frac{\mu_0 N^2 \pi a^2}{\ell}$$

where N denotes number of turns, a is radius, and ℓ is length of the solenoid. On substituting the given values, we get

$$\begin{aligned} L &= \frac{(4\pi \times 10^{-7}) \text{ Hm}^{-1} (1000)^2 \pi (0.0125)^2 \text{ m}^2}{0.25\text{m}} \\ &= 2.47 \times 10^{-3} \text{ H} \end{aligned}$$

The inductive reactance at a frequency of 100Hz is

$$\begin{aligned} X_L &= \omega L = 2\pi \left(100 \frac{\text{rad}}{\text{s}} \right) (2.47 \times 10^{-3}) \text{ H} \\ &= 1.55\Omega \end{aligned}$$

Thus, inductive reactance of this solenoid at 100Hz is comparable to the intrinsic (ohmic) resistance R . In a circuit diagram, it would be shown as

$$L = 2.47 \text{ H} \text{ and } R = 1.00 \Omega$$



You may now like to test your understanding of these ideas.

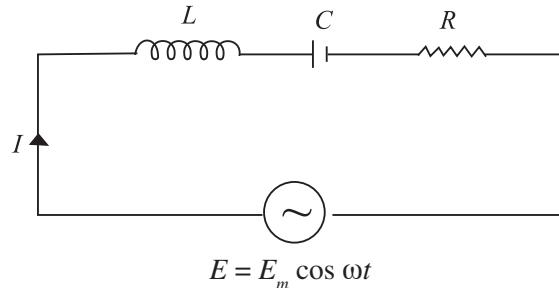


INTEXT QUESTIONS 19.8

1. Describe the role of Lenz's law when an ideal inductor is connected to an ac generator.
2. In section 19.3.1, self-inductance was characterised as electrical inertia. Using this as a guide, why would you expect current in an inductor connected to an ac generator to decrease as the self-inductance increases?

19.3.4 Series LCR Circuit

Refer to Fig. 19.22. It shows a circuit having an inductor L , a capacitor C and a resistor R in series with an ac source, providing instantaneous emf $E = E_m \sin \omega t$. The current through all the three circuit elements is the same in amplitude and phase but potential differences across each of them, as discussed earlier, are not in the same phase. Note that

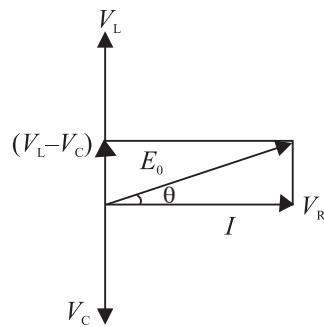
**Notes****Fig. 19.22:** A series **LCR** circuit

- (i) The potential difference across the resistor $V_R = I_0 R$ and it will be in-phase with current.
- (ii) Amplitude of P.D. across the capacitor $V_C = I_0 X_C$ and it lags behind the current by an angle $\pi/2$ and (iii) amplitude of P.D. across the inductor $V_L = I_0 X_L$ and it leads the current by an angle $\pi/2$.

Due to different phases, we can not add voltages algebraically to obtain the resultant peak voltage across the circuit. To add up these voltages, we draw a phasor diagram showing proper phase relationship of the three voltages (Fig.19.23). The diagram clearly shows that voltages across the inductor and capacitor are in opposite phase and hence net voltage across the reactive components is $(V_L - V_C)$. The resultant peak voltage across the circuit is therefore given by

$$\begin{aligned} E_0 &= \sqrt{(V_L - V_C)^2 + V_R^2} \\ &= \sqrt{I_0^2 \{ (X_L - X_C)^2 + R^2 \}} \end{aligned}$$

or
$$\frac{E_0}{I_0} = \sqrt{(X_L - X_C)^2 + R^2}$$

**Fig. 19.23:** Phasor diagram
of voltages across
LCR.

The opposition to flow of current offered by a *LCR* circuit is called its *impedance*. The impedance of the circuit is given by

$$Z = \frac{E_{\text{rms}}}{I_{\text{rms}}} = \frac{E_0}{I_0} = \sqrt{(X_L - X_C)^2 + R^2} = \sqrt{\left(2\pi v L - \frac{1}{2\pi v C}\right)^2 + R^2} \quad (19.27)$$

Hence, the rms current across an *LCR* circuit is given by

$$I_{\text{rms}} = \frac{E_{\text{rms}}}{Z}$$

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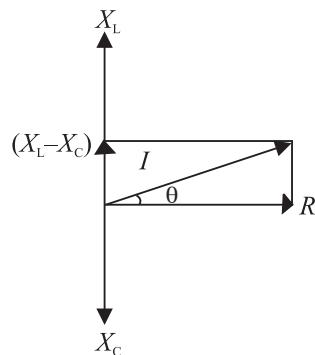


Fig. 19.24 : Phasor diagram for Z

Also from Fig.19.23 it is clear that in LCR circuit, the emf leads (or lags) the current by an angle ϕ , given by

$$\tan\phi = \frac{V_L - V_C}{V_R} = \frac{X_L I_0 - X_C I_0}{R I_0} = \frac{X_L - X_C}{R} \quad (19.28)$$

This means that R , X_L , X_C and Z can also be represented on a phasor diagram similar to voltage (Fig.19.24).

Resonance

You now know that inductive reactance (X_L) increases and capacitive reactance (X_C) decreases with increase in frequency of the applied ac source. Moreover, these are out of phase. Therefore, there may be a certain frequency v_r for which $X_L = X_C$:

$$\begin{aligned} \text{i.e. } 2\pi v_r L &= \frac{1}{2\pi v_r C} \\ \Rightarrow v_r &= \frac{1}{2\pi\sqrt{LC}} \end{aligned} \quad (19.29)$$

This frequency is called *resonance frequency* and at this frequency, impedance has minimum value : $Z_{\min} = R$. The circuit now becomes purely resistive. Voltage across the capacitor and the inductor, being equal in magnitude, annul each other. Since a resonant circuit is purely resistive, the net voltage is in phase with current ($\phi = 0$) and maximum current flows through the circuit. The circuit is said to be in resonance with applied ac. The graphs given in Fig.19.25 show the variation of peak value of current in an LCR circuit with the variation of the frequency of the applied source. The resonance frequency of a given LCR circuit is independent of resistance. But as shown in Fig.19.25, the peak value of current increases as resistance decreases.

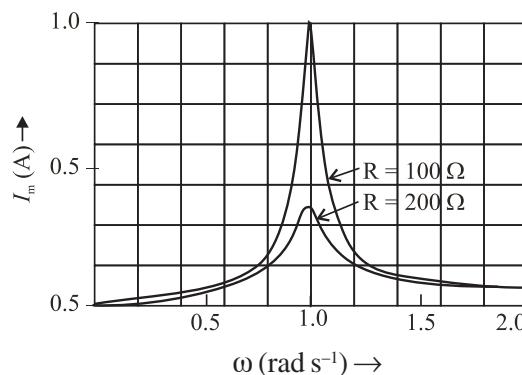


Fig.19.25 : Variation of peak current in a LCR circuit with frequency for (i) $R = 100 \Omega$, and (ii) $R = 200 \Omega$



The phenomenon of resonance in *LCR* circuits is utilised to tune our radio/TV receivers to the frequencies transmitted by different stations. The tuner has an inductor and a variable capacitor. We can change the natural frequency of the *L-C* circuit by changing the capacitance of the capacitor. When natural frequency of the tuner circuit matches the frequency of the transmitter, the intercepting radio waves induce maximum current in our receiving antenna and we say that particular radio/TV station is tuned to it.

Power in a LCR Circuit

You know that a capacitor connected to an ac source reversibly stores and releases electric energy. There is no net energy delivered by the source. Similarly, an inductor connected to an ac source reversibly stores and releases magnetic energy. There is no net energy delivered by the source. However, an ac generator delivers a net amount of energy when connected to a resistor. Hence, when a resistor, an inductor and a capacitor are connected in series with an ac source, it is still only the resistor that causes net energy transfer. We can confirm this by calculating the power delivered by the source, which could be a generator.

The instantaneous power is the product of the voltage and the current drawn from the source. Therefore, we can write

$$P = VI$$

On substituting for V and I , we get

$$\begin{aligned} P &= V_m \cos \omega t \left[\frac{V_m}{Z} \cos (\omega t + \phi) \right] \\ &= \frac{V_m^2}{Z}, \frac{2 \cos \omega t \cos (\omega t + \phi)}{2} \\ &= \frac{V_m^2}{2Z} [\cos \phi + \cos (\omega t + \frac{\phi}{2})] \end{aligned} \quad (19.30)$$

The phase angle ϕ and angular frequency ω play important role in the power delivered by the source. If the impedance Z is large at a particular angular frequency, the power will be small at all times. This result is consistent with the idea that impedance measures how the combination of elements impedes (or limits) ac current. Since the average value of the second term over one cycle is zero, the average power delivered by the source to the circuit is given by

$$\text{Average Power} = \frac{V_m^2}{2Z} \cos \phi \quad (19.31)$$

$$= \frac{V_m}{\sqrt{2}} \cdot \frac{V_m}{\sqrt{2}Z} \cos \phi = V_{\text{rms}} I_{\text{rms}} \cos \phi \quad (19.32)$$

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$\cos\phi$ is called *power factor* and is given by

$$\cos\phi = \frac{R}{Z} = \frac{R}{\sqrt{R^2 + (X_L - X_C)^2}} \quad (19.33)$$

The power factor delimits the maximum average power per cycle provided by the generator. In a purely resistive circuit (or in a resonating circuit where $X_L = X_C$), $Z = R$, so that $\cos\phi = \frac{R}{R} = 1$. That is, when $\phi = 0$, the average power

dissipated per cycle is maximum: $P_m = V_{rms} I_{rms}$.

On the other hand, in a purely reactive circuit, i.e., when $R = 0$, $\cos\phi = 0$ or $\phi = 90^\circ$ and the average power dissipated per cycle $P = 0$. That is, the current in a pure inductor or pure capacitor is maintained without any loss of power. Such a current, therefore, is called *wattless current*.

19.4 POWER GENERATOR

One of the most important sources of electrical power is called **generator**. A *generator* is a device that converts mechanical energy into electrical energy with the help of magnetic field. No other source of electric power can produce as large amounts of electric power as the generator. A conductor or a set of conductors is rotated in a magnetic field and voltage is developed across the rotating conductor due to electromagnetic induction. The energy for the rotation of the conductors can be supplied by water, coal, diesel or gas or even nuclear fuel. Accordingly, we have hydro-generators, thermal generators, and nuclear reactors, respectively.

There are two types of generators

- alternating current generator or A.C. generator also called alternators.
- direct current generator or D.C. generator or dynamo.

Both these generators work on the principle of electromagnetic induction.

19.4.1 A.C. Generator or Alternator

A generator basically consists of a loop of wire rotating in a magnetic field. Refer to Fig. 19.26. It shows a rectangular loop of wire placed in a uniform magnetic field. As the loop is rotated along a horizontal axis, the magnetic flux through the loop changes. To see this, recall that the magnetic flux through the loop, as shown in Fig. 19.26, is given by

$$\phi(t) = \mathbf{B} \cdot \hat{\mathbf{n}} A$$

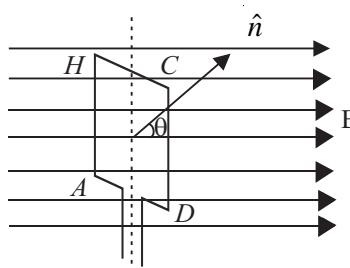


Fig. 19.26 : A loop of wire rotating in a magnetic field.



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where \mathbf{B} is the field, $\hat{\mathbf{n}}$ is a unit vector normal to the plane of the loop of area A . If the angle between the field direction and the loop at any instant is denoted by θ , $\phi(t)$ can be written as

$$\phi(t) = AB \cos \theta$$

When we rotate the loop with a constant angular velocity ω , the angle θ changes as

$$\theta = \omega t \quad (19.34)$$

$$\therefore \phi(t) = AB \cos \omega t$$

Now, using Faraday's law of electromagnetic induction, we can calculate the emf induced in the loop :

$$\varepsilon(t) = -\frac{d\phi}{dt} = \omega AB \sin \omega t \quad (19.35)$$

The emf induced across a coil with N number of turns is given by

$$\begin{aligned} \varepsilon(t) &= N \omega AB \sin \omega t \\ &= \varepsilon_0 \sin \omega t \end{aligned} \quad (19.35a)$$

That is, when a rectangular coil rotates in a uniform magnetic field, the induced emf is sinusoidal.

An A.C. generator consists of four main parts (see in Fig.19.27 : (i) Armature, (ii) Field magnet, (iii) Slip-rings, (iv) Brushes.

An armature is a coil of large number of turns of insulated copper wire wound on a cylindrical soft iron drum. It is capable of rotation at right angles to the magnetic field on a rotor shaft passing through it along the axis of the drum. This drum of soft iron serves two purpose : it supports the coil, and increases magnetic induction through the coil. A field magnet is provides to produce a uniform and permanent radial magnetic field between its pole pieces.

Slip Rings provide alternating current generated in armature to flow in the device connected across them through brushes. These are two metal rings to which the two ends of the armatures are connected. These rings are fixed to the shaft. They are insulated from the shaft as well as from each other. Brushes are two flexible metal or carbon rods [B_1 and B_2 (Fig. 19.27)], which are fixed and constantly in touch with revolving rings. It is with the help of these brushes that the current is passed on from the armature and rings to the main wires which supply the current to the outer circuit.

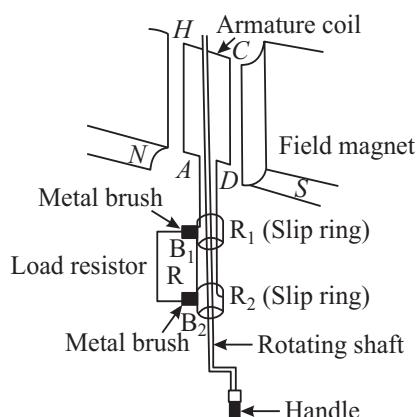


Fig.19.27 : Schematics of an ac generator

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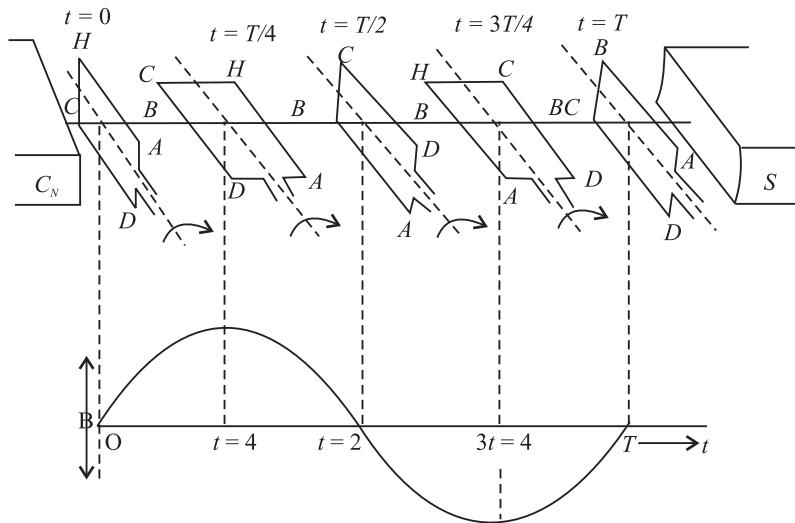


Fig. 19.28 : Working principle of an ac generator

The principle of working of an ac generator is illustrated in Fig.19.28.

Suppose the armature coil $AHCD$ rotates in the anticlockwise direction. As it rotates, the magnetic flux linked with it changes and the current is induced in the coil. The direction of the induced current is given by Fleming's right hand rule. Considering the armature to be in the vertical position and its rotation in anticlockwise direction, the wire AH moves downward and DC moves upwards, the direction of induced emf is from H to A and D to C i.e., in the coil it flows along $DCHA$. In the external circuit the current flows along B_1RB_2 as shown in Fig.19.28(a). This direction of current remains the same during the first half turn of the armature. However, during the second half revolution (Fig.19.28(b)), the wire AH moves upwards while the wires CD moves downwards. The current flows in the direction $AHCD$ in the armature coil i.e., the direction of induced current in the coil is reversed. In the external circuit direction is B_2RB_1 . Therefore, the direction of the induced emf and the current changes after every half revolution in the external circuit also. Hence, the current thus produced alternates in each cycle (Fig. 19.28(c)).

The arrangement of slip rings and brushes creates problems of insulation and sparking when large output powers are involved. Therefore, in most practical generators, the field is rotated and the armature (coil) is kept stationary. In such a generator, armature coils are fixed permanently around the inner circumference of the housing of the generator while the field coil pole pieces are rotated on a shaft within the stationary armature.

19.4.2 Dynamo (DC Generator)

A *dynamo* is a machine in which mechanical energy is changed into electrical energy in the form of direct current. You must have seen a dynamo attached to a bicycle for lighting purpose. In automobiles, dynamo has a dual function for lighting



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and charging the battery. The essential parts of dynamo are (i) field magnet, (ii) armature, (iii) commutator split rings and (iv) brushes.

Armatures and field magnets differ in dynamo and alternator. In the dynamo, the field magnets are stationary and the armature rotates while in an alternator, armature is stationary (stator) and the field magnet (rotor) rotates.

In a dynamo, ac waveform or the sine wave produced by an a.c. generator is converted into d.c. form by the split ring commutator. Each half of the commutator is connected permanently to one end of the loop and the commutator rotates with the loop. Each brush presses against one segment of the commutator. The brushes remain stationary while the commutator rotates. The brushes press against opposite segments of the commutator and every time the voltage reverses polarity, the split rings change position. This means that one brush always remains positive while the other becomes negative, and a d.c. fluctuating voltage is obtained across the brushes.

A dynamo has almost the same parts as an ac dynamo but it differs from the latter in one respect: In place of slip ring, we put two split rings R_1 and R_2 which are the two half of the same ring, as shown in Fig. 19.29(a). The ends of the armature coil are connected to these rings and the ring rotates with the armature and changes the contact with the brushes B_1 and B_2 . This part of the dynamo is known as **commutator**.

When the coil is rotated in the clockwise direction, the current produced in the armature is a.c. but the commutator changes it into d.c. in the outer circuit. In the first half cycle, Fig. 19.29(a), current flows along $DCHA$. The current in the external circuit flows along B_1LB_2 . In the second half, Fig. 19.29(b), current in the armature is reversed and flows along $AHCD$ and as the

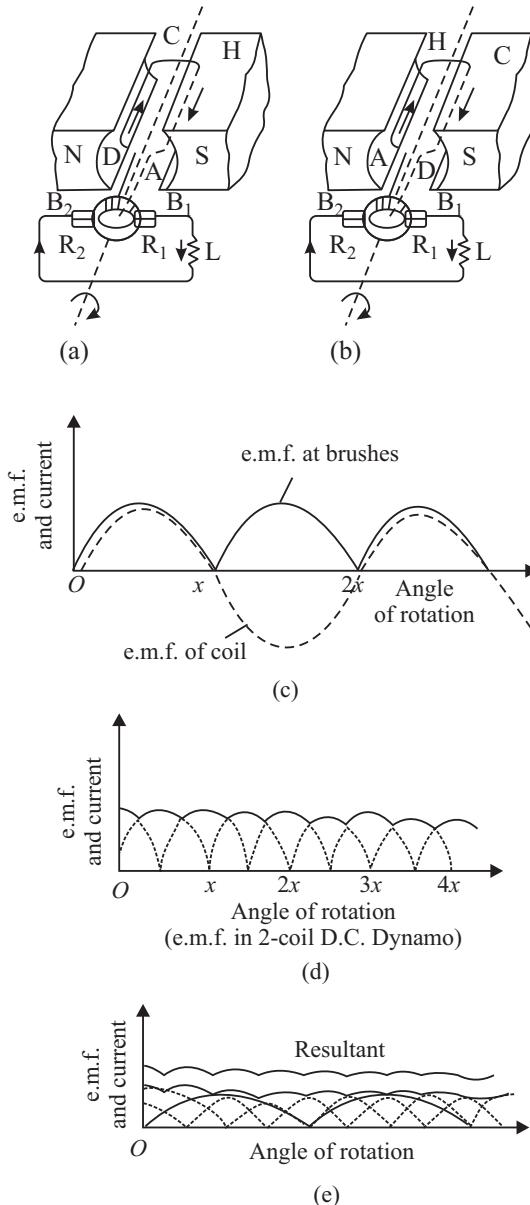


Fig. 19.29 : A dc generator

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ring R_1 comes into contact with B_1 to B_2 . Thus, current in the external circuit always flows in the same direction. The current produced in the outer circuit is graphically represented in Fig.19.29(c) as the coil is rotated from the vertical position, perpendicular to the magnetic lines of force. The current generated by such a simple d.c. dynamo is unidirectional but its value varies considerably and even falls to zero twice during each rotation of the coil.

One way of overcoming this variation would be to use two coils, mutually at right angles, and to divide the commutator ring into four sections, connected to the ends of the coils. In such a case, both these coils produce emf of the same type but they differ in phase by $\pi/2$. The resultant current or emf is obtained by superposition of the two, as shown in Fig. 19.29(d). In this way, the fluctuations are considerably reduced. Similarly, in order to get a steady current, we use a large number of coils, each consisting of good many turns. The commutator ring is divided into as many segments as the number of ends of coils, so that the coils work independently and send current into the outer circuit. The resultant current obtained is shown in Fig.19.29(e) which is practically parallel to the time axes.



INTEXT QUESTIONS 19.9

1. Distinguish between an ac and dc generator.
2. Name the essential parts of a generator?
3. Why do we use a commutator in a dc generator?
4. Where do you find the use of dynamo in daily life?

Low Voltage and Load Shedding

For normal operation of any electrical device, proper voltage is essential. If the voltage supplied by the electric supply company is less than the desired value, we face the problem of low voltage. In fact, low voltage is not as harmful to the appliance as the high voltage. However, due to **low voltage**, most of the appliances do not work properly. To overcome this, use voltage stabilizers. If the low voltage is within the range of the stabilizer, you will get constant voltage. You can use CVT (constant voltage transformers) also to get constant voltage.

As you know, the electricity generated at a power station is transmitted at high voltage to city sub-station. At the sub-station, voltage is reduced using a step down transformer. In order to avoid the danger of burning off the transformers, the supply undertakings try to keep the load on the transformer within the specified rating. If the transformer through which you receive the



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voltage is heavily loaded (more than the specified value), the supplier will either shed the load by cutting the supply from the power source, or request the consumers to decrease the load by switching off the (heating or cooling) appliances of higher wattages. This process is known as **load shedding**.

In case of load shedding, you can use inverters. Inverters are low frequency oscillator circuits which convert direct current from battery to alternating current of desired value and frequency (230V and 50Hz).

19.5 TRANSFORMER

Transformer is a device that changes (increases or decreases) the magnitude of alternating voltage or current based on the phenomenon of electromagnetic induction. A transformer has at least two windings of insulated copper wire linked by a common magnetic flux but the windings are electrically insulated from one another. The transformer windings connected to a supply source, which may be an ac main or the output of a generator, is called **primary winding**. The transformer winding connected to the load R_L is called the secondary winding. In the secondary winding, emf is induced when a.c. is applied to the primary. The primary and secondary windings, though electrically isolated from each other, are magnetically coupled with each other.

Basically, a transformer is a device which transfers electric energy (or power) from primary windings to secondary windings. The primary converts the changing electrical energy into magnetic energy. The secondary converts the magnetic energy back into electric energy.

An ideal transformer is one in which

- the resistance of the primary and secondary coils is zero;
- there is no flux leakage so that the same magnetic flux is linked with each turn of the primary and secondary coils; and
- there is no energy loss in the core.

Fig. 19.30 illustrates the configuration of a typical transformer. It consists of two coils, called primary and secondary, wound on a core (transformer). The coils, made of insulated copper wire, are wound around a ring of iron made of isolated laminated sheets instead of a solid core. The laminations minimize eddy currents in iron. Energy loss in a transformer can be reduced by using the laminations of “soft” iron for the core and thick high conductivity wires for the primary and secondary windings.

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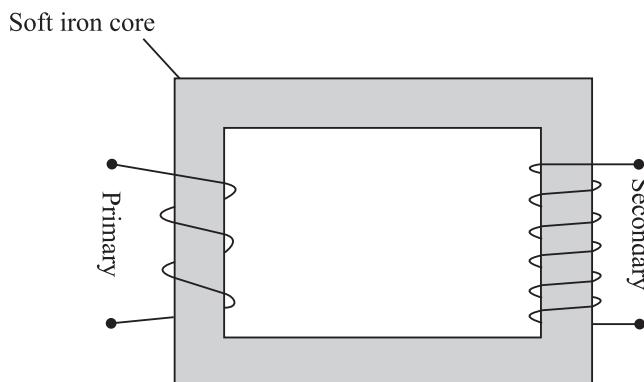


Fig.19.30 : A schematic representation of a transformer

We now discuss the working of a transformer in the following two cases:

(a) Secondary an open circuit : Suppose the current in the primary changes the flux through the core at the rate $d\phi/dt$. Then the induced (back) emf in the primary with N_p turns is given by

$$E_p = -N_p \frac{d\phi}{dt}$$

and the induced emf in the secondary windings of N_s turns is

or $E_s = -N_s \frac{d\phi}{dt}$

$$\frac{E_p}{E_s} = \frac{N_p}{N_s} \quad (19.36)$$

(b) Secondary not an open circuit : Suppose a load resistance R_L is connected across the secondary, so that the secondary current is I_s and primary current is I_p . If there is no energy loss from the system, we can write

$$\text{Power input} = \text{Power output}$$

or $E_p I_p = E_s I_s$

so that $\frac{I_p}{I_s} = \frac{E_s}{E_p} = \frac{N_p}{N_s} = k.$ (19.37)

Thus when the induced emf becomes k times the applied emf, the induced current is $\frac{1}{k}$ times the original current. In other words, what is gained in voltage is lost in current.



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19.5.1 Types of transformers

There are basically two types of transformers.

(i) A **step-up transformer** increases the voltage (decreases the current) in secondary windings. In such transformers (Fig.19.31a) the number of turns in secondary is more than the number of turns in primary.

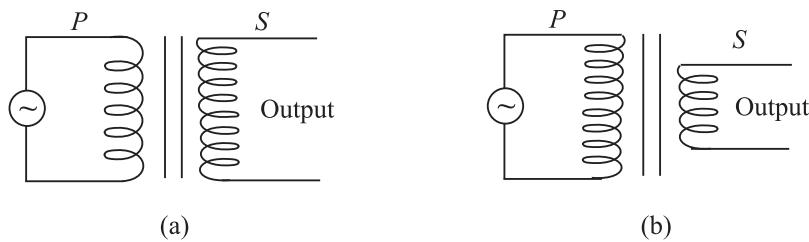


Fig. 19.31 : Iron cored a) step-up, and b) stepdown transformers

(ii) A **step-down transformer** decreases the voltage (increases the current) in the secondary windings. In such transformers (Fig 19.31b), the number of turns in secondary is less than the number of turns in the primary.

19.5.2 Efficiency of Transformers

While discussing the theory of the transformers we considered an ideal transformer in which there is no power loss. But in practice, some energy is always converted into heat in the core and the windings of the transformer. As a result, the electrical energy output across the secondary is less than the electrical energy input. The efficiency of a transformer is given by

$$\eta = \frac{\text{Energy output}}{\text{Energy input}} \times 100\%$$

$$= \frac{\text{Power output}}{\text{Power input}} \times 100\%$$

The efficiency of a transformer is less than 100%.

In a transformer the energy losses result from

- (a) Resistive heating in copper coils - *cooper loss*,
- (b) Eddy current losses in form of heating of iron core - *Eddy current loss*.
- (c) Magnetization heating of the core during repeated reversal of magnetization - *hysteresis loss*.
- (d) Flux leakage from the *core*.



Electrical Power Transmission

You have learnt how electricity is generated using ac or dc generators. You must have come across small units of generating sets in shops, offices and cinema halls. When power goes off, the mains is switched over to generator. In commercial use, generators which produce power of million of watts at about 15kV (kilo volt) is common. These generating plants may be hundreds of kilometers away from your town. Very large mechanical power (kinetic energy) is, therefore, necessary to rotate the rotor which produces magnetic field inside enormously large coils. The rotors are rotated by the turbines. These turbines are driven by different sources of energy.

To minimise loss of energy, power is transmitted at low current in the transmission lines. For this power companies step up voltage using transformers. At a power plant, potential difference is raised to about 330kV. This is accompanied by small current. At the consumer end of the transmission lines, the potential difference is lowered using step down transformers.

You may now like to know how high potential difference used to transmit electrical power over long distances minimises current. We explain this with an example. Suppose electrical power P has to be delivered at a potential difference V by supply lines of total resistance R . The current $I = P/V$ and the loss in the lines is $I^2R = P^2R/V^2$. It means that greater V ensures smaller loss. In fact, doubling V quarters the loss.

Electrical power is, thus, transmitted more economically at high potential difference. But this creates insulation problems and raises installation cost. In a 400kV supergrid, currents of 2500 A are typical and the power loss is about 200kW per kilometer of cable, i.e., 0.02% (percent) loss per kilometer. The ease and efficiency with which alternating potential differences are stepped-up and stepped-down in a transformer and the fact that alternators produce much higher potential difference than d.c. generators (25kV compared with several thousand volts), are the main considerations influencing the use of high alternating rather than direct potential in most situations. However, due to poor efficiency and power thefts, as a nation, we lose about} Rs. 50,000 crore annually.

Example 19.7 : What is the efficiency of a transformer in which the 1880 W of primary power provides for 1730 W of secondary power?

Solution : Given $P_{\text{pri}} = 1880\text{W}$ and $P_{\text{sec}} = 1730\text{W}$. Hence

$$\text{Efficiency} = \frac{P_{\text{sec}}}{P_{\text{pri}}} \times 100$$



Notes

Thus, the transformer is 92% efficient.

Example 19.8 : A transformer has 100 turns in its primary winding and 500 turns in its secondary windings. If the primary voltage and current are respectively 120V and 3A, what are the secondary voltage and current?

Solution : Given $N_1 = 100$, $N_2 = 500$, $V_1 = 120\text{V}$ and $I_1 = 3\text{A}$

$$V_2 = \frac{N_2}{N_1} \times V_1 = \frac{500\text{turns}}{100\text{turns}} \times 120\text{ V} = 600\text{V}$$

$$I_2 = \frac{N_1}{N_2} \times I_1 = \frac{100\text{turns}}{500\text{turns}} \times 3\text{A} = 0.6\text{A}$$



INTEXT QUESTIONS 19.10

1. Can a transformer work on dc? Justify your answer.
2. Why does step-up transformer have more turns in the secondary than in the primary?
3. Is the secondary to primary current ratio same as the secondary to primary voltage ratio in a transformer?
4. Toy trains often use a transformer to supply power for the trains and controls. Is this transformer step-up or step-down?



WHAT YOU HAVE LEARNT

- A current is induced in a coil of wire if magnetic flux linking the surface of the coil changes. This is known as the phenomenon of **electromagnetic induction**.
- The induced emf ϵ in a single loop is given by **Faraday's law**:

$$\epsilon = \frac{d\phi_B}{dt}$$

where ϕ_B is the magnetic flux linking the loop.

- According to **Lenz's Law**, the induced emf opposes the cause which produces it.

MODULE - 5

Electricity and
Magnetism



Notes

Electromagnetic Induction and Alternating Current

- Induced closed loops of current are set up on the body of the conductor (usually a sheet) when it is placed in a changing magnetic field. These currents are called eddy currents.
- If the current changes in a coil, a self-induced emf exists across it.
- For a long, tightly wound solenoid of length ℓ , cross - sectional area A , having N number of turns, the self-inductance is given by

$$L = \frac{\mu_0 N^2 A}{\ell}$$

- Current in an LR circuit takes some time to attain maximum value.
- The changing currents in two nearby coils induce emf mutually.
- In an LC circuit, the charge on the capacitor and the current in the circuit oscillate sinusoidally with the angular frequency ω_0 given by

$$\omega_0 = \frac{1}{\sqrt{LC}}$$

- In an ac circuit, the voltage across the source is given by $V = V_m \cos \omega t$ and current $I = I_m \cos (\omega t + \phi)$
- In a purely resistive ac circuit, the voltage and current are in phase.

The average power in such a circuit is $P_{av} = \frac{I_m^2 R}{2}$

- In a purely capacitive ac circuit, the current leads the voltage by 90° . The average power in such a circuit is zero.
- In a purely inductive ac circuit, the current lags the voltage by 90° . The average power in such a circuit is zero.

- In a series LCR circuit, $I_m = \frac{V_m}{Z} = \frac{V_m}{[R^2 + (X_L - X_C)^2]^{1/2}}$,
where Z is the impedance of circuit : $Z = [R^2 + (X_L - X_C)^2]^{1/2}$
- For $X_L - X_C = 0$, an ac circuit is purely resistive and the maximum current $I_m = V_m/R$. The circuit is said to be in resonance at $\omega_o = 1/\sqrt{LC}$.
- The average power $P_{av} = V_{rms} \cdot I_{rms} = I_{rms}^2 R$.
- A generator converts mechanical energy into electrical energy. It works on the principle of electromagnetic induction.



Notes

- A transformer is a static electrical device which converts an alternating high voltage to low alternating voltage and vice versa.
- The transformers are of two types: Step-up to increase the voltage, and Step-down : to decrease the voltage.
- The secondary to primary voltage ratio is in the same proportion as the secondary to primary turns ratio i.e.

$$\frac{V_s}{V_p} = \frac{N_s}{N_p}$$

- Main sources of power losses in a transformer are heating up of the windings and eddy current
- For transmission of power from a power station to our homes, transformers and transmission lines are used.



TERMINAL EXERCISES

1. Each loop in a 250-turn coil has face area $S = 9.0 \times 10^{-2} \text{ m}^2$. (a) What is the rate of change of the flux linking each turn of the coil if the induced emf in the coil is 7.5V? (b) If the flux is due to a uniform magnetic field at 45° from the axis of the coil, calculate the rate of change of the field to induce that emf.
2. (a) In Fig.19.32 what is the direction of the induced current in the loop when the area of the loop is decreased by pulling on it with the forces labelled F? B is directed into the page and perpendicular to it.
(b) What is the direction of the induced current in the smaller loop of Fig.19.31b when a clockwise current as seen from the left is suddenly established in the larger loop, by a battery not shown?

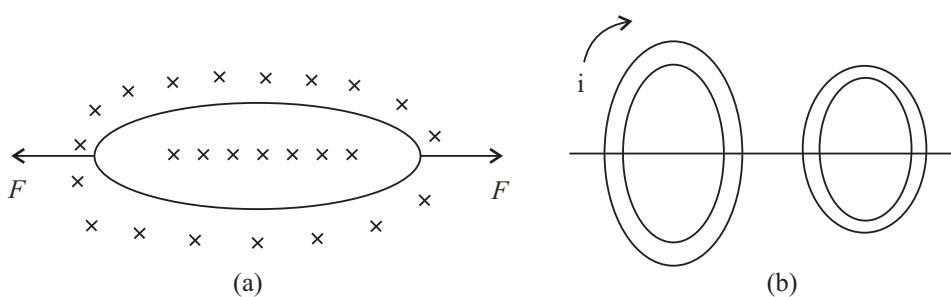


Fig. 19.32

MODULE - 5

Electricity and
Magnetism



Notes

Electromagnetic Induction and Alternating Current

3. (a) If the number of turns in a solenoid is doubled, by what amount will its self-inductance change?
(b) Patrol in a vehicle's engine is ignited when a high voltage applied to a spark plug causes a spark to jump between two conductors of the plug. This high voltage is provided by an ignition coil, which is an arrangement of two coils wound tightly one on top of the other. Current from the vehicle's battery flows through the coil with fewer turns. This current is interrupted periodically by a switch. The sudden change in current induces a large emf in the coil with more turns, and this emf drives the spark. A typical ignition coil draws a current of 3.0 A and supplies an emf of 24kV to the spark plugs. If the current in the coil is interrupted every 0.10ms, what is the mutual inductance of the ignition coil?
4. (a) Why is the rms value of an ac current always less than its peak value?
(b) The current in a $2.5\mu\text{F}$ capacitor connected to an ac source is given by
$$I = -4.71 \sin 377t \mu\text{A}$$
Calculate the maximum voltage across the capacitor.
5. (a) Calculate the capacitive reactance (for $C = 2 \mu\text{F}$) and the inductive reactance (for $L = 2 \text{ mH}$) at (i) 25Hz and (ii) 50Hz.
(b) Calculate the maximum and rms currents in a $22 \mu\text{H}$ inductor connected to a 5V (rms) 100MHz generator.
6. A series LCR circuit with $R = 580\Omega$, $L = 31\text{mH}$, and $C = 47 \text{ nF}$ is driven by an ac source. The amplitude and angular frequency of the source are 65 V and 33 krad/s. Determine (a) the reactance of the capacitor, (b) the reactance of the inductor, (c) the impedance of the circuit, (d) the phase difference between the voltage across the source and the current, and (e) the current amplitude. Does current lead behind or lag the voltage across the source?
7. What is electromagnetic induction? Explain Faraday's laws of electromagnetic induction.
8. State Lenz's law. Show that Lenz's law is a consequence of law of conservation of energy.
9. What is self-induction? Explain the physical significance of self-inductance.
10. Distinguish between the self-inductance and mutual-inductance. On what factors do they depend?
11. How much e.m.f. will be induced in a 10H inductor in which the current changes from 10A to 7A in $9 \times 10^{-2}\text{s}$?
12. Explain why the reactance of a capacitor decreases with increasing frequency, whereas the reactance of an inductor increases with increasing frequency?
13. What is impedance of an LCR series circuit? Derive an expression for power dissipated in a.c. LCR circuit.

**Notes**

14. Suppose the frequency of a generator is increased from 60Hz to 120Hz. What effect would this have on output voltage?
15. A motor and a generator basically perform opposite functions. Yet some one makes a statement that a motor really acts as a motor and a generator at the same time? Is this really true?
16. A light bulb in series with an A.C. generator and the primary winding of a transformer glows dimly when the secondary leads are connected to a load, such as a resistor, the bulb in the primary winding will brighten, why?
17. If the terminals of a battery are connected to the primary winding of transformer, why will a steady potential differences not appear across the secondary windings.
18. The power supply for a picture tube in a colour television (TV) set typically requires 15,000V A.C. How can this potential difference be provided if only 230V are available at a household electric outlet?
19. Would two coils acts as transformer without an iron core? If so, why not omit the core to save money?
20. An ac source has a 10-volt out-put. A particular circuit requires only a 2V A.C. input. How would you accomplish this? Explain.
21. A person has a single transformer with 50 turns on one part of the core and 500 turns on the other. Is this a step-up or a step-down transformer? Explain.
22. Some transformers have various terminals or “taps” on the secondary so that connecting to different tap puts different functions of the total number of secondary windings into a circuit? What is the advantage of this?
23. A transformer in an electric welding machine draws 3A from a 240V A.C. power line and delivers 400A. What is the potential difference across the secondary of the transformer?
24. A 240-V, 400W electric mixer is connected to a 120-V power line through a transformer. What is the ratio of turns in the transformer? and How much current is drawn from the power line?
25. The primary of a step-up transformer having 125 turns is connected to a house lighting circuit of 220 V_{ac} . If the secondary is to deliver 15,000 volts, how many turns must it have?
26. The secondary of a step-down transformer has 25 turns of wire and primary is connected to a 220V ac. line. If the secondary is to deliver 2.5 volt at the out-put terminals, how many turns should the primary have?
27. The primary of a step-down transformer has 600 turns and is connected to a 120V ac line. If the secondary is to supply 5 volts at its terminal and electron current of 3.5A, find the number of turns in the secondary and the electron current in the primary?

MODULE - 5

Electricity and
Magnetism



Notes

Electromagnetic Induction and Alternating Current

28. A step-up transformer with 352 turns in the primary is connected to a 220V ac line. The secondary delivers 10,000 volts at its terminal and a current of 40 milliampere.
- How many turns are in the secondary?
 - What is the current in the primary?
 - What power is drawn from the line?



ANSWERS TO INTEXT QUESTIONS

19.1

1. $N = 1000, r = 5 \times 10^{-2} \text{m}$ and $B_1 = 10 \text{T}$ $B_2 = 0 \text{T}$

a) For $t = 1 \text{s}$,

$$\begin{aligned}|e| &= N \frac{(B_2 - B_1)}{t} \pi r^2 \\&= 10^3 \times \frac{10 \times \pi \times 25 \times 10^{-4}}{1} \\&= 25\pi \text{V} \\&= 25 \times 3.14 = 78.50 \text{V}\end{aligned}$$

b) For $t = 1 \text{ms}$

$$\begin{aligned}|e| &= \frac{10^3 \times 10 \pi \times 25 \times 10^{-4}}{10^{-3}} \\&= 78.5 \times 10^3 \text{V}\end{aligned}$$

2. Since $\phi = A + Dt^2$, $e_1 = \frac{d\phi}{dt} = 2Dt$

$$\begin{aligned}\therefore e &= Ne_1 = 2N Dt \\&= 2 \times 250 \times 15t = 7500t\end{aligned}$$

For $t = 0$, $e_1 = 0$ and hence $e = 0 \text{V}$

For $t = 3 \text{s}$, $e = 22500 \text{V}$

3. $\phi = \mathbf{B} \cdot \mathbf{S} = BS \cos\theta$

**Notes**

$$|e| = N \frac{d\phi}{dt}$$

$$|e| = \left| NS \frac{dB}{dt} \cos\theta \right| \because \theta \text{ is constt}$$

(a) $|e|$ is max.when $\cos\theta = 1$, $\theta = 0$, i.e., The coil is normal to the field.(b) $|e|$ is min.when $\theta = 90^\circ$, i.e. coil surface is parallel to the field.**19.2**

- As we look on the coil from magnet side Anticlockwise for both A and B.
- In all the loops except loop E there is a change in magnetic flux. For each of them the induced current will be anticlockwise
- Yes, there is an induced current in the ring. The bar magnet is acted upon by a repulsive force due to the induced current in the ring.
- To minimise loss of energy due to eddy currents.

19.3

$$\begin{aligned} 1. \quad e &= L \frac{dI}{dt} = \omega \frac{N^2 A}{\ell} \frac{(I_2 - I_1)}{t} \\ &= \frac{4\pi \times 10^{-7} \times \pi \times 10^{-2} \times (2.5 - 0)}{1 \times 10^{-3}} \\ &= 10^{-6} \text{ V} \end{aligned}$$

- Because, current in the two parallel strands flow in opposite directions and oppose the self induced currents and thus minimize the induction effects.

$$\begin{aligned} 3. \quad 3.5 \times 10^{-3} &= 9.7 \times 10^{-3} \times \frac{dI}{dt} \\ &= \frac{dI}{dt} = \frac{3.5}{9.7} = 0.36 \text{ A s}^{-1} \end{aligned}$$

19.4

- Because, the inductor creates an inertia to the growth of current by inducing a back emf

MODULE - 5

Electricity and
Magnetism



Notes

Electromagnetic Induction and Alternating Current

$$2. 2.2 \times 10^{-3} = \frac{L}{R}$$
$$\Rightarrow L = 2.2 \times 68 \times 10^{-3} \text{ H}$$
$$= 150 \text{ mH}$$

19.5

1. (a) If i_1 is increasing, the flux emerging out of the first coil is also increasing. Therefore, the induced current in the second coil will oppose this flux by a current flowing in clockwise sense as seen by O. Therefore B will be positive and A negative.
(b) If i_2 is decreasing, flux emerging out of the first coil is decreasing. To increase it the induced current should flow in out anticlockwise sense leaving C at positive potential and D at negative.
2. No, the mutual inductance will decrease. Because, when the two coils are at right angles coupling of flux from one coil to another coil will be the least.

19.6

1. It actually does but we can not detect it, because the frequency of our domestic ac is 50Hz. Our eye can not detect changes that take place faster than 15 times a second.

$$2. (i) I_{\text{rms}} = \frac{E_{\text{rms}}}{R} = \frac{220}{25} \frac{\text{V}}{\Omega} = 8.8 \text{ A.}$$

$$(ii) \text{ Peak value of current } I_{\text{m}} = \sqrt{2} I_{\text{rms}} = 1.4 \times 8.8 = 12.32 \text{ A.}$$

$$\begin{aligned} \text{Instantaneous current} &= I_0 \sin 2\pi\nu t \\ &= 12.32 \sin 100\pi t \end{aligned}$$

- (iii) Average value of current over integral number of cycles will be zero.
3. Since an ac current varies sinusoidally, its average value over a complete cycle is zero but rms value is finite.

19.7

1. Capacitive reactance $X_C = \frac{1}{2\pi\nu C}$. As C increases X_C decreases and I increases.
2. A charged capacitor takes some time in getting discharged. As frequency of source increases it starts charging the capacitor before it is completely

**Notes**

- discharged. Thus the maximum charge on capacitor and hence maximum current flowing through the capacitor increases though V_m is constant.
3. Because the energy stored in the capacitor during a charging half cycle is completely recovered during discharging half cycle. As a result energy stored in the capacitor per cycle is zero.
 4. Capacitative reactance $X_C = \frac{1}{2\pi vC}$ as v increases X_C decreases. This is so because on capacitor plates now more charge accumulates.

19.8

1. In accordance with Lenz's law a back emf is induced across the inductor when ac is passed through it. The back emf $e = -L \frac{dI}{dt}$.
2. $I_{rms} = \frac{V_{rms}}{X_L}$ frequency increases, $X_L (= 2\pi vL)$ increases, hence I_{rms} decreases.

19.9

1. (i) The a.c. generator has slip rings whereas the d.c. generator has a split rings commutator.
(ii) a.c. generator produces current voltage in sinusoidal form but d.c. generator produces current flowing in one direction all through.
2. Four essential parts of a generator are armature, field magnet, slip rings and brushes.
3. The commutator converts a.c. wave form to d.c. wave form.
4. Attached to the bicycle for lighting purpose.

19.10

1. No, because the working of a transformer depends on the principle of electromagnetic induction, which requires time varying current.
2. Because the ratio of the voltage in primary and secondary coils is proportional to the ratio of number of their turns.
3. No, they are reciprocal to each other.
4. Step-down transformer.

Answers To Problems in Terminal Exercise

1. (a) $3 \times 10^{-2} \text{ W}_b \text{ s}^{-1}$ (b) 0.47 T s^{-1}
4. (b) $5 \times 10^{-2} \text{ V}$

MODULE - 5

Electricity and
Magnetism



Notes

Electromagnetic Induction and Alternating Current

5. (a) (i) $\frac{1}{\pi} \times 10^4 \Omega$ (ii) $\frac{1}{2\pi} \times 10^4 \Omega$
(b) (i) $0.1 \pi \Omega$ (ii) $0.2 \pi \Omega$
6. (a) $6.7 \times 10^2 \Omega$ (b) 99Ω (c) 813.9Ω (d) 4 rad
(e) 0.16 A (f) Current lags
11. 333.3 V 23. 1.8 A .
24. $1 : 2$, $\frac{10}{3} \text{ A}$. 25. 8522 turns
26. 2200 turns 27. 25 turns, $\frac{1}{7} \text{ A}$.
28. (a) 16000 turns, (b) $\frac{20}{11} \text{ A}$ (c) 400 W

SENIOR SECONDARY COURSE

PHYSICS

STUDENT'S ASSIGNMENT – 5

Maximum Marks: 50

Time : 1½ Hours

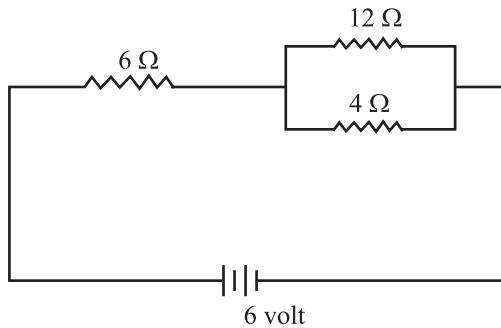
INSTRUCTIONS

- Answer All the questions on a separate sheet of paper
- Give the following information on your answer sheet:
 - Name
 - Enrolment Number
 - Subject
 - Assignment Number
 - Address
- Get your assignment checked by the subject teacher at your study centre so that you get positive feedback about your performance.

Do not send your assignment to NIOS

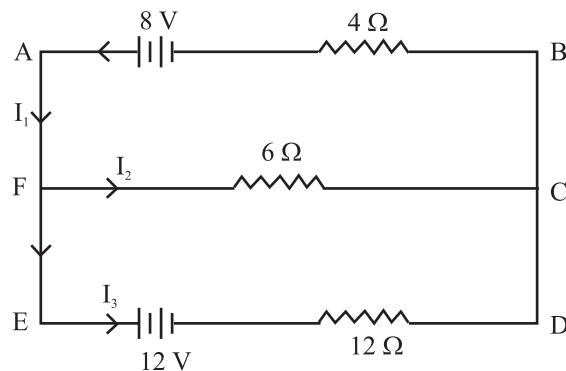
1. Two point charges 10 MC and – 5 MC are kept certain distance apart, along X-axis. If electrostatic force on 10 MC charge is 90 N along +X axis, then what will be the magnitude and direction of electrostatic force on – 5 MC charge? (1)
2. At large distances from an electric dipole, how does the field strength vary with distance? (1)
3. What is the shape of an equipotential surface around a point charge? (1)
4. A copper wire is stretched to double its length. What will be the effect on its specific resistance? (1)
5. Write the sequence of colour bands on a carbon resistor whose value is $4752\ 47\Omega \pm 10\%$. (1)
6. What is the value of angle of dip at the magnetic equator? (1)
7. A wire held vertical in front of an observer is carrying around in the upward direction. With respect to the observer, what will be the direction of the magnetic field at a point behind the wire? (1)
8. Resonant frequency f a series LCR circuit is 1000 Hz. What will be the nature of the circuit at a frequency of 1200 Hz? (1)
9. Calculate electric field strength midway between two point charges 10 MC and – 10 MC kept 2 cm apart in vacuum. (2)
10. A cell of $e.m.f$ 2.0 volt and internal resistance of $4\ \Omega$. Calculate the terminal potential difference of the cell. (2)

11. For the circuit show here, calculate power dissipation in 6Ω resistor. (2)



12. Explain the principle of operation of a step-up transformer with regard to turns ratio. (3)

13. State kirchloffs laws and use them to find values of I_1 , I_2 , and I_3 shown in the given network: (4)



14. State Gauss' theorem in electrostatics. Use it to find field due to a line harge of linear charge density λ . at a distance y from it.

15. (a) Show that when two capacitors of capacitance C_1 and C_2 are joined in series, their net capacitance is given by $C = \frac{1}{C_1} + \frac{1}{C_2}$.

- (b) Find the value of capacitance that must be put in series with a 12 MF capacitor to get a net capacitance of 3 MF. (2 + 2)

16. State prior savarts law and use it to find magnetic field at the centre of a circular loop haring N turns of radius R each and carrying current I.

17. Obtain an expression for tongue experienced by the coil of a galvanometer when current I is passed through it? What purpose does radial magnetic field serve in a moving coil galvanometer?

18. State Faraday's laws of electromagnetic induction and Lenz's law. An air coil is connected in series with a bulb and a battery of suitable e.m.f. What will happen to the brightness of the bulb when an iron rod is inserted in the coil? (4)

19. What is a dielectric? How does it behave in an electrostatic field? How is introduction of a dielectric slab between the plates of an air-filled parallel plate capacitor able to increase its capacitance? Explain with suitable labelled diagrams. (5)

20. A series LCR circuit with $R = 80 \Omega$, $L = 100 \text{ mH}$ and $C = 25\text{MF}$ is driven by an ac source of $e.m.f \ \varepsilon = 100 \sin(1000t)$ volt. Find
- X_p , X_c and Z .
 - rms current in the circuit
 - power dissipation in the circuit
 - resonant frequency
 - impedance of the circuit at resonance
- (5)

Ans :

- 90N along $+x$ axis.
- $E \times \frac{1}{r^3}$
- Spherical, with point at its centre
- Specific resistance will remain unchanged
- Yellow, violet, black, silver
- Zero
- To the left of the observer.
- Inductive
- $1.8 \times 10^9 \text{ NC}^{-1}$

10. $V = \varepsilon - Ir, I = \frac{\varepsilon}{R+r}$ so, $V = 1.6$ volt

11. Current in the circuit, $I = \frac{2}{3}$ ampere

$$P = I^2 R = \left(\frac{2}{3}\right)^2 (6) \text{ watt} = \frac{8}{3} \text{ watts}$$

13. $I_1 = 0.5\text{A}$, $I_2 = 1.0\text{A}$, $I_3 = -0.5\text{A}$

14. $\varepsilon = \frac{\lambda}{2\pi\epsilon_0 r}$

15. (b) 4MF

20. (a) $X_l = 100 \Omega$ $X_c = L_{10} \Omega$, $Z = 100 \Omega$

(b) $I_{rm.s} = 0.707$ ampere

(c) 40 w

(d) 100 Hz

(e) 80 ohm

MODULE - VI
OPTICS AND OPTICAL INSTRUMENTS

- 20 Reflection and Refraction of Light
- 21 Dispersion and Scattering of light
- 22 Wave Phenomena and Light
- 23 Optical Instruments



REFLECTION AND REFRACTION OF LIGHT

Light makes us to see things and is responsible for our visual contact with our immediate environment. It enables us to admire and adore various beautiful manifestations of mother nature in flowers, plants, birds, animals, and other forms of life. Can you imagine how much shall we be deprived if we were visually impaired? Could we appreciate the brilliance of a diamond or the majesty of a rainbow? Have you ever thought how light makes us see? How does it travel from the sun and stars to the earth and what is it made of? Such questions have engaged human intelligence since the very beginning. You will learn about some phenomena which provide answers to such questions.

Look at light entering a room through a small opening in a wall. You will note the motion of dust particles, which essentially provide simple evidence that light travels in a straight line. An arrow headed straight line represents the direction of propagation of light and is called a ray; a collection of rays is called a **beam**. The ray treatment of light constitutes **geometrical optics**. In lesson 22, you will learn that light behaves as a wave. But a wave of short wavelength can be well approximated by the ray treatment. When a ray of light falls on a mirror, its direction changes. This process is called *reflection*. But when a ray of light falls at the boundary of two dissimilar surfaces, it bends. This process is known as *refraction*. You will learn about reflection from mirrors and refraction from lenses in this lesson. You will also learn about *total internal reflection*. These phenomena find a number of useful applications in daily life from automobiles and health care to communication.



OBJECTIVES

After studying this lesson, you should be able to:

- explain reflection at curved surfaces and establish the relationship between the focal length and radius of curvature of spherical mirrors;

MODULE - 6

Optics and Optical Instruments



Notes

Reflection and Refraction of Light

- state sign convention for spherical surfaces;
- derive the relation between the object distance, the image distance and the focal length of a mirror as well as a spherical refractive surface;
- state the laws of refraction;
- explain total internal reflection and its applications in everyday life;
- derive Newton's formula for measuring the focal length of a lens;
- describe displacement method to find the focal length of a lens; and
- derive an expression for the focal length of a combination of lenses in contact.

20.1 REFLECTION OF LIGHT FROM SPHERICAL SURFACES

In your earlier classes, you have learnt the laws of reflection at a plane surface. Let us recall these laws :

Law 1 –The incident ray, the reflected ray and the normal to the reflecting surface at the point of incidence always lie in the same plane.

Law 2 –The angle of incidence is equal to the angle of reflection :

$$\angle i = \angle r$$

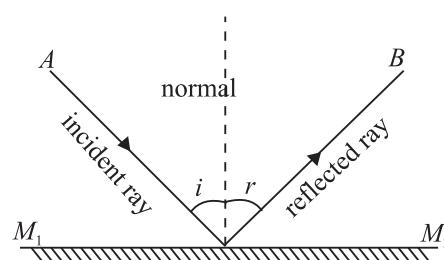


Fig. 20.1 : Reflection of light from a plane surface

These are illustrated in Fig. 20.1. Though initially stated for plane surfaces, these laws are also true for spherical mirrors. This is because a spherical mirror can be regarded as made up of a large number of extremely small plane mirrors. A well-polished spoon is a familiar example of a spherical mirror. Have you seen the image of your face in it? Fig. 20.2(a) and 20.2 (b) show two main types of spherical mirrors.

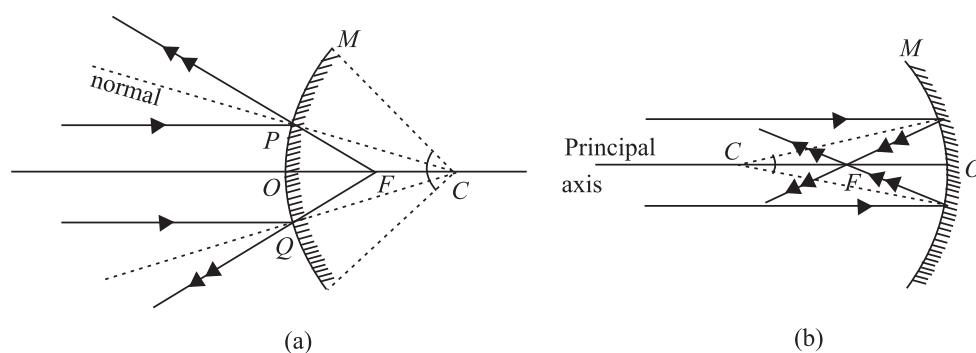


Fig. 20.2 : Spherical mirrors : a) a convex mirror, and b) a concave mirror



Notes

Note that the reflecting surface of a convex mirror curves outwards while that of a concave mirror curves inwards. We now define a few important terms used for spherical mirrors.

The centre of the sphere, of which the mirror is a part, is called the *centre of curvature* of the mirror and the radius of this sphere defines its *radius of curvature*. The middle point *O* of the reflecting surface of the mirror is called its *pole*. The straight line passing through *C* and *O* is said to be the *principal axis* of the mirror. The circular outline (or periphery) of the mirror is called its *aperture* and the angle ($\angle MCM'$) which the aperture subtends at *C* is called the *angular aperture* of the mirror. Aperture is a measure of the size of the mirror.

A beam of light incident on a spherical mirror parallel to the principal axis converges to or appears to diverge from a common point after reflection. This point is known as *principal focus* of the mirror. The distance between the pole and the principal focus gives the *focal length* of the mirror. A plane passing through the focus perpendicular to the principal axis is called the *focal plane*.

We will consider only small aperture mirrors and rays close to the principal axis, called **paraxial rays**. (The rays away from the principal axis are called **marginal** or **parapheral rays**.)



INTEXT QUESTIONS 20.1

1. Answer the following questions :
 - (a) Which mirror has the largest radius of curvature : plane, concave or convex?
 - (b) Will the focal length of a spherical mirror change when immersed in water?
 - (c) What is the nature of the image formed by a plane or a convex mirror?
 - (d) Why does a spherical mirror have only one focal point?
2. Draw diagrams for concave mirrors of radii 5cm, 7cm and 10cm with common centre of curvature. Calculate the focal length for each mirror. Draw a ray parallel to the common principal axis and draw reflected rays for each mirror.
3. The radius of curvature of a spherical mirror is 30cm. What will be its focal length if (i) the inside surface is silvered? (ii) outside surface is silvered?
4. Why are dish antennas curved?

20.1.1 Ray Diagrams for Image Formation

Let us again refer to Fig. 20.2(a) and 20.2(b). You will note that

- the ray of light through centre of curvature retraces its path.

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- the ray of light parallel to the principal axis, on reflection, passes through the focus; and
- the ray of light through F is reflected parallel to the principal axis.

To locate an image, any two of these three rays can be chosen. The images are of two types : real and virtual.

Real image of an object is formed when reflected rays actually intersect. These images are inverted and can be projected on a screen. They are formed on the same side as the object in front of the mirror (Fig. 20.3(a)).

Virtual image of an object is formed by reflected rays that appear to diverge from the mirror. Such images are always erect and virtual; these cannot be projected on a screen. They are formed behind the mirror (Fig. 20.3(b)).

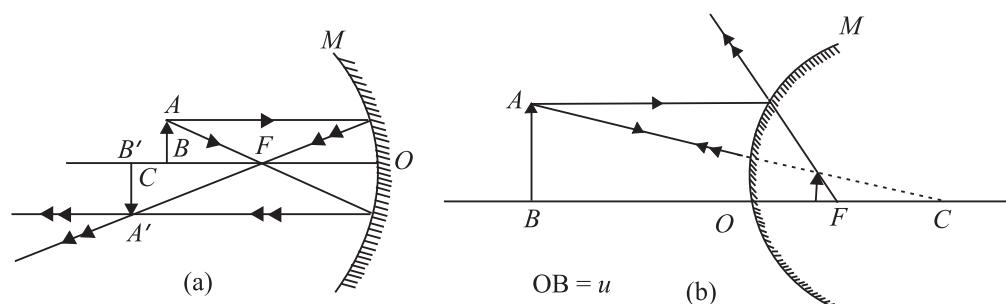


Fig. 20.3 : Image formed by a) concave mirror, and b) convex mirror

20.1.2 Sign Convention

We follow the sign convention based on the cartesian coordinate system. While using this convention, the following points should be kept in mind:

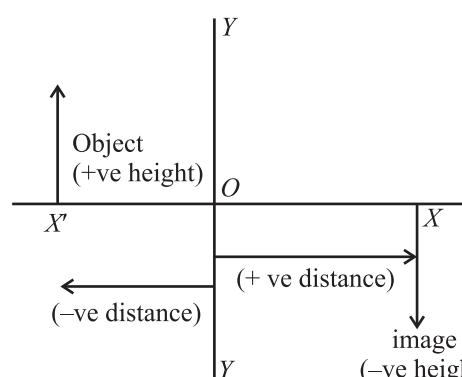


Fig. 20.4 : Sign convention

1. All distances are measured from the pole (O) of the mirror. The object is always placed on the left so that the incident ray is always taken as travelling from left to right.
2. All the distances on the left of O are taken as negative and those on the right of O as positive.
3. The distances measured above and normal to the principal axis are taken as positive and the downward distances as negative.

The radius of curvature and the focal length of a concave mirror are negative and those for a convex mirror are positive.

20.2 DERIVATION OF MIRROR FORMULA

We now look for a relation between the object distance (u), the image distance (v) and the focal length f of a spherical mirror. We make use of simple geometry to arrive at a relation, which surprisingly is applicable in all situations. Refer to Fig. 20.5, which shows an object AB placed in front of a concave mirror. The mirror produces an image $A'B'$.

AX and AY are two rays from the point A on the object AB , M is the concave mirror while XA' and YA' are the reflected rays.

Using sign conventions, we can write

$$\text{object distance, } OB = -u,$$

$$\text{focal length, } OF = -f,$$

$$\text{image distance, } OB' = -v,$$

$$\text{and radius of curvature } OC = -2f$$

Consider ΔABF and ΔFOY . These are similar triangles. We can, therefore, write

$$\frac{AB}{OY} = \frac{FB}{OF} \quad (20.1)$$

Similarly, from similar triangles ΔXOF and $\Delta B'A'F$, we have

$$\frac{XO}{A'B'} = \frac{OF}{FB'} \quad (20.2)$$

But $AB = XO$, as AX is parallel to the principal axis. Also $A'B' = OY$. Since left hand sides of Eqns. (20.1) and (20.2) are equal, we equate their right hand sides. Hence, we have

$$\frac{FB}{OF} = \frac{OF}{FB'} \quad (20.3)$$

Putting the values in terms of u , v and f in Eqn. (20.3), we can write

$$\frac{-u - (-f)}{-f} = \frac{-f}{-v - (-f)}$$

$$\frac{-u + f}{-f} = \frac{-f}{-v + f}$$

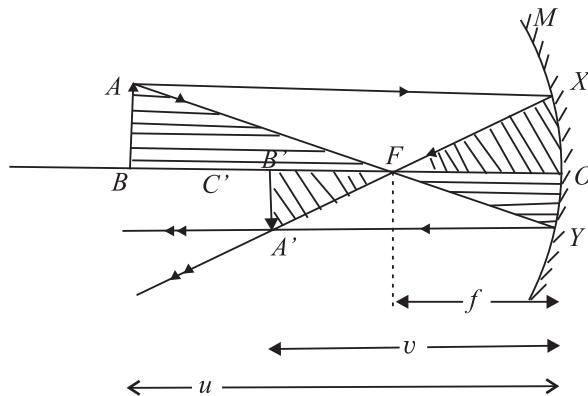


Fig. 20.5 : Image formation by a concave mirror: mirror formula

In optics it is customary to denote object distance by v . You should not confuse it with velocity.

Notes



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Reflection and Refraction of Light

On cross multiplication, we get

$$uv - uf - vf + f^2 = f^2$$

or

$$uv = uf + vf$$

Dividing throughout by uvf , we get the desired relation between the focal length and the object and image distances :

$$\frac{1}{f} = \frac{1}{v} + \frac{1}{u} \quad (20.4)$$

We next introduce another important term **magnification**. This indicates the ratio of the size of image to that of the object :

$$m = \frac{\text{size of the image}}{\text{size of the object}} = \frac{h_2}{h_1}$$

But

$$\frac{A'B'}{AB} = \frac{-v}{-u}$$

$$\therefore m = -\frac{h_2}{h_1} = \frac{v}{u} \quad (20.5)$$

Since a real image is inverted, we can write

$$m = \frac{A'B'}{AB} = -\frac{v}{u} \quad (20.5b)$$

To solve numerical problems, remember the following steps :-

1. For any spherical mirror, use the mirror formula:

$$\frac{1}{f} = \frac{1}{v} + \frac{1}{u}$$

2. Substitute the numerical values of the given quantities with proper signs.
3. Do not give any sign to the quantity to be determined; it will automatically be obtained with the relevant sign.
4. Remember that the linear magnification is negative for a real image and positive for a virtual image.
5. It is always better to draw a figure before starting the (numerical) work.



INTEXT QUESTIONS 20.2

1. A person standing near a mirror finds his head look smaller and his hips larger. How is this mirror made?
2. Why are the shaving mirrors concave while the rear view mirrors convex? Draw ray diagrams to support your answer.



Notes

3. As the position of an object in front of a concave mirror of focal length 25cm is varied, the position of the image also changes. Plot the image distance as a function of the object distance letting the latter change from $-x$ to $+x$. When is the image real? Where is it virtual? Draw a rough sketch in each case.
4. Give two situations in which a concave mirror can form a magnified image of an object placed in front of it. Illustrate your answer by a ray diagram.
5. An object 2.6cm high is 24cm from a concave mirror whose radius of curvature is 16cm. Determine (i) the location of its image, and (ii) size and nature of the image.
6. A concave mirror forms a real image four times as tall as the object placed 15cm from it. Find the position of the image and the radius of curvature of the mirror.
7. A convex mirror of radius of curvature 20cm forms an image which is half the size of the object. Locate the position of the object and its image.
8. A monkey gazes in a polished spherical ball of 10cm radius. If his eye is 20cm from the surface, where will the image of his eye form?

20.3 REFRACTION OF LIGHT

When light passes obliquely from a rarer medium (air) to a denser medium (water, glass), there is a change in its direction of propagation. ***This bending of light at the boundary of two dissimilar media is called refraction.***

When a ray of light is refracted at an interface, it obeys the following two laws :

Law I : The incident ray, the refracted ray and the normal to the surface at the point of incidence always lie in the same plane.

Law II : The ratio of the sine of the angle of incidence to the sine of the angle of refraction is constant for a given pair of media. It is independent of the angle of incidence when light propagates from a rarer to a denser medium. Moreover, for a light of given colour, the ratio depends only on the pair of media.

This law was pronounced by the Dutch scientist Willebrord van Roijen Snell and in his honour is often referred to as ***Snell's law***. According to Snell's law

$$\frac{\sin i}{\sin r} = \mu_{12}$$

where μ_{12} is a constant, called the *refractive index* of second medium with respect to the first medium, and determines how much bending would take place at the interface separating the two media. It may also be expressed as the ratio of the

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Table 20.1 : Refractive indices of some common materials

| Medium | μ |
|-------------------|-------|
| Vacuum/air | 1 |
| Water | 1.33 |
| Ordinary glass | 1.50 |
| Crown glass | 1.52 |
| Dense flint glass | 1.65 |
| Diamond | 2.42 |

Reflection and Refraction of Light

velocity of light in the first medium to the velocity of light in the second medium

$$\mu_{12} = \frac{c_1}{c_2}$$

Refractive indices of a few typical substances are given in Table 20.1. Note that these values are with respect to air or vacuum. The medium having larger refractive index is optically denser medium while the one having smaller refractive index is called rarer medium. So water is denser than air but rarer than glass. Similarly, crown glass is denser than ordinary glass but rarer than flint glass.

If we consider refraction from air to a medium like glass, which is optically denser than air [Fig. 20.6 (a)], then $\angle r$ is less than $\angle i$. On the other hand, if the ray passes from water to air, $\angle r$ is greater than $\angle i$ [Fig. 20.6(b)]. That is, the refracted ray bends towards the normal on the air–glass interface and bends away from the normal on water–air interface.

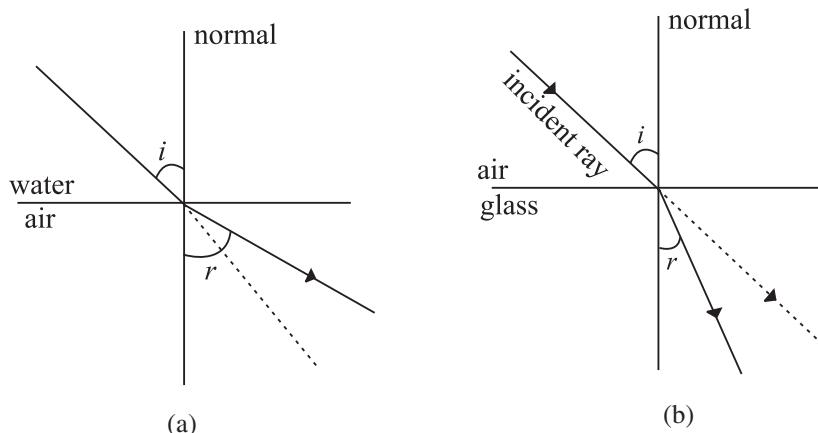


Fig. 20.6 : a) Refraction on air–glass interface, and b) refraction on water–air interface

Willebrord Van Rijen Snell (1580 – 1626)

Willebrord Snell was born in 1580 in Lieden. He began to study mathematics at a very young age. He entered the University of Leiden and initially studied law. But, soon he turned his attention towards mathematics and started teaching at the university by the time he was 20. In 1613, Snell succeeded his father as Professor of Mathematics.



He did some important work in mathematics, including the method of calculating the approximate value of π by polygon. His method of using 96



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sided polygon gives the correct value of π up to seven places while the classical method only gave this value upto two correct places. Snell also published some books including his work on comets. However, his biggest contribution to science was his discovery of the laws of refraction. However, he did not publish his work on refraction. It became known only in 1703, seventy seven years after his death, when Huygens published his results in "Dioptrics".

20.3.1 Reversibility of light

Refer to Fig. 20.6(b) again. It illustrates the principle of reversibility. It appears as if the ray of light is retracing its path. It is not always necessary that the light travels from air to a denser medium. In fact, there can be any combination of transparent media. Suppose, light is incident at a water-glass interface. Then, by applying Snell's law, we have

$$\frac{\sin i_w}{\sin i_g} = \mu_{wg} \quad (20.6)$$

Now, let us consider separate air-glass and air-water interfaces. By Snell's law, we can write

$$\frac{\sin i_a}{\sin i_g} = \mu_{ag}$$

and

$$\frac{\sin i_a}{\sin i_w} = \mu_{aw}$$

On combining these results, we get

$$\mu_{ag} \sin i_g = \mu_{aw} \sin i_w \quad (20.7)$$

This can be rewritten as

$$\frac{\sin i_w}{\sin i_g} = \frac{\mu_{ag}}{\mu_{aw}} \quad (20.8)$$

On comparing Eqns. (20.6) and (20.8), we get

$$\mu_{wg} = \frac{\mu_{ag}}{\mu_{aw}} \quad (20.9)$$

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This result shows that when light travels from water to glass, the refractive index of glass with respect to water can be expressed in terms of the refractive indices of glass and water with respect to air.

Example 20.1 : A ray of light is incident at an angle of 30° at a water-glass interface. Calculate the angle of refraction. Given $\mu_{\text{ag}} = 1.5$, $\mu_{\text{aw}} = 1.3$.

Solution : From Eqn. (20.8), we have

$$\frac{\sin i_w}{\sin i_g} = \frac{\mu_{\text{ag}}}{\mu_{\text{aw}}}$$

$$\frac{\sin 30^\circ}{\sin i_g} = \frac{1.5}{1.3}$$

or
$$\sin i_g = \left(\frac{1.3}{1.5}\right) \times \frac{1}{2}$$

$$= 0.4446$$

or
$$i_g = 25^\circ 41'$$

Example 20.2 : Calculate the speed of light in water if its refractive index with respect to air is $4/3$. Speed of light in vacuum = $3 \times 10^8 \text{ ms}^{-1}$.

Solution : We know that

$$\mu = \frac{c}{v}$$

or
$$v = \frac{c}{\mu}$$

$$= \frac{(3 \times 10^8 \text{ ms}^{-1})}{4/3}$$

$$= \frac{3 \times 10^8 \times 3}{4}$$

$$= 2.25 \times 10^8 \text{ ms}^{-1}$$

Example 20.3 : The refractive indices of glass and water are 1.52 and 1.33 respectively. Calculate the refractive index of glass with respect to water.

Solution : Using Eqn. (20.9), we can write

$$\mu_{\text{wg}} = \frac{\mu_{\text{ag}}}{\mu_{\text{aw}}} = \frac{1.52}{1.33} = 1.14$$



INTEXT QUESTIONS 20.3

- What would be the lateral displacement when a light beam is incident normally on a glass slab?
- Trace the path of light if it is incident on a semicircular glass slab towards its centre when $\angle i < \angle i_c$ and $\angle i > \angle i_c$.
- How and why does the Earth's atmosphere alter the apparent shape of the Sun and Moon when they are near the horizon?
- Why do stars twinkle?
- Why does a vessel filled with water appear to be shallower (less deep) than when without water? Draw a neat ray diagram for it.
- Calculate the angle of refraction of light incident on water surface at an angle of 52° . Take $\mu = 4/3$.

Notes



20.4 TOTAL INTERNAL REFLECTION



ACTIVITY 20.1

Take a stick, cover it with cycle grease and then dip it in water or take a narrow glass bottle, like that used for keeping Homeopathic medicines, and dip it in water. You will observe that the stick or the bottle shine almost like silver. Do you know the reason? This strange effect is due to a special case of refraction. We know that when a ray of light travels from an optically denser to an optically rarer medium, say from glass to air or from water to air, the refracted ray bends away from the normal. This means that the angle of refraction is greater than the angle of incidence. What happens to the refracted ray when the angle of incidence is increased? The bending of refracted ray also increases. However, the maximum value of the angle of refraction can be 90° . ***The angle of incidence in the denser medium for which the angle of refraction in rarer medium, air in this case, equals 90° is called the critical angle, i_c .*** The refracted ray then moves along the boundary separating the two media. If the angle of incidence is greater than the critical angle, the incident ray will be reflected back in the same medium, as shown in Fig. 20.7(c). Such a reflection is called ***Total Internal Reflection*** and the incident ray is said to be totally internally reflected. For total internal reflection to take place, the following two conditions must be satisfied.

- Light must travel from an optically denser to an optically rarer medium.
- The angle of incidence in the denser medium must be greater than the critical angle for the two media.

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The glass tube in water in Activity 20.1 appeared silvery as total internal reflection took place from its surface.

An expression for the critical angle in terms of the refractive index may be obtained readily, using Snell's law. For refraction at the glass-air interface, we can write

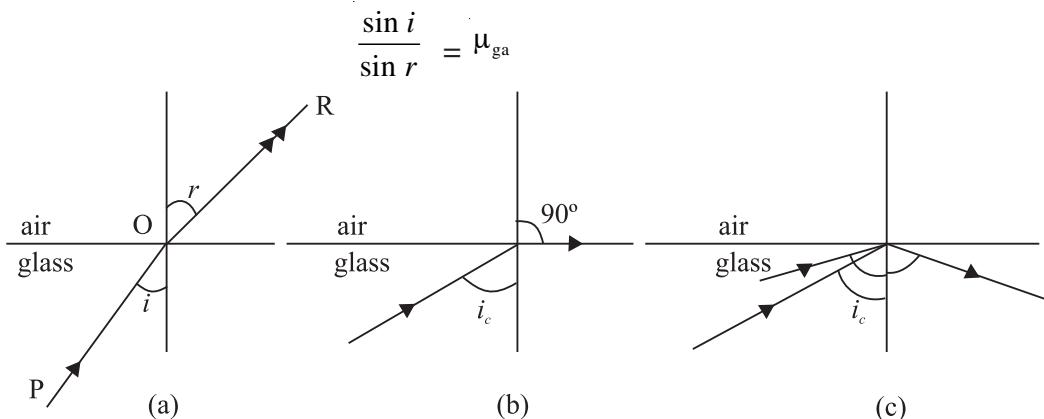


Fig. 20.7 : Refraction of light as it travels from glass to air for a) $i < i_c$, b) $i = i_c$ and c) $i > i_c$

Table 20.2 :Critical angles for a few substances

| Substance | μ | Critical angle |
|-------------------|-------|----------------|
| Water | 1.33 | 48.75° |
| Crown glass | 1.52 | 41.14° |
| Diamond | 2.42 | 24.41° |
| Dense flint glass | 1.65 | 37.31° |

Putting $r = 90^\circ$ for $i = i_c$, we have

$$\frac{\sin i_c}{\sin 90^\circ} = \mu_{ga}$$

or

$$\sin i_c = \mu_{ga}$$

Hence

$$\mu_{ag} = \frac{1}{\mu_{ga}} = \frac{1}{\sin i_c}$$

The critical angles for a few substances are given in Table 20.2

Example 20.4 : Refractive index of glass is 1.52. Calculate the critical angle for glass air interface.

Solution : We know that

$$\mu = 1/\sin i_c$$

$$\sin i_c = 1/\mu = \frac{1}{1.52}$$

\therefore

$$i_c = 42^\circ$$

Much of the shine in transparent substances is due to total internal reflection. Can you now explain why diamonds sparkle so much? This is because the critical angle is quite small and most of the light entering the crystal undergoes multiple internal reflections before it finally emerges out of it.

In ordinary reflection, the reflected beam is always weaker than the incident beam, even if the reflecting surface is highly polished. This is due to the fact that some



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20.4.1 Applications of Refraction and Total Internal Reflection

There are many manifestations of these phenomena in real life situations. We will consider only a few of them.

(a) Mirage : Mirage is an optical illusion which is observed in deserts or on tarred roads in hot summer days. This, you might have observed, creates an illusion of water, which actually is not there.

Due to excessive heat, the road gets very hot and the air in contact with it also gets heated up. The densities and the refractive indices of the layers immediately above the road are lower than those of the cooler higher layers. Since there is no abrupt change in medium (see Fig. 20.9), a ray of light from a distant object, such as a tree, bends more and more as it passes through these layers. And when it falls on a layer at an angle greater than the critical angle for the two consecutive layers, total internal reflection occurs. This produces an inverted image of the tree giving an illusion of reflection from a pool of water.

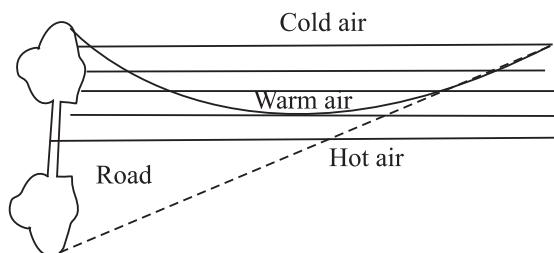


Fig. 20.8 : Formation of mirage

Totally Reflecting Prisms : A prism with right angled isosceles triangular base or a totally reflecting prism with angles of 90° , 45° and 45° is a very useful device for reflecting light.

Refer to Fig. 20.9(a). The symmetry of the prism allows light to be incident on O at an angle of 45° , which is greater than the critical angle for glass i.e. 42° . As a result, light suffers total internal reflection and is deviated by 90° .

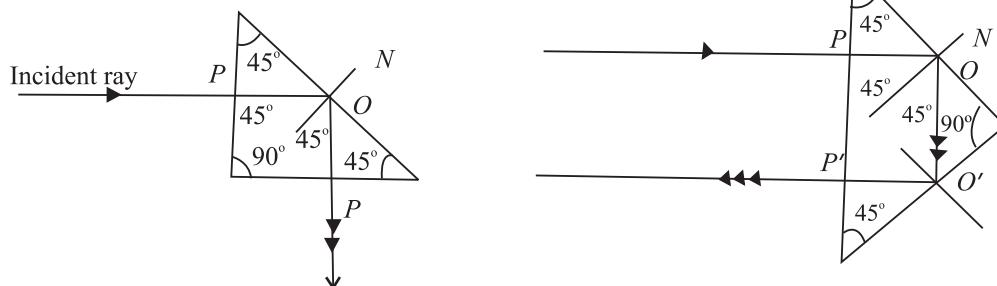


Fig. 20.9 : Totally reflecting prisms

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Reflection and Refraction of Light

Choosing another face for the incident rays, it will be seen (Fig. 20.9(b)) that the ray gets deviated through 180° by two successive total internal reflections taking place at O and O' .

Optical Fibres

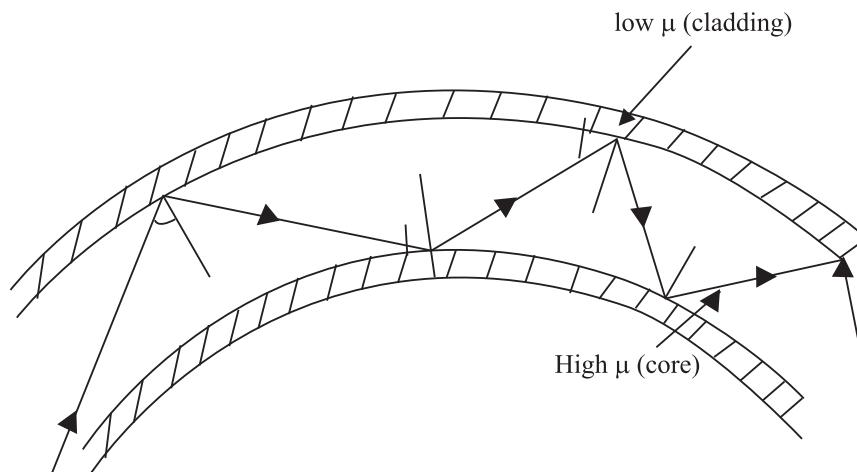


Fig. 20.10 : Multiple reflection in an optical fibre

An optical fibre is a hair-thin structure of glass or quartz. It has an inner core which is covered by a thin layer (called **cladding**) of a material of slightly lower refractive index. For example, the refractive index of the core is about 1.7 and that of the cladding is 1.5. This arrangement ensures total internal reflection. You can easily understand it, if you recall the conditions for total internal reflection.

When light is incident on one end of the fibre at a small angle, it undergoes multiple total internal reflections along the fibre (Fig. 20.10). The light finally emerges with undiminished intensity at the other end. Even if the fibre is bent, this process is not affected. Today optical fibres are used in a big way. A flexible light pipe using optical fibres may be used in the examination of inaccessible parts of the body e.g. laproscopic examination of stomach, urinary bladder etc. Other medical applications of optical fibres are in neurosurgery and study of bronchi. Besides medical applications, optical fibres have brought revolutionary changes in the way we communicate now. Each fibre can carry as many as 10,000 telephone messages without much loss of intensity, to far off places. That is why millions of people across continents can interact simultaneously on a fibre optic network.



INTEXT QUESTIONS 20.4

1. Why can't total internal reflection take place if the ray is travelling from a rarer to a denser medium?
2. Critical angle for glass is 42° . Would this value change if a piece of glass is immersed in water? Give reason for your answer.



Notes

20.5 REFRACTION AT A SPHERICAL SURFACE

We can study formation of images of objects placed around spherical surfaces such as glass marbles (Kanchas), water drops, glass bottle, etc. For measuring distances from spherical refracting surfaces, we use the same sign convention as applicable for spherical mirrors. Refer to Fig. 20.11.

SPS' is a convex refracting surface separating two media, air and glass. C is its centre of curvature. P is a point on SPS' almost symmetrically placed. You may call it the *pole*. CP is then the *principal axis*. O is a point object. OA is an incident ray and AB is the refracted ray. Another ray OP falls on the surface normally and goes undeviated after refraction. PC and AB appear to come from I . Hence I is the virtual image of O .

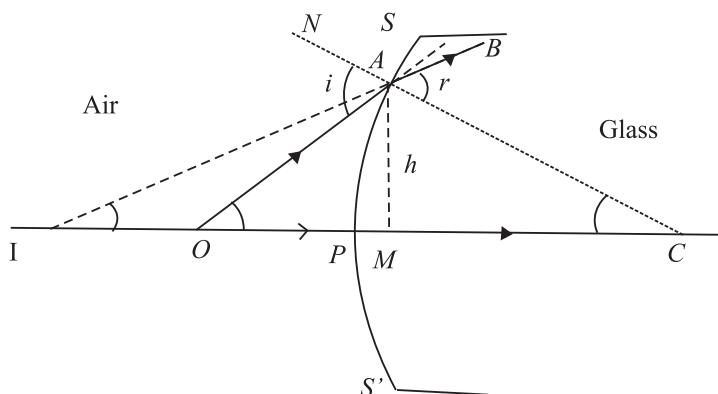


Fig. 20.11 : Refraction at a spherical surface

Let $\angle OAN = i$, the angle of incidence and $\angle CAB = r$, the angle of refraction. Using the proper sign convention, we can write

$$PO = -u; PI = -v; PC = +R$$

Let α , β , and γ be the angles subtended by OA , IA and CA , respectively with the principal axis and h the height of the normal dropped from A on the principal axis. In ΔOCA and ΔICA , we have

$$i = \alpha + \gamma \quad (i \text{ is exterior angle}) \quad (20.10)$$

and

$$r = \beta + \gamma \quad (r \text{ is exterior angle}) \quad (20.11)$$

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From Snell's law, we recall that

$$\frac{\sin i}{\sin r} = \mu$$

where μ is the refractive index of the glass surface with respect to air. For a surface of small aperture, P will be close to A and so i and r will be very small ($\sin i \approx i$, $\sin r \approx r$). The above equation, therefore, gives

$$i = \mu r \quad (20.12)$$

Substituting the values of i and r in Eqn. (20.12) from Eqns. (20.10) and (20.11) respectively, we get

$$\begin{aligned} \alpha + \gamma &= \mu(\beta + \gamma) \\ \text{or} \quad \alpha - \mu\beta &= \gamma(\mu - 1) \end{aligned} \quad (20.13)$$

As α , β and γ are very small, we can take $\tan \alpha \approx \alpha$, and $\tan \beta \approx \beta$, and $\tan \gamma \approx \gamma$. Now referring to ΔOAM in Fig. 20.11, we can write

$$\alpha \approx \tan \alpha = \frac{AM}{MO} = \frac{AP}{PO} = \frac{h}{-u} \quad (\text{if } M \text{ is very near to } P)$$

$$\beta \approx \tan \beta = \frac{AM}{MI} = \frac{AM}{PI} = \frac{h}{-v}$$

$$\text{and} \quad \gamma \approx \tan \gamma = \frac{AM}{MC} = \frac{AM}{PC} = \frac{h}{R}$$

Substituting for α , β and γ in Eqn. (20.13), we get

$$\begin{aligned} \frac{h}{-u} - \frac{\mu h}{v} &= (\mu - 1) \frac{h}{R} \\ \text{or} \quad \frac{\mu}{v} - \frac{1}{u} &= \frac{\mu - 1}{R} \end{aligned} \quad (20.14)$$

This important relationship correlates the object and image distances to the refractive index μ and the radius of curvature of the refracting surface.

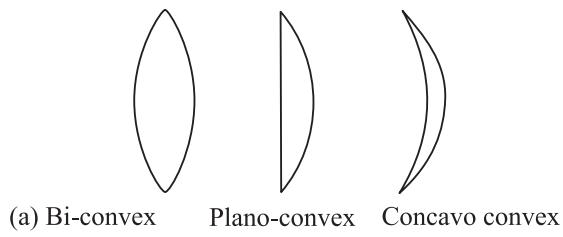
20.5.1 Reflection through lenses

A lens is a thin piece of transparent material (usually glass) having two surfaces, one or both of which are curved (mostly spherical). You have read in your earlier classes that lenses are mainly of two types, namely, convex lens and concave lens. Each of them is sub-divided into three types as shown in Fig. 20.12. Thus, you can have plano-convex and plano-concave lenses too.

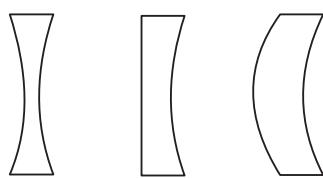


Notes

Thin lens : If the thickness of a lens is negligible in comparison to the radii of curvature of its curved surfaces, the lens is referred to as a thin lens. We will deal with thin lenses only.



Principal axis is the line joining the centres of curvature of two surfaces of the lens.



(a) Bi-convex Plano-convex Concavo convex

(b) Biconcave Plano-concave Convex concavo

Fig. 20.12 : Types of lenses

Optical centre is the point at the center of the lens situated on the principal axis. The rays passing through the optical centre do not deviate.

Principal focus is the point at which rays parallel and close to the principal axis converge to or appear to diverge. It is denoted by F (Fig. 20.13). Rays of light can pass through a lens in either direction. So every lens has two principal foci, one on its either side.

Focal length is the distance between the optical centre and the principal focus. In Fig. 20.13, OF is focal length (f). As per the sign convention, OF is positive for a convex lens and negative for a concave lens.

Focal plane is the plane passing through the focus of a lens perpendicular to its principal axis.

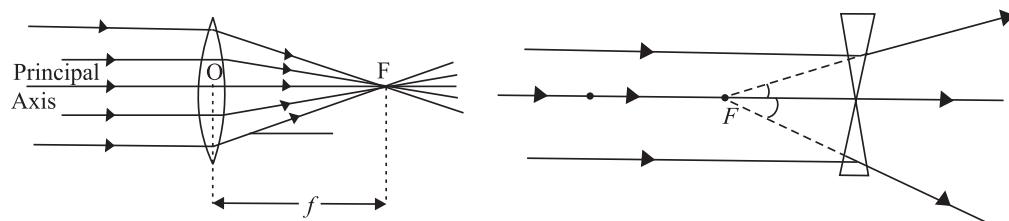


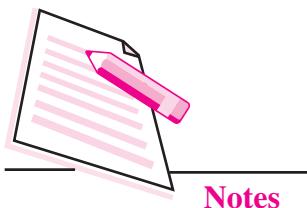
Fig. 20.13 : Foci of a) convex, and b) concave lenses

20.5.2 Lens Maker's Formula and Magnification

You can now guess that the focal length must be related to the radius of curvature and the refractive index of the material of the lens. Suppose that a thin convex lens L is held on an optical bench (Fig. 20.14). Let the refractive index of the material of the lens with respect to air be μ and the radii of curvatures of its two surfaces be R_1 and R_2 , respectively. Let a point object be situated on the principal

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Since the lens used is actually thin, points A and B may be considered very close to point a and hence $C_1 A$ is taken equal to $C_1 Q$ and $C_2 B$ as $C_2 Q$.

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axis at P . C_1 and C_2 are the centres of curvature of the curved surfaces 1 and 2, respectively.

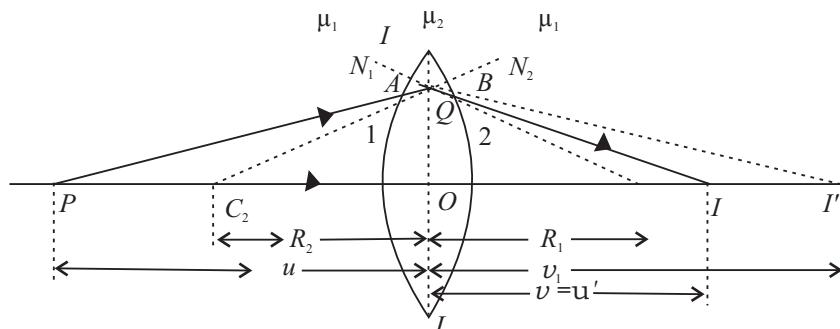


Fig. 20.14 : Point image of a point object formed by a thin double convex lens

A ray from P strikes surface 1 at A . $C_1 N_1$ is normal to surface 1 at the point A . The ray PA travels from the rarer medium (air) to the denser medium (glass), and bends towards the normal to proceed in the direction AB . The ray AB would meet the principal axis $C_2 C_1$ at the point I' *in the absence of the surface 2*. Similarly, another ray from P passing through the optical centre O passes through the Point I' . I' is thus the virtual image of the object P .

Then object distance $OP = u$ and image distance $OI' = v_1$ (say). Using Eqn. (20.14) we can write

$$\frac{\mu}{v_1} - \frac{1}{u} = \frac{\mu - 1}{R_1} \quad (20.15)$$

Due to the presence of surface 2, the ray AB strikes it at B . $C_2 N_2$ is the normal to it at point B . As the ray AB is travelling from a denser medium (glass) to a rarer medium (air), it bends away from the normal $C_2 N_2$ and proceeds in the direction BI and meets another ray from P at I . Thus I is image of the object P formed by the lens. It means that image distance $OI = v$.

Considering point object O , its virtual image is I' (due to surface 1) and the final image is I . I' is the virtual object for surface 2 and I is the final image. Then for the virtual object I' and the final image I , we have, object distance $OI' = u' = v_1$ and image distance $OI = v$.

On applying Eqn. (20.12) and considering that the ray AB is passing from *glass to air*, we have

$$\frac{(1/\mu)}{v} + \frac{1}{v_1} = \frac{(1/\mu)-1}{R_2}$$

$$\text{or, } \frac{1}{\mu v} - \frac{1}{v_1} = \frac{1-\mu}{\mu R_2}$$



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Multiplying both sides by μ , we get

$$\frac{1}{v} - \frac{\mu}{v_1} = \frac{\mu - 1}{R_2} \quad (20.16)$$

Adding Eqns. (20.15) and (20.16), we have

$$\frac{1}{v} - \frac{1}{u} = (\mu - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right) \quad (20.17)$$

If $u = \infty$, that is the object is at infinity, the incoming rays are parallel and after refraction will converge at the focus ($v = f$). Then Eqn. (20.17) reduces to

$$\frac{1}{f} = (\mu - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right) \quad (20.18)$$

This is called lens maker's formula.

From Eqns. (20.17) and (20.18), we can conclude that

- The focal length of a lens depends on the radii of curvature of spherical surfaces. Focal length of a lens of larger radii of curvature will be more.
- Focal length of a lens is smaller if the refractive index of its material is high.

In case a lens is dipped in water or any other transparent medium, the value of μ changes and you can actually work out that focal length will increase. However, if the density of the medium is more than that of the material of the lens, say carbon disulphide, the rays may even diverge.

20.5.3 Newton's Formula

Fig. 20.5.3 shows the image of object AB formed at $A'B'$ by a convex lens L . F_1 and F_2 are the first and second principal foci respectively.

Let us measure the distances of the object and image from the first focus and second focus respectively. Let x_1 be the distance of object from the first focus and x_2 be the distance of image from the second focus and f_1 and f_2 the first and second focal lengths, respectively as shown in Fig. 20.5.3.

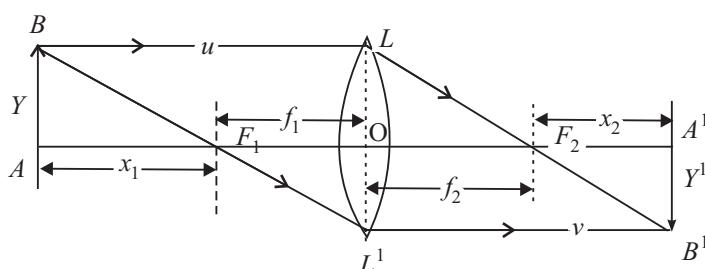


Fig. 20.5.3

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Now, in similar “s, ABF_1 and $OL'F_1$

$$\frac{-y'}{y} = \frac{-f_1}{-x_1}$$

Also from similar Δ s, OLF_2 and $A'B'F_2$

$$\frac{-y'}{y} = \frac{x_2}{f_2}$$

Comparing these two equations we get

$$x_1x_2 = f_1f_2$$

for $f_1 \equiv f_2 \approx f$ (say), then $x_1x_2 = +f^2$

$$\text{or } f = \sqrt{x_1x_2}$$

This relation is called Newton's formula and can be conveniently used to measure the focal length.

20.5.4 Displacement Method to find the Position of Images (Conjugates points)

In the figure 20.5.4, $A'B'$ is the image of the object AB as formed by a lens L . $OA = u$ and $OA' = v$.

The principle of reversibility of light rays tells us that if we move the lens towards the right such that $AO = v$, then again the image will be formed at the same place.

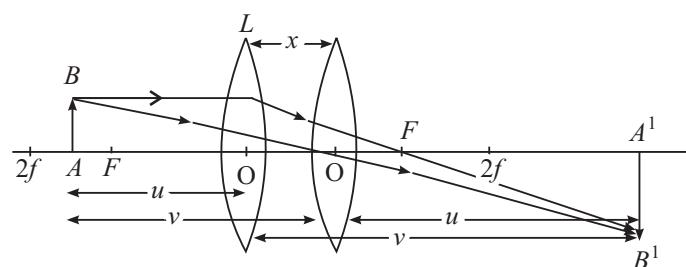


Fig. 20.5.4

Thus $AA' = D = u + v$... (i)

and the separation between the two positions of the lens:

$$OO' = x = (v - u) \quad \dots \text{(ii)}$$



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Adding (i) and (ii) we get

$$v = \frac{x+D}{2}$$

and, subtracting (ii) from (i) we get

$$u = \frac{D-x}{2}$$

Substituting these values in the lens formula, we get.

$$\frac{1}{f} = \frac{1}{v} - \frac{1}{(-u)}$$

$$\frac{1}{f} = \frac{2}{x+D} + \frac{2}{D-x} = \frac{2}{D+x} + \frac{2}{D-x}$$

$$\frac{1}{f} = \frac{2(D-x+D+x)}{D^2-x^2}$$

$$\frac{1}{f} = \frac{4D}{D^2-x^2}$$

or $f = \frac{D^2-x^2}{4D}$

Thus, keeping the positions of the object and screen fixed we can obtain equally clear, bright and sharp images of the object on the screen corresponding to the two positions of the lens. This again is a very convenient way of finding f of a lens.

20.6 FORMATION OF IMAGES BY LENSES

The following properties of the rays are used in the formation of images by lenses:

- A ray of light through the optical centre of the lens passes undeviated.
- A parallel ray, after refraction, passes through the principal focus.
- A ray of light through F or F' is rendered parallel to the principal axis after refraction.

Any two of these rays can be chosen for drawing ray diagrams.

The lens formula $\frac{1}{f} = \frac{1}{v} - \frac{1}{u}$ suggests the dependence of the image distance (v)

on the object distance (u) and the focal length (f) of the lens.

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The magnification of a lens is defined as the ratio of the height of the image formed by the lens to the height of the object and is denoted by m :

$$m = \frac{I}{O} = \frac{v}{u}$$

where I is height of the image and O the height of the object.

Example 20.5 : The radii of curvature of a double convex lens are 15cm and 30cm, respectively. Calculate its focal length. Also, calculate the focal length when it is immersed in a liquid of refractive index 1.65. Take μ of glass = 1.5.

Solution : From Eqn. (20.18) we recall that

$$\frac{1}{f} = (\mu - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right)$$

Here $R_1 = +15\text{cm}$, and $R_2 = -30\text{cm}$. On substituting the given data, we get

$$\frac{1}{f} = (1.5 - 1) \left(\frac{1}{15} - \frac{1}{-30} \right)$$

$$\Rightarrow f = 20\text{ cm}$$

When the lens is immersed in a liquid, μ will be replaced by μ_{lg} :

$$\begin{aligned}\mu_{\text{lg}} &= \frac{\mu_{\text{ag}}}{\mu_{\text{al}}} \\ &= \frac{1.5}{1.65} = \frac{10}{11}\end{aligned}$$

Therefore

$$\begin{aligned}\frac{1}{f_l} &= (\mu_{\text{lg}} - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right) \\ &= \left(\frac{10}{11} - 1 \right) \left(\frac{1}{15} - \frac{1}{-30} \right) \\ &= -\frac{1}{110} \\ \therefore f &= -110\text{cm}\end{aligned}$$

As f is negative, the lens indeed behaves like a concave lens.

20.7 POWER OF A LENS

A practical application of lenses is in the correction of the defects of vision. You may be using spectacles or seen other learners, parents and persons using



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spectacles. However, when asked about the power of their lens, they simply quote a positive or negative number. What does this number signify? This number is the power of a lens in dioptre. The power of a lens is defined as the reciprocal of its focal length in metre:

$$P = \frac{1}{f}$$

The S1 unit of power of a lens is m^{-1} . Dioptre is only a commercial unit generally used by opticians. The power of a convex lens is positive and that of a concave lens is negative. Note that greater power implies smaller focal length. Using lens maker's formula, we can relate power of a lens to its radii of curvature:

$$\frac{1}{f} = (\mu - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right)$$

or $P = (\mu - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right)$

Example 20.6 : Calculate the radius of curvature of a biconvex lens with both surfaces of equal radii, to be made from glass ($\mu = 1.54$), in order to get a power of +2.75 dioptre.

Solution : $P = (\mu - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right)$

$$P = +2.75 \text{ dioptre}$$

$$\mu = 1.54$$

$$R_1 = R$$

and $R_2 = -R$

Substituting the given values in lens makers formula, we get

$$2.75 = (0.54) \left(\frac{2}{R} \right)$$

$$R = \frac{0.54 \times 2}{2.75}$$

$$= 0.39 \text{ m}$$

$$= 39 \text{ cm}$$

20.8 COMBINATION OF LENSES

Refer to Fig. 20.15. Two thin convex lenses A and B having focal lengths f_1 and f_2 , respectively have been kept in contact with each other. O is a point object placed on the common principal axis of the lenses.

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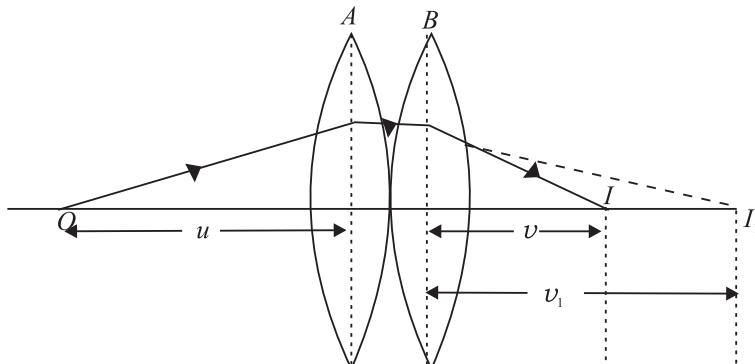


Fig. 20.15 : Two thin convex lenses in contact

Note that lens *A* forms the image of object *O* at *I*₁. This image serves as the **virtual** object for lens *B* and the final image is thus formed at *I*. If *v* be the object distance and *v*₁ the image distance for the lens *A*, then using the lens formula, we can write

$$\frac{1}{v_1} - \frac{1}{u} = \frac{1}{f_1} \quad (20.19)$$

If *v* is the final image distance for the lens *B*, we have

$$\frac{1}{v} - \frac{1}{v_1} = \frac{1}{f_2} \quad (20.20)$$

Note that in writing the above expression, we have taken *v*₁ as the object distance for the thin lens *B*.

Adding Eqns. (20.19) and (20.20), we get

$$\frac{1}{v} - \frac{1}{u} = \frac{1}{f_1} + \frac{1}{f_2} \quad (20.21)$$

If the combination of lenses is replaced by a single lens of focal length *F* such that it forms the image of *O* at the same position *I*, then this lens is said to be equivalent to both the lenses. It is also called the **equivalent lens** for the combination. For the equivalent lens, we can write

$$\frac{1}{v} - \frac{1}{u} = \frac{1}{F} \quad (20.22)$$

where

$$\frac{1}{F} = \frac{1}{f_1} + \frac{1}{f_2}. \quad (20.23)$$



Notes

If P is power of the equivalent lens and P_1 and P_2 are respectively the powers of individual lenses, then

$$P = P_1 + P_2 \quad (20.24)$$

Note that Eqns.(20.23) and (20.24) derived by assuming two thin convex lenses in contact also hold good for any combination of two thin lenses in contact (the two lenses may both be convex, or concave or one may be concave and the other convex).

Example 20.7 : Two thin convex lenses of focal lengths 20cm and 40cm are in contact with each other. Calculate the focal length and the power of the equivalent lens.

Solution : The formula for the focal length of the combination $\frac{1}{F} = \frac{1}{f_1} + \frac{1}{f_2}$ gives

$$\frac{1}{F} = \frac{1}{20} + \frac{1}{40}$$

$$= \frac{3}{40}$$

or $F = \frac{40}{3} = 13.3\text{cm} = 0.133\text{m}$

Power of the equivalent lens is

$$P = \frac{1}{F} = \frac{1}{0.133} = +7.5 \text{ dioptrre.}$$



INTEXT QUESTIONS 20.5

1. On what factors does the focal length of a lens depend?
2. A lens, whose radii of curvature are different, is used to form the image of an object placed on its axis. If the face of the lens facing the object is reversed, will the position of the image change?
3. The refractive index of the material of an equi-double convex lens is 1.5. Prove that the focal length is equal to the radius of curvature.
4. What type of a lens is formed by an air bubble inside water?
5. A lens when immersed in a transparent liquid becomes invisible. Under what condition does this happen?
6. Calculate the focal length and the power of a lens if the radii of curvature of its two surfaces are +20cm and -25cm ($\mu = 1.5$).

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7. Is it possible for two lenses in contact to produce zero power?
8. A convex lens of focal length 40cm is kept in contact with a concave lens of focal length 20cm. Calculate the focal length and the power of the combination.

Defects in image formation

Lenses and mirrors are widely used in our daily life. It has been observed that they do not produce a point image of a point object. This can be seen by holding a lens against the Sun and observing its image on a paper. You will note that it is not exactly circular. Mirrors too do not produce a perfect image. The defects in the image formation are known as **aberrations**. The aberrations depend on (i) the quality of lens or mirror and (ii) the type of light used.

Two major aberrations observed in lenses and mirrors, are (a) **spherical aberration** and (b) **chromatic aberration**. These aberrations produce serious defects in the images formed by the cameras, telescopes and microscopes etc.

Spherical Aberration

This is a monochromatic defect in image formation which arises due to the sphericity and aperture of the refracting or reflecting surfaces. The paraxial rays and the marginal rays form images at different points I_p and I_m respectively (Fig. 20.16)

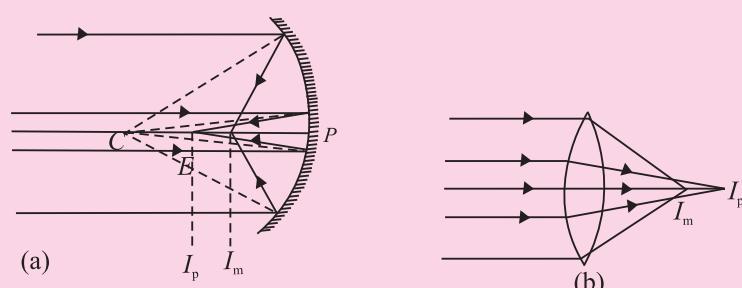


Fig. 20.16 : Spherical aberration in a) spherical mirror, and b) lens. I_p is image formed by the paraxial rays and I_m that formed by the marginal rays.

The **spherical aberration** in both mirrors and lenses can be reduced by allowing only the paraxial rays to be incident on the surface. It is done by using stops. Alternatively, the paraxial rays may be cut-off by covering the central portion, thus allowing only the marginal or peripheral rays to form the image. However, the use of stops reduces the brightness of the image.

A much appreciated method is the use of elliptical or parabolic mirrors.



Notes

The other methods to minimize spherical aberration in lenses are : use of plano convex lenses or using a suitable combination of a convex and a concave lens.

Chromatic Aberration in Lenses

A convex lens may be taken as equivalent to two small-angled prisms placed base to base and the concave lens as equivalent to such prisms placed vertex to vertex. Thus, a polychromatic beam incident on a lens will get dispersed. The parallel beam will be focused at different coloured focii. This defect of the image formed by spherical lenses is called **chromatic aberration**. It occurs due to the dispersion of a polychromatic incident beam (Fig. 20.17). Obviously the red colour is focused farther from the lens while the blue colour is focused nearer the lens (in a concave lens the focusing of the red and blue colours takes place in the same manner but on the opposite side of it).

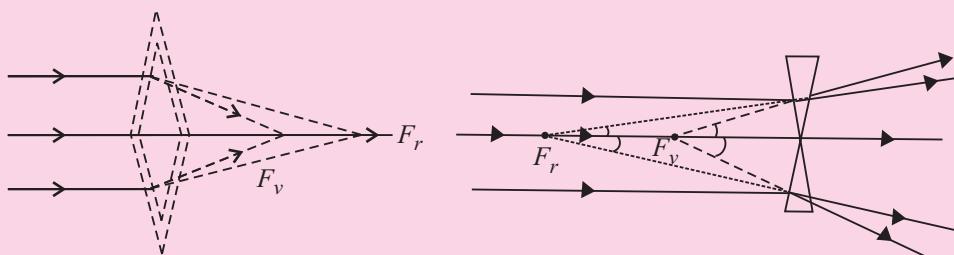


Fig. 20.17: Chromatic aberration

To remove this defect we combine a convergent lens of suitable material and focal length when combined with a divergent lens of suitable focal length and material. Such a lens combination is called an **achromatic doublet**. The focal length of the concave lens can be found from the necessary condition for achromatism given by

$$\frac{\omega_1}{f_1} + \frac{\omega_2}{f_2} = 0$$



WHAT YOU HAVE LEARNT

- Real image is formed when reflected rays actually intersect after reflection. It can be projected on a screen.
- The focal length is half of the radius of curvature.

$$f = \frac{R}{2}$$

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The object and image distances are related to magnification as

$$m = \frac{v}{u}$$

- Refraction of light results in change in the speed of light when it travels from one medium to another. This causes the rays of light to bend towards or away from the normal.
- The refractive index μ determines the extent of bending of light at the interface of two media.
- Snell's law is mathematically expressed as

$$\frac{\sin i}{\sin r} = \mu_{12}$$

where i is the angle of incidence in media 1 and r is the angle of refraction in media 2.

- Total internal reflection is a special case of refraction wherein light travelling from a denser to a rarer media is incident at an angle greater than the critical angle:

$$\mu = \frac{1}{\sin i_c}$$

- Any transparent media bounded by two spherical surfaces or one spherical and one plane surface forms a lens.
- Images by lenses depend on the focal length and the distance of the object from it.
- Convex lenses are converging while concave lenses are diverging.

$$\frac{1}{f} = (\mu - 1) \left(\frac{1}{R_1} - \frac{1}{R_2} \right)$$

$$m = \frac{v}{u}$$

$$\text{and } \frac{1}{f} = \frac{1}{v} - \frac{1}{u}$$

are simple relationships between the focal length (f), the refractive index, the radii of curvatures (R_1, R_2), the object distance (u) and the image distance (v).

- Newton's formula can be used to measure the focal length of a lens.
- Displacement method is a very convenient way of finding focal length of a lens.



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- Power of a lens indicates how diverging or converging it is:

$$P = \frac{1}{f}$$

Power is expressed in dioptre. (or m^{-1} in SI units)

- The focal length F of an equivalent lens when two their lenses of focal lengths f_1 and f_2 one kept in contact is given by

$$\frac{1}{F} = \frac{1}{f_1} + \frac{1}{f_2}$$



TERMINAL EXERCISES

1. List the uses of concave and convex mirrors.
2. What is the nature and position of image formed when the object is at (i) infinity (ii) $2f$ (iii) f in case of concave mirror and convex mirror.
3. List the factors on which lateral displacement of an incident ray depends as it suffers refraction through a parallel-sided glass slab? Why is the lateral displacement larger if angle of incidence is greater. Show this with the help of a ray diagram.
4. State conditions for total internal reflection of light to take place.
5. How is $+1.5$ dioptre different from -1.5 dioptre? Define dioptre.
6. Why does the intensity of light become less due to refraction?
7. A lamp is 4m from a wall. Calculate the focal length of a concave mirror which forms a five times magnified image of the lamp on the wall. How far from the wall must the mirror be placed?
8. A dentist's concave mirror has a radius of curvature of 30cm. How far must it be placed from a cavity in order to give a virtual image magnified five times?
9. A needle placed 45cm from a lens forms an image on a screen placed 90cm on the other side of the lens. Identify the type of the lens and determine its focal length. What is the size of the image, if the size of the needle is 5.0cm?
10. An object of size 3.0cm is placed 14cm in front of a concave lens of focal length 21 cm. Describe the nature of the image by the lens. What happens if the object is moved farther from the lens?
11. An object is placed at a distance of 100cm from a double convex lens which forms a real image at a distance of 20cm. The radii of curvature of the surfaces

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of a lens are 25cm and 12.5 cm respectively. Calculate the refractive index of the material of the lens.

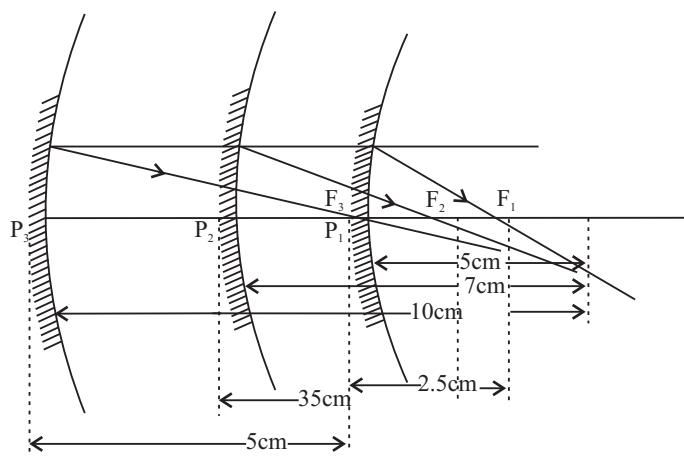
12. A ray of light is travelling from diamond to glass. Calculate the value of the critical angle for the ray, if the refractive index of glass is 1.51 and that of diamond 2.47.
13. A small object is placed at a distance of 15cm from two coaxial thin convex lenses in contact. If the focal length of each lens is 20cm. Calculate the focal length and the power of the combination and the distance between the object and its image.
14. While finding the focal length of a convex lens, an object was kept at a distance of 65.0 cm from the screen. Two positions of the lens for which clear image of the object was formed on the screen were obtained. The distance between these two positions was found to be 15 cm. Calculate the focal length of the given lens.



ANSWERS TO INTEXT QUESTIONS

20.1

1. (a) plane mirror (its radius of curvature is infinitely large).
(b) No. The focal length of a spherical mirror is half of its radius of curvature ($f \sim R/2$) and has nothing to do with the medium in which it is immersed.
(c) Virtual
(d) This is because the rays parallel to the principal axis converge at the focal point F ; and the rays starting from F , after reflection from the mirror, become parallel to the principal axis. Thus, F serves both as the first and the second focal point.
2. Focal lengths : 2.5cm, 3.5cm, 5cm.

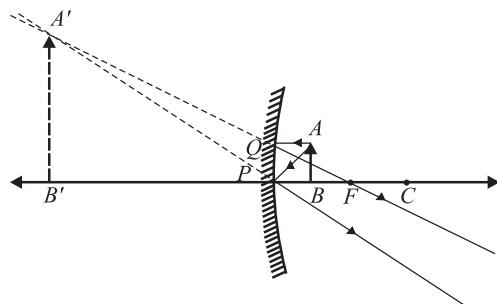


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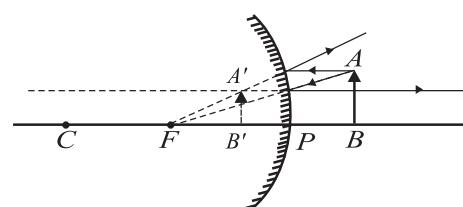
3. $f = -15\text{cm}$; $f = +15\text{cm}$.
4. The dish antennas are curved so that the incident parallel rays can be focussed on the receiver.

20.2

1. The upper part of the mirror must be convex and its lower part concave.
2. Objects placed close to a concave mirror give an enlarged image. Convex mirrors give a diminished erect image and have a larger field of view.

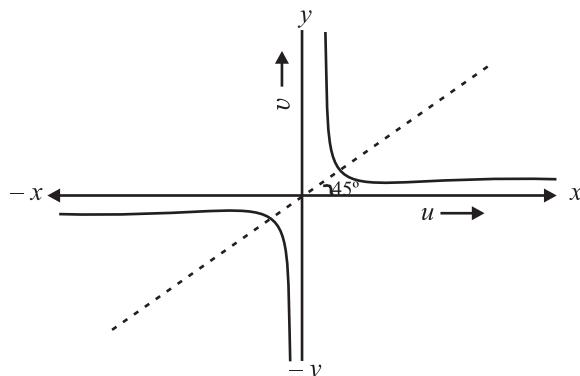


(a) Image formed by concave mirror

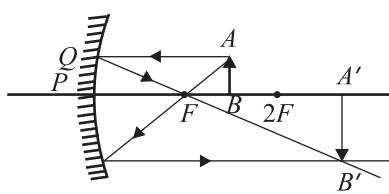
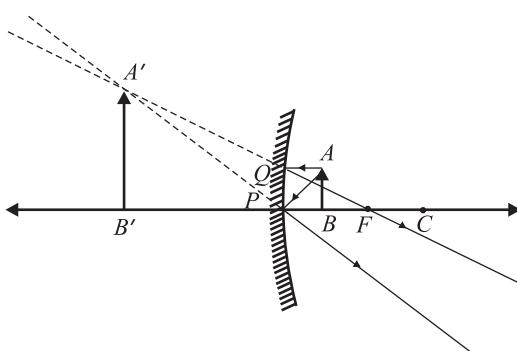


(b) Image formed by convex mirror

3. for $|u| > f$, we get real image; $u = -2f$ is a special case when an object kept at the centre of curvature of the mirror forms a real image at this point itself ($v = -2f$). For $u < f$, we get virtual image.



4. When (i) $u < f$, and (ii) $f < u < 2f$.



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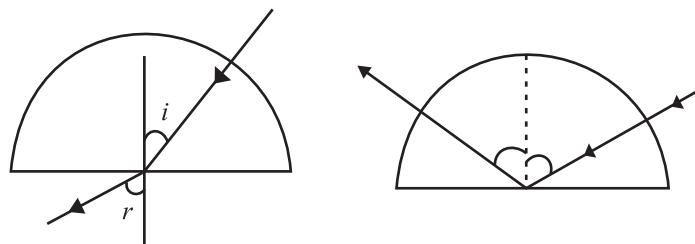
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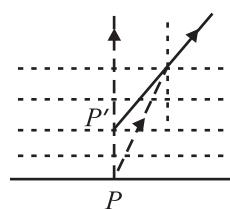
5. (i) 12cm in front of mirror, real and inverted, (ii) 0.8cm
6. $v = -60\text{cm}$, $R = -24\text{cm}$
7. $u = -10\text{cm}$, $v = +5\text{cm}$
8. $v = 4\text{cm}$

20.3

1. No lateral displacement.



2. $\angle r > \angle i$ when $\angle i < \angle c$ Total internal reflection where $\angle i > \angle c$
3. The density of air and hence its refractive index decrease as we go higher in altitude. As a result, the light rays from the Sun, when it is below the horizon, pass from the rarer to the denser medium and bend towards the normal, till they are received by the eye of the observer. This causes the shape to appear elongated.
4. Due to the change in density of the different layers of air in the atmosphere, μ changes continuously. Therefore, the refractive index of air varies at different levels of atmosphere. This along with air currents causes twinkling of stars.
5. Due to refraction point P appears at P' .

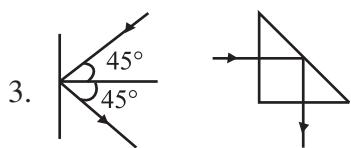


6. 36.2°

20.4

1. Total internal reflection cannot take place if the ray travels from a rarer to a denser medium as the angle of refraction will always be less than the angle of incidence.
2. Yes the critical angle will change as

$$\mu_{ag} = \frac{1}{\sin i_c} \quad \mu_{wg} = \frac{\mu_{ag}}{\mu_{aw}}$$



The intensity in the second case is more due to total internal reflection.

4. 20cm, $i_c = \sin^{-1} 0.8$

20.5

2. No. Changing the position of R_1 and R_2 in the lens maker's formula does not affect the value of f . So the image will be formed in the same position.
3. Substitute $R_1 = R$; $R_2 = -R$ and $\mu = 1.5$ in the lens maker's formula. You will get $f = R$.
4. Concave lens. But it is shaped like a convex lens.
5. This happens when the refractive index of the material of the lens is the same as that of the liquid.
6. $f = 22.2$ cm and $P = 4.5$ dioptre
7. Yes, by placing a convex and a concave lens of equal focal length in contact.
8. -40cm , -2.5 dioptre

Answers to Problems in Terminal Exercise

7. $f = -0.83$, 5m.
8. 12cm
9. $f = 30\text{cm}$, size of image = 10cm, converging lens
10. The image is erect, virtual and diminished in size, and located at 8.4cm from the lens on the same side as the object. As the object is moved away from the lens, the virtual image moves towards the focus of the lens but never beyond and progressively diminishes in size.
11. $\mu = 1.5$
12. 37.7°
13. 10cm, 10 dioptre, 45 cm.
14. $f = 15.38$ cm



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Notes

21

DISPERSION AND SCATTERING OF LIGHT

In the previous lesson you have learnt about reflection, refraction and total internal reflection of light. You have also learnt about image formation by mirrors and lenses and their uses in daily life. When a narrow beam of ordinary light is refracted by a prism, we see colour bands. This phenomenon has to be other than reflection or refraction. *The splitting of white light into its constituent colours or wavelengths by a medium is called dispersion.* In this lesson, you will study about this phenomenon. A beautiful manifestation of this phenomenon in nature is in the form of rainbow. You will also learn in this lesson about the phenomenon of scattering of light, which gives sky its blue colour and the sun red colour at sunrise and sunset. Elementary idea of Raman effect will also be discussed in this lesson.



OBJECTIVES

After studying this lesson, you should be able to :

- *explain dispersion of light;*
- *derive relation between the angle of deviation (δ), angle of prism (A) and refractive index of the material of the prism (μ);*
- *relate the refractive index with wavelength and explain dispersion through a prism;*
- *explain formation of primary and secondary rainbows;*
- *explain scattering of light and list its applications; and; and*
- *explain Raman effect.*

21.1 DISPERSION OF LIGHT

Natural phenomena like rings around planets (halos) and formation of rainbow etc. cannot be explained by the rectilinear propagation of light. To understand



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such events, light is considered as having wave nature. (You will learn about it in the next lesson.) As you know, light waves are transverse electromagnetic waves which propagate with speed $3 \times 10^8 \text{ ms}^{-1}$ in vacuum. Of the wide range of electromagnetic spectrum, the visible light forms only a small part. Sunlight consists of seven different wavelengths corresponding to seven colours. Thus, colours may be identified with their wavelengths. You have already learnt *that the speed and wavelength of waves change when they travel from one medium to another*. The speed of light waves and their corresponding wavelengths also change with the change in the medium. The speed of a wave having a certain wavelength becomes less than its speed in free space when it enters an optically denser medium.

The refractive index μ has been defined as the ratio of the speed of light in vacuum to the speed of light in the medium. It means that the refractive index of a given medium will be different for waves having wavelengths $3.8 \times 10^{-7} \text{ m}$ and $5.8 \times 10^{-7} \text{ m}$ because these waves travel with different speeds in the same medium. This ***variation of the refractive index of a material with wavelength is known as dispersion***. This phenomenon is different from refraction. In free space and even in air, the speeds of all waves of the visible light are the same. So, they are not separated. (Such a medium is called a non-dispersive medium.) But in an optically denser medium, the component wavelengths (colours) travel with different speeds and therefore get separated. Such a medium is called *dispersive medium*. Does this suggest that light will exhibit dispersion whenever it passes through an optically denser medium. Let us learn about it now.

21.1.1 Dispersion through a Prism

The separation of colours by a medium is not a sufficient condition to observe dispersion of light. These colours must be widely separated and should not mix up again after emerging from the dispersing medium. A glass slab (Fig. 21.1) is not suitable for observing dispersion as the rays of the emergent beam are very close and parallel to the incident beam

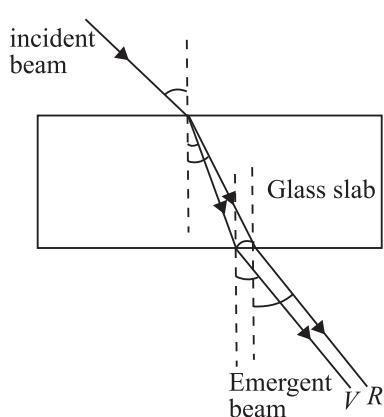


Fig. 21.1 : Passage of light through a glass slab

Newton used a prism to demonstrate dispersion of light. Refer to Fig. 21.2. White light from a slit falls on the face AB of the prism and light emerging from face AC is seen to split into different colours. Coloured patches can be seen on a screen. The face AC increases the separation between the rays refracted at the face AB . The incident white light PQ thus splits up into its component seven colours : Violet, indigo, blue, green, yellow, orange and red (VIBGYOR). The wavelengths travelling with different speeds are refracted through different angles and are thus

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separated. This *splitting of white light into component colours is known as dispersion*. *MR* and *MV* correspond to the red and violet light respectively. These colours on the screen produce the *spectrum*.

The bending of the original beam *PQN* along *MR* and *MV* etc. is known as *deviation*. The angle between the emergent ray and the incident ray is known as the **angle of deviation**. Thus δ_v and δ_r represent the angles of deviation for violet light and red light, respectively.

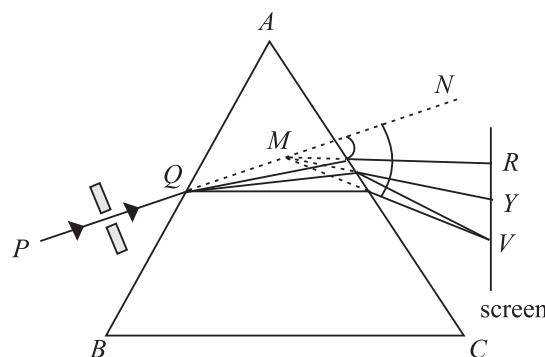


Fig. 21.2 : Dispersion of light by a prism

Read the following example carefully to fix the ideas on variation of the refractive index with the wavelength of light.

Example 21.1: A beam of light of average wavelength 600nm, on entering a glass prism, splits into three coloured beams of wavelengths 384 nm, 589 nm and 760 nm respectively. Determine the refractive indices of the material of the prism for these wavelengths.

Solution : The refractive index of the material of the prism is given by

$$\mu = \frac{c}{v}$$

where c is speed of light in vacuum, and v is speed of light in the medium (prism).

Since velocity of a wave is product of frequency and wavelength, we can write

$$c = v\lambda_a \quad \text{and} \quad v = v\lambda_m$$

where λ_a and λ_m are the wavelengths in air and medium respectively and v is the frequency of light waves. Thus

$$\mu = \frac{v\lambda_a}{v\lambda_m} = \frac{\lambda_a}{\lambda_m}$$

For 384 nm wavelength, the refractive index is

$$\mu_1 = \frac{600 \times 10^{-9} \text{ m}}{384 \times 10^{-9} \text{ m}} = 1.56$$



Notes

For wave length of 589 nm :

$$\mu_2 = \frac{600 \times 10^{-9} \text{ m}}{58.9 \times 10^{-9} \text{ m}} = 1.02$$

and for 760nm wavelength :

$$\mu_3 = \frac{600 \times 10^{-9} \text{ m}}{760 \times 10^{-9} \text{ m}} = 0.8$$

We have seen that the refractive index of a material depends on

- the nature of the material, and
- the wavelength of light.

An interesting outcome of the above example is that the variation in wavelength ($\Delta\lambda = \lambda_2 - \lambda_1$) produces variation in the refractive index ($\Delta\mu = \mu_2 - \mu_1$). The ratio

$\frac{\Delta\mu}{\Delta\lambda}$ is known as the spectral *dispersive power of the material of prism*.

21.1.2 The Angle of Deviation

We would now establish the relation between the angle of incidence i , the angle of deviation δ and the angle of prism A . Let us consider that a monochromatic beam of light PQ is incident on the face AB of the principal section of the prism ABC [Fig.21.3]. On refraction, it goes along QR inside the prism and emerges along RS from face AC . Let $\angle A \equiv \angle BAC$ be the refracting angle of the prism. We draw normals NQ and MR on the faces AB and AC , respectively and produce them backward to meet at O . Then you can easily convince yourself that $\angle NQP = i$, $\angle MRS = e$, $\angle RQO = r_1$, and $\angle QRO = r_2$ are the angle of incidence, the angle of emergence and the angle of refraction at the faces AB and AC , respectively. The angle between the emergent ray RS and the incident ray PQ at D is known as the angle of deviation (δ).

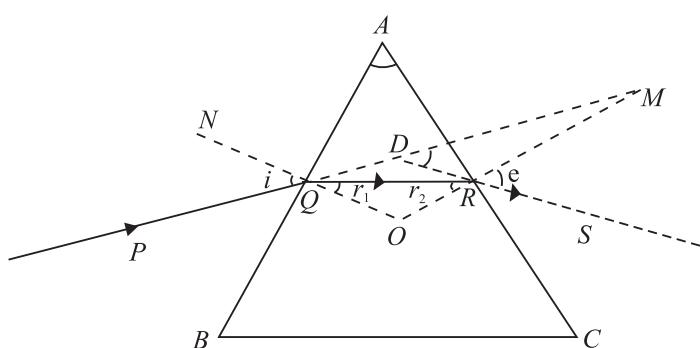


Fig. 21.3 : Refraction through a prism

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Since $\angle MDR = \angle \delta$, As it is the external angle of the triangle QDR , we can write

$$\begin{aligned}\angle \delta &= \angle DQR + \angle DRQ \\ &= (\angle i - \angle r_1) + (\angle e - \angle r_2) \\ \text{or} \quad \angle \delta &= (\angle i + \angle e) - (\angle r_1 + \angle r_2)\end{aligned}\quad (21.1)$$

You may recall that the sum of the internal angles of a quadrilateral is equal to 360° . In the quadrilateral $AQOR$, $\angle A + \angle Q + \angle O + \angle R = 360^\circ$. Since NQ and MR are normals on faces AB and AC , respectively. Therefore

$$\begin{aligned}\angle QAR + \angle QOR &= 180^\circ \\ \text{or} \quad \angle A + \angle QOR &= 180^\circ\end{aligned}\quad (21.2)$$

But in $\triangle QOR$

$$\begin{aligned}\angle OQR + \angle QRO + \angle QOR &= 180^\circ \\ \text{or} \quad \angle r_1 + \angle r_2 + \angle QOR &= 180^\circ\end{aligned}\quad (21.3)$$

On comparing Eqns. (21.2) and (21.3), we have

$$\angle r_1 + \angle r_2 = \angle A \quad (21.4)$$

Combining this result with Eqn. (21.1), we have

$$\begin{aligned}\angle \delta &= (\angle i + \angle e) - \angle A \\ \text{or} \quad \angle i + \angle e &= \angle A + \angle \delta \quad (21.5)\end{aligned}$$

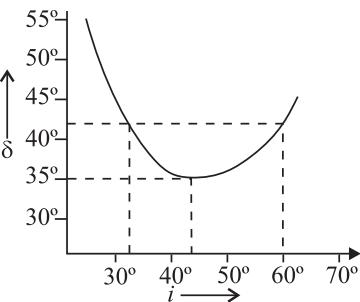


Fig. 21.4 : Plot between angle of incidence i and angle of deviation δ

Angle of Minimum Deviation

If we vary the angle of incidence i , the angle of deviation δ also changes; it becomes minimum for a certain value of i and again starts increasing as i increases further (Fig. 21.4). The minimum value of the angle of deviation is called *angle of minimum deviation* (δ_m). It depends on the material of the prism and the wavelength of light used. In fact, one angle of deviation may be obtained corresponding to two values of the angles of incidence. Using the principle of reversibility of light, we find that the second value of angle of incidence corresponds to the angle of emergence (e). In the minimum deviation position, there is only one value of the angle of incidence. So we have

$$\angle e = \angle i$$

Using this fact in Eqn.(21.5) and replacing δ by δ_m , we have

$$\angle i = \frac{\angle A + \angle \delta_m}{2} \quad (21.6)$$

Applying the principle of reversibility of light rays and under the condition $\angle e = \angle i$, we can write $\angle r_1 = \angle r_2 = \angle r$, say



Notes

On substituting this result in Eqn. (21.4), we get

$$\angle r = \frac{\angle A}{2} \quad (21.7)$$

The light beam inside the prism, under the condition of minimum deviation, passes symmetrically through the prism and is parallel to its base. The refractive index of the material of the prism is therefore given by

$$\mu = \frac{\sin i}{\sin r} = \frac{\sin\left(\frac{A + \delta_m}{2}\right)}{\sin\frac{A}{2}} \quad (21.8)$$

The refractive index μ can be calculated using Eqn.(21.8) for a monochromatic or a polychromatic beam of light. The value of δ_m is different for different colours. It gives a unique value of the angle of incidence and the emergent beam is brightest for this incidence.

For a prism of small angle A , keeping i and r small, we can write

$$\sin i = i, \sin r = r, \text{ and } \sin e = e$$

Hence

$$\mu = \frac{\sin i}{\sin r_1} = \frac{i}{r_1} \text{ or } i = \mu r_1$$

$$\text{Also } \mu = \frac{\sin e}{\sin r_2} = \frac{e}{r_2} \quad \text{or} \quad e = \mu r_2$$

Therefore,

$$\angle i + \angle e = \mu (\angle r_1 + \angle r_2)$$

Using this result in Eqns. (26.4) and (26.5), we get

$$\begin{aligned} \mu \angle A &= \angle A + \angle \delta \\ \text{or } \angle \delta &= (\mu - 1)\angle A \end{aligned} \quad (21.9)$$

We know that μ depends on the wavelength of light. So deviation will also depend on the wavelength of light. That is why δ_v is different from δ_r . Since the velocity of the red light is more than that of the violet light in glass, the deviation of the red light would be less as compared to that of the violet light.

$$\delta_v > \delta_r$$

This implies that $\mu_v > \mu_r$. This change in the refractive index of the material with the wavelength of light is responsible for dispersion phenomenon.

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21.1.3 Angular Dispersion and Dispersive Power

The difference between the angles of deviation for any two wavelengths (colours) is known as the *angular dispersion* for those wavelengths. The angular dispersion between the red and violet wavelengths is $\delta_V - \delta_R$. In the visible part of the spectrum, the wavelength of the yellow colour is nearly the average wavelength of the spectrum. The deviation for this colour δ_Y may, therefore, be taken as the average of all deviations.

The *ratio of the angular dispersion to the mean deviation is taken as the dispersive power* (ω) of the material of the prism :

$$\omega = \frac{\delta_V - \delta_R}{\delta_Y}$$

We can express this result in terms of the refractive indices using Eqn. (21.9) :

$$\begin{aligned}\omega &= \frac{(\mu_V - 1) \angle A - (\mu_R - 1) \angle A}{(\mu_V - 1) \angle A} \\ &= \frac{\mu_V - \mu_R}{\mu_Y - 1} = \frac{\Delta \mu}{\mu - 1}\end{aligned}\quad (21.10)$$

Example 21.2 : The refracting angle of a prism is $30'$ and its refractive index is 1.6. Calculate the deviation caused by the prism.

Solution : We know that $\delta = (\mu - 1) \angle A$

On substituting the given data, we get

$$\delta = (1.6 - 1) \times \frac{1^\circ}{2} = \frac{0.6}{2} = 0.3^\circ = 18'$$

Example 21.3 : For a prism of angle A , the angle of minimum deviation is $A/2$. Calculate its refractive index, when a monochromatic light is used. Given $A = 60^\circ$

Solution : The refractive index is given by

$$\mu = \frac{\sin\left(\frac{A + \delta_m}{2}\right)}{\sin(A/2)}$$

Now $\delta_m = A/2$ so that

$$\mu = \frac{\sin\left(\frac{A + A/2}{2}\right)}{\sin(A/2)} = \frac{\sin\left(\frac{3}{4}A\right)}{\sin\left(\frac{A}{2}\right)} = \frac{\sin\left(\frac{3}{4}A\right)}{\sin\left(\frac{A}{2}\right)} = \sqrt{2} = 1.4$$



INTEXT QUESTIONS 21.1

- Most ordinary gases do not show dispersion with visible light. Why?
- With your knowledge about the relative values of μ for the component colours of white light, state which colour is deviated more from its original direction?
- Does dispersion depend on the size and angle of the prism?
- Calculate the refractive index of an equilateral prism if the angle of minimum deviation is equal to the angle of the prism.



Notes

Rainbow formation

Dispersion of sunlight through suspended water drops in air produces a spectacular effect in nature in the form of rainbow on a rainy day. With Sun at our back, we can see a brighter and another fainter rainbow. The brighter one is called the **primary rainbow** and the other one is said to be **secondary rainbow**. Sometimes we see only one **rainbow**. The bows are in the form of coloured arcs whose common centre lies at the line joining the Sun and our eye. Rainbow can also be seen in a fountain of water in the evening or morning when the sun rays are incident on the water drops at a definite angle.

Primary Rainbow

The primary rainbow is formed by two refractions and a single internal reflection of sunlight in a water drop. (See Fig. 21.5(a)). Descartes explained that rainbow is seen through the rays which have suffered minimum deviation. Parallel rays from the Sun suffering deviation of $137^{\circ}29'$ or making an angle of $42^{\circ}31'$ at the eye with the incident ray, after emerging from the water drop, produce bright shining colours in the bow. Dispersion by water causes different colours (red to violet) to make their own arcs which lie within a cone of 43° for red and 41° for violet rays on the outer and inner sides of the bow (Fig. 21.5 (b)).

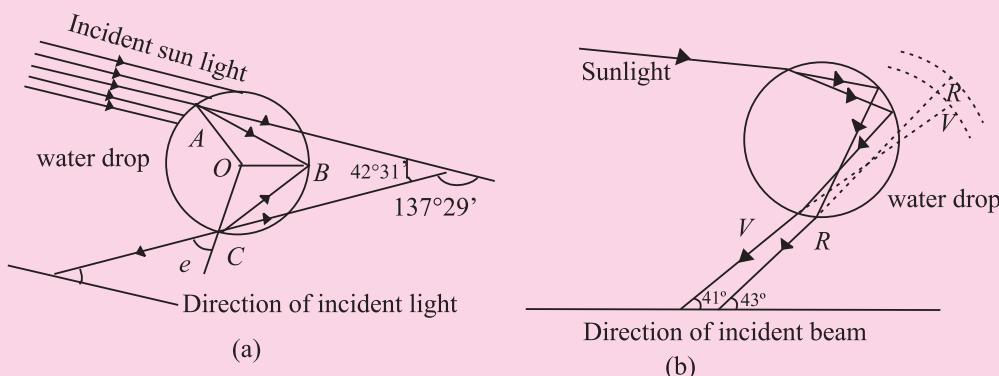


Fig. 21.5 : (a) A ray suffering two refractions and one internal reflection in a drop of water. Mean angle of minimum deviation is $137^{\circ}29'$, and (b) dispersion by a water drop.

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Secondary Rainbow

The secondary rainbow is formed by two refractions and two internal reflections of light on the water drop. The angles of minimum deviations for red and violet colours are 231° and 234° , respectively, so they subtend a cone of 51° for the red and 54° for the violet colour. From Fig.21.6 it is clear that the red colour will be on the inner and the violet colour on the outer side of the bow.

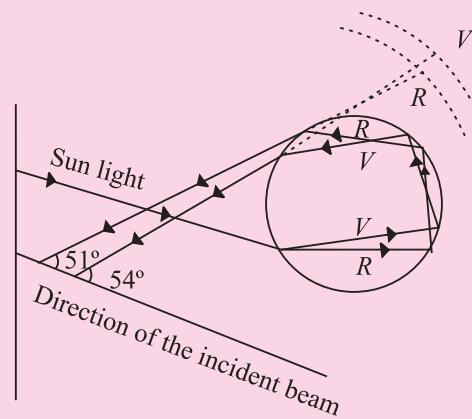


Fig. 21.6 : Formation of the secondary rainbow

The simultaneous appearance of the primary and secondary rainbows is shown in Fig.21.7. The space between the two bows is relatively dark. Note that the secondary rainbow lies above the primary bow.

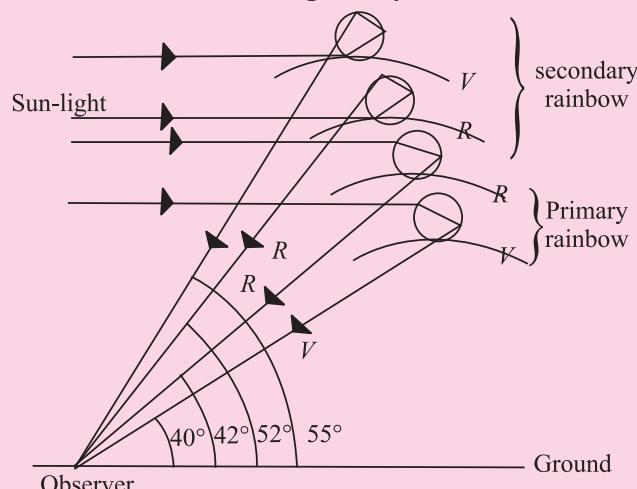


Fig. 21.7 : Simultaneous formation of the primary and secondary rainbow.

21.2 SCATTERING OF LIGHT IN ATMOSPHERE

On a clear day when we look at the sky, it appears blue. But the clouds appear white. Similarly, production of brilliant colours when sunlight passes through jewels and crystals also attracts our attention. You may like to know : How and why does it happen? These phenomena can be explained in terms of *scattering of light*. A solution of dust or particle-free benzene exposed to sunlight gives brilliant blue colour when looked sideways.

21.2.1 Scattering of Light

This phenomenon involves interaction of radiation with matter. Tiny dust particles are present in Earth's atmosphere. When sunlight falls on them, it gets diffused in



Notes

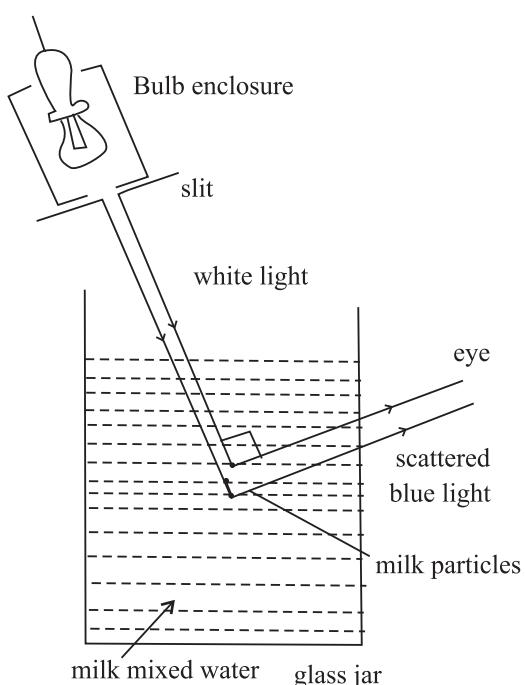


Fig. 21.8 : The scattering of light from milk particles

Let us perform a simple activity.



ACTIVITY 21.1

Take a glass jar or a trough, fill it with water and add a little milk to it. Now allow a narrow beam of light from a white bulb to fall on it. Observe the light at 90° . You will see a bluish beam through water. This experiment shows that after scattering, the wavelengths of light become a peculiarly different in a given direction (Fig. 21.14).

The phenomenon of scattering is a two step process : absorption of light by the scattering particle and then instant re-emission by it in all possible directions. Thus, this phenomenon is different from reflection. The scattered light does not obey the laws of reflection. It is important to note that the size of the particle must be less than the wavelength of light incident on it. A bigger sized particle will scatter all the wavelengths equally. The intensity of scattered light is given by *Rayleigh's law* of scattering. According to this law, ***the intensity of scattered light is inversely proportional to the fourth power of its wavelength***:

$$I \propto \frac{1}{\lambda^4}$$

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Here I is intensity and λ is wavelength of the scattered light. Thus, when white light is incident on the scattering particle, the blue light is scattered the most and the red light is scattered the least.

Example 21.4 : Waves of wavelength 3934Å, 5890Å and 6867Å are found in the scattered beam when sunlight is incident on a thin layer of chimney smoke. Which of these is scattered more intensely?

Solution : The intensity of scattered light is given by

$$I \propto \frac{1}{\lambda^4}$$

Since 3934Å is the smallest wavelength, it will be scattered most intensely.

On the basis of scattering of light, we can explain why sky appears blue, clouds appear white and the sun appears red at sunrise as well as at sunset.



**C.V. Raman
(1888 – 1970)**

Chandra Shekhar Venkat Raman is the only Indian national to receive Nobel prize (1930) in physics till date. His love for physics was so intense that he resigned his job of an officer in Indian finance department and accepted the post of Palit Professor of Physics at the Department of Physics, Calcutta University. His main contributions are : Raman effect on scattering of light, molecular diffraction of light, mechanical theory of bowed strings, diffraction of X-rays, theory of musical instruments and physics of crystals.

As Director of Indian Institute of Science, Bangalore and later as the founder Director of Raman Research Institute, he did yeoman's to Indian science and put it on firm footings in pre-independence period.

(A) Blue Colour of the Sky

We know that scattering of light by air molecules, water droplets or dust particles present in the atmosphere can be explained in accordance with Rayleigh's law. The shorter wavelengths are scattered more than the longer wavelengths. Thus, the blue light is scattered almost six times more intensely than the red light as the wavelength of the blue light is roughly 0.7 times that of the red. The scattered light becomes rich in the shorter wavelengths of violet, blue and green colours. On further scattering, the violet light does not reach observer's eye as the eye is comparatively less sensitive to violet than blue and other wavelengths in its neighbourhood. So, when we look at the sky far away from the sun, it appears blue.

**Notes**

Example 21.5 : What will be the colour of the sky for an astronaut in a spaceship flying at a high altitude.

Solution : At a high altitude, in the absence of dust particle and air molecules, the sunlight is not scattered. So, the sky will appear black.

(B) White colour of the clouds

The clouds are formed by the assembly of small water drops whose size becomes more than the average wavelength of the visible light (5000\AA). These droplets scatter all the wavelengths with almost equal intensity. The resultant scattered light is therefore white. So, a thin layer of clouds appears white. What about dense clouds?

(C) Red colour of the Sun at Sunrise and Sunset

We are now able to understand the red colour of the Sun at sunrise and sunset. In the morning and evening when the Sun is near the horizon, light has to travel a greater distance through the atmosphere. The violet and blue wavelengths are scattered by dust particles and air molecules at an angle of about 90° . The sunlight thus becomes devoid of shorter wavelengths and the longer wavelength of red colour reaches the observer (Fig. 21.9). So the Sun appears to us as red.

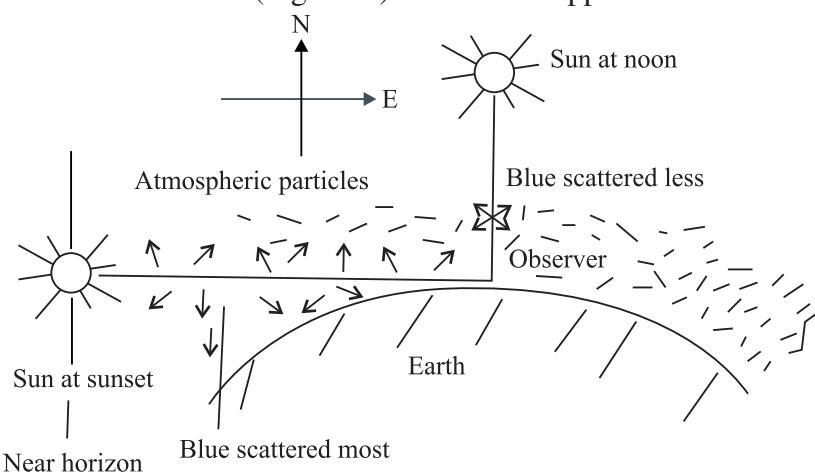


Fig. 21.9 : Red colour of the sun at sunset and sunrise (blue is scattered away).

At noon, the Sun is overhead and its distance from the observer is comparatively less. The blue colour is also scattered less. This results in the Sun appearing white, as a matter of fact, crimson.

21.2.2 Raman effect

When light radiation undergoes scattering from a transparent substance (solid, liquid or gas) then the frequency of the scattered radiation may be greater or less than the frequency of the incident radiation. This phenomenon is known as Raman effect as it was first observed by C. V. Raman in 1926. An analogue

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of this optical phenomenon was observed earlier by A. H. Compton in connection with the scattering of X-rays. The spectrum of the scattered radiation is known as Raman spectrum. This has lines having frequency greater than the frequency of the incident radiation (known as anti-Stokes' lines) as also lines having frequency less than the frequency of the incident radiation (called Stokes' lines).

A simple explanation of Raman effect can be given as follows. When light radiation interacts with a substance three possibilities may arise. In the first possibility, the light radiation interacting with the substance does not undergo any change of energy. Hence, its frequency remains unchanged. In the second possibility, the light radiation may impart some of its energy to the substance. As a result, the energy of the light radiation decreases. This leads to a decrease in the frequency of the scattered radiation (corresponding to Stokes' lines). In the third possibility, the incident radiation may interact with the substance which is already in the excited state. In the process, the radiation gains energy resulting into increase in its frequency (corresponding to anti-Stokes' lines).

Raman effect has lot of applications in various fields. C. V. Raman was awarded Nobel prize in physics for this discovery in 1930.



INTEXT QUESTIONS 21.2

1. Why dense clouds appear black?
2. Why does the sky appear deep blue after rains on a clear day?
3. Can you suggest an experiment to demonstrate the red colour of the Sun at sunrise and sunset?
4. The photographs taken from a satellite show the sky dark. Why?
5. What are anti stokes' lines?



WHAT YOU HAVE LEARNT

- Light of single wavelength or colour is said to be monochromatic but sunlight, which has several colours or wavelengths, is polychromatic.
- The splitting of light into its constituent wavelengths on entering an optically denser medium is called dispersion.
- A prism is used to produce dispersed light, which when taken on the screen, forms the spectrum.
- The angle of deviation is minimum if the angles of incidence and emergence become equal. In this situation, the beam is most intense for that colour.



Notes

- The angle of deviation and refractive index for a small-angled prism are connected by the relation $\delta = (\mu - 1)A$.
- The rainbow is formed by dispersion of sunlight by raindrops at definite angles for each colour so that the condition of minimum deviation is satisfied.
- Rainbows are of two types : primary and secondary. The outer side of the primary rainbow is red but the inner side is violet. The remaining colours lie in between to follow the order (VIBGYOR). The scheme of colours gets reversed in the secondary rainbow.
- The blue colour of the sky, the white colour of clouds and the reddish colour of the Sun at sunrise and sunset are due to scattering of light. The intensity of scattered light is inversely proportional to the fourth power of the wavelength $(I \propto \frac{1}{\lambda^4})$. This is called Rayleigh's law. So the blue colour is scattered more than the red.
- When light radiation undergoes scattering from a transparent substance, then frequency of scattered radiation may be greater or less than frequency of incident radiation. This phenomenon is known as Raman effect.



TERMINAL EXERCISE

- For a prism, show that $i + e = A + \delta$.
- Would you prefer small-angled or a large-angled prism to produce dispersion. Why?
- Under what condition is the deviation caused by a prism directly proportional to its refractive index?
- Explain why the sea water appears blue at high seas.
- The angle of minimum deviation for a 60° glass prism is 39° . Calculate the refractive index of glass.
- The deviation produced for red, yellow and violet colours by a crown glass are 2.84° , 3.28° and 3.72° respectively. Calculate the dispersive power of the glass material.
- Calculate the dispersive power for flint glass for the following data : $\mu_C = 1.6444$, $\mu_D = 1.6520$ and $\mu_F = 1.6637$, where C, D & F are the Fraunhofer nomenclatures.
- A lens can be viewed as a combination of two prisms placed with their bases together. Can we observe dispersion using a lens. Justify your answer.
- Human eye has a convex lens. Do we observe dispersion with unaided eye?

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ANSWERS TO INTEXT QUESTIONS

21.1

1. The velocity of propagation of waves of different wavelengths of visible light is almost the same in most ordinary gases. Hence, they do not disperse visible light. Their refractive index is also very close to 1.
2. Violet, because $\lambda_r > \lambda_v$ and the velocity of the red light is more than that of the violet light inside an optically denser medium.
3. No
4. $\mu = \frac{\sin 60^\circ}{\sin 30^\circ} = \sqrt{3} = 1.732$

21.2

1. It absorbs sunlight
2. It becomes clear of dust particles and bigger water molecules. The scattering now takes place strictly according to Rayleigh's law.
3. We can take sodium thiosulphate solution in a round bottom flask and add a small quantity of sulphuric acid. On illuminating this solution with a high power bulb, we can see a scenario similar to the colour of the sun at sunrise and sunset.
4. At very high altitudes no centres (particles) of scattering of sunlight are present. So the sky appears dark.
5. The spectral lines having frequency greater than the frequency of incident radiation are known as anti stokes' lines.

Answers to Problems in Terminal Exercise

5. 1.5
6. 0.27
7. 0.03

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22



Notes

WAVE PHENOMENA AND LIGHT

In the preceding two lessons of this module, you studied about reflection, refraction, dispersion and scattering of light. To understand these, we used the fact that light travels in a straight line. However, this concept failed to explain redistribution of energy when two light waves were superposed or their bending around corners. These observed phenomena could be explained only on the basis of wave nature of light. Christian Huygens, who was a contemporary of Newton, postulated that light is a wave and the wave theory of light was established beyond doubt through experimental observations on interference and diffraction. In this lesson, you will also learn about polarisation, which conclusively proved that light is a wave and transverse in nature.



OBJECTIVES

After studying this lesson, you should be able to :

- state Huygens' principle and apply it to explain wave propagation;
- explain the phenomena of interference and diffraction of light;
- explain diffraction of light by a single-slit; and
- show that polarisation of light established its wave nature; and
- derive Brewster's law.

22.1 HUYGENS' PRINCIPLE

Huygens' postulated that light is a wave, which travels through a hypothetical medium called ether. This hypothetical medium has the strange property of

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Wave phenomena and Light

occupying all space, including vacuum! The vibrations from the source of light propagate in the form of waves and the energy carried by them is distributed equally in all directions.

The concept of wavefront is central **Huygens' principle**. Let us first understand what a wavefront is with the help of a simple activity.

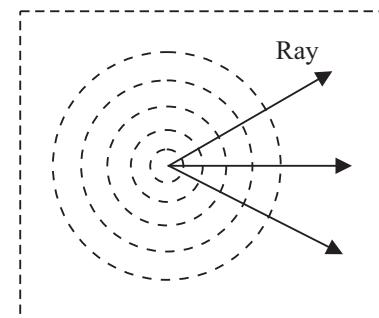


Fig. 22.1: Circular wavefronts on the surface of water



ACTIVITY 22.1

Take a wide based trough full of water and drop a small piece of stone in it. What do you observe? You will see that circular ripples due to the up and down motion of water molecules spread out from the point where the stone touched the water surface. If you look carefully at these ripples, you will notice that each point on the circumference of any of these ripples is in the same state of motion i.e., each point on the circumference of a ripple oscillates with the same amplitude and in the same phase. In other words, we can say that *the circumference of a ripple is the locus of the points vibrating in the same phase at a given instant and is known as the wavefront*. Therefore, the circular water ripples spreading out from the point of disturbance on the water surface represent a **circular wavefront**. Obviously, the distance of every point on a wavefront is the same from the point of disturbance, i.e., the source of waves.

For a point source emitting light in an isotropic medium, the locus of the points where all waves are in the same phase, will be a sphere. Thus, a point source of light **emits spherical wavefronts**. (In two dimensions, as on the water surface, the wavefronts appear circular.) Similarly, a line source of light emits **cylindrical wavefronts**. *The line perpendicular to the wavefront at a point represents the direction of motion of the wavefront at that point. This line is called the ray of light and a collection of such rays is called a beam of light.* When the source of light is at a large distance, any small portion of the wavefront can be considered to be a **plane wavefront**.

The **Huygens' principle** states that

- Each point on a wavefront becomes a source of secondary disturbance which spreads out in the medium.
- The position of wavefront at any later instant may be obtained by drawing a forward common envelop to all these secondary wavelets at that instant.
- In an isotropic medium, the energy carried by waves is transmitted equally in all directions.



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- If the initial shape, position, the direction of motion and the speed of the wavefront is known, its position at a later instant can be ascertained by geometrical construction. Note that the wavefront does not travel in the backward direction.

To visualise Huygens' construction, you may imagine a point source at the centre of a hollow sphere. The outer surface of this sphere acts as a primary wavefront. If this sphere is enclosed by another hollow sphere of larger radius, the outer surface of the second hollow sphere will act as a secondary wavefront. (The nearest mechanical analogue of such an arrangement is a football.) If the second sphere is further enclosed by another sphere of still bigger radius, the surface of the outermost (third) sphere becomes secondary wavefront and the middle (second) sphere acts as the primary wavefront. In two dimensions, the primary and secondary wavefronts appear as concentric circles.

22.1.1 Propagation of Waves

Now let us use Huygens' principle to describe the propagation of light waves in the form of propagation of wavefronts. Fig. 22.2 shows the shape and location of a plane wavefront AB at the time $t = 0$. You should note that the line AB lies in a plane perpendicular to the plane of the paper. Dots represented by a, b, c, \dots , on the wavefront AB are the sources of secondary wavelets. All these sources emit secondary wavelets at the same time and they all travel with the same speed along the direction of motion of the wavefront AB . In Fig. 22.2, the circular arcs represent the wavelets emitted from a, b, c, \dots taking each point as center. *These wavelets have been obtained by drawing arcs of radius, $r = vt$, where v is the velocity of the wavefront and t is the time at which we wish to obtain the wavefront.* The tangent, CD , to all these wavelets represents the new wavefront at time $t = T$.

Let us take another example of Huygens' construction for an expanding circular wavefront. Refer to Fig. 22.3, which indicates a circular wavefront, centred at O , at time $t = 0$. Position A, B, C, \dots represent point sources on this wavefront. Now to draw the wavefront at a later time $t = T$, what would you do? You should draw arcs from the points A, B, C, \dots , of radius equal to the speed of the expanding wavefront multiplied by T . These arcs will

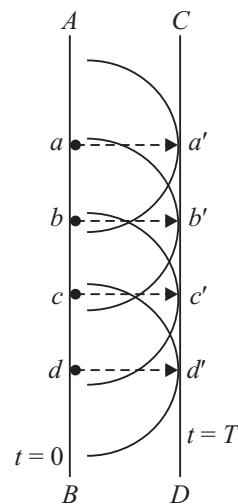


Fig. 22.2: Construction of a plane wave front

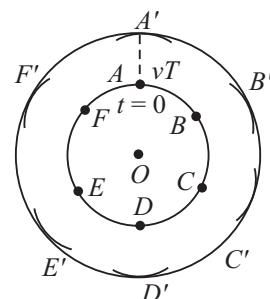


Fig. 22.3: Construction of circular wavefront using Huygens' principle

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represent secondary wavelets. The tangents drawn to these arcs will determine the shape and location of the expanding circular wavefront at time T .

We hope you have now understood the technique of Huygens' construction. Now, you may like to know the *physical significance of Huygens' construction*. By determining the shape and location of a wavefront at a subsequent instant of time with the help of its shape and location at an earlier instant, we are essentially describing the propagation of the wavefront. Therefore, Huygens' construction enables us to describe wave motion.



INTEXT QUESTIONS 22.1

1. What is the relative orientation of a wavefront and the direction of propagation of the wave?
2. A source of secondary disturbance is emitting wavelets at an instant $t = 0$ s. Calculate the ratio of the radii of wavelets at $t = 3\text{s}$ and $t = 6\text{s}$.

22.2 INTERFERENCE OF LIGHT

Let us first perform a simple activity:



ACTIVITY 22.2

Prepare a soap solution by adding some detergent powder to water. Dip a wire loop into the soap solution and shake it. When you take out the wire loop, you will find a thin film on it. Bring this soap film near a light bulb and position yourself along the direction of the reflected light from the film. You will observe beautiful colours. Do you know the reason? To answer this question, we have to understand the phenomenon of *interference of light*. In simple terms, *interference of light refers to redistribution of energy due to superposition of light waves from two coherent sources*. The phenomenon of interference of light was first observed experimentally by Thomas Young in 1802 in his famous two-slit experiment. This experimental observation played a significant role in establishing the wave theory of light. The basic theoretical principle involved in the phenomenon of interference as well as diffraction of light is the *superposition principle*.

22.2.1 Young's Double Slit Experiment

Young's experimental set up is shown schematically in Fig. 22.4. In his experiment, sunlight was allowed to pass through a pin hole S and then, at some distance away, through two pin holes S_1 and S_2 equidistant from S and close to each other.



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According to Huygens' wave theory of light, spherical wavefronts would spread out from the pin hole S which get divided into two wavefronts by S_1 and S_2 . If S is illuminated by a monochromatic source of light, such as sodium, these act as coherent sources and in-phase waves of equal amplitude from these sources superpose as they move beyond S_1S_2 . As a consequence of superposition (of the two sets of identical waves from S_1 and S_2), redistribution of energy takes place and a pattern consisting of alternate bright and dark fringes is produced on the screen such as placed at C . Let us now learn the explanation of the observed fringe pattern in the Young's interference experiment.

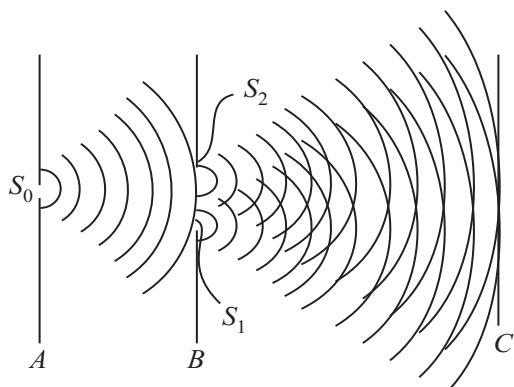


Fig. 22.4: Schematic arrangement of Young's double-slit experiment

Eugene Thomas Young (1773-1829)



Born on 16 June, 1773, Eugene Thomas Young will always be known for his study on the human ear, the human eye, how it focuses and on astigmatism. His research on colour blindness led him to the three component theory of colour vision. Working on human ears and eyes, he dedicated much time to the speed of sound and light. He knew that if two sound waves of equal intensity reached the ear 180° out of phase, they cancelled out each other's effect and no sound was heard. It occurred to him that a similar interference effect should be observed with two light beams, if light consisted of waves. This led Young to devise an experiment, now commonly referred to as the Young's double-slit experiment.

In his later years, Young devoted most of his time deciphering the Egyptian hieroglyphics found on the Rosetta stone discovered in the Nile Delta in 1799.

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(a) Constructive Interference: You may recall from the superposition principle that some points on the screen C will have maximum displacement (or amplitude) because the crests due to one set of waves coincide with the crests due to another set of waves. In other words, at this point, the waves arrive in-phase and hence the total amplitude is much higher than the amplitude of individual waves. The same holds true for the points where the troughs due to one set of waves coincide with the troughs due to another set. Such points will appear bright because the intensity of light wave is proportional to the square of the amplitude. Superposition of waves at these points leads to what is known as ***constructive interference***.

(b) Destructive Interference: The points where the crests due to one set of waves coincide with the troughs due to the other set and vice-versa, the total amplitude is zero. It is so because the waves reach these points completely out of phase. Such points appear dark on the screen. These points correspond to ***destructive interference***.

(c) Intensity of fringes: To analyse the interference pattern, we calculate the intensity of the bright and dark fringes in the interference pattern for harmonic waves. Refer to Fig. 22.5, which is schematic representation of the geometry of Young's experiment. The phenomenon of interference arises due to superposition of two harmonic waves of same frequency and amplitude but differing in phase. Let the phase difference between these two waves be δ . We can write y_1 and y_2 , the displacements at a fixed point P due to the two waves, as

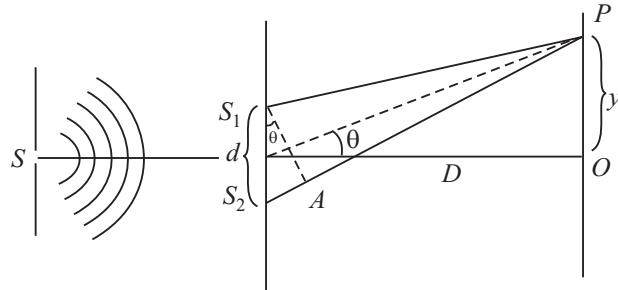


Fig. 22.5: Geometry of Young's double slit experiment

$$y_1 = a \sin \omega t$$

$$\text{and} \quad y_2 = a \sin (\omega t + \delta)$$

where δ signifies the phase difference between these waves. Note that we have not included the spatial term because we are considering a fixed point in space.

According to the principle of superposition of waves, the resultant displacement is given by

$$\begin{aligned} y &= y_1 + y_2 \\ &= a \sin \omega t + a \sin (\omega t + \delta) \\ &= a [\sin \omega t + \sin (\omega t + \delta)] \\ &= 2a \sin \left(\omega t + \frac{\delta}{2} \right) \cos \left(-\frac{\delta}{2} \right) \\ &= A \sin \left(\omega t + \frac{\delta}{2} \right) \end{aligned}$$



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where amplitude of the resultant wave is given by

$$A = 2a \cos(\delta/2).$$

The intensity of the resultant wave at point P can be expressed as

$$\begin{aligned} I &\propto A^2 \\ &\propto 4a^2 \cos^2(\delta/2) \end{aligned} \quad (22.1)$$

To see the dependence of intensity on the phase difference between the two waves, let us consider the following two cases.

Case 1: When the phase difference, $\delta = 0, 2\pi, 4\pi, \dots, 2n\pi$

$$\begin{aligned} I &= 4a^2 \cos^2 0 \\ &= 4a^2 \end{aligned}$$

Case 2: When, $\delta = \pi, 3\pi, 5\pi, \dots, (2n+1)\pi$

$$\begin{aligned} I &= 4a^2 \cos^2(\delta/2) \\ &= 0 \end{aligned}$$

From these results we can conclude that when phase difference between superposing waves is an integral multiple of 2π , the two waves arrive at the screen ‘in-phase’ and the resultant intensity (or the brightness) at those points is more than that due to individual waves (which is equal to $4a^2$). On the other hand, when phase difference between the two superposing waves is an odd multiple of π , the two superposing waves arrive at the screen ‘out of phase’. Such points have zero intensity and appear to be dark on the screen.

(d) Phase Difference and Path Difference

It is obvious from the above discussion that to know whether a point on the screen will be bright or dark, we need to know the phase difference between the waves arriving at that point. The phase difference can be expressed in terms of the path difference between the waves during their journey from the sources to a point on the observation screen. You may recall that waves starting from S_1 and S_2 are in phase. Thus, whatever phase difference arises between them at the point P is because of the different paths travelled by them upto observation point from S_1 and S_2 . From Fig. 22.5, we can write the path difference as

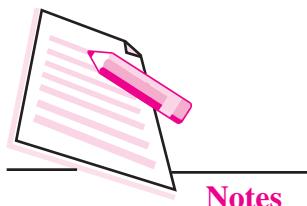
$$\Delta = S_2P - S_1P$$

We know that path difference of one wavelength is equivalent to a phase difference of 2π . Thus, the relation between the phase difference δ and the path difference Δ is

$$\Delta = \left(\frac{\lambda}{2\pi} \right) \delta \quad (22.2)$$

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From Eqn. (22.1) we note that bright fringes (corresponding to constructive interference) are observed when the phase difference is $2n\pi$. Using this in Eqn. (22.2) we find that the path difference for observing bright fringes is

$$(\Delta)_{\text{bright}} = \left(\frac{\lambda}{2\pi} \right) 2n\pi = n\lambda; n = 0, 1, 2, \dots \quad (22.3)$$

Similarly, for dark fringes, we get

$$\begin{aligned} (\Delta)_{\text{dark}} &= (\lambda/2\pi) (2n+1) \pi \\ &= (2n + 1) \frac{\pi}{2}; n = 0, 1, 2, \dots \end{aligned} \quad (22.4)$$

Having obtained expressions for the bright and dark fringes in terms of the path difference and the wavelength of the light used, let us now relate path difference with the geometry of the experiment, i.e., relate Δ with the distance D between the source and the screen, separation between the pin holes (d) and the location of the point P on the screen. From Fig. 22.5 we note that

$$\Delta = S_2 P - S_1 P = S_2 A = d \sin \theta$$

Assuming θ to be small, we can write

$$\sin \theta \approx \tan \theta \approx \theta$$

and

$$\sin \theta = x / D$$

Therefore, the expression for path difference can be rewritten as

$$\Delta = d \sin \theta = x \frac{d}{D} \quad (22.5)$$

On substituting Eqn. (22.5) in Eqns. (22.2) and (22.3), we get

$$\frac{d}{D} (x_n)_{\text{bright}} = n\lambda$$

or $(x_n)_{\text{bright}} = \frac{n\lambda D}{d}; n = 0, 1, 2, \dots \quad (22.6)$

and $\frac{d}{D} (x_n)_{\text{dark}} = (n + \frac{1}{2})\lambda$

or $(x_n)_{\text{dark}} = (n + \frac{1}{2}) \frac{\lambda D}{d}; n = 0, 1, 2, \dots \quad (22.7)$

Eqns. (22.6) and (22.7) specify the positions of the bright and dark fringes on the screen.

(e) Fringe width

You may now ask: How wide is a bright or a dark fringe? To answer this question, we first determine the location of two consecutive bright (or dark) fringes. Let us



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first do it for bright fringes. For third and second bright fringes, from Eqn. (22.6), we can write

$$(x_3)_{\text{bright}} = 3 \frac{\lambda D}{d}$$

and

$$(x_2)_{\text{bright}} = 2 \frac{\lambda D}{d}$$

Therefore, fringe width, β is given by

$$\beta = (x_3)_{\text{bright}} - (x_2)_{\text{bright}} = \frac{\lambda D}{d} \quad (22.8)$$

You should convince yourself that the fringe width of an interference pattern remains the same for any two consecutive value of n . Note that fringe width is directly proportional to linear power of wavelength and distance between the source plane and screen and inversely proportional to the distance between the slits. In actual practice, fringes are so fine that we use a magnifying glass to see them.

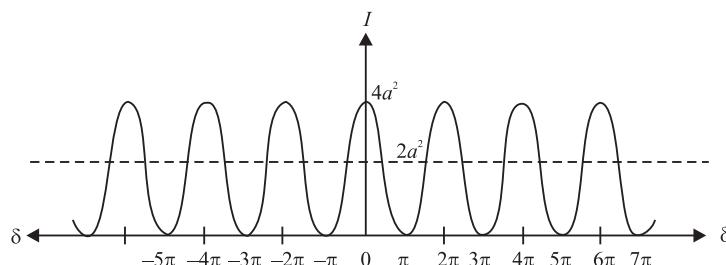


Fig. 22.6: Intensity distribution in an interference pattern

Next let us learn about the intensity of bright and dark fringes in the interference pattern. We know that when two light waves arrive at a point on the screen out of phase, we get dark fringes. You may ask : Does this phenomenon not violate the law of conservation of energy because energy carried by two light waves seem to be destroyed? It is not so; the energy conservation principle is not violated in the interference pattern. Actually, the energy which disappears at the dark fringes reappears at the bright fringes. You may note from Eqn. (22.1) that the intensity of the bright fringes is four times the intensity due to an individual wave. Therefore, in an interference fringe pattern, shown in Fig. (22.6), **the energy is redistributed and it varies between $4a^2$ and zero**. Each beam, acting independently, will contribute a^2 and hence, in the absence of interference, the screen will be uniformly illuminated with intensity $2a^2$ due to the light coming from two identical sources. This is the average intensity shown by the broken line in Fig. 22.6.

You have seen that the observed interference pattern in the Young's experiment can be understood qualitatively as well as quantitatively with the help of wave

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theory of light. To be sure that you have good understanding, answer the following intext questions.



INTEXT QUESTIONS 22.2

1. On what factors does the resultant displacement at any point in the region of superposition of two waves depend?
2. In Young's experiment, how is the constructive interference produced on the screen?
3. If we replace the pinholes S_1 and S_2 by two incandescent light bulbs, can we still observe the bright and dark fringes on the screen?
4. What are coherent sources? Can our eyes not act as coherent sources?

22.3 DIFFRACTION OF LIGHT

In earlier lessons, you were told that rectilinear propagation is one of the characteristics of light. The most obvious manifestation of the rectilinear propagation of light is in the formation of shadow. But, if you study formation of shadows carefully, you will find that, as such, these are not sharp at the edges. For example, the law of rectilinear propagation is violated when the light passes through a very narrow aperture or falls on an obstacle of very small dimensions. At the edges of the aperture or the obstacle, light bends into the shadow region and does not propagate along a straight line. ***This bending of light around the edges of an obstacle is known as diffraction.***

Before discussing the phenomenon of diffraction of light in detail, you may like to observe diffraction of light yourself. Here is a simple situation. Look at the street light at night and almost close your eyes. What do you see? The light will appear to streak out from the lamp/tube. This happens due to the diffraction (bending) of light round the corners of your eyelids.

Another way to observe diffraction is to use a handkerchief. Hold it close to your eyes and look at the Sun or a lamp. You will observe circular fringes, which form due to diffraction of light by small apertures formed by crissed-crossed threads.

In the above situations, the dimensions of the diffracting obstacle/aperture are very small. To observe diffraction, either of the following conditions must be satisfied:

- a) *The size of the obstacle or the aperture should be of the order of the wavelength of the incident wave.*



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- b) The separation between the obstacle or aperture and the screen should be considerably larger (a few thousand times) than the size of the obstacle or aperture.

On the basis of the above observations, it is easy to understand why we normally do not observe diffraction of light and why light appears to travel in a straight line. You know that the wavelength of light is of the order of 10^{-6} m. Therefore, *to observe diffraction of light, we need to have obstacles or aperture having dimensions of this order!*

22.3.1 Diffraction at a Single Slit

Let us see how diffraction pattern appears for a simple opening like a single slit. Refer to Fig. 22.7. It shows the experimental arrangement for producing diffraction pattern. S is a monochromatic source of light. It is placed on the focal plane of a converging lens so that a plane wavefront is incident on a narrow slit. Another converging lens focusses light from different portions of the slit on the observation screen.

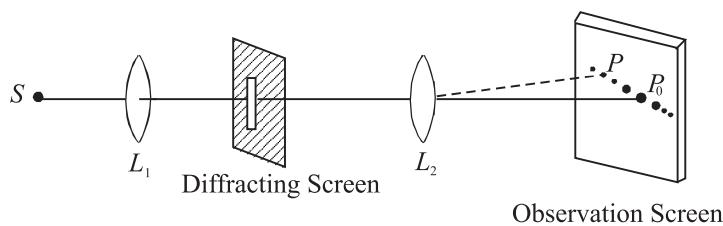


Fig. 22.7: Schematic representation of single slit diffraction

The salient features of the actual diffraction pattern produced by a single vertical slit from a point source as shown in Fig. 22.8 are :

- A horizontal streak of light along a line normal to the length of the slit.
- The horizontal pattern is a series of bright spots.

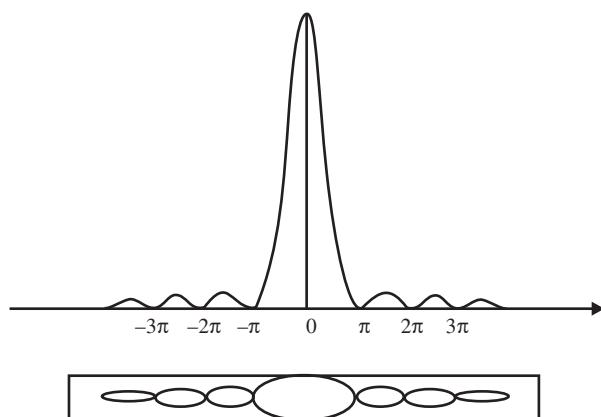


Fig. 22.8 : Observed diffraction pattern single of slit

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- The spot at the centre is the brightest. On either side of this spot, we observe a few more symmetrically situated bright spots of diminishing intensity. The central spot is called *principal maxima* and other spots are called *secondary maxima*.
- The width of the central spot is twice the width of other spots.

To understand the theoretical basis of these results, we note that according to Huygens' wave theory, plane wavefronts are incident on the barrier containing the slit. As these wavefronts fall on the barrier, only that part of the wavefront passes through the slit which is incident on it. This part of the wavefront continues to propagate to the right of the barrier. However, the shape of the wavefront does not remain plane beyond the slit.

Refer to Fig. 22.9 which shows that each point of the aperture such as QPR ... Q' form a series of coherent sources of secondary wavelets. In the central part of the wavefront to the right of the barrier, the wavelet emitted from the point *P*, say, spreads because of the presence of wavelets on its both sides emitted from the points such as *Q* and *R*. Since the shape of the wavefront is determined by the tangent to these wavelets, the central part of the wavefront remains plane as it propagates. But for the wavelets emitted from points *Q* and *Q'* near the edges of the slit, there are no wavelets beyond the edges with which these may superpose. Since the superposition helps to maintain the shape of the wavefront as plane, the absence of such superposing wavelets for the wavelets emitted from the points near the edges allows them to deviate from their plain shape. In other words, the wavelets at the edges tend to spread out. As a result, the plane wavefront incident on a thin aperture of finite size, after passing through it does not remain plane.

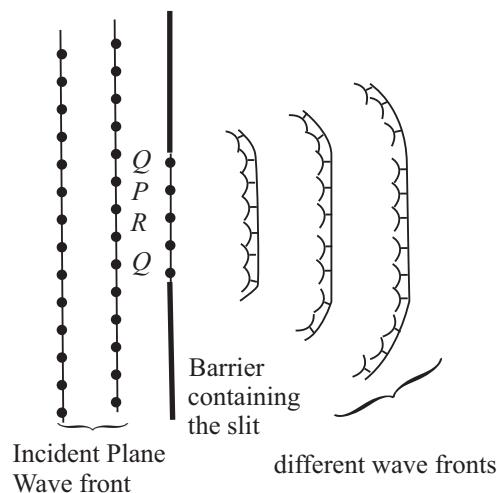


Fig. 22.9: Huygen's construction for diffraction of light from a narrow slit

To understand the *intensity distribution* of the single-slit diffraction pattern, we determine the nature of the superposition of waves reaching the screen. In order to apply Huygens' principle, let us divide the width '*a*' of the slit into, say, 100



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equal parts. Each of these can be considered as a sources of secondary wavelets. The wavelets emanating from these points spread out into the region to the right of the slit. Since the plane wavefront is incident on the slit, initially all points on it are in phase. Therefore, the wavelets emitted by these points are all in phase at the time of leaving the slit. Now let us consider the effect of the superposition of these wavelets at point O on the screen. The symmetry of the Fig. 22.10 suggests that the wavelets emitted from source of 1 and 100 will reach O in phase. It is so because both the wavelets travel equal path length. When they started their journey from the respective points on the slit, they were in phase. Hence they arrive at O in phase and superpose in such a manner as to give resultant amplitude much more than that due to the individual wavelets from the source 1 and 100. Similarly, for each wavelet from source 2 to 50, we have a corresponding wavelet from the source 99 to 51 which will produce constructive interference causing enhancement in intensity at the center O . Thus the point O will appear bright on the screen.

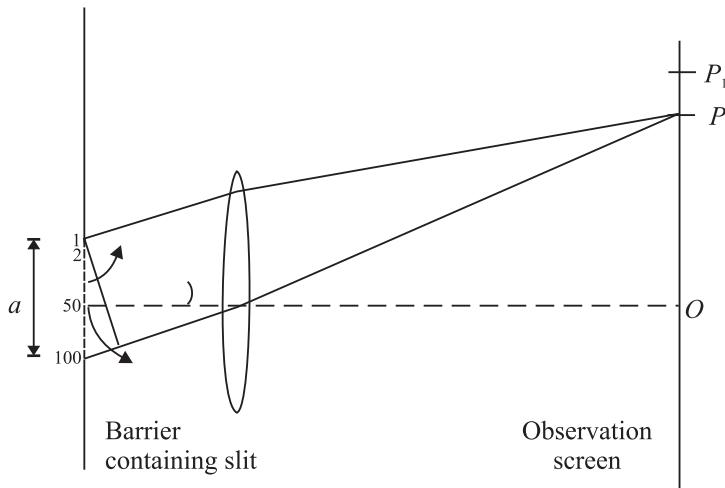


Fig. 22.10: Schematic representation of single slit diffraction

Now let us consider an off-axis point P on the observation screen. Suppose that point P is such that the *path difference* between the extreme points i.e. sources 1 and 100 is equal to λ . Thus the path difference between the wavelets from source 1 and 51 will be nearly equal to $(\lambda/2)$.

You may recall from the interference of light that the waves coming from the sources 1 and 51 will arrive at P out of phase and give rise to destructive interference. Similarly, wavelets from the sources 2 and 52 and all such pair of wavelets will give rise to destructive interference at the point P . Therefore, we will have minimum intensity at point P . Similarly, we will get minimum intensity for other points for which the path difference between the source edges is equal to 2λ . We can imagine that the slit is divided into four equal points and we can, by similar pairing of 1 and 26, 2 and 27, ... show that first and second quarters have a path difference of $\lambda/2$ and cancel each other. Third and fourth quarters cancel

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each other by the same argument so that the resultant intensity will be minimum, and so on. *We can therefore conclude that when the path difference between the extreme waves diffracted by the extreme points in a particular direction is an integral multiple of λ , the resultant diffracted intensity in that direction will be zero.*

Let us now find intensity at a point P' which lies between the points P and P_1 and the *path difference between waves diffracted from extreme points is $3\lambda/2$* . We divide the wavefront at the slit into 3 equal parts. In such a situation, secondary wavelets from the corresponding sources of two parts will have a path difference of $\lambda/2$ when they reach the point P and cancel each other. However, wavelets from the third portion of the wavefront will all contribute constructively (presuming that practically the path difference for wavelets from this part is zero) and produce brightness at P' . Since only one third of the wavefront contributes towards the intensity at P' as compared to O , where the whole wavefront contributes, the intensity at P' is considerably less than that of the intensity at O . The point P' and all other similar points constitute secondary maxima.

However, you must note here that this is only a qualitative and simplified explanation of the diffraction at a single slit. You will study more rigorous analysis of this phenomena when you pursue higher studies in physics.



INTEXT QUESTIONS 22.3

1. Does the phenomenon of diffraction show that the light does not travel along a straight line path?
2. Distinguish between interference and diffraction of light.
3. Why are the intensity of the principal maximum and the secondary maxima of a single slit diffraction not the same?

22.4 POLARISATION OF LIGHT

In the previous two sections of this lesson, you learnt about the phenomena of interference and diffraction of light. While discussing these phenomena, we did not bother to know the nature of light waves; whether these were longitudinal or transverse. However, polarisation of light conclusively established that light is a transverse wave.

To understand the phenomenon of polarisation, you can perform a simple activity.



Notes

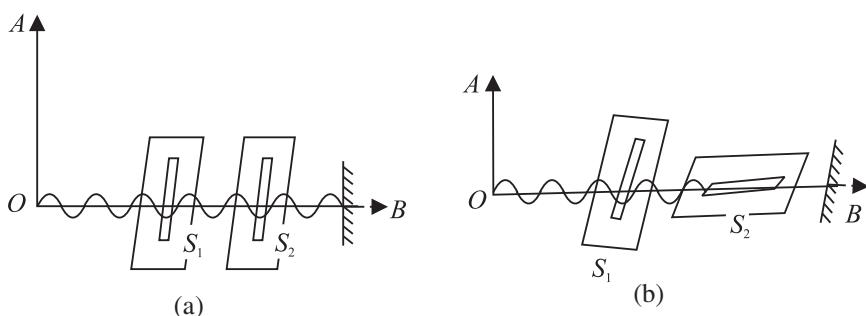


Fig. 22.11 : Transverse wave on a rope passing through a) two vertical slits, and b) one vertical and one horizontal slit

Take two card boards having narrow vertical slits S_1 and S_2 and hold them parallel to each other. Pass a length of a string through the two slits, fix its one end and hold the other in your hand. Now move your hand up and down and sideways to generate waves in all directions. You will see that the waves passing through the vertical slit S_1 will also pass through S_2 , as shown in Fig. 22.11(a). Repeat the experiment by making the slit S_2 horizontal. You will see no waves beyond S_2 . It means that waves passing through S_1 cannot pass through the horizontal slit S_2 . This is because the vibrations in the wave are in a plane at right angles to the slits S_2 , as shown in Fig. 22.11(b).

This activity can be repeated for light by placing a source of light at O and replacing the slits by two polaroids. You will see light in case(a) only. This shows that light has vibrations confined to a plane. It is said to be *linearly polarised* or *plane polarised* after passing through the first polaroid (Fig. 22.12).

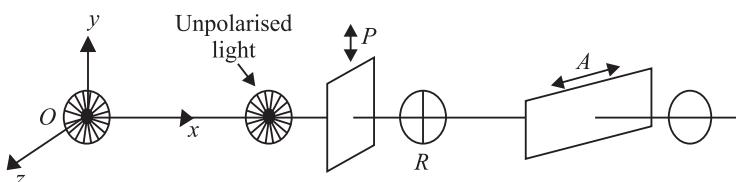


Fig. 22.12 : Schematics of the apparatus for observing polarisation of light

When an unpolarised light falls on glass, water or any other transparent material, the reflected light is, in general, partially plane polarised. Fig. 22.13 shows unpolarised light AO incident on a glass plate. The reflected light is shown by OR and the transmitted wave by OT . When the light is incident at polarising angle, the polarisation is complete. At this angle, the reflected and transmitted rays are at right angles to each other.

The polarising angle depends on the refractive index of the material of glass plate on which the (unpolarised) beam of light is incident. The relation between r and

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Wave phenomena and Light

i_p is obtained by using Snell's law (refer Fig. 22.13):

$$\begin{aligned}\mu &= \frac{\sin i_p}{\sin r} = \frac{\sin i_p}{\sin(90 - i_p)} \\ &= \frac{\sin i_p}{\cos i_p} = \tan i_p.\end{aligned}$$

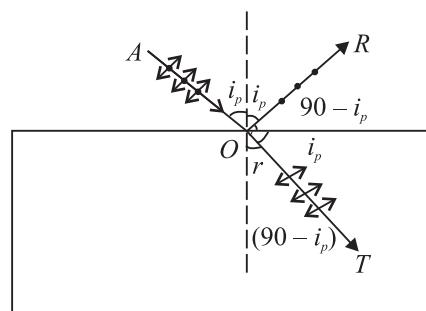


Fig. 22.13 : Polarisation of reflected and refracted light

This is known as **Brewster's law**. It implies that polarising angle i_p depends on the refractive index of the material. For air water interface, $i_p = 53^\circ$. It means that when the sun is 37° above the horizontal, the light reflected from a calm pond or lake will be completely linearly polarised. Brewster's law has many applications in daily life. Glare caused by the light reflected from a smooth surface can be reduced by using polarising materials called *polaroids*, which are made from tiny crystals of quinine iodosulphate; all lined up in the same direction in a sheet of nitro cellulose. Such crystals (called dichoric) transmit light in one specific plane and absorb those in a perpendicular plan. Thus, *polaroid* coatings on sunglasses reduce glare by absorbing a component of the polarized light. Polaroid discs are used in photography as 'filters' in front of camera lens and facilitate details which would otherwise be hidden by glare. Polarimeters are used in sugar industry for quality control.



INTEXT QUESTIONS 22.4

1. Polarisation of light is the surest evidence that light is a transverse wave. Justify.
2. Is it correct to say that the direction of motion of a wave may not lie in the plane of polarisation?
3. Suppose a beam of unpolarised light is incident on a set of two *polaroids*. If you want to block light completely with the help of these polaroids, what should be the angle between the transmission axes of these polaroids?
4. Do sound waves in air exhibit polarization?



WHAT YOU HAVE LEARNT

- According to the Huygens' wave theory, light propagates in the form of wavefronts.
- The locus of all particles of the medium vibrating in the same phase at any instant of time is called the wavefront
- If two light sources emit light waves of the same frequency, same amplitude and move along the same path maintaining a constant phase difference between them, they are said to be coherent.
- When waves from two coherent sources superpose, a redistribution of energy takes place at different points. This is called the interference of light.
- For constructive interference, phase difference $\Delta = 2n\pi$ and for destructive interference, phase difference $\Delta = (2n + 1)\pi$.
- The bending of light near the corners of an obstacle or aperture is called diffraction of light.
- The phenomenon in which vibrations of light get confined in a particular plane containing the direction of propagation is called polarisation of light.



Notes



TERMINAL EXERCISE

1. Explain in brief the theories describing the nature of light.
2. What is a wavefront? What is the direction of a beam of light with respect to the associated wavefront? State the Huygens' principle and explain the propagation of light waves.
3. Obtain the laws of reflection on the basis of Huygens' wave theory.
4. What is the principle of superposition of waves? Explain the interference of light.
5. Describe Young's double slit experiment to produce interference. Deduce an expression for the width of the interference fringes.
6. What would happen to the interference pattern obtained in the Young's double slit experiment when
 - (i) one of the slits is closed;
 - (ii) the experiment is performed in water instead of air;
 - (iii) the source of yellow light is used in place of the green light source;
 - (iv) the separation between the two slits is gradually increased;

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Wave phenomena and Light

- (v) white light is used in place of a monochromatic light;
- (vi) the separation between the slits and the screen is increased;
- (vii) two slits are slightly moved closer; and
- (viii) each slit width is increased.
7. In Young's experimental set-up, the slit separation is 2 mm and the distance between the slits and the observation screen is 100 cm. Calculate the path difference between the waves arriving at a point 5 cm away from the point where the line dividing the slits touches the screen.
8. With the help of Huygens' construction, explain the phenomenon of diffraction.
9. How would you demonstrate that the light waves are transverse in nature?
10. Distinguish between the polarized and unpolarized lights.
11. State and explain Brewster's law.
12. The polarising angle for a medium is 60° . Calculate the refractive index.
13. For a material of refractive index 1.42, calculate the polarising angle for a beam of unpolarised light incident on it.



ANSWERS TO INTEXT QUESTIONS

22.1

1. Perpendicular to each other ($\theta = \pi/2$) 2. $1/2$

22.2

1. On the amplitude of the waves and the phase difference between them.
2. When the phase difference between the two superposing beams is an integral multiple of 2π , we obtain constructive interference.
3. No, it is so because two independent sources of light will emit light waves with different wavelengths, amplitudes and the two set of waves will not have constant phase relationship. Such sources of light are called incoherent sources. For observing interference of light, the sources of light must be coherent. When the light waves are coming from two incoherent sources, the points on the screen where two crests or two trough superpose at one instant to produce brightness may receive, at the other instant, the crest of the wave from one source and trough from the other and produce darkness. Thus, the whole screen will appear uniformly illuminated if the pinholes S_1 and S_2 are replaced by two incandescent light bulbs.



Notes

4. Coherent sources should emit waves

- (a) of same frequency and wavelength,
- (b) in phase or having constant phase difference, and
- (c) same amplitude and period.

Moreover, these should be close. Our eyes may not meet this criterion.

22.3

1. Yes
2. Interference is the superposition of secondary waves emanating from two different secondary sources whereas diffraction is the superposition of secondary waves emanating from different portions of the same wavefronts.
3. Due to the increasing path difference between wavelets.

22.4

1. No. Because, in a longitudinal wave, the direction of vibrations is the same as the direction of motion of the wave.
2. No. 3. 90° or 270° 4. No.

Answers to Problems in Terminal Exercise

7. 0.1 mm 12. 1.73 13. 54°

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Notes

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OPTICAL INSTRUMENTS

We get most of the information about the surrounding world through our eyes. But as you know, an unaided eye has limitations; objects which are too far like stars, planets etc. appear so small that we are unable to see their details. Similarly, objects which are too small, e.g. pollen grains, bacteria, viruses etc. remain invisible to the unaided eyes. Moreover, our eyes do not keep a permanent record of what they see, except what is retained by our memory. You may therefore ask the question: How can we see very minute and very distant objects? The special devices meant for this purpose are called **optical instruments**.

In this lesson you will study about two important optical instruments, namely, a microscope and a telescope. As you must be knowing, a microscope magnifies small objects while a telescope is used to see distant objects. The design of these appliances depends on the requirement. (The knowledge of image formation by the mirrors and lenses, which you have acquired in Lesson 20, will help you understand the working of these optical instruments.) The utility of a microscope is determined by its magnifying power and resolving power. For a telescope, the keyword is *resolving power*. You must have read about Hubble's space telescope, which is being used by scientists to get details of far off galaxies and search for a life-sustaining planet beyond our solar system.



OBJECTIVES

After studying this lesson, you should be able to:

- explain the working principle of simple and compound microscopes;
- derive an expression for the magnifying power of a microscope;
- distinguish between linear and angular magnifications;
- explain the working principle of refracting and reflecting telescopes; and
- calculate the resolving powers of an eye, a telescope and a microscope.



Notes

23.1 MICROSCOPE

In Lesson 20 you have learnt about image formation by mirrors and lenses. If you take a convex lens and hold it above this page, you will see images of the alphabets/words. If you move the lens and bring it closer and closer to the page, the alphabets printed on it will start looking enlarged. This is because their enlarged, virtual and erect image is being formed by the lens. That is, it is essentially acting as a magnifying glass or simple microscope. You may have seen a doctor, examining measles on the body of a child. Watch makers and jewellers use it to magnify small components of watches and fine jewellery work. You can take a convex lens and try to focus sunlight on a small piece of paper. You will see that after some time, the piece of paper starts burning. A convex lens can, therefore, start a fire. That is why it is dangerous to leave empty glass bottles in the woods. The sunlight falling on the glass bottles may get focused on dry leaves in the woods and set them on fire. Sometimes, these result in wild fires, which destroy large parts of a forest and/or habitation. Such fires are quite common in Australia, Indonesia and U.S.

As a simple microscope, a convex lens is satisfactory for magnifying small nearby objects up to about twenty times their original size. For large magnification, a compound microscope is used, which is a combination of basically two lenses. In a physics laboratory, a magnifying glass is used to read vernier scales attached to a travelling microscope and a spectrometer.

While studying simple and compound microscopes, we come across scientific terms like (i) near point, (ii) least distance of distinct vision, (iii) angular magnification or magnifying power, (iv) normal adjustment etc. Let us first define these.

- (i) **Near point** is the distance from the eye for which the image of an object placed there is formed (by eye lens) on the retina. The near point varies from person to person and with the age of an individual. At a young age (say below 10 years), the near point may be as close as 7-8 cm. In the old age, the near point shifts to larger values, say 100-200 cm or even more. That is why young children tend to keep their books so close whereas the aged persons keep a book or newspaper far away from the eye.
- (ii) **Least distance of distinct vision** is the distance upto which the human eye can see the object clearly without any strain on it. For a normal human eye, this distance is generally taken to be 25 cm.
- (iii) **Angular magnification** is the ratio of the angle subtended by the image at the eye (when the microscope is used) to the angle subtended by the object at the unaided eye when the object is placed at the least distance of distinct vision. It is also called the magnifying power of the microscope.

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Optical Instruments

- (iv) **Normal Adjustment:** When the image is formed at infinity, least strain is exerted on the eye for getting it focused on the retina. This is known as normal adjustment.
- (v) **Linear magnification** is the ratio of the size of the image to the size of the object.
- (vi) **Visual angle** is the angle subtended by the object at human eye.

23.1.1 A Simple Microscope

When a convex lens of short focal length is used to see magnified image of a small object, it is called a simple microscope.

We know that when an object is placed between the optical center and the focus of a convex lens, its image is virtual, erect, and magnified and on the same side as the object. In practice, such a lens is held close to eye and the distance of the object is adjusted till a clear image is formed at the least distance of distinct vision. This is illustrated in Fig. 23.1, which shows an object AB placed between F and O . Its virtual image $A'B'$ is formed on the same side as the object. The position of the object is so adjusted that the image is formed at the least distance of distinct vision (D).

Magnifying power of a simple microscope

Magnifying power of an optical instrument is the ratio of the angle subtended by the image at the eye to the angle subtended by the object seen directly, when both lie at the least distance of distinct vision or the near point. It is also called angular magnification and is denoted by M . Referring to Fig. 23.1(a) and (b), the angular magnification of simple

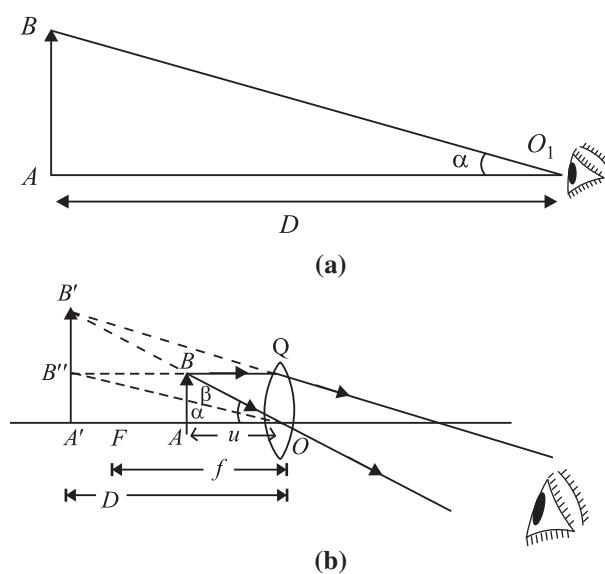


Fig.23.1 : Angular magnification of a magnifying glass



Notes

microscope is given by $M = \frac{\angle A'OB'}{\angle AOB} = \frac{\beta}{\alpha}$. In practice, the angles α and β are small. Therefore, you can replace these by their tangents, i.e. write

$$M = \frac{\tan \beta}{\tan \alpha} \quad (23.1)$$

From $\Delta s A'OB'$ and ΔAOB , we can write $\tan \beta = \frac{A'B'}{A'O} = \frac{A'B'}{D}$ and

$\tan \alpha = \frac{A'B''}{A'O} = \frac{AB}{D}$. On putting these values of $\tan \beta$ and $\tan \alpha$ in Eqn. (23.1), we get

$$M = \frac{A'B'}{D} / \frac{AB}{D} = \frac{A'B'}{AB}$$

Since $\Delta s AOB$ and $A'OB'$ in Fig 23.1(b) are similar, we can write

$$\frac{A'B'}{AB} = \frac{A'O}{AO} \quad (23.2)$$

Following the standard sign convention, we note that

$$A'O = -D$$

and

$$AO = -u$$

Hence, from Eqn. (23.2), we obtain

$$\frac{A'B'}{AB} = \frac{D}{u} \quad (23.3)$$

If f is the focal length of the lens acting as a simple microscope, then using the lens formula $\left(\frac{1}{v} - \frac{1}{u} = \frac{1}{f}\right)$ and noting that $v = -D$, $u = -u$ and $f = f$, we get

$$\frac{1}{-D} - \frac{1}{-u} = \frac{1}{f}$$

$$\text{or } -\frac{1}{D} + \frac{1}{u} = \frac{1}{f}$$

Multiplying both the sides by D , and rearranging term, you can write

$$\frac{D}{u} = 1 + \frac{D}{f} \quad (23.4)$$

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On combining Eqns. (23.3) and (23.4), we get

$$\frac{A'B'}{AB} = 1 + \frac{D}{f}$$

or $M = 1 + \frac{D}{f}$ (23.5)

From this result we note that lesser the focal length of the convex lens, greater is the value of the angular magnification or magnifying power of the simple microscope.

Normal Adjustment : In this case, the image is formed at infinity. The magnifying power of the microscope is defined as the ratio of the angle subtended by the image at the eye to the angle subtended by the object at the unaided eye when the object is placed at D . Fig 23.2(a) shows that the object is placed at the least distance of distinct vision D .

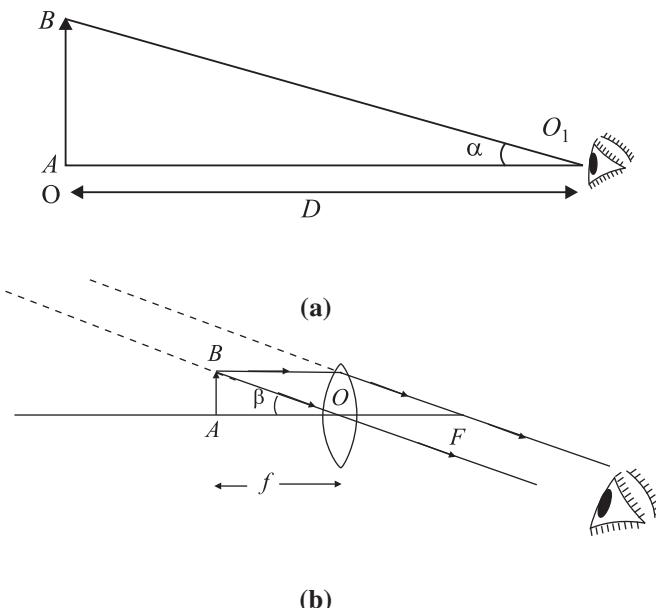


Fig.23.2 : Image formation for normal adjustment

The angles subtended by the object and the image at the unaided eye are α and β , respectively. The magnifying power is defined as

$$M = \frac{\beta}{\alpha}$$

In practice, the angles α and β are small, and, as before, replacing these by their tangents, we get

$$M = \frac{\tan \beta}{\tan \alpha}$$



Notes

i.e.

$$= \frac{AB}{AO} / \frac{AB}{AO_l}$$

$$= \frac{AO_l}{AO} = \frac{D}{f}$$

or

$$M = \frac{D}{f}$$

(23.6)

You may note that in the normal adjustment, the viewing of the image is more comfortable. To help you fix your ideas, we now give a solved example. Read it carefully.

Example 23.1: Calculate the magnifying power of a simple microscope having a focal length of 2.5 cm.

Solution : For a simple microscope, the magnifying power is given by [Eqn. (23.5)] :

$$M = 1 + \frac{D}{f}$$

Putting $D = 25$ cm and $f = 2.5$ cm, we get

$$M = 1 + \frac{25}{2.5} = 1 + 10 = 11$$

23.1.2 A Compound Microscope

A compound microscope consists of two convex lenses. A lens of short aperture and short focal length faces the object and is called the **objective**. Another lens of short focal length but large aperture facing the eye is called the **eye piece**. The objective and eye piece are placed coaxially at the two ends of a tube.

When the object is placed between F and $2F$ of the objective, its a real, inverted and magnified image is formed beyond $2F$ on the other side of the objective. This image acts as an object for the eye lens, which then acts as a simple microscope. The eye lens is so adjusted that the image lies between its focus and the optical center so as to form a magnified image at the least distance of distinct vision from the eye lens.

Magnifying Power of a compound microscope

Magnifying power of a compound microscope is defined as the ratio of the angle subtended by the final image at the eye to the angle subtended by the object at unaided eye, when both are placed at the least distance of distinct vision. It is denoted by M . By referring to Fig. 23.3, we can write

$$M = \frac{\beta}{\alpha}$$

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Since the angles α and β are small, these can be replaced by their tangents, so that

$$M = \frac{\tan \beta}{\tan \alpha}$$

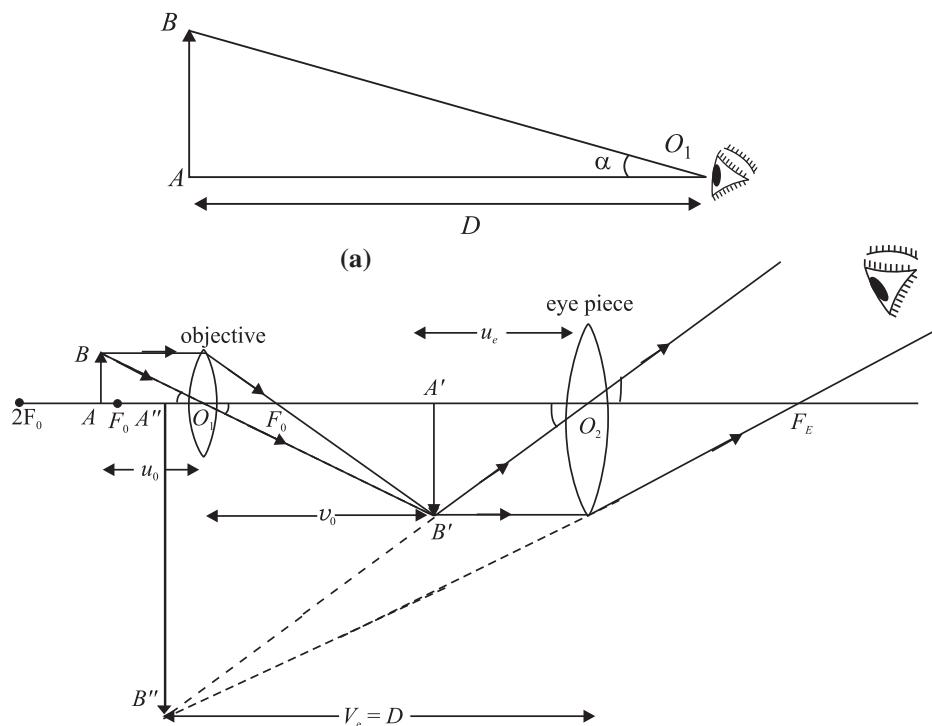


Fig.23.3 : Image formation by a compound microscope when the final image is formed at the least distance of distinct vision.

$$M = \frac{A'B'}{D} / \frac{AB}{D}$$

$$\Rightarrow M = \frac{A''B''}{AB} = \frac{A''B''}{A'B'} \cdot \frac{A'B'}{AB}$$

From similar $\Delta s A''B''O_2$ and $A'B'O_2$, we can write

$$\frac{A''B''}{A'B'} = \frac{A''O_2}{A'O_2} = \frac{D}{u_e}$$

Also from similar $\Delta s A'B'O_1$ and ABO , we have

$$\frac{A'B'}{AB} = \frac{v_o}{u_o}$$

Note that $m_e = \frac{A''B''}{A'B'}$ defines magnification produced by eye lens and $m_o =$

$\frac{A'B'}{AB}$ denotes magnification produced by the objective lens. Hence



Notes

From Lesson 20, you may recall the lens formula. For eye lens, we can write

$$\frac{1}{v_e} - \frac{1}{u_e} = \frac{1}{f_e}$$

Multiply on both sides by v_e to get

$$\frac{v_e}{v_e} - \frac{v_e}{u_e} = \frac{v_e}{f_e}$$

$$\Rightarrow \frac{v_e}{u_e} = 1 - \frac{v_e}{f_e}$$

Since f_e is positive and $v_e = -D$ as per sign convention, we can write

$$m_e = \frac{v_e}{u_e} = 1 + \frac{D}{f_e} \quad (23.8)$$

On combining Eqns. (23.7) and (23.8), we get

$$M = \frac{v_o}{u_o} \times \left(1 + \frac{D}{f_e} \right)$$

In practice, the focal length of an objective of a microscope is very small and object AB is placed just outside the focus of objective. That is

$$\therefore u_o \approx f_0$$

Since the focal length of the eye lens is also small, the distance of the image $A'B'$ from the object lens is nearly equal to the length of the microscope tube i.e.

$$v_o \approx L$$

Hence, the relation for the magnifying power in terms of parameters related to the microscope may be written as

$$M = \frac{L}{f_0} \left(1 + \frac{D}{f_e} \right) \quad (23.10)$$

Magnifying power in normal adjustment : In this case the image is formed at infinity. As discussed earlier, the magnifying power of the compound microscope may be written as

$$\begin{aligned} M &= m_o \times m_e \\ &= \frac{v_o}{u_o} \left(\frac{D}{f_e} \right) \end{aligned}$$

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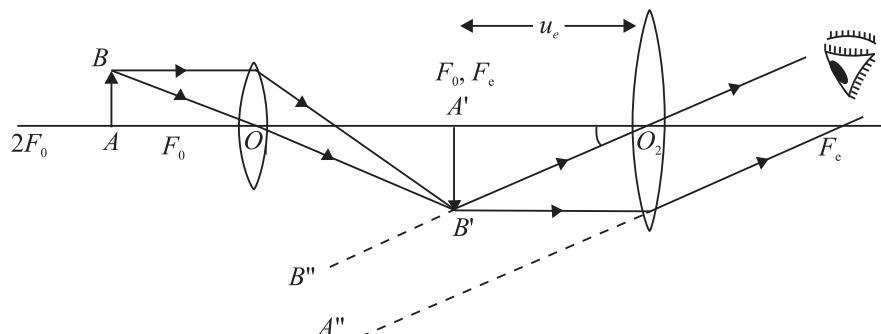


Fig. 23.4 : Compound microscope in normal adjustment

You may now like to go through a numerical example.

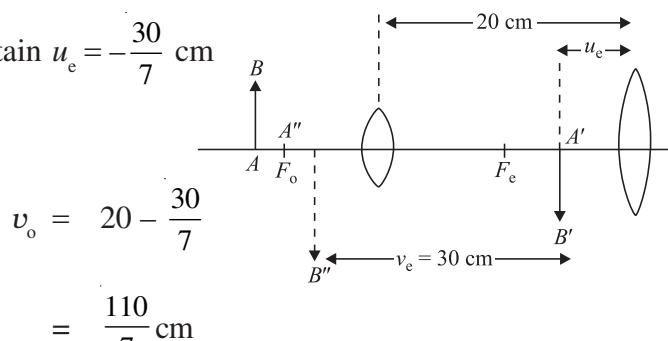
Example 23.2 : A microscope has an objective of focal length 2 cm, an eye piece of focal length 5 cm and the distance between the centers of two lens is 20 cm. If the image is formed 30 cm away from the eye piece, find the magnification of the microscope.

Solution : For the objective, $f_o = 2$ cm and $f_e = 5$ cm. For the eyepiece, $v_e = -30$ cm and $f_e = 5$ cm. We can calculate v_o using the relation

$$\frac{1}{v_o} - \frac{1}{u_o} = \frac{1}{f_o}$$

On solving, you will easily obtain $u_o = -\frac{30}{7}$ cm

For the objective lens



$$v_o = 20 - \frac{30}{7} = \frac{110}{7} \text{ cm}$$

Using the formula

$$\frac{1}{v_o} - \frac{1}{u_o} = \frac{1}{f_o}$$

we have

$$\frac{1}{110/7} - \frac{1}{u_o} = \frac{1}{2}$$

or

$$u_o = -\frac{110}{48} \text{ cm}$$



Notes

The magnifying power of the objective

$$m_o = \frac{v_o}{u_o} = \frac{110/7}{-110/48} = -\frac{48}{7}$$

The magnification due to the eyepiece is

$$m_e = \frac{v_e}{u_e} = \frac{-30/1}{-30/7} = 7$$

Therefore, the magnification of the microscope is given by

$$\begin{aligned} M &= (m_o) (m_e) \\ &= \left(-\frac{48}{7}\right) (7) = -48 \end{aligned}$$



INTEXT QUESTIONS 23.1

- What is the nature of images formed by a (i) simple microscope (ii) Compound microscope?
- Differentiate between the magnifying power and magnification?
- The magnifying power of a simple microscope is 11. What is its focal length?
- Suppose you have two lenses of focal lengths 100 cm and 4 cm respectively. Which one would you choose as the eyepiece of your compound microscope and why?
- Why should both the objective and the eyepiece of a compound microscope have short focal lengths?

23.2 TELESCOPES

Telescopes are used to see distant objects such as celestial and terrestrial bodies. Some of these objects may not be visible to the unaided eye. The visual angle subtended by the distant objects at the eye is so small that the object cannot be perceived. The use of a telescope increases the visual angle and brings the image nearer to the eye. Mainly two types of telescopes are in common use : refracting telescope and reflecting telescope. We now discuss these.

23.2.1 Refracting Telescope

The refracting telescopes are also of two types :

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Notes

Optical Instruments

- **Astronomical telescopes** are used to observe heavenly or astronomical bodies.
- **Terrestrial telescopes** are used to see distant objects on the earth. So it is necessary to see an erect image. Even Galilean telescope is used to see objects distinctly on the surface of earth.

An astronomical telescope produces a virtual and erect image. As heavenly bodies are round, the inverted image does not affect the observation. This telescope consists of a two lens system. The lens facing the object has a large aperture and large focal length (f_o). It is called the *objective*. The other lens, which is towards the eye, is called the *eye lens*. It has a small aperture and short focal length (f_e). The objective and eye-piece are mounted coaxially in two metallic tubes.

The objective forms a real and inverted image of the distant object in its focal plane. The position of the lens is so adjusted that the final image is formed at infinity. (This adjustment is called normal adjustment.) The position of the eyepiece can also be adjusted so that the final image is formed at the least distance of distinct vision.

(a) **When the final image is formed at infinity** (Normal adjustment), the paraxial rays coming from a heavenly object are parallel to each other and they make an angle α with the principal axis. These rays after passing through the objective, form a real and inverted image in the focal plane of objective. In this case, the position of the eyepiece is so adjusted that the final image is formed at infinity.

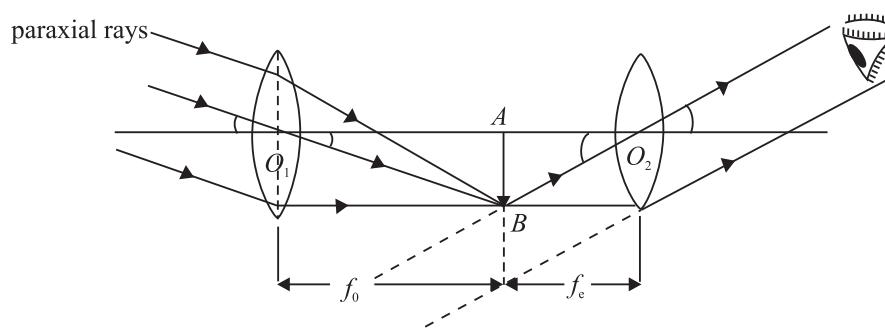


Fig 23.6 : Working principle of an astronomical telescope

Magnifying power of a telescope is defined as the ratio of the angle subtended by the image at the eye as seen through the telescope to the angle subtended by the object at objective when both the object and the image lie at infinity. It is also called **angular magnification** and is denoted by M . By definition,

$$M = \frac{\beta}{\alpha}$$



Notes

$$\begin{aligned}
 M &= \frac{\tan \beta}{\tan \alpha} \\
 &= \frac{AB/AO_2}{AB/AO_1} = \frac{AO_1}{AO_2} \\
 &= \frac{f_o}{f_e}
 \end{aligned} \tag{23.11}$$

It follows that the magnifying power of a telescope in normal adjustment will be large if the objective is of large focal length and the eyepiece is of short focal length. The length of telescope in normal adjustment is $(f_o + f_e)$

(b) When the final image is formed at the least distance of distinct vision, the paraxial rays coming from a heavenly object make an angle α with the principal axis. After passing through the objective, they meet on the other side of it and form a real and inverted image AB . The position of the eyepiece is so adjusted that it finally forms the image at the least distance of distinct vision.

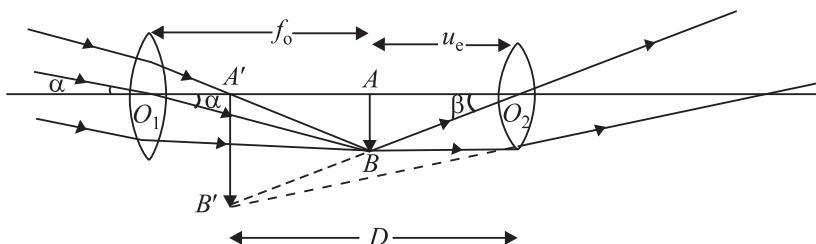


Fig 23.7 : Image formed by a telescope at D

Magnifying power: It is defined as the ratio of the angle subtended at the eye by the image formed at D to the angle subtended by the object lying at infinity:

$$\begin{aligned}
 M &= \frac{\beta}{\alpha} \\
 &\approx \frac{\tan \beta}{\tan \alpha} \\
 &= \frac{AB/AO_2}{AB/AO_1} = \frac{AO_1}{AO_2} \\
 &= \frac{f_o}{u_e}
 \end{aligned} \tag{23.12}$$

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Since $\frac{1}{v_e} - \frac{1}{u_e} = \frac{1}{f_e}$ for the eyepiece, we can write

$$\frac{1}{u_e} = \frac{1}{v_e} - \frac{1}{f_e}$$

$$= -\frac{1}{f_e} \left(1 - \frac{f_e}{v_e} \right)$$

or

$$M = \frac{f_o}{u_e} = -\frac{f_o}{f_e} \left(1 - \frac{f_e}{v_e} \right) \quad (23.13)$$

Applying the new cartesian sign convention $f_o = +f_o$, $v_e = -D$, $f_e = +f_e$, we can write

$$M = -\frac{f_o}{f_e} \left(1 + \frac{f_e}{D} \right) \quad (23.14)$$

The negative sign of magnifying power of the telescope suggests that the final image is inverted and real. The above expression tells that the magnifying power of a telescope is larger when adjusted at the least distance of distinct vision to the telescope when focused for normal adjustment.

Example 23.3: The focal length of the objective of an astronomical telescope is 75 cm and that of the eyepiece is 5 cm. If the final image is formed at the least distance of distinct vision from the eye, calculate the magnifying power of the telescope.

Solution:

Here $f_o = 75$ cm, $f_e = 5$ cm, $D = 25$ cm

$$M = -\frac{f_o}{f_e} \left(1 + \frac{f_e}{D} \right) = \frac{-75}{5} \left(1 + \frac{5}{25} \right) = -18$$

23.2.2 Reflecting telescope

A reflecting telescope is used to see distant stars and possesses large light-gathering power in order to obtain a bright image of even a faint star deep in space. The objective is made of a concave mirror, having large aperture and large focal length. This concave mirror, being parabolic in shape, is free from spherical aberration.

Before the reflected rays of light meet to form a real, inverted and diminished image of a distant star at the focal plane of concave mirror, they are intercepted and reflected by a plane Mirror $M_1 M_2$ inclined at an angle of 45° to the principal



Notes

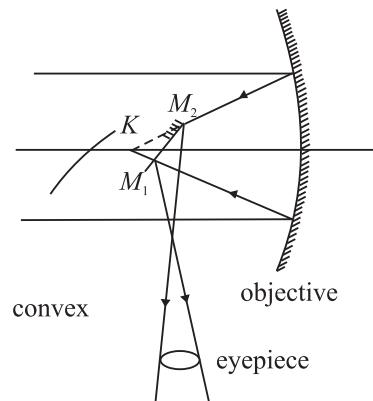


Fig 23.8 : Newtonian Reflector

axis of the concave mirror. This plane mirror deviates the rays and the real image is formed in front of the eye piece, which is at right angle to the principal axis of concave mirror. The function of the eye-piece is to form a magnified, virtual image of the star enabling eye to see it distinctly.

If f_o is the focal length of the concave mirror and f_e is the focal length of eye piece, the magnifying power of the reflecting telescope is given by

$$M = \frac{f_o}{f_e}$$

Further, if D is the diameter of the objective and d is the diameter of the pupil of the eye, the brightness ratio is given by

$$B = D^2/d^2$$

The other form of the reflecting telescope is shown in Fig 23.9. It was designed by **Cassegrain**. In this case the objective has a small opening at its center. The rays from the distant star, after striking the concave mirror, are made to intercept at A_2 and the final image is viewed through the eyepiece.

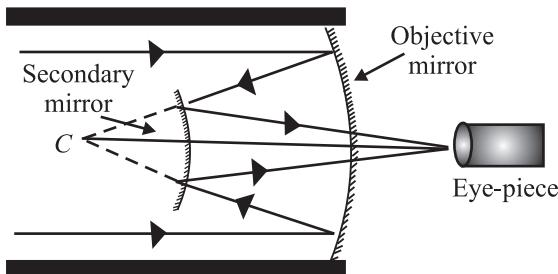


Fig 23.9 : Cassegrain reflector

There are several advantages of a reflecting telescope over a refracting telescope.

- Since the objective is not a lens, the reflecting telescopes are free from chromatic aberration. Thus rays of different colours reaching the objective from distant stars are focussed at the same point.
- Since the spherical mirrors are parabolic mirrors, free from spherical aberration, they produce a very sharp and distinct image.
- Even a very faint star can be seen through the reflecting telescope because they have large aperture and have large light-gathering power. The brightness of the image is directly proportional to the area of the objective :

$$B \propto \frac{\pi D^2}{4}$$

where D is the diameter of the objective of the telescope. If d is the diameter of the pupil of the eye then brightness of the telescope B is defined as the ratio

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of light gathered by the telescope to that gathered by the unaided eye from the distant object

$$B = \frac{\pi D^2 / 4}{\pi d^2 / 4} = \frac{D^2}{d^2}$$

- In reflecting type of telescopes, there is negligible absorption of light.
- Large apertures of reflecting telescope enable us to see minute details of distant stars and explore deeper into space. That is why in recent years, astronomers have discovered new stars and stellar systems. You should look out for such details in science magazines and news dailies.



INTEXT QUESTIONS 23.2

1. How would the magnification of a telescope be affected by increasing the focal length of:
 - (a) the objective _____

 - (b) the eye piece _____

2. If the focal length of the objective of a telescope is 50 cm and that of the eyepiece is 2 cm. What is the magnification?
3. State one difference between the refracting and reflecting telescope.
4. What is normal adjustment?
5. If the telescope is inverted, will it serve as a microscope?

23.3 RESOLVING POWER : THE RAYLEIGH'S CRITERION

In earlier lessons, you have seen that the image of a point source is not a point, but has a definite size and is surrounded by a diffraction pattern. Similarly, if there are two point sources very close to each other, the two diffraction patterns formed by the two sources may overlap and hence it may be difficult to distinguish them as separate by the unaided eye. The resolving power of an optical instrument is its ability to resolve (or separate) the images of two point objects lying close to each other. Rayleigh suggested that two images can be seen as distinct when the first minimum of the diffraction pattern due to one object falls on the central maximum of the other. This is called *Rayleigh's criterion*.



Notes

If we assume that the pupil of our eye is about 2 mm in diameter, two points can be seen distinctly separate if they subtend an angle equal to about one minute of arc at the eye. **The reciprocal of this angle is known as the resolving power of the eye.**

Now let us calculate the resolving power of common optical instruments. We begin our discussion with a telescope.

23.3.1 Resolving Power of a Telescope

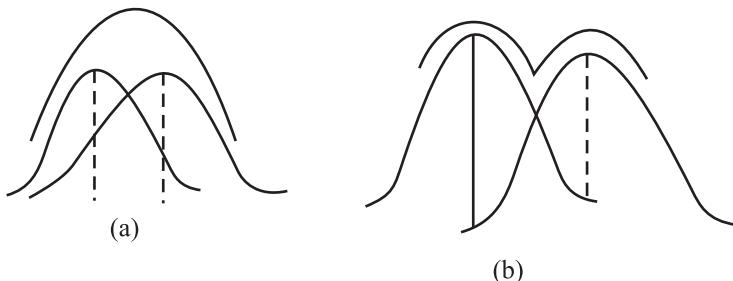


Fig. 23.10 : Rayleigh's criterion for resolution **a)** when the angular separation is less than θ , the two points are seen as one, and **b)** when the angular separation is more than θ , the two points are distinctly visible.

The resolving power of a telescope is its ability to form separate images of two distant point objects situated close to each other. It is measured in terms of the angle subtended at its objective by two close but distinct objects whose images are just seen in the telescope as separate. This angle is called the **limit of resolution** of the telescope. If the angle subtended by two distinct objects is less than this angle, the images of the objects can not be resolved by the telescope. The smaller the value of this angle, higher will be the resolving power of the telescope. Thus, the reciprocal of the limit of resolution gives the resolving power of the telescope.

If λ is the wavelength of light, D the diameter of the telescope objective, and θ the angle subtended by the point object at the objective, the limit of resolution of the telescope is given by (Rayleigh's criterion)

$$\theta = \frac{1.22\lambda}{D}$$

Hence, the resolving power of the telescope.

$$(R.P)_T = \frac{1}{\theta} = \frac{D}{1.22\lambda} \quad (23.15)$$

From Eqn. (23.15) it is clear that to get a high resolving power, a telescope with large aperture objective or light of lower wavelength has to be used.

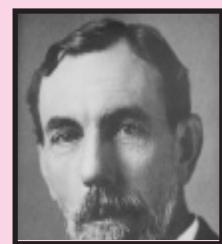
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Lord Rayleigh (1842 – 1919)

Born to the second Baron Rayleigh of Terling place, Witham in the country of Essex, England, John strutt had a very poor health in his childhood. Due to this he had a disrupted schooling. But he had the good luck of having Edward Rath and Stokes as his teachers. As a result, he passed his tripos examination in 1865 as senior Wrangler and become the first recipient of Smiths prize.

In addition to the discovery of Argon, for which he was awarded Nobel prize (1904), Rayleigh did extensive work in the fields of hydrodynamics, thermodynamics, optics and mathematics. His travelling wave theory, which suggested that elastic waves can be guided by a surface, paved way for researches in seismology and electronic signal processing. During the later years of his life, he also showed interest in psychiatry research. Lunar feature-crater Rayleigh and planetary feature crater Rayleigh on Mars are a tribute to his contributions.

Example 23.4: A telescope of aperture 3 cm is focussed on a window at 80 metre distance fitted with a wiremesh of spacing 2 mm. Will the telescope be able to observe the wire mesh? Mean wavelength of light $\lambda = 5.5 \times 10^{-7}$ m.

Solution: Given $\lambda = 5.5 \times 10^{-7}$ m and $D = 3$ cm = 3×10^{-2} m

Therefore, the limit of resolution

$$\theta = \frac{1.22\lambda}{D} = \frac{1.22 \times 5.5 \times 10^{-7} \text{ m}}{3 \times 10^{-2} \text{ m}} = 2.236 \times 10^{-5} \text{ rad}$$

The telescope will be able to resolve the wiremesh, if the angle subtended by it on the objective is equal to or greater than θ , the limit of resolution. The angle subtended by the wiremesh on the objective

$$\alpha = \frac{\text{spacing of wiremesh}}{\text{distance of the objective from the wiremesh}}$$

$$= \frac{2 \text{ mm}}{80 \text{ m}} = \frac{2 \times 10^{-3}}{80 \text{ m}} = 2.5 \times 10^{-5} \text{ rad.}$$

As the angle 2.5×10^{-5} radian exceeds the limit of a resolution ($= 2.236 \times 10^{-5}$ radian), the telescope will be able to observe the wire mesh.



Notes

23.3.2 Resolving Power of a Microscope

The resolving power of a microscope represents its ability to form separate images of two objects situated very close to each other. The resolving power of a microscope is measured in terms of the smallest linear separation between the two objects which can just be seen through the microscope as separate. This smallest linear separation between two objects is called **the limit of resolution of the microscope**.

The smaller the value of linear separation, the higher will be the resolving power of the microscope. Thus, **the reciprocal of the limit of resolution gives the resolving power of the microscope**.

If λ is the wavelength of light used to illuminate the object, θ is the half angle of the cone of light from the point object at the eye and n is the refractive index of the medium between the object and the objective, the limit of resolution of the microscope is given by

$$d = \frac{\lambda}{2n \sin \theta} \quad (23.16)$$

Thus the resolving power of microscope will be

$$(R.P)_m = \frac{2n \sin \theta}{\lambda}$$

(23.17) The expression $2n \sin \theta$ is called numerical aperture (N.A). The highest value of N.A of the objective obtainable in practice is 1.6, and for the eye, N.A is 0.004.

It is clear from Eqn. (23.17) that the resolving power of a microscope can be increased by increasing the numerical aperture and decreasing the wavelength of the light used to illuminate the object. That is why ultraviolet microscopes and an electron microscope have a very high resolving power.

Applications in Astronomy

The astronomical (or optical) telescope can be used for observing stars, planets and other astronomical objects. For better resolving power, the optical telescopes are made of objectives having a large aperture (objective diameter). However, such big lenses are difficult to be made and support. Therefore, most astronomical telescopes use reflecting mirrors instead of lenses. These can be easily supported as a mirror weighs less as compared to a lens of equivalent optical quality.

The astronomical telescopes, which are ground-based, suffer from blurring of images. Also, ultraviolet, x-ray, gamma-ray etc. are absorbed by the earth's

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surface. They cannot be studied by ground-based telescopes. In order to study these rays coming from astronomical objects, telescopes are mounted in satellites above the Earth's atmosphere. NASA's Hubble space telescope is an example of such telescope. Chandra X-ray observation, Compton x-ray observation and Infrared telescopes have recently been set up in space.



INTEXT QUESTIONS 23.3

1. How can the resolving power of a telescope be improved?
2. What is the relationship between the limit of resolution and the resolving power of the eye?
3. If the wavelength of the light used to illuminate the object is increased, what will be the effect on the limit of resolution of the microscope?
4. If in a telescope objective is made of larger diameter and light of shorter wavelength is used, how would the resolving power change?



WHAT YOU HAVE LEARNT

- The angle subtended by an object at the human eye is called as the visual angle.
- The angular magnification or magnifying power of a microscope is the ratio of the angle subtended by the image at the eye to the angle subtended by the object when both are placed at the near point.
- Linear magnification is defined as the ratio of the size of the image to the size of the object.
- The magnifying power of a simple microscope is $M = 1 + \frac{D}{f}$, where D is least distance of distinct vision and f is focal length of the lens.
- In a compound microscope, unlike the simple microscope, magnification takes place at two stages. There is an eye piece and an objective both having short focal lengths. But the focal length of the objective is comparatively shorter than that of the eye piece.
- The magnifying power of a compound microscope is given as

$$M = m_0 \times m_e$$

But $m_e = 1 + \frac{D}{f}$. Therefore



Notes

$$M = \frac{v_o}{u_o} \left(1 + \frac{D}{f_e} \right)$$

where v_o is distance between the image and the objective, u_o is object distance from the objective, D is the least distance of distinct vision ($= 25\text{cm}$) and f_e is focal length of the eye-piece.

- Telescope is used to see the distant objects which subtend very small visual angle at the eye. The use of a telescope increases the visual angle at the eye. The far-off object appears to be closer to the eye and can be seen easily.
- Two types of telescopes are used (i) Refracting (ii) Reflecting.
- The objective of the refracting telescope is a converging lens. But the objective in a reflecting telescope is a spherical mirror of large focal length. There are several advantages of reflecting telescope over a refracting telescope.

The magnifying power of a telescope is

$$M = f_o/f_e$$

where f_o is focal length of the objective and f_e is focal length of the eyepiece.



TERMINAL EXERCISE

1. What is the difference between simple and compound microscopes? Derive an expression for the magnification of a compound microscope.
2. Distinguish between the refracting and reflecting telescope. Draw a ray diagram for the Newton's telescope.
3. Derive an equation for the magnifying power of a refracting telescope.
4. What do you mean by the least distance of distinct vision? What is its value for a normal eye?
5. Can we photograph the image formed by a compound microscope? Explain your answer.
6. Define the resolving power of an optical instrument. What is the value of limit of resolution for a normal eye?
7. What are the main differences in the design of a compound microscope and a terrestrial telescope?
8. The eyepiece of a telescope has a focal length of 10 cm. The distance between the objective and eye piece is 2.1 m. What is the angular magnification of the telescope?

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9. The image formed by a microscope objective of focal length 4 mm is 18 cm from its second focal point. The eyepiece has a focal length of 3.125 cm. What is the magnification of the microscope?
10. The objective of a telescope has a diameter three times that of a second telescope. How much more amount of light is gathered by the first telescope as compared to the second?



ANSWERS TO INTEXT QUESTIONS

23.1

1. Image formed by a simple microscope is virtual erect and magnified. whereas the image formed by a compound microscope is real, inverted and magnified.
2. Magnifying power is the ratio of the angle subtended by the image at eye piece to the angle subtended by the object placed at the near point. Magnification is the ratio of the size of image to the size of object.
3. $M = 11, m = 1 + \frac{D}{f}$. Putting $D = 25$ cm, we get $f = 2.5$ cm
4. If you choose the lens with 4 cm focal length, the magnifying power will be high because $m = \frac{f_o}{f_e}$
5. The magnifying power of a compound microscope is given by $M = \frac{-L}{f_o} \left(1 + \frac{D}{f_e} \right)$
Obviously, M will have a large value, if both f_o and f_e are small.

23.2

1. (a) Objective of large focal length increases the magnifying power of the telescope.
(b) Magnification is reduced by increasing the focal length of eyepiece.
2. Magnification $m = \frac{f_o}{f_e} = \frac{50 \text{ cm}}{2 \text{ cm}} = 25$
3. The objectives of a telescope is a spherical mirror of large focal length instead of converging lens as in a refracting telescope.
4. A telescope is said to be in normal adjustment, if the final image is formed at infinity.
5. No

23.3

1. By taking a large aperture or by using a light of lower wavelength.
2. The limit of resolution of an eye is inversely proportional to its resolving power. Limit of resolution will also be increased.
3. Since resolving power of telescope is given by $R.P = \frac{D}{1.22\lambda}$, it would increase.

**Notes****Answers To Problems in Terminal Exercise**

8. 21 9. 400
10. 9 times.

SENIOR SECONDARY COURSE

REFLECTION AND REFRACTION OF LIGHT

STUDENT'S ASSIGNMENT – 6

Maximum Marks: 50

Time : 1½ Hours

INSTRUCTIONS

- Answer All the questions on a separate sheet of paper
- Give the following information on your answer sheet:
 - Name
 - Enrolment Number
 - Subject
 - Assignment Number
 - Address
- Get your assignment checked by the subject teacher at your study centre so that you get positive feedback about your performance.

Do not send your assignment to NIOS

1. Radius of curvature of a concave mirror is 30 cm . What is its focal length? (1)
2. Refractive index of glass is 5.5. What is the speed of light in glass? (1)
3. Find in dioptres the power of a concave lens of focal length 25 cm . (1)
4. Out of amplitude, wavelength, speed and frequency which characteristic of wave remains unaffected in passing from one medium to another. (1)
5. In passing through a prism which suffers greater angular deviation red light or blue light? (1)
6. When an object is placed at a distance R in front of a lens its inverted image of the same size is formed on the other side of the lens. Is the lens convex or concave. (1)
7. The focal length of the objective of a telescope is 10 cm and that of its eye piece is f_e . What is the length of the telescope tube in its normal adjustment? (1)
8. Name the optical phenomenon responsible for the blue colour of the sky? Does the sky appear blue on moon also. (1)
9. Draw ray diagrams to show, how a totally reflecting prism may turn a ray of light through (i) 90° (ii) 180° . (2)
10. Focal length of an equiconvex lens, made of a material of refractive index 1.5 is 20 cm . Calculate the radius of curvature of its surfaces. (2)

11. When a ray of light is incident air glass interface making an angle 30° from the surface of separation the reflected light is found to be 100% plane polarized. Find the refractive index of glass with respect to air. (2)
12. Draw a labelled ray diagram showing image formation by a compound microscope. (2)
13. A cylindrical container is 20 cm deep. Water ($v = \frac{4}{3}$) is filled in the container upto the brim. By how many centimeters will its bottom appear as raised up. (4)
14. Distinguish clearly between interference and diffraction of light. (4)
15. Explain the chromotic aberration in lenses? Describe a method of minimizing this defect. (4)
16. What is meant by the resoluing power of an optical instrument? Drawing suitable diagrams explain Rayhigh's criterian for the limit of resolution of two very close spectral lines. Write formula for the resolving power of (i) microscope (ii) telescope.
17. For an equiangular prism show that refractive index $\left(n = \frac{\sin(A + \delta)/2}{\sin A/2} \right)$ where δ is the angle of minimum deviation. (4)
18. State huygens principle for the propagation of wave front. Draw (i) incident and reflected wave front for light incident on a concave minor from a very distant source. (ii) incident and refracted wave fronts for light incident on a convex lens from a point object situated at $2f$. (5)
19. Draw a neat diagram showing image formation of a point object by a convex lens. Derive lens maker's formula. (5)
20. Describe young's double slit experiment. Derive expression for fringewidth of interference pattern formed. (5)

MODULE - VII

ATOMS AND NUCLEI

- 24 Structure of Atom
- 25 Dual Nature of Radiation and Matter
- 26 Nuclei and Radioactivity
- 27 Nuclear Fission and Fusion

24

 Notes

STRUCTURE OF ATOM

So far you have studied about mechanical, thermal, electrical and magnetic properties of matter. Have you ever thought as to why do different materials have different properties? That is, why does chalk break so easily but a piece of aluminium flattens on impact? Why do some metals start conducting current when light falls on them? And so on. To understand such properties of materials, we recall that atoms are building blocks of all forms of matter. That is, despite its appearance being continuous, matter has definite structure on microscopic level which is beyond the reach of our sense of seeing. This suggests that to discover answers to above said questions, you need to know the structure of the atom.

Our understanding of the structure of atom has evolved over a period of time. In this lesson, we have discussed different atomic models. Starting with Rutherford's model based on his classic scattering experiment, in this lesson we have discussed Bohr's model of atom that explains the electronic structure. Bohr's theory also helps us to explain the atomic spectrum of hydrogen atom.



OBJECTIVES

After studying this lesson, you should be able to:

- describe Rutherford's scattering experiment and its findings;
- explain Rutherford's atomic model and state its shortcomings.
- calculate the radius of Bohr's first orbit and velocity of an electron in it;
- derive an expression for the energy of an electron in a hydrogen atom; and
- draw the energy level diagram of a hydrogen atom and explain its spectrum.
- describe about the production, properties, types and uses of x-rays; and
- define Mosley law and Duane-Hunt law.



Notes

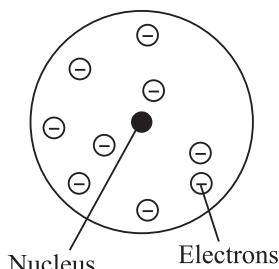


Fig. 24.1 : Plum-pudding model of atom

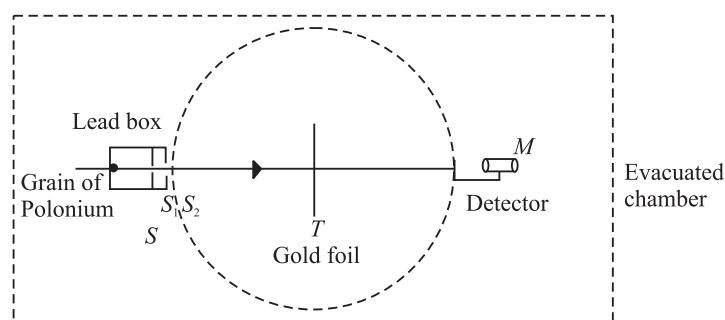
The Concept of Atom

The concept of atom is as old as human civilization. In ancient times Democritus in Greece and Kanad in India tried to explain the changes around us in terms of particles. But the exact theory of atom was presented by John Dalton, an English Chemist in 1808. He described atom as the smallest, indivisible particle endowed with all the properties of element and takes part in chemical reactions. Dalton's atom was an ultimate particle having no structure. This idea was accepted by the scientists in the nineteenth century as they knew nothing about the structure of atoms. The discovery of electrons by J.J. Thomson in 1897, while studying discharge of electricity through gases at low pressures, provided the first insight that atoms have a structure and negatively charged electrons are constituents of all atoms. Since the atom as a whole is neutral, it must also have equal amount of positive charge. Moreover, since electrons were thousands of times lighter than the atom, it was thought that the positively charged constituent of atoms carried the entire mass. On the basis of his experiments, Thomson suggested the *plum pudding model* of atom (Fig. 24.1). According to this, an atom is a tiny, uniformly charged positive ball in which negatively charged electrons are suitably placed to make it neutral. It seemed perfectly reasonable then.

Our understanding of the structure of atom since the times of Thomson has improved considerably. Due to the pioneering works of Lord Rutherford, Niels Bohr, James Chadwick, Pauli, Schrodinger and others. In fact, our concept of new world of sub-atomic particles came into existence and has led to the invention of epoch making new technologies, like micro-electronics and nanotechnology.

24.1 RUTHERFORD'S EXPERIMENT ON SCATTERING OF A-PARTICLES

On the advice of Lord Rutherford, two of his students Geiger and Marsden performed an experiment in which a beam of α -particles was bombarded on a thin gold foil. The experimental arrangement used by them is shown in Fig. 24.2.

Fig. 24.2 : Schematics of experimental set up used for α -particle scattering

A well collimated fine pencil of α -particles from a source S was made to fall on a thin gold foil (T). The scattered α -particles were received on a ZnS fluorescent screen, which produced a visible flash of light when struck by an α -particle (and acted as detector), backed by a low power microscope (M). The detector was capable of rotation on a circular scale with T at the centre. The whole apparatus was enclosed in an evacuated chamber to avoid collisions of α -particles with air molecules. It was expected that if Thomson model was correct, most of the particles would go straight through the foil, with only minor deviation from the original path.

Geiger and Marsden observed that most of the α -particles suffered only small deflections, as expected. But a few got deflected at large angles (90° or more). Some of them (1 in 8000) even got deflected at 180° . Fig. 24.3 presents the experimental results. The large angle scattering of α -particles could not be explained on the basis of Thomson model of atom.

To explain large angle scattering, Lord Rutherford suggested the nuclear model of atom. He argued that α -particles which pass at a large distance from the nucleus experience negligible coulombian repulsive force and hence, pass almost undeflected. However, closer to the nucleus a α -particle comes, greater force of repulsion it experiences and hence gets deflected at a greater angle. A few α -particles which proceed for a head-on collision towards the nucleus are scattered back by 180° along its direction of approach, as indicated by α -particle 4 in Fig. 24.4.

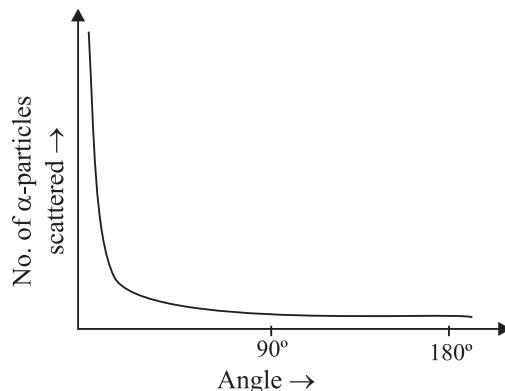


Fig. 24.3 : Experimental result of Rutherford's experiment

Notes

According to Thomson model, α -particles should experience weak force due to electrons. However, since α -particles are about 7000 times heavier than electrons and travelled at high speed, large angle scattering required strong repulsive force to be exerted.

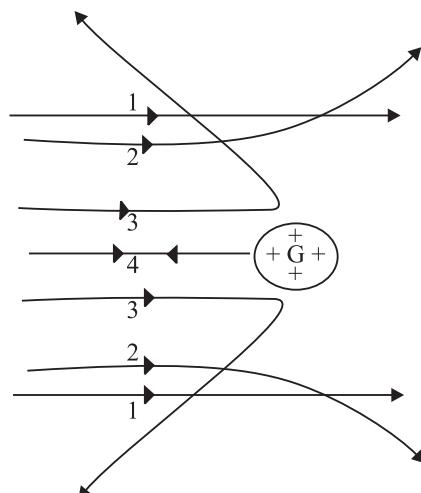


Fig. 24.4 : Paths traversed by α -particles scattered by a gold foil

MODULE - 7

Atoms and Nuclei

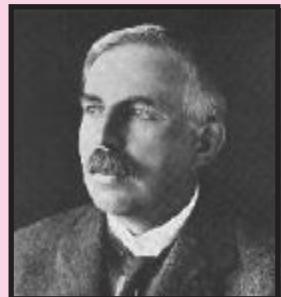


Notes

Structure of Atom

Lord Rutherford

(1871–1937)



Born in New Zealand, Rutherford studied under J.J. Thomson at the Cavendish Laboratory in England. His pioneering work on atom is a defining landmark. He developed Becquerel's discovery of radioactivity into an exact science and documented proof that the atoms of heavier elements, which had been thought to be immutable, actually disintegrate (decay) into various forms of radiation. In 1898, Rutherford discovered

that two quite separate types of emissions came from radioactive atoms and he named them alpha and beta rays. Beta rays were soon shown to be high speed electrons. In 1907, he showed that the alpha particle was a helium atom stripped of its electrons. He and his assistant, Hans Geiger, developed Rutherford-Geiger detector to electrically detecting particles emitted by radioactive atoms. With this he could determine important physical constants such as Avogadro's number, the number of atoms or molecules in one gram-mole of material.

In 1911, Rutherford proposed the nuclear model of atom; that almost entire mass of an atom is concentrated in a nucleus 10^{-5} times the atom itself and electrons revolve around it. This second great work won him the **Nobel Prize** in chemistry in 1908.

Eminent Indian physicist, educationist and philosopher Dr. D.S. Kothari was one of his students and worked on pressure ionisation in stars.

24.1.1 Nuclear Model of Atom

Rutherford argued that large angle scattering of α -particles can be explained only by stipulating the presence of a hard, positively charged core of atom. Thus he proposed a new model of atom with following characteristics:

- The entire charge and most of the mass of the atom is confined in a very small ($\sim 10^{-15}$ m) central region, called the *nucleus*.
- The negatively charged electrons revolve at a distance around it such that the atom as a whole is electrically neutral and stable.

The nuclear model of atom proposed by Rutherford faced some difficulties. Some of the consequences of Rutherford's model contradicted experimental observations.

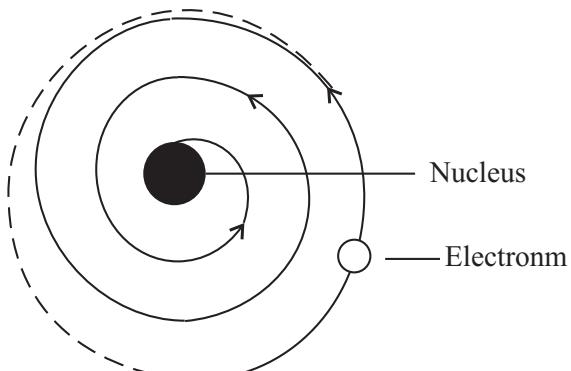


Fig. 24.5 : Motion of electrons in Rutherford's nuclear model of atom

(i) Stability of the atom : We know that electrons are negatively charged. These are attracted by the nucleus and get accelerated. An accelerated charged particle, according to classical wave theory, emits electromagnetic radiations. Hence, the revolving electrons should lose energy eventually and spiral into the nucleus (Fig. 24.5). This would have made the atom short-lived and contradicted the observed stability of matter.

(ii) Frequency of electromagnetic radiation : The electron spiralling towards the nucleus will emit electromagnetic radiations of all frequencies giving rise to a continuous spectrum. But experiments show that atoms emit radiations of certain well defined frequencies only (line spectra).

From the above discussion, you may be tempted to conclude that nuclear model of atom could not explain the experimental facts. Nevertheless, it contributed significantly to our understanding and was the first landmark in the right direction.



INTEXT QUESTIONS 24.1

1. Choose the correct answers :
 - (a) In Rutherford's scattering experiment, target was bombarded with (i) β -rays (ii) γ -rays (iii) α -rays.
 - (b) The nucleus is surrounded by :
(i) electrons (ii) protons (iii) α -particles
 - (c) The large angle scattering of α -particles indicated the presence of (i) some positively charged hard core inside the atom (ii) some porous core inside the atom (iii) negatively charged core.
2. Name two experimental observations that could not be explained by Rutherford's model.



Notes



Notes

Bohr quantised energy as well as angular momentum to explain hydrogen spectrum; Planck quantised only energy to explain black body radiation.

24.2 BOHR'S MODEL OF HYDROGEN ATOM

To overcome the difficulties of Rutherford model, in 1913, Neils Bohr proposed a model of atomic structure, based on quantum ideas proposed by Max Planck. Bohr's model not only described the structure of atom but also accounted for its stability. It proved highly successful in explaining the observed spectrum of the hydrogen atom. Let us learn about it now.

Bohr's Postulates

Bohr started with the planetary model of atom. However, to overcome the problems that plagued Rutherford model, Bohr made several assumptions. These are known as Bohr's postulates. There are four postulates.

- Electrons in an atom move in circular orbits around the nucleus with the centripetal force supplied by the Coulomb force of attraction between the electron and the nucleus. Mathematically, we can write*

$$\frac{mv^2}{r} = \frac{1}{4\pi\epsilon_0} \frac{Ze^2}{r^2} \quad (24.1)$$

where Z denotes the number of positive charges in the nucleus.

- Of the infinite number of possible circular orbits, only those orbits are allowed for which the value of orbital angular momentum of the electron is an integral multiple of $h/2\pi$:*

$$|\mathbf{L}| = mvr = \frac{nh}{2\pi} \quad (24.2)$$

where \mathbf{L} is the orbital angular momentum, equal to mvr for a circular orbit. Here h is Planck's constant and n is an integer.

- An electron moving in an allowed orbit does not radiate any energy. In these allowed orbits, the energy of the electron is constant. These orbits are called stationary states.*

Note that an electron can move in a stationary state but its energy is constant.

- Energy is emitted by an atom only when its electron “falls” from an allowed higher energy level E_f to another allowed lower level E_i . The change in energy is the energy of the emitted photon. Similarly, an electron only absorbs radiation when it “jumps” to a higher energy level from a lower energy level. The change in energy of an electron can be related to the frequency or wavelength of the emitted or absorbed photon:*

For emission

$$\Delta E = E_i - E_f = h\nu \quad (24.3a)$$

For absorption

$$\Delta E = E_f - E_i = h\nu \quad (24.3b)$$

where ν is the frequency of the emitted photon.

Niels Henrik David Bohr

(1885-1962)



Niels Bohr was born in Copenhagen, Denmark. He grew up in an atmosphere most favourable to the development of his genius. His father was an eminent physiologist and was largely responsible for awakening his interest in physics while he was still at school. In the spring of 1912, he worked in Rutherford's laboratory in Manchester. He studied the structure of atoms on the basis of Rutherford's nuclear model of atom. He succeeded in working out and presenting a picture of atomic structure that explained atomic spectra of hydrogen atom.

In 1916, he was appointed Professor of Theoretical Physics at Copenhagen University, and in 1920 (until his death in 1962), he became head of the Institute for Theoretical Physics, established for him at that university.

Recognition of his work on the structure of atom came with the award of the **Nobel Prize** in Physics in 1922.

Notes

Note that these postulates beautifully combine classical and quantum ideas. For example, the first postulate is in accordance with classical physics while other postulates use quantum physics.

24.2.1 Energy Levels

To calculate the energy of an electron in n th orbit of radius r_n , we rewrite Eqn. (24.1) as

$$\frac{mv_n^2}{r_n} = \frac{1}{4\pi\epsilon_0} \frac{Ze^2}{r_n^2}$$

where v_n is speed of the electron in its orbit.

On multiplying both sides of this equation by mr_n^3 we get

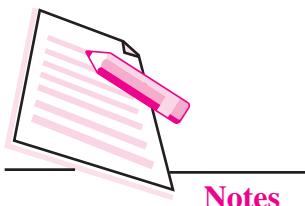
$$m^2 v_n^2 r_n^2 = \frac{1}{4\pi\epsilon_0} m Z e^2 r_n$$

On combining this result with Eqn. (24.2), we get

$$m^2 v_n^2 r_n^2 = n^2 \frac{h^2}{4\pi^2} = \frac{m}{4\pi\epsilon_0} Z e^2 r_n \quad (24.4)$$

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Atoms and Nuclei



Notes

Structure of Atom

On re-arranging terms, we get an expression for the radius of the n th orbit :

$$r_n = 4\pi\epsilon_0 \frac{n^2 h^2}{4\pi^2 m Z e^2}$$

$$= \frac{n^2 h^2 \epsilon_0}{Z e^2 m \pi} \quad n = 1, 2, 3, \dots \quad (24.5)$$

Note that radius of an orbit is directly proportional to second power of the number of orbit. It means that radius is more for higher orbits. Moreover, the relative values of the radii of permitted orbits are in the ratio $1^2, 2^2, 3^2, 4^2, \dots$, ie. in the ratio $1 : 4 : 9 : 16$ and so on. For hydrogen atom ($Z = 1$), the radius of its inner most orbit is called **Bohr radius**. It is denoted by a_0 and its magnitude is 5.3×10^{-11} m. In terms of a_0 , the radii of other orbits are given by the relation

$$r_n = n^2 a_0$$

It shows that the spacing between consecutive orbits increases progressively. On inserting the value of r_n from Eqn. (24.5) in Eqn. (24.2), we get an expression for the speed of the electron in the n th orbit :

$$v_n = \frac{nh}{2\pi m r_n} = \frac{nh}{2\pi m} \cdot \frac{Ze^2 m \pi}{n^2 h^2 \epsilon_0}$$

$$= \frac{1}{2} \frac{Ze^2}{\epsilon_0 nh} \quad (24.6)$$

From Lesson 16 you will recall that potential energy of a negative charge (electron in this case) in bringing it from infinity to a point at a distance r in a field of positive charge (nucleus in this case) is obtained by summing (integrating) the product of Coulomb force and distance :

$$U = -\frac{1}{4\pi\epsilon_0} \int_{r_n}^{\infty} \frac{Ze^2}{r^2} dr$$

$$= \frac{1}{4\pi\epsilon_0} \left[\frac{Ze^2}{r} \right]_{r_n}^{\infty}$$

$$= -\frac{1}{4\pi\epsilon_0} \frac{Ze^2}{r_n} \quad (24.7)$$

since potential energy of the electron at infinity will be zero.

It readily follows from Eqn. (24.1) that

$$\frac{1}{4\pi\epsilon_0} \frac{Ze^2}{r_n} = mv_n^2$$

Hence, potential energy of an electron in n^{th} orbit

$$U = -mv_n^2 \quad (24.8)$$

Since kinetic energy

$$K.E = \frac{1}{2}mv_n^2 \quad (24.9)$$

the total energy of the electron in nth orbit is given by

$$\begin{aligned} E &= K.E + U \\ &= \frac{1}{2}mv_n^2 - mv_n^2 \\ &= -\frac{1}{2}mv_n^2 \end{aligned}$$

Combining this result with Eqn. (24.6), we get

$$\begin{aligned} E &= -\frac{m}{2} \left(\frac{2\pi Ze^2}{4\pi\epsilon_0 nh} \right)^2 \\ &= -\frac{m}{8\epsilon_0^2} \frac{Z^2 e^4}{n^2 h^2} \end{aligned} \quad (24.10)$$

$$= \frac{RZ^2}{n^2}; n = 1, 2, 3\dots \quad (24.11)$$

where

$$R = \frac{me^4}{8\epsilon_0^2 h^2} \quad (24.12)$$

is called Rydberg constant. From Eqn. (24.11) we note that

- energy of an electron in various allowed orbits is inversely proportional to the square of the number of orbit; and
- the energy in an orbit is negative, which implies that the electron is bound to the nucleus.

Putting the standard values of $m = 9.11 \times 10^{-31} \text{ kg}$, $e = 1.6 \times 10^{-19} \text{ C}$, $\epsilon_0 = 0.85 \times 10^{-11} \text{ C}^2 \text{N}^{-1} \text{m}^{-2}$, and $h = 6.62 \times 10^{-34} \text{ Js}$ in Eqn. (24.12), we obtain $R = 2.17 \times 10^{-18} \text{ J} = 13.6 \text{ eV}$, since $1 \text{ eV} = 1.6 \times 10^{-19} \text{ J}$. On using this result in Eqn. (24.11), we find that the energy of an electron in n^{th} orbit of hydrogen atom (in eV) is given by

$$E_n = -\frac{13.6}{n^2} \quad (24.13)$$



Notes

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Atoms and Nuclei



Notes

Structure of Atom

Thus every orbit can be specified with a definite energy; the energy of the first orbit being the lowest:

$$E_1 = -13.6 \text{ eV}$$

and the highest energy state

$$E_{\infty} = 0$$

It means that different orbits represent different energy levels -13.6 eV to 0 . This is depicted in Fig. 24.6. $E = 0$ signifies that the electron is free.

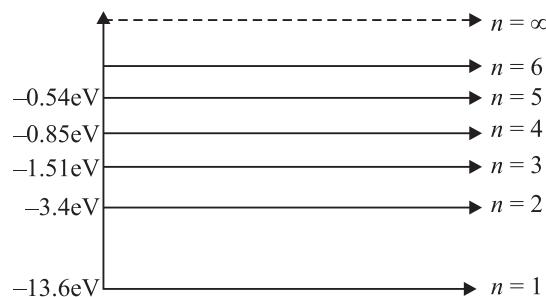


Fig. 24.6: Energy levels in hydrogen atom

According to Bohr's fourth postulate, the frequency v_{mn} of the emitted (absorbed) radiation when the electron falls (jumps) from the n th state to the m th state is given by

$$v_{mn} = \frac{RZ^2}{h} \left(\frac{1}{m^2} - \frac{1}{n^2} \right) \quad (24.14)$$

Fraunhofer Lines

The spectrum of sunlight, when examined carefully by a high power spectroscope, is found to be crossed by a large number of dark lines spread over the length of the continuous spectrum. Wollaston observed these lines in the year 1802. But their existence was studied by Fraunhofer on the basis of Kirchoff's laws and named these as Fraunhofer lines. The main body of the sun emits continuous spectrum but the atmosphere of comparatively much cooler vapours and gases in the Sun's atmosphere, called the chromosphere ($\sim 6000^\circ \text{C}$), absorb light corresponding to certain wavelengths. These appear as dark lines in the continuous spectrum of the sun.

Kirchhoff compared the absorbed wavelengths with the wavelengths emitted by various elements present on the earth and identified 60 terrestrial (existing on earth) elements present in the outer atmosphere of sun, e.g. oxygen, hydrogen, sodium, iron, calcium etc.



INTEXT QUESTIONS 24.2

1. Which of Bohr's postulates "fit" with classical physics and which support the ideas of quantum physics?
2. According to Bohr, why did an atom not collapse while its electrons revolved around the nucleus?
3. According to Bohr, what is happening in the atom when a photon of light is
 - (i) emitted
 - (ii) absorbed?
4. Write the energy of the first three orbits of hydrogen atom on the basis of Bohr's model.
5. An atom is excited to an energy level E_1 from its ground state energy level E_0 . What will be the wavelength of the radiation emitted?
6. In case of hydrogen atom, the radius of the electron in its n th orbit is proportional to
 - (i) $1/n$
 - (ii) $1/n^2$
 - (iii) n
 - (iv) n^2
7. The total energy E_n of the electron in the n th orbit of hydrogen atom is proportional to
 - (i) e^4
 - (ii) e^3
 - (iii) e^2
 - (iv) e

Notes



24.3 HYDROGEN SPECTRUM

Refer to Fig. 24.7. It shows frequency spectrum of hydrogen atom. As may be noted, the line spectrum of hydrogen consists of many lines in different regions of the spectrum. The various lines in a particular region of spectrum are found to have a pattern and may be represented by a common formula. So they are said to form a series. Let us about the series of hydrogen spectrum

Lyman series was discovered by in 1906. According to Bohr, this series arises when an electron jumps to the first orbit ($m = 1$) from an higher orbit ($n = 2, 3, 4\dots$). The frequencies of various spectral lines of this series are given by

$$\nu_{1n} = \frac{R}{h} \left(\frac{1}{1^2} - \frac{1}{n^2} \right)$$

where n is natural number greater than one.

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Atoms and Nuclei



Notes

Structure of Atom

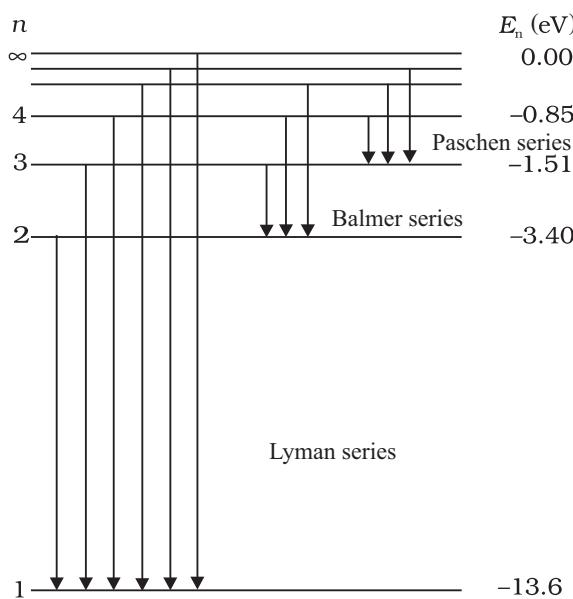


Fig. 24.7 : Energy Level diagram showing emission of various spectral series in hydrogen atom

Balmer series was discovered in 1885 in the visible region. According to Bohr, in this series, electron jumps to the second orbit ($m = 2$) from higher orbits ($n = 3, 4, 5\dots$). The frequencies of various spectral lines of the series are given by

$$v_{2n} = \frac{R}{h} \left(\frac{1}{2^2} - \frac{1}{n^2} \right); n > 2$$

Paschen series was discovered in 1908 in the near infra-red region. The existence of this series can be explained by assuming that electrons jump to third orbit ($m = 3$) from higher orbits ($n = 4, 5, 6\dots$). The frequencies of various spectral lines in the region are given by

$$v_{3n} = \frac{R}{h} \left(\frac{1}{3^2} - \frac{1}{n^2} \right); n > 3$$

Brackett series was discovered in mid infra-red region. In this series, electrons jump to fourth orbit ($n = 4$) from higher orbits ($n = 5, 6\dots$). Therefore, the frequencies of various spectral lines in the region are given by

$$v_{4n} = \frac{R}{h} \left(\frac{1}{4^2} - \frac{1}{n^2} \right); n > 4$$

Pfund series was discovered in far infra-red region. According to Bohr, this series is obtained when electron jumps to fifth orbit ($n_1 = 5$) from any higher orbit ($n = 6, 7\dots$). The frequencies of various spectral lines of the series are given by

$$v_{5n} = \frac{R}{h} \left(\frac{1}{5^2} - \frac{1}{n^2} \right); n > 5$$



Notes

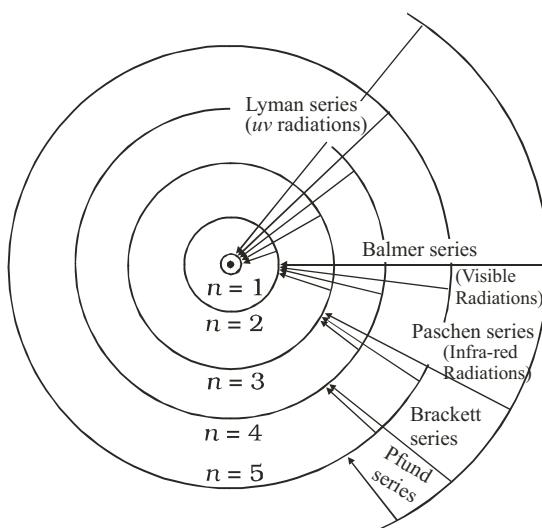


Fig. 24.8 : Permitted orbits in an atom of hydrogen and transitions leading to spectral lines of various series.



INTEXT QUESTIONS 24.3

- The negative total energy of an orbital electron means that it a) has emitted a photon, b) is bound to the nucleus, c) is in stable equilibrium, d) satisfies Bohr's postulate

$$L = \frac{nh}{2\pi}$$

- An electron jumps to the fourth orbit. When the electron jumps back to the lower energy level, the number of spectral lines emitted will be a) 6, b) 8, c) 5, d) 3.
- Lymann series of spectral lines are emitted when electron jump from higher orbits to the orbit
a) first, b) second, c) third, d) fourth.
- Which physical property of electron was quantized by Bohr?
- An electron jumps from third orbit to first orbit. Calculate the change in angular momentum of an electron?



24.4 X-RAYS

X-rays are produced when fast moving electrons are suddenly stopped by a heavy metal in a glass tube having extremely low pressure. The electrons emitted by the hot filament are focussed on the target which is made up of metal of high melting point and high atomic number as shown in Fig. 24.4.

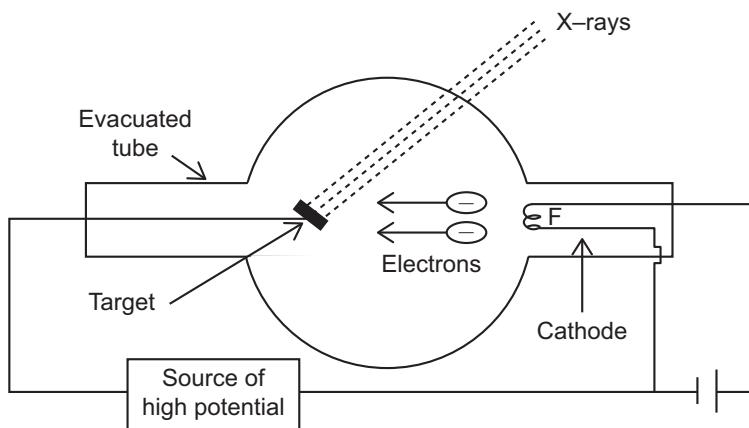


Fig. 24.4

When the electrons approach the target, 5% of their energy gets converted into X-rays and rest of the energy gets converted into heat, which is kept under control by the circulating cold water. The tube has extremely low pressure so that the electrons emitted from the hot filament (*F*) may directly hit the target without suffering collisions in between.

The intensity of the X-rays is controlled by adjusting the filament current and the quality is controlled by the accelerating voltage applied between the filament and target. This voltage usually ranges between 10 kV and 1 MV.

Properties of X-rays

X-rays show the following properties:

- They affect the photographic plate
- They cause fluorescence in certain chemical compounds.
- They ionize the gases.
- They show no reflection in mirrors, no refraction in glass, no diffraction with the conventional gratings but when refined techniques are used with atomic layers of crystals, they show all these familiar phenomena of light.
- They do not get deviated by electric or magnetic field.



Notes

X-rays are of two types :

- 1. Continuous X-rays:** All the X-ray tubes emit X-rays of all wavelengths beyond certain wavelength λ_{\min} . Some of the important features of the spectrum of continuous X-rays are as follows:
 - (i) The intensity of X-rays increases at all wavelengths as the voltage across the tube is increased.
 - (ii) The shortest wavelength λ_{\min} emitted is sharply defined and it depends on the voltage applied.
 - (iii) As the voltage is increased, the wavelength at which the maximum emission occurs shifts towards the shortest wavelength side as shown in Fig. 24.5.

Continuous X-rays are produced when the kinetic energy of the incident electrons is transformed into electromagnetic radiation upon collision with atoms. Before being stopped, electrons make several collisions and produce photons of all frequencies.

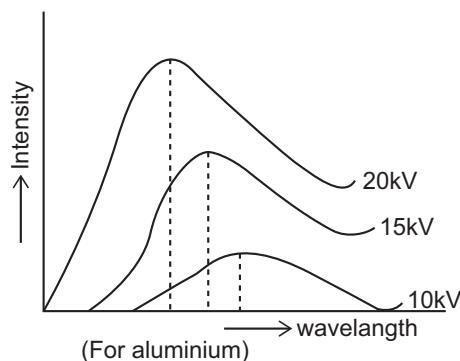


Fig. 24.5

Photon of largest frequency is produced when an electron makes head-on collision with the atom and loses all its energy at once. For such a collision, the photon frequency v_{\max} is highest and corresponding wavelength λ_{\min} is lowest.

$$\text{Thus, } eV = h\nu_{\max} = \frac{hc}{\lambda_{\min}}$$

This is known as Duane – Hunt Law.

- 2. Characteristic X-rays:** In addition to continuous X-rays, X-rays tubes emit radiations which are characteristic of the target used. It is observed

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Atoms and Nuclei



Notes

Structure of Atom

that on the continuous spectrum, at certain wavelengths, large amounts of energy are radiated. The positions of these lines do not depend on the voltage applied but depend only on the nature of the target used.

Mosley's law

Mosley investigated the characteristic X-rays of a large number of elements. He found that some specific characteristic lines appeared in the spectra of all elements but at slightly differing wavelengths (Fig. 24.6). Each characteristic line obeyed a specific equation. For example, K₂ lines obey the following relation

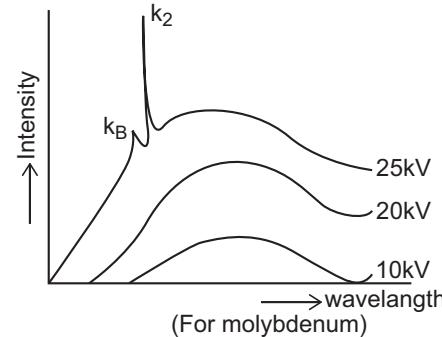


Fig. 24.6



WHAT YOU HAVE LEARNT

- Rutherford's scattering experiment indicated the presence of small central region inside the atom where all the positive charge and most of the mass of the atom is concentrated. The region was named as the nucleus.
- Electrons revolve around the nucleus and total negative charge is equal to the total positive charge of the nucleus.
- Rutherford's model of atom could not explain satisfactorily the observed stability of the atom and the electromagnetic radiation emitted by the atoms.
- A satisfactory model of an atom was suggested by Niels Bohr based on four postulates.
- Permissible orbits for electrons are those for which angular momentum ($I\omega$) = $nh/2\pi$
- Emission (absorption) of energy takes place when electron jumps from a higher orbit to a lower orbit (from a lower to a higher orbit).
- The radii of the permitted orbits in which the electron is free to revolve around the nucleus of the hydrogen atom are given by

$$a_n = \frac{n^2 h^2}{4\pi^2 m k e^2} = \frac{n^2 h^2 \epsilon_0}{Z e^2 m \pi}$$

For hydrogen atom, the radius of the first permitted orbit is $a = 0.53\text{\AA}$.

- The energy of the electron in the n^{th} orbit of the hydrogen atom is given by

$$E_n = -\frac{e^4 m}{8\pi^2 \epsilon_0^2 n^2}$$

The negative sign of total energy indicates that the electron is bound to the nucleus.

- The frequency of the photon emitted when the electron moves from the energy level E_i to E_f is given by :

$$\nu_{mn} = \frac{R}{h} \left[\frac{1}{m^2} - \frac{1}{n^2} \right]$$

- x-rays are produced when fast moving electrons are suddenly stopped by a heavy metal.
- x-rays are of two types (i) continuous and (ii) characteristic.
- Duane-Hunt law $eV = h\nu_{\max} = \frac{hc}{\lambda_{\min}}$
- Mosley law $\bar{v} = R \left(\frac{1}{1^2} - \frac{1}{2^2} \right) (z-1)^2$



TERMINAL EXERCISE

- In Rutherford's scattering experiment why do most of the α -particles pass straight through the target foil?
- In Rutherford's α -particle scattering experiment, what observation led him to predict the existence of nucleus?
- Why did Rutherford assume that electrons revolve in circular orbits around the nucleus?
- What is the ratio of the energies of the hydrogen atom in its first excited state to that its second excited state?
- What is the SI unit of Rydberg's constant?
- The Rydberg constant for hydrogen is 1096700 m^{-1} . Calculate the short and long wavelength limits of Lyman series.
- How many times does the electron of H-atom go round the first orbit in 1s ?
- Describe Rutherford's scattering experiment and discuss its findings and limitations.
- State the postulates of Bohr's model of atom.



Notes

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Atoms and Nuclei



Notes

Structure of Atom

10. Derive an expression of the energy of the electron in the n th orbit of hydrogen atom.
11. Choose the correct answers
 - (a) The total energy E_n of the electron in the n th orbit of hydrogen atom is proportional to.
 - (i) $1/n^4$ (ii) $1/n^2$ (iii) $1/n^3$ (iv) $1/n$
 - (b) The energy required to remove an electron from $n = 1$ to $n = \infty$ in the case of hydrogen atom is
 - (i) 13.6 V (ii) 13.6 eV (iii) 13.6 MeV (iv) 13.6 keV
 - (c) For hydrogen atom when the electron jumps from higher energy level $n = 5, 6, 7, \dots$ etc. to the energy level $n = 4$, a set of spectral lines are obtained. These are called
 - (i) Balmer Series
 - (ii) Bracket Series
 - (iii) Paschen Series
 - (iv) Lyman Series
12. Calculate the radius of the third and fourth permitted orbits of electron in the hydrogen atom.
13. The energy transition in H-atom occurs from $n = 3$ to $n = 2$ energy level. Given $R = 1.097 \times 10^7 \text{ m}^{-1}$.
 - (i) What is the wavelength of the emitted radiations?
 - (ii) Will this radiation lie in the range of visible light?
 - (iii) To which spectral series does this transition belong?
14. The ionisation potential of hydrogen is 13.6 volt. What is the energy of the atom in $n = 2$ state?
15. How are x-rays produced? Draw a labelled diagram to illustrate?
16. List the properties of x-rays and compare them with that of visible light.
17. How are continuous x-rays produced? What is the condition for the production of photons of highest frequency?



ANSWERS TO INTEXT QUESTIONS

24.1

1. a (iii), b (ii), c (i)
2. It could not explain the large angle scattering of particles from the gold foils as observed by Rutherford.

24.2

1. Bohr's first postulate is from classical physics; remaining three are from quantum physics.
2. Because the orbits are stationary.
3. (i) Electron falls from higher to lower energy state.
(ii) Electron is excited to some higher energy state.
4. $E_1 = -13.6\text{ eV}$, $E_2 = 3.4 \text{ eV}$, $E_3 = -1.51\text{ eV}$
5. $\lambda = \frac{hc}{E_i - E_0}$
6. (iv)

24.3

1. (b)
2. (a) Number of spectral lines emitted = $\frac{1}{2}n(n-1) = \frac{1}{2} \times 4(4-1) = 6$
3. (a)
4. Angular momentum of revolving electron.
5. From the n th states with principal quantum number n calculate the number of wavelengths observed in the spectrum from a hydrogen sample.

Answers to Problems in Terminal Exercise

- | | |
|--|---|
| 4. 9 : 4 | 6. $\lambda_s = 911.4 \text{ \AA}$, $\lambda_l = 1215 \text{ \AA}$ |
| 7. $6.57 \times 10^{15} \text{ Hz}$. | 11. (a) (i), (b) (ii), (c) (iii), (d) (i), (e) (v) |
| 13. (i) 6563 \AA , (ii) visible (iii) Balmer Series. | 14. -3.4 eV . |

**Notes**

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Atoms and Nuclei



Notes

25

DUAL NATURE OF RADIATION AND MATTER

You must have seen films in cinema halls. The picture on the screen is produced by passing light through films which have the scenes shot on them. But have you ever thought as to how the sound is reproduced in the cinema? The sound is also recorded on the side of the film as a sound track. The light beam passing through this sound track falls on a photocell, which converts it into electrical pulses. These electrical pulses are converted into sound. In this lesson you will study the effect which governs the working of a photocell. It is called the *photoelectric effect*. It is also used in burglar alarm to detect intruders. Einstein's explanation of photoelectric effect led de Broglie (read as de Broy) to the wave-particle duality, i.e. matter exhibits wave as well as particle properties.

You now know that a particle is characterized by properties such as definite position, size, mass, velocity, momentum, etc. Its motion is described by Newton's laws of motion. On the other hand, a wave is characterized by properties such as periodicity in space-time, wavelength, amplitude, frequency, wavevelocity, etc. It transports energy, but no matter. That is, it extends in space unlike a particle, which is localised. The term wave-particle duality refers to the behaviour where both wave-like and particle-like properties are exhibited under different conditions by the same entity. His arguments were simple: Nature likes simplicity and loves symmetry. So if wave-particle duality can be exhibited by light, it should be exhibited by matter as well. You will learn about his explanation of matter waves in sec. 25.4.



OBJECTIVES

After studying this lesson, you should be able to:

- *explain photoelectric effect;*
- *describe the experimental arrangement to study photoelectric effect;*

- state the laws of photoelectric emission;
- interpret the graphs between frequency of radiation and retarding potential;
- write deBroglie wavelength of matter waves associated with a particle of momentum \mathbf{p} ; and
- describe the experimental arrangement for the verification of matter waves.

25.1 PHOTOELECTRIC EFFECT

In 1887, while working on propagation of electromagnetic waves, Hertz discovered that air in a spark gap became a better conductor when it was illuminated by ultraviolet rays. Further experiments by him showed that zinc became positively charged on irradiation by ultraviolet rays. In 1900, Leonard showed that electrons were emitted from a metal surface when light of sufficiently high frequency falls on it. This phenomenon is known as *photoelectric effect* and the electrons so emitted are called photoelectrons.

The emission of electrons from metals irradiated by light of a frequency greater than a certain characteristic frequency is called *photoelectric effect*.

25.1.1 Experimental Arrangement to Study Photoelectric Effect

Refer to Fig. 25.1. It shows a schematic diagram of the apparatus that can be used to study this phenomenon.

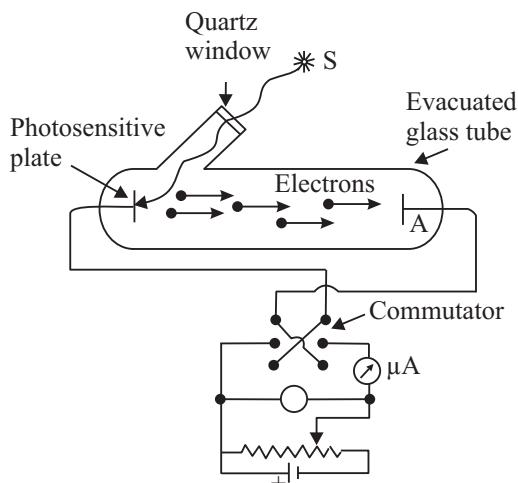


Fig. 25.1: Experimental arrangement for observing the photoelectric effect

A metallic cup C called photo cathode is sealed inside an evacuated tube along with another metal plate A , which is used to collect photoelectrons emitted by C . These electrodes are connected to a battery and microammeter circuit, as shown in Fig. 25.1. The battery is so connected that the voltage on plate A is positive with respect to C . If the battery terminals are reversed, the voltage of the plate A will become negative relative to C .

Notes

The emission of electrons from metals can also take place when they are heated. This is known as **thermionic emission**.

Note that electrons gain energy from thermal energy in thermionic emission.

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Notes

Dual Nature of Radiation and Matter

To study the effect of intensity of incident light on the number of photoelectrons emitted by C , the collector plate A is kept at positive potential relative to C .

Keeping the frequency of incident light and the value of accelerating potential fixed, the photoelectrons emitted per unit area from the emitting surface vary linearly with the intensity of light, as shown in Fig. 25.2 (a).

Case-I: Plate A positive relative to C

Let us first consider the case when the plate A is at a positive potential relative to C . When light of high frequency is incident on the emitter, it starts emitting electrons. Since A is at a higher potential relative to C , the emitted electrons experience an attractive force. When we increase the voltage on A , the kinetic energy of the photoelectrons increases. The current in the outer circuit shown by the microammeter depends on the number of electrons reaching the plate A . If we keep on increasing the voltage, a stage comes when all the emitted electrons are collected by the plate. The current is said to have *saturated* at this stage. If the voltage on the plate is increased further, the current remains constant in magnitude. This behaviour of current with respect to plate voltage is shown in Fig. 25.2(b). The voltage V_s is called *saturating voltage*.

Case-II: Plate A negative relative to C

If C is at a positive potential relative to the plate A and light of a proper frequency is incident on the emitter, photoelectrons emitted by C will experience retarding potential, which impedes their motion towards A . Some of the electrons emitted from C may still reach the plate. This gives rise to current, which is registered by the microammeter. What does this mean? If the p.d between A and C only provides the force which makes the electrons move towards the plate, then none of the electrons should have reached the plate. Since such electrons have overcome the retarding potential while moving against it to reach the plate, they have some initial kinetic energy. This is also confirmed by observed results. For any incident light of particular frequency, if the retarding potential is gradually increased in magnitude, a stage is ultimately reached when none of the electrons reach the plate and the current becomes zero.

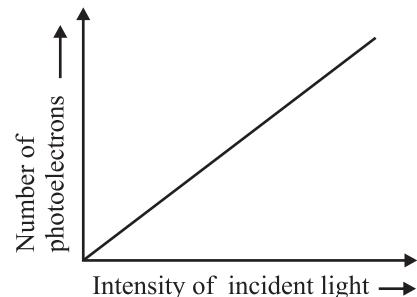


Fig. 25.2 : (a) Variation of number of photoelectrons with intensity of incident light

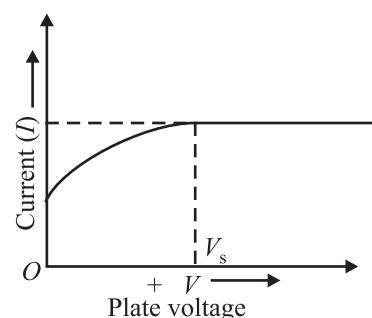


Fig. 25.2 : (b) Dependence of photoelectric current on plate voltage.

The minimum retarding potential for which the photoelectric current becomes zero for a particular frequency of incident light is called the stopping potential, V_0 for that frequency.

The work done by an electron W against the stopping potential V_0 is eV_0 where e is electronic charge. This work is done by the electron at the expense of its kinetic energy. So, we can write

$$eV_0 = \frac{1}{2} m v_{\max}^2 \quad (25.1)$$

The stopping potential V_0 was found by Millikan to depend on the frequency of the incident light. A plot of the stopping potential (V_0) versus the frequency of the incident light (v) is shown in Fig. 25.3. You will note that there is a minimum cut-off frequency v_0 below which ejection of electrons is not possible. It is called *threshold frequency*.

To study the effect of frequency of incident light on stopping potential, Millikan adjusted the intensity of light at a fixed value for various frequencies and studied the variation of photoelectric current with anode potential. He obtained different values of stopping potential for different frequencies of incident light. Moreover, the stopping potential is more negative for higher frequencies, as shown in Fig. 25.4. This implies that if the frequency of the incident light increases, the maximum kinetic energy of the photoelectrons also increases. Therefore, with increasing frequency, greater retarding potential is required to completely stop the movement of photoelectrons towards the anode.

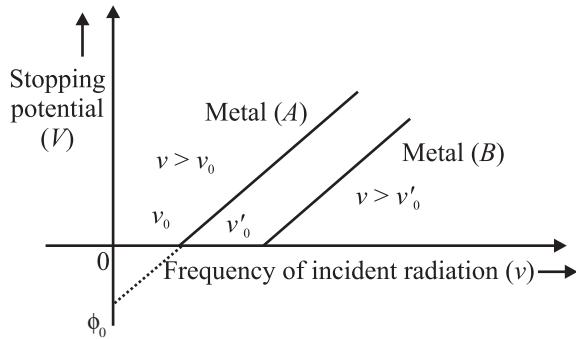


Fig. 25.3 : Stopping potential versus frequency of incident light.

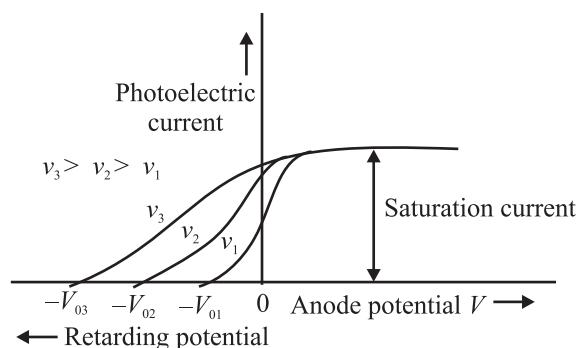


Fig. 25.4 : Photo electric current

This experiment also established that there exists a minimum cut-off frequency v_0 for which stopping potential is zero. Moreover, photo emission begins as soon as light is incident on the material, i.e. photo emission is instantaneous, even if the incident light is dim. Now it is known that time lag between incident light and emission of photoelectrons from the emitter is of the order of 10^{-9} s.



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These observations can be summarised as follows :

- *The maximum velocity of photoelectrons increases with frequency of incident light and depends on the nature of emitter material.*
- *The maximum velocity of photoelectrons does not depend on the intensity of incident light.*
- *For every material, there exists a threshold frequency below which no photoelectrons are emitted.*
- *For a particular frequency, the number of photoelectrons emitted per unit area of the emitting surface is proportional to the intensity of the incident light.*
- *There is practically no time lapse ($\sim 10^{-9}$ s) between the incidence of light on the metal and emission of electrons from it. In other words, photoelectric emission is an instantaneous process.*



INTEXT QUESTIONS 25.1

1. State whether the following statements are true or false :
 - (a) In thermionic emission, electrons gain energy from photons.
 - (b) The maximum velocity of photoelectron is independent of the frequency of incident radiation.
 - (c) There exists a frequency v_0 below which no photoelectric effect takes place.
2. Refer to Fig. 25.3 and interpret the intercepts on x and y -axes and calculate the slope.
3. Draw a graph showing the variation of stopping potential ($-V_0$) with the intensity of incident light.

25.2 EINSTEIN'S THEORY OF PHOTOELECTRIC EMISSION

In 1905, Einstein proposed a simple but revolutionary explanation for the photoelectric effect. He assumed that light consists of bundles of energy, called photons and viewed photoelectric effect as a collision between a photon and a bound electron.

The energy E of a single photon is given by

$$E = hv \quad (25.2)$$



Robert A. Millikan (1868-1953)

Robert Andrews Millikan was born on March 22, 1868 in U.S.A. During his undergraduate course, his favourite subjects were Greek and Mathematics. But after his graduation in 1891, he took, for two years, a teaching post in elementary physics. In this period, he developed interest in the subject. He received his Ph.D. (1895) for research on polarization of light emitted by incandescent surfaces.

Millikan spent a year (1895-1896) in Germany, at the Universities of Berlin and Göttingen. He returned at the invitation of A.A. Michelson to take appointment as his assistant at the newly established Ryerson Laboratory at the University of Chicago (1896). He became Professor at that University in 1910, a post which he retained till 1921. As a scientist, Millikan made numerous momentous discoveries in the fields of electricity, optics, and molecular physics. His earliest major success was the accurate determination of the charge carried by an electron, using the elegant "falling-drop method". He also proved that this quantity was a constant for all electrons demonstrating the quantised nature of charge.

He also verified experimentally Einstein's photoelectric equation, and made the first direct photoelectric determination of Planck's constant h . Throughout his life, Millikan remained a prolific author, making numerous contributions to scientific journals. He was awarded the Nobel Prize in Physics in 1923.

where v is the frequency of the incident light and h is Planck's constant. Let us now assume that a photon of energy hv is incident on the metal surface. Suppose ϕ_0 is the energy needed for an electron to come out of the metal surface. As you have studied earlier, this energy is also called the *work function* of the conductor. *The work function of a conductor is the minimum energy required by an electron to come out of the conductor surface.*

The typical values of work function for a few metals are given (in eV.) in Table 25.1, along with the corresponding threshold frequency (v_0).

What do you think would happen when a photon of energy E ($> \phi_0$) strikes the metal surface? We expect that out of the total energy E , an amount ϕ_0 would be used up by the electron to come out of the metal surface. The difference in energy, i.e. $(E - \phi_0)$, would then be imparted to the emitted electron in the form of kinetic energy. (The electron may lose some energy in internal collisions before it escapes from the metal surface.) Mathematically, we can write

$$hv = \phi_0 + K_{\max} \quad (25.3)$$



Notes

Table 25.1: Work function and threshold frequencies of some typical metals

| Metal | ϕ_0 (eV) | v (Hz) |
|-----------|---------------|------------------------|
| Sodium | 2.5 | 6.07×10^{14} |
| Potassium | 2.3 | 5.58×10^{14} |
| Zinc | 3.4 | 8.25×10^{14} |
| Iron | 4.8 | 11.65×10^{14} |
| Nickel | 5.9 | 14.32×10^{14} |

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Albert Einstein (1879-1955)



Albert Einstein was born in Wurttemberg, Germany, on March 14, 1879. In 1901, he acquired Swiss citizenship and, as he was unable to find a teaching post, he accepted a position as technical assistant in the Swiss Patent Office. During his stay at the Patent office, in his spare time, he produced much of his remarkable work, including the theory of photoelectric effect and the special theory of relativity. In 1909 he became Professor Extraordinary at Zurich. In 1911 he accepted the post of Professor of Theoretical Physics at Prague but returned to Zurich in the following year to fill a similar post. In 1914, he was appointed Director of the Kaiser Wilhelm Physical Institute and Professor in the University of Berlin. He became a German citizen in 1914. He was awarded the Nobel Prize in Physics in 1921 for his theory of photoelectric effect, though he is more famous for his theory of relativity. He remained in Berlin until 1933, when he renounced his citizenship for political reasons and immigrated to take the position of Professor of Theoretical Physics at Princeton in USA.

He became a US citizen in 1940 and retired from his post in 1945. He spent the later years of his life working on General Theory of Relativity and Unification of basic Forces. Einstein nurtured scientific humanism. He protested to President Roosevelt against the use of nuclear bombs for destruction of humanity. He is considered the greatest scientist to have ever walked on this planet and named scientist of the millennium.

Let us now see how observed results can be explained on the basis of this theory.
Let us take

$$\phi_0 = h\nu_0.$$

Then Eqn. (25.3) takes the form

$$K_{\max} = \frac{1}{2}mv^2 = h(v - \nu_0) \quad (25.4)$$

This equation implies that

- For ν_{\max} to be positive, no emission can take place for $v < \nu_0$. That is, the incident light must have frequency above the threshold frequency.
- K_{\max} is linearly proportional to $(v - \nu_0)$
- An increase in the intensity of incident light of frequency v corresponds to an increase in the number of photons. Each and every photon has same

energy; there is no increase in the energy of photoelectrons. However, the no. of emitted electrons and hence photocurrent will increase with increase in intensity.

- Since photoelectric effect is produced by collisions between photons and electrons, the energy transfer from photons is instantaneous, i.e. there is almost no time lag.
- Since work function is a characteristic property of a material, v_0 is independent of the intensity of incident light.



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We see that Einstein's theory of the photoelectric effect successfully explains its physical origin.

To understand these concepts and get an idea about the values of physical parameters, go through the following examples carefully.

Example 25.1: Sodium has a work function of 2.3 eV. Calculate (i) its threshold frequency, (ii) the maximum velocity of photoelectrons produced when sodium is illuminated by light of wavelength 5×10^{-7} m, (iii) the stopping potential for light of this wavelength. Given $h = 6.6 \times 10^{-34}$ J s, $c = 3 \times 10^8$ m s $^{-1}$, 1eV = 1.6×10^{-19} J, and mass of electron $m = 9.1 \times 10^{-31}$ kg.

Solution: (i) The threshold frequency is given as $hv_0 = \phi_0$. Here, $h = 6.6 \times 10^{-34}$ Js and $\phi_0 = 2.3$ eV = $2.3 \times 1.6 \times 10^{-19}$ J.

∴

$$v_0 = \frac{\phi_0}{h}$$

$$= \frac{2.3 \times 1.6 \times 10^{-19} \text{ J}}{6.6 \times 10^{-34} \text{ J s}} = 5.6 \times 10^{14} \text{ Hz}$$

(ii) From Einstein's photoelectric equation, we know that

$$hv = \phi_0 + K_{\max} = \phi_0 + \frac{1}{2} mv_{\max}^2,$$

Since $v = \frac{c}{\lambda}$, we can write

$$E = h \times \frac{c}{\lambda} = \phi_0 + \frac{1}{2} (mv_{\max}^2)$$

where c is velocity of light and λ is its wavelength. On substituting the given values, we get

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$$\therefore E = \frac{(6.6 \times 10^{-34} \text{ Js})(3 \times 10^8 \text{ ms}^{-1})}{5 \times 10^{-7} \text{ m}}$$

$$= 3.96 \times 10^{-19} \text{ J}$$

$$\Rightarrow 3.96 \times 10^{-19} = 2.3 \times 1.6 \times 10^{-19} + \frac{1}{2} m v_{\max}^2$$

$$= 3.68 \times 10^{-19} + \frac{1}{2} m v_{\max}^2$$

$$\therefore v_{\max}^2 = \frac{2 \times 0.28 \times 10^{-19}}{m} = \frac{2 \times 0.28 \times 10^{-19}}{9.1 \times 10^{-31}}$$

$$\therefore v_{\max} = \sqrt{\frac{0.56 \times 10^{-19} \text{ J}}{9.1 \times 10^{-31} \text{ kg}}} = 2.5 \times 10^5 \text{ m s}^{-1}$$

(iii) The stopping potential V_0 is given as

$$eV_0 = \frac{1}{2} m v_{\max}^2$$

$$\therefore V_0 = \frac{0.28 \times 10^{-19} \text{ J}}{1.6 \times 10^{-19} \text{ J V}^{-1}} = 0.18 \text{ V}$$

You may now like to answer some simple questions



INTEXT QUESTIONS 25.2

1. Calculate the momentum of a photon of frequency v .
2. If the wavelength of an electromagnetic radiation is doubled, how will be the energy of the photons change?
3. The intensity of incident radiation is doubled. How will it affect the kinetic energy of emitted photoelectrons.

25.3 PHOTOELECTRIC TUBE

You have studied the photoelectric effect in detail now. We know that when light of a frequency above v_0 is incident on a material, electrons are emitted. Their kinetic energies are different. We also know that flow of electrons constitutes current.

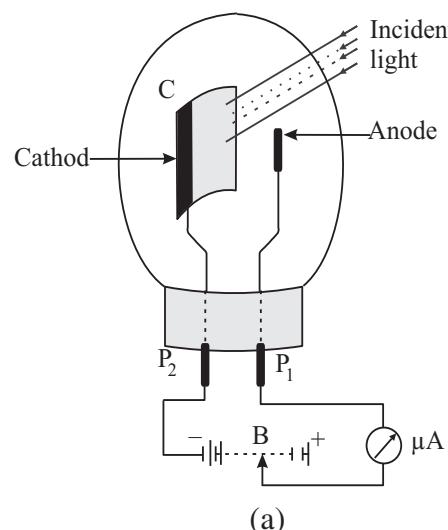
The photoelectric tube is based on photoelectric effect.

A photoelectric tube consists of an evacuated glass vessel which contains a semi-cylindrical cathode and an anode in the form of a straight wire. The cathode is coated with a metal of low work function to ensure emission of photoelectrons when light of a pre-decided frequency is incident on it. The threshold frequency above which a phototube responds determines the choice of this coating.

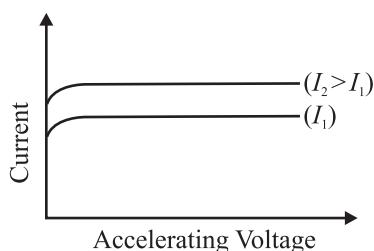
The anode is usually made of nickel or platinum. Electrical connections P_1 and P_2 are brought out on to the surface of the glass vessel. A battery and a microammeter are connected between the anode and the cathode to provide the accelerating voltage. The arrow on the battery indicates that the voltage applied by it can be varied. The microammeter placed in the circuit measures the current passing through it (Fig. 25.5a).

To understand the working of a photoelectric tube, let us suppose that light of frequency higher than the threshold frequency is incident on the cathode. Some photoelectrons will be emitted even when accelerating potential between the cathode and the anode is zero. As you know, these electrons come out of the cathode with definite kinetic energy and reach the anode giving rise to a current, which is indicated by the microammeter. Let us now apply some accelerating voltage and see what happens. Obviously, more electrons will reach the anode and increase the current. This is shown in Fig. 25.5(b).

As we keep on increasing the voltage between the cathode and the anode of the photoelectric tube, current also increases. However, at high voltage, the current saturates to a fixed value as shown in Figure 25.5(b). ***The value of saturation current is determined by the intensity (I) of incident light.*** The magnitude of saturation current is of the order of nano-ampere ($\sim 10^{-9}$ A). It is seen that ***if the intensity of the light is increased, the saturation current also increases, as shown in the Fig. 25.5. (b)***



(a)



(b)

Fig. 25.5: Variation of current and accelerating voltage



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25.3.1 Applications

Photoelectric cells find wide applications in processes wherever light energy has to be transformed into equivalent electric current.

(i) Reproduction of sound in cinematography (motion pictures): One of the most important applications of photoelectric cells is in reproduction of sound in films. A sound track is a track on the film of uniform width whose intensity varies in accordance with the audio frequency variations. Light is made to pass through this film and is then made to fall on the cathode of a photoelectric cell. The current developed in the circuit of the photoelectric cell is proportional to the audio frequency and the variations in current are also in accordance with the variations in the audio frequency. This current is then made to pass through a resistance. The voltage across the resistance is then suitably amplified and fed to a loudspeaker. The loudspeaker reproduces the sound as was originally recorded on the sound track. You will learn more about it on the optional module on photography and video-recording.

(ii) Transmitting pictures over large distances : Photo electric tubes are also used in systems that transmit pictures over large distances. The technique of transmission of pictures to large distances is called **photo-telegraphy**.

(iii) Other Uses : Many types of systems used for counting articles or living beings are based on photoelectric tubes. These are also used in burglar alarms, fire alarms, detectors used for detection of traffic law defaulters, in television camera for scanning, telecasting scenes and in industry for detecting minor flaws or holes in metal sheets.



INTEXT QUESTIONS 25.3

1. State whether the following statements are true or false:
 - (a) The cathode in a phototube is biased positively with respect to anode.
 - (b) The saturation current in a phototube depends on the frequency of incident radiation.
 - (c) The saturation current in a photodiode increases with intensity of incident light.
2. State three applications of photoelectric tube.
3. A phototube is illuminated by a small bright source placed 100 cm away. When the same source of light is 50 cm away, what will be the effect on the number of electrons emitted by photo cathode?

In the previous section, you have studied Einstein's theory of photoelectric effect and learnt that light consists of photons. You have also learnt that the phenomena

of interference and diffraction can be explained on the basis of wave theory of light. This duality in the nature of light came to be accepted by the physicists in the early 20th century. Thinking about the wave-particle duality of light, de Broglie asked himself the question : If light exhibits dual nature, will particles of matter also not act like waves? Successful resolution of this question led to de Broglie hypothesis.

25.4 THE DE BROGLIE HYPOTHESIS

As a young graduate student, de Broglie argued with a great amount of insight that since nature loves symmetry and simplicity in physical phenomena, ordinary “particles” such as electrons, and protons should also exhibit wave characteristics under certain circumstances. His argument runs as follows : Light is an electromagnetic radiation and exhibits wave-particle duality. Therefore, Einstein’s mass-energy equivalence relation ($E = mc^2$), which essentially treats light as quantum of photon, a particle, can hold only if matter also exhibits wave character. He therefore proposed that the wavelength and frequency of matter waves should be determined by the momentum and energy of the particle in exactly the same way as for photons : $E = pc$ and the associated wavelength λ of a particle having momentum p is given by

$$\lambda = \frac{h}{p} \quad (25.5)$$

Since the momentum of such a particle is given by $p = mv$, we can write

$$\lambda = \frac{h}{mv} \quad (25.6)$$

λ is called deBroglie wavelength. Eqn. (25.5) is a complete statement of wave-particle duality. It implies that a particle with a momentum p can exhibit wave-like properties and the wavelength of the associated matter waves is h/p . The converse is also true, i.e., a wave of wavelength λ can exhibit particle-like properties and the momentum of the wave-matter is h/λ .

This hypothesis, submitted as Ph.D Thesis was initially rejected by the examiners. However, soon, experimental evidence proved de Broglie’s argument. This has a very important inspirational lesson for us : We must keenly analyse every statement and try to seek experimental evidence.

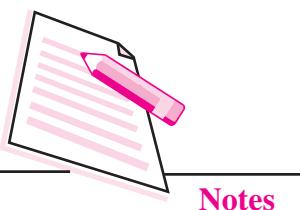
The actual wavelength of anything macroscopic is incomprehensively small, as you can see by calculating it for a cricket ball. The case is quite different for elementary particles such as the electron. An electron has energy E when accelerated through potential difference V . Hence, we can write



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$$\frac{1}{2}mv^2 = qV \quad (25.7)$$

or

$$v = \sqrt{\frac{2qV}{m}}$$

so that

$$mv = p = \sqrt{2qmV} \quad (25.8)$$

On combining this result with Eqn.(25.5), we find that de-Broglie wavelength is given by

$$\lambda = \frac{h}{p} = \frac{h}{\sqrt{2qmV}} \quad (25.9)$$

The constants appearing in Eqn. (25.9) have the values: $h = 6.625 \times 10^{-34}$ Js, $q = 1.602 \times 10^{-19}$ C and $m = 9.11 \times 10^{-31}$ kg. On substituting these values in Eqn.(25.9), we obtain

$$\begin{aligned} \lambda &= \frac{6.625 \times 10^{-34} \text{ Js}}{\sqrt{2 \times (1.602 \times 10^{-19} \text{ C}) \times (9.11 \times 10^{-31} \text{ kg}) \times \sqrt{V}}} \\ &= \frac{12.3}{\sqrt{V}} \times 10^{-10} \text{ m} \\ &= \frac{12.3}{\sqrt{V}} \text{ Å} \end{aligned} \quad (25.10)$$

It means that if an electron is accelerated through a potential difference of 100V, its wavelength will be given by

$$\begin{aligned} \lambda &= \frac{12.3}{\sqrt{100}} \text{ Å} \\ &= 1.23 \text{ Å} \end{aligned}$$

This is also the wavelength of an electron of energy 100eV. You can easily verify this using the relation

$$\lambda = \frac{h}{(2meE)^{1/2}}$$

The wavelength of matter waves associated with 100eV electrons lies in the X-ray region and is of the same order as the interatomic separation in a solid. We therefore expect these to undergo diffraction by a crystal lattice.

The first experimental evidence of matter waves came from the work of Davisson and Germer, who were studying scattering of electrons by crystals. Let us learn about it now.



Louis Victor de Broglie

(1892-1987)

Louis de Broglie was born at Dieppe, France on 15th August, 1892. He first studied the arts and took his degree in history in 1910. Then, as his liking for science prevailed, he studied for a science degree, which he gained in 1913. In 1924 at the Faculty of Sciences at Paris University, he submitted a thesis *Recherches sur la Théorie des Quanta* (Researches on the quantum theory), which gained him his doctor's degree. This thesis contained a series of important findings, which he had obtained in the course of about two years. The ideas set out in that work served the basis for developing *wave mechanics*, a theory which has transformed our knowledge of physical phenomena on the atomic scale.

In 1929 he was awarded the Nobel Prize for Physics for his discovery of the wave nature of electrons.

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25.4.1 Experimental Evidence for Existence of de Broglie Waves

The schematic diagram of Davisson-Germer experiment is shown in Fig.25.6. The set up consists of a filament F, which serves as a source of electrons. The electrons emitted from this filament are made to pass through a set of metal diaphragms having a number of slits. The electrons emitted by the filament come out in various directions. The metal diaphragms serve to

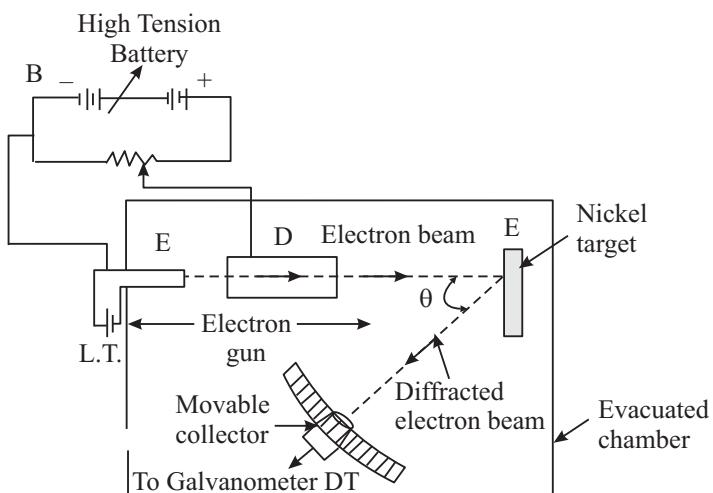


Fig. 25.6: Experimental set up to verify the existence of matter waves

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collimate these electrons. Only those electrons which are able to pass through the slits in the various diaphragms are able to come out.

Note that the energy of the collimated stream of electrons is controlled by changing the magnitude of the accelerating voltage. The beam of electrons is made to fall perpendicularly on a single crystal of nickel. The set-up also contains a detector D_t which can be placed at any particular angle with respect to the normal to the target crystal. This detector determines the intensity of the reflected beam. Note that there is nothing special in the choice of nickel.

Fig. 25.7 shows a plot of detector current versus kinetic energy of incident electrons for $\theta = 50^\circ$. As may be noted, the detector current shows a maxima for electrons of kinetic energy 54 eV. If you calculate the wavelength of these electrons, you will get

$$\lambda = \frac{6.62 \times 10^{-34} \text{ Js}}{[2 \times (9.1 \times 10^{-31} \text{ kg}) \times 54 \times 1.6 \times 10^{-19} \text{ J}]^{1/2}}$$
$$= 1.67 \text{ \AA}$$

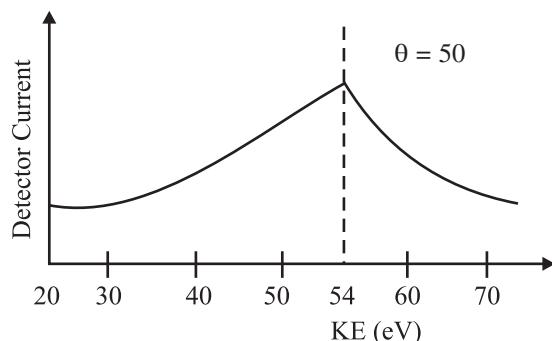


Fig. 25.7: Plot of detector current versus kinetic energy of electrons

25.4.2 Applications of de Broglie Waves

We now know that very small values of wavelength can be achieved by increasing the kinetic energy of electrons. From Lesson 23 you may recall that resolving power of an optical microscope depends on the wavelength of light used. In fact, the resolution increases with decreasing wavelength. Can you guess as to what would happen if a beam of very energetic electrons is used in a microscope instead of photons? Well, obviously you could obtain very high resolution and magnification by lowering the deBroglie wavelength associated with the electrons. This technique is used in electron microscopes. A comparison of structure and working of electron microscope with optical microscope is shown in Fig. 25.8.

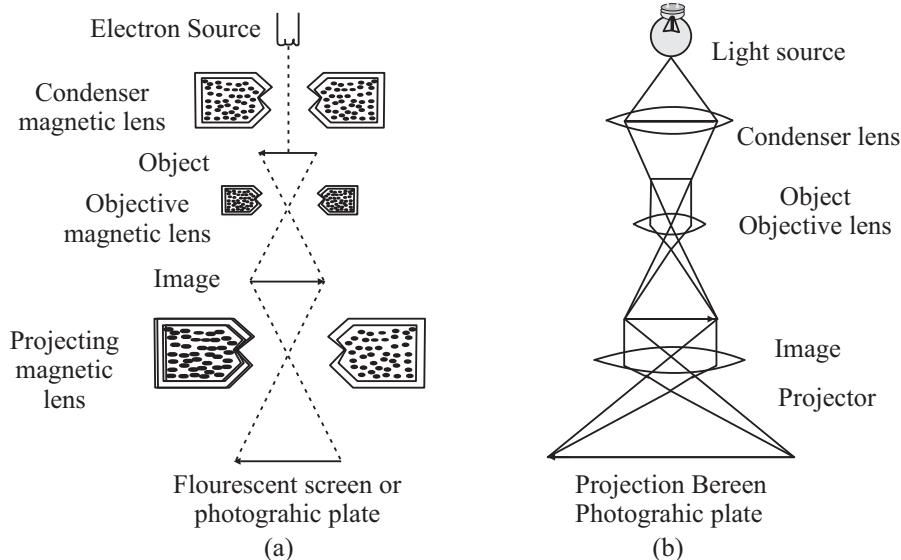


Fig. 25.8: a) Electron microscope, and b) Optical microscope

Notes

Story of Davisson and Germer's experiment

Germer had recorded in his notebook that he discovered a crack in the vacuum trap in the electron scattering apparatus on Feb. 5, 1925 when he was working with Clinton Davisson at Westren Electric, New York, U.S.A. This was not the first time their equipment had broken, and not the first time they had “resurrected” their precious nickel crystal by **heating it in vacuum and hydrogen**.



This particular break and the subsequent method of repair, however, had a crucial role to play in the later discovery of electron diffraction. By 6 April 1925, the repairs had been completed and the tube put back into operation. During the following weeks, as the tube was run through the usual series of tests, results very similar to those obtained four years earlier were obtained. Then suddenly, in the middle of May, unprecedented results began to appear. This puzzled Davisson and Germer so much that they halted the experiments a few days later, cut open the tube and examined the target (with microscopist F. F. Lucas) to see if they could detect the cause of new observations. What they found was this: the polycrystalline form of nickel target had been changed by the extreme heating until it had formed about ten crystal facets in the area

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from which the incident electron beam was scattered. Davisson and Germer surmised that the new scattering pattern must have been caused by the new crystal arrangement of the target. In other words, they concluded that it was the arrangement of the atoms in the crystals, not the structure of the atoms that was responsible for the new intensity pattern of scattered electrons.

During the summer of 1926, Davisson and his wife had planned a vacation trip to relax and visit relatives in England. Something was to happen on this particular trip. Theoretical physics was undergoing fundamental changes at this time. In the early months of 1926, Erwin Schrodinger's remarkable series of papers on wave mechanics appeared, following Louis de Broglie's papers of 1923-24 and Albert Einstein's quantum gas paper of 1925. These papers were the subject of lively discussions at the Oxford meeting of the British Association for the Advancement of Science.

Davisson, who generally kept abreast of recent developments in his field but appears to have been largely unaware of these developments in quantum mechanics, attended this meeting. He was surprised when he heard a lecture by Born in which his own and Kunsman's (platinum target) curves of 1923 were cited as confirmatory evidence for de Broglie's electron waves!

Davisson shared the 1937 Nobel Prize for Physics with G.P. Thomson (son of J.J. Thomson).

Electron Microscope

Electron microscopes are scientific instruments that use a beam of highly energetic electrons to examine objects on a very fine scale. This examination can yield the following information:

The surface features of an object or “how it looks”, its texture; direct relation between these features and material properties (hardness, reflectivity, etc.), the shape and size of the particles making up the object; direct relation between these structures and materials properties (ductility, strength, reactivity, etc.), the elements and compounds that the object is composed of and the relative amounts of them; direct relationship between composition and material properties (melting point, reactivity, hardness, etc.). How are the atoms arranged in the object?

Electron microscopes were developed due to the limitations of optical microscopes, which are limited to $500\times$ or $1000\times$ magnification and a resolution of 0.2 micrometers. In the early 1930's, this theoretical limit had been reached and there was a scientific desire to see the finer details of the interior structures of organic cells (nucleus, *mitochondria*, etc.). This required $10,000\times$ plus magnification which was just not possible using the microscopes available at that time.

The Transmission Electron Microscope (TEM) was the first Electron Microscope to be developed and is patterned exactly on the Light Transmission Microscope, except that a focused beam of electrons is used instead of light to image the specimen and gain information about its structure and composition. It was developed by Max Knoll and Ernst Ruska in Germany in 1931.

Transmission Electron Microscope (TEM)

A TEM works much like a slide projector. A projector throws a beam of light on the slide. As the light passes through the slide, it is affected by the structures and objects on the slide. As a result, only certain parts of the light beam are transmitted through certain parts of the slide. This transmitted beam is then projected onto the viewing screen, forming an enlarged image of the slide.

TEMs work in the same way, except that they shine a beam of electrons through the specimen. Whatever part is transmitted is projected onto a phosphor screen for the user to see.

The electron gun produces a stream of monochromatic electrons. This stream is focused to a small, thin, coherent beam by the use of condenser lenses 1 and 2. The first lens (usually controlled by the “spot size knob”) largely determines the “spot size”; the general size range of the spot that strikes the sample. The second lens usually controlled by the “intensity or brightness knob” actually changes the size of the spot on the sample; changing it from a wide dispersed spot to a pinpoint beam. The beam is restricted by the condenser **aperture**, knocking out high angle electrons (those far from the optic axis). The beam strikes the **sample specimen** and parts of it are transmitted.

This transmitted portion is focused by the objective lens into an image. Optional Objective and Selected Area metal **apertures** can restrict the beam; the Objective aperture enhances contrast by blocking out high-angle diffracted electrons, the Selected Area aperture enables the user to examine the periodic diffraction of electrons by ordered arrangements of atoms in the sample.

The image is passed down the column through the intermediate and projector lenses, being enlarged all the way.

The image strikes the phosphor image screen and light is generated, allowing the user to see the image. The darker areas of the image represent those areas of the sample that fewer electrons were transmitted through (they are thicker or denser). The lighter areas of the image represent those areas of the sample that more electrons were transmitted through (they are thinner or less dense).

Example 25.2: An electron is accelerated through a potential difference of 182 V. Calculate its associated wavelength.

Solutions: We know that deBroglie wavelength, $\lambda = \frac{h}{p} = \frac{12.3}{\sqrt{V}} \text{ Å}$. Here $V = 182 \text{ V}$.



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$$\therefore \lambda = \frac{12.3}{\sqrt{182}} \text{ Å} = \frac{12.3}{13.5} = 0.91 \text{ Å}$$

Example 25.3: Calculate the maximum kinetic energy of the emitted photoelectrons when light of frequency $v = 10^{15}$ Hz is incident on a zinc plate. The work function of zinc is 3.4 eV.

Solution: From Einstein's relation, we recall that

$$hv = \phi_0 + K_{\max}$$

For this problem, $h = 6.625 \times 10^{-34}$ Js, $v = 10^{15}$ Hz, $E = hv = 6.625 \times 10^{-34} \times 10^{15}$ $= 6.625 \times 10^{-19}$ J and $\phi_0 = 3.4$ eV $= 3.4 \times 1.602 \times 10^{19}$ J $= 5.4468 \times 10^{-19}$ J

$$\therefore K_{\max} = E - \phi_0 = (6.625 - 5.447) \times 10^{-19} \text{ J} = 1.178 \times 10^{-19} \text{ J}$$



INTEXT QUESTIONS 25.4

1. State whether the following statement are true or false:
 - (a) According to deBroglie, stationary particles exhibit wave-like characteristics.
 - (b) Matter waves are the same thing as deBroglie waves.
 - (c) Very poor resolution can be obtained in a microscope using energetic electrons by lowering deBroglie wavelengths associated with electrons.
2. A 50 g ball rolls along a table with a speed of 20 cm s⁻¹. How large is its associated wavelength? Given $h = 6.625 \times 10^{-34}$ Js.
3. Why can we not observe de Broglie wavelength associated with a cricket ball?



WHAT YOU HAVE LEARNT

- Emission of electrons from a metal when light of proper frequency incident on its surface is called photoelectric emission.
- In photoelectric emission, electrons gain energy from light.
- The stopping potential increases with increase in frequency of incident light.
- There exists a frequency v_0 for every material below which no photoelectric effect takes place.
- The maximum velocity of photoelectrons increases with increasing frequency of incident light but is independent of the intensity of incident light.



Notes



TERMINAL EXERCISE

1. In photoelectric emission, what happens to the incident photons?
2. What is the difference between a photon and a matter particle?
3. Why is the wave nature of matter not apparent in daily life?
4. How is velocity of photoelectrons affected if the wavelength of incident light is increased?
5. The threshold frequency of a metal is 5×10^{14} Hz. Can a photon of wavelength 6000 Å emit an energetic photoelectron?
6. Does the threshold frequency for a metal depend on the incident radiations?
7. What are the various uses of photocell?
8. What was the aim of Davisson and Germer's experiment? On what principle does it depend?
9. Describe the experiment used for studying the photoelectric effect.
10. Explain the terms (a) Saturation voltage and (b) Stopping potential.
11. State the laws of photoelectric emission.
12. Describe the salient features of Einstein's theory of photoelectric effect.
13. Explain Einstein's relation: $h\nu = E_0 + K_{\max}$
14. Calculate the wavelength associated with electrons moving with a velocity $v = 1 \times 10^8$ ms⁻¹. Take mass of electron = 9.1×10^{-31} kg and $h = 6.6 \times 10^{-34}$ J.s.
15. Describe an experiment which verifies the existence of deBroglie waves.

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16. Show that the deBroglie wavelength associated with electrons accelerated through a potential V is given by the relation;

$$\lambda = \frac{12.3}{\sqrt{V}} \text{ Å}$$

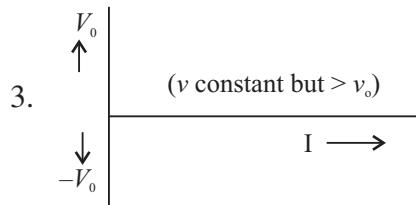


ANSWERS TO INTEXT QUESTIONS

25.1

1. (a) False (b) False (c) True
2. x – intercept gives the threshold frequency
 y – intercept gives $e \times$ work function (ϕ_0)

$$V_0 = \frac{h}{e} v - \frac{h}{e} v_0. \text{ slope of graph gives } \frac{h}{e}$$



25.2

1. $\lambda = \frac{h}{p} \Rightarrow p = \frac{h}{\lambda} = \frac{h}{c/v} = \frac{hv}{c}$
2. $E = hc/\lambda$
If λ is doubled, E will become half
3. It is unchanged.

25.3

1. (a) False (b) False (c) True
2. (i) Reproduction of sound in films,
(ii) Transmitting pictures over great distances.
(iii) Thief detecting system.
3. Number of photo electrons will increase by a factor of 4.

25.4

1. (a) false, (b) True (c) True

2. $P = mv$ and $\lambda = \frac{h}{P}$

Here $m = 50\text{g} = 0.05\text{kg}$ and $v = 20 \text{ cms}^{-1} = 0.02\text{ms}^{-1}$

$$\therefore \lambda = 6.6 \times 10^{-32}\text{m}$$

3. From Eqn.(25.14) it is clear that if mass m is large, the value of λ will be small. Same is the case with cricket ball.
4. 7.25\AA



Notes

Answers to Problems in Terminal Exercise

14. 7.25\AA

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Notes

26

NUCLEI AND RADIOACTIVITY

So far you have learnt that atom is the smallest entity that acts as the building block of all matter. It consists of an extremely small *central core*, called the *nucleus*, around which electrons revolve in certain specified orbits. Though nucleus is very tiny, it is *amazingly complex and you may like to know more about it*. The march towards our understanding the physics of nuclei began towards the end of nineteenth century with the chance discovery of the natural phenomenon of radioactivity; disintegration of atomic nuclei to attain stability. This discovery provided us tools to probe the structure of nucleus : What is its size and mass? What does it contain? What forces make its constituent particles cling together and why?

In fact, the α -particles used by Geiger and Marsden to ‘see’ what was inside an atom were obtained from naturally occurring radioactive element ^{214}Bi . These investigations opened up very fertile and new avenues of research. A lot of good new physics of the atom began to emerge out and changed the course of developments in a short span of time. You will learn about these now.



OBJECTIVES

After studying this lesson, you should be able to :

- *determine the number of neutrons and protons in nuclei of different atoms;*
- *calculate the sizes of atomic nuclei;*
- *explain the nature of forces between nucleons;*
- *explain the terms ‘mass defect’ and ‘binding energy’;*
- *draw binding energy per nucleon curve and discuss the stability of atomic nuclei;*
- *discuss the phenomenon of radioactivity, and identify the three types of radioactive radiations;*

- explain the growth and decay of radioactivity in a sample;
- calculate the half-life, and decay constant of a radioactive substance; and
- explain the uses of radioactivity in various fields.

26.1 THE ATOMIC NUCLEUS

Soon after the discovery of nucleus in an atom by Rutherford in 1911, physicists tried to study as to what resides inside the nucleus. The discovery of neutron by James Chadwick in 1932 gave an impetus to these searches as it clearly suggested to the scientific world that the building blocks of the nucleus are the protons and the neutrons.

26.1.1 Charge and Mass

The atomic nucleus contains two types of particles, **protons and neutrons**. While *protons* are *positively charged*, *neutrons* are *neutral*. The *electrons*, which revolve in certain specified orbits around the nucleus, are *negatively charged* particles. The magnitude of charge on a proton in a nucleus is exactly equal to the magnitude of charge on an electron. Further, the number of protons in a nucleus is also equal to the number of electrons so that the atom is as a whole is electrically neutral.

Neutrons and protons are collectively referred to as *nucleons*. Their combined number in a nucleus, that is the number of nucleons, is called the *mass number*. It is denoted by A . The number of protons in a nucleus (or the number of electrons in an atom) is called the *atomic number*. It is denoted by Z . The number of neutrons in a nucleus is usually denoted by $N = A - Z$. Usually $N \geq Z$. The difference ($N - Z$) increases as A increases. Note that for a lithium nucleus containing 3 protons and 4 neutrons, the atomic number Z is 3, and the mass number A is 7.

Protons are slightly lighter than neutrons and almost the entire mass of an atom is concentrated in its nucleus. The mass of a nucleus is nearly equal to the product of A and the mass of a proton (or that of a neutron). Since mass of a proton is 1.67×10^{-27} kg, and A lies between 1 and 240 for most nuclei, the masses of nuclei vary roughly between 1.67×10^{-27} kg and 4.0×10^{-25} kg.

The charge of a nucleus is equal to Ze , where e is the fundamental unit of charge (that is the magnitude of charge on an electron). You may recall that it is equal to 1.6×10^{-19} C. For naturally occurring nuclei, Z varies from 1 to 92, while for transuranic elements (i.e. the artificially produced elements), Z varies from 93 to 105.

26.1.2 Size

The sizes of atomic nuclei are usually quoted in terms of their radii. Many nuclei are nearly spherical in shape and the radius R is given approximately by the formula

$$R = r_0 A^{1/3}$$



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Here r_0 is the unit nuclear radius and its numerical value is taken as 1.2 fermi, a unit of length in honour of famous physicist Enrico Fermi. It is equal to 10^{-15} m. The radius of the lightest nucleus (hydrogen) is thus about $1.2f$, as A for hydrogen is one. The radius of the heaviest naturally occurring nucleus (uranium) is approximately $7.5f$, as $A = 238$. You may note here that since the volume of any spherical object of radius r is equal to $(4/3)\pi R^3$, the volume of a nucleus is proportional to A , the mass number.

Can you now guess the volume of a nucleus relative to that of an atom? Knowing that the sizes of the nucleus and of the atom are approximately 10^{-15} m and 10^{-10} m, respectively, the volume of an atom is roughly 10^{+5} times the volume of a nucleus. To enable you to visualise these dimensions, the volume of a nucleus relative to atom is something like the volume of a bucket of water relative to the volume of water in Bhakra Dam.

You may now also like to know the order of magnitude of the density of nuclear matter. If we consider the lightest nucleus, hydrogen, whose mass is 1.673×10^{-27} kg and the radius is 1.2×10^{-15} m, and take it to be spherical, the density can be calculated using the relation

$$d_H = \frac{M_H}{\frac{4\pi}{3} R_H^3} = \frac{1.673 \times 10^{-27} \text{ kg}}{\frac{4\pi}{3} \times (1.2 \times 10^{-15} \text{ m})^3} = 2.3 \times 10^{17} \text{ kg m}^{-3}.$$

For oxygen, $R_0 = 3 \times 10^{-15}$ m and $M_0 = 2.7 \times 10^{26}$ kg, so that

$$d_0 = 2.39 \times 10^{17} \text{ kg m}^{-3}$$

That is, the densities of hydrogen and oxygen are of the same order. You may recall that the density of water is 10^3 and density of mercury is $13.6 \times 10^3 \text{ kg m}^{-3}$. It means that nuclear matter is extremely densely packed. To give you an idea of these magnitudes, if our earth were such a densely packed mass ($= 6 \times 10^{24}$ kg), it would be a sphere of radius 184 m only. Similarly, the radius of nuclear sphere, whose mass will be equal to the mass of our sun will be 10 km!

26.1.3 Notation

The nucleus of an atom is represented by the chemical symbol of the element, with the A value as its superscript and Z value as its subscript; both on the left hand side of the chemical symbol. Thus if the chemical symbol of an element is, say, X, its nucleus is represented by ${}^A_Z X$. For example, for the nucleus of chlorine, which has 17 protons and 18 neutrons, we write ${}^{35}_{17}\text{Cl}$. Note that 35 here is mass number.

The atoms of different elements can have the same mass number, though they may have different number of protons. *Atoms having the same A value but different Z values are called Isobars.* Thus argon with $A = 40$ and $Z = 18$ is an isobar of calcium which has $A = 40$ and $Z = 20$. Note that isobars have different chemical properties since these are determined by Z . *Atoms of the same element having the same Z value but different A values are called isotopes.* Thus, chlorine with $Z = 17$ and $A = 35$, and chlorine with $Z = 17$ and $A = 37$, are isotopes of some element, chlorine. Since isotopes have same Z value, they show identical chemical properties. Note that isotopes differ in the number of neutrons in their nuclei. Atoms having the same number of neutrons in their nuclei are called the *isotones*. Thus, sodium with $A = 23$ and $Z = 11$ is an isotope of magnesium with $A = 24$ and $Z = 12$.

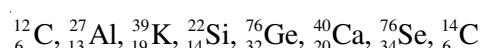
Example 26.1 : Calculate the number of electrons, protons, neutrons and nucleons in an atom of $^{238}_{92}\text{U}$.

Solution : $^{238}_{92}\text{U}$ symbolises uranium, which has 92 protons and 238 nucleus. Hence Atomic number $Z = 92$ = number of protons

Mass number $A = 238$ = number of (protons + neutrons) = Number of nucleons

$$\begin{aligned} \text{Number of neutrons} &= A - Z \\ &= 238 - 92 \\ &= 146. \end{aligned}$$

Example 26.2 : Select the pairs of Isotopes, Isobars and Isotones in the following list.



Solution : Isotopes – (Same Z - value) : $^{12}_6\text{C}$ and $^{14}_6\text{C}$

Isotones – [Same $A - Z$ values] : $\left[^{27}_{13}\text{Al} \text{ and } ^{28}_{14}\text{Si} \right], \left[^{39}_{19}\text{K} \text{ and } ^{40}_{20}\text{Ca} \right]$

Isobars – (Same A values) : $^{76}_{32}\text{Ge}$ and $^{76}_{34}\text{Se}$



INTEXT QUESTIONS 26.1

1. Make groups of Isotopes, Isobars and Isotones from the following collection of different atoms:



Notes

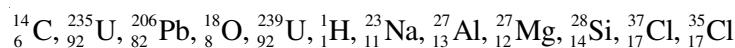
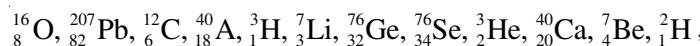
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2. Fill in the blanks :

- Neutron is than proton.
 - The total number of protons and neutrons in an atom is called the number of that atom.
 - The protons and neutrons together are called
 - The number of neutrons in $^{27}_{13} \text{Al}$ =
 - The number of protons in $^{28}_{14} \text{Si}$ =
 - Two atoms are said to belong to different elements if their numbers are different.
3. Which number cannot be different in two atoms of the same element – mass number, atomic number, neutron number?

26.1.4 Unified Atomic Mass

It has been experimentally determined that mass of proton (m_p) is 1836 times the mass of electron (m_e), and the mass of neutron (m_n) is 1840 m_e . Since the mass of an electron is negligibly small compared to the mass of a nucleon, the mass of an atom is effectively due to the mass of its nucleons. However, the neutron is slightly heavier than the proton. It is, therefore, desirable to choose a standard to express the masses of all the atoms (and also that of protons and neutrons). Now a days, atomic masses are expressed in terms of the actual mass of $^{12}_6 \text{C}$ isotope of carbon. The unit of atomic mass, abbreviated as u, is defined as $(1/12)^{\text{th}}$ of the actual mass of $^{12}_6 \text{C}$. We know that the value of the mass of a carbon atom is $1.99267 \times 10^{-26} \text{kg}$. Hence

$$\begin{aligned} 1\text{u} &= (1/12) \times \text{mass of one carbon atom with } A=12 \\ &= (1/12) \times (1.99267 \times 10^{-26} \text{kg}) \\ &= 1.660565 \times 10^{-27} \text{kg} \\ &= 1.66 \times 10^{-27} \text{ kg} \end{aligned}$$

Since mass of a proton (m_p) is $1.6723 \times 10^{-27} \text{kg}$, and mass of a neutron (m_n) is $1.6747 \times 10^{-27} \text{kg}$, we can express these in terms of u :

$$m_p = \frac{1.6723 \times 10^{-27}}{1.6606 \times 10^{-27}} \text{u} = 1.00727 \text{ u}$$

and

$$m_n = \frac{1.6747 \times 10^{-27}}{1.6606 \times 10^{-27}} \text{ u} = 1.00865 \text{ u}$$

Can you now express the mass of an electron ($m_e = 9.1 \times 10^{-31} \text{ kg}$) in terms of u? Since we will use nuclear masses in u, it is quite useful to know its energy – equivalent. To do so, we use *Einstein's mass-energy equivalence relation*, viz

$$\text{Energy} = \text{mass} \times c^2$$

where c is velocity of light in vacuum. Thus

$$\begin{aligned} 1\text{u} &= (1.66 \times 10^{-27}\text{kg}) (2.9979 \times 10^8 \text{ms}^{-1})^2 \\ &= 14.92 \times 10^{-11}\text{J} \\ &= \frac{14.92 \times 10^{-11}}{1.60 \times 10^{-13}} \text{MeV} \\ &= 931.3 \text{MeV} \end{aligned}$$

Note that joule (J) is too big a unit for use in nuclear physics. That is why we have expressed u in MeV (million electron volts). 1MeV is the energy gained by an electron when accelerated through a potential difference of one million volts. It is equal to $1.6 \times 10^{-13}\text{J}$.

26.1.5 Mass Defect and Binding Energy

The mass of the nucleus of an atom of any element is always found to be less than the sum of the masses of its constituent nucleons. This difference in mass is called **mass-defect**. For example, the nucleus of deuterium isotope of hydrogen has one proton and one neutron. The measured masses of these particles are $1.6723 \times 10^{-27}\text{kg}$ and $1.6747 \times 10^{-27}\text{kg}$, respectively. It means that total mass of a proton and a neutron is $3.34709 \times 10^{-27}\text{kg}$. But the mass of deuterium nucleus is $3.34313 \times 10^{-27}\text{kg}$. It means that the measured mass of deuterium nucleus is $3.96242 \times 10^{-30}\text{kg}$ less than the measured masses of a proton plus a neutron. So we say that mass defect in the case of deuterium is $3.96242 \times 10^{-30}\text{ kg}$. Let us denote it by Δm . Mathematically, for an atom denoted by ${}^A_Z X$, we can write

$$\begin{aligned} \text{Sum of the masses of the nucleons} &= Zm_p + (A-Z)m_n \\ \therefore \Delta m &= [Zm_p + (A-Z)m_n] - M \end{aligned} \tag{26.1}$$

where M is actual mass of nucleus.



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Energy equivalent of mass defect is obtained by using mass-energy equivalence relation:

$$BE = \Delta m c^2 \text{ joules} \quad (26.2)$$

For Deuterium

$$\begin{aligned} BE &= (3.96242 \times 10^{-30} \text{ kg}) \times (2.998 \times 10^8 \text{ ms}^{-1})^2 \\ &= 35.164 \times 10^{-14} \text{ kg m}^2 \text{ s}^{-2} \\ &= 3.5164 \times 10^{-13} \text{ J} \\ &= 2.223 \times 10^6 \text{ eV} \end{aligned}$$

since $1 \text{ eV} = 1.602 \times 10^{-19} \text{ J}$.

This means that we have to supply atleast 2.223 MeV energy to free the constituent nucleons – proton and neutron – of deuterium nucleus. You can generalise this result to say that *mass defect appears as energy which binds the nucleons together*. This is essentially used up in doing work against the forces which make the nucleons to cling.

Binding Energy per nucleon, $B = \Delta m c^2/A$

$$\text{or } B = \frac{[Zm_p + (A-Z)m_n - M]c^2}{A} \quad (26.3)$$

For $^{12}_6\text{C}$, $Z = 6$ and $A = 12$. Therefore $(A-Z) = 12 - 6 = 6$. Also $M = 12 \text{ u}$; ($1 \text{ u} = 931.3 \text{ MeV}$)

Therefore,

$$\begin{aligned} B &= \frac{[6m_p + 6m_n - 12] \times 931.3}{12} \text{ MeV} \\ &= 7.41 \text{ MeV} \end{aligned}$$

where we have used $m_p = 1.00727 \text{ u}$ and $m_n = 1.00865 \text{ u}$.

It suggests that on breaking the nucleus of carbon atom, nearly 90 MeV energy will be released, which can be used for various purposes. This is obtained in nuclear fission of a heavy atom like $^{238}_{92}\text{U}$. You will learn about it in the next lesson. This is also the source of energy in an atom bomb.

The value of B is found to increase to about 8.8 MeV as we move from helium ($A = 4$) to iron ($A = 56$); thereafter it decreases gradually and drops to about 7.6 MeV for uranium ($A = 238$). Fig.26.2 shows the variation of binding energy per nucleons with mass number.



Notes

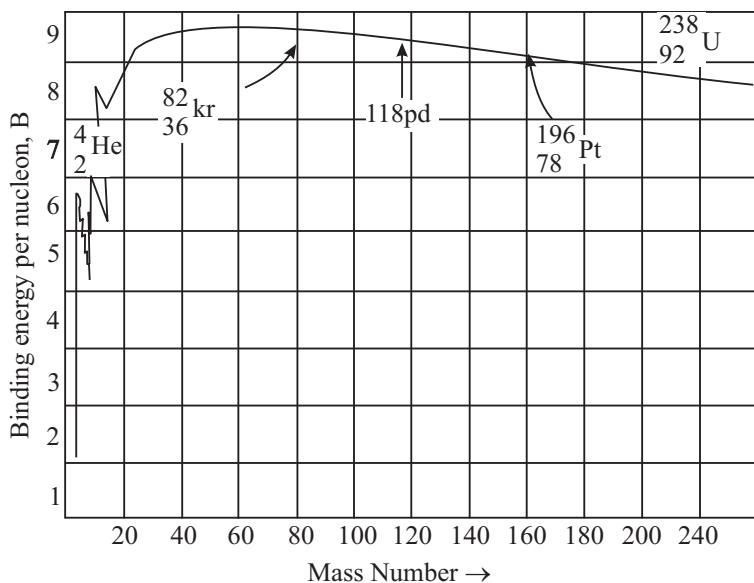


Fig. 26.2 : The variation of binding energy per nucleon with mass number

Note that binding energy curve shows sharp peaks for ${}^4\text{He}$, ${}^{12}\text{C}$, ${}^{16}\text{O}$ and ${}^{20}\text{Ne}$. Moreover, B is small indicating that light nuclei with $A < 20$ are less stable. For example, the value of B for heavy hydrogen (${}^2\text{He}$) is only 1.1 MeV per nucleon.

The subsidiary peaks occurring at ${}^2\text{He}$, ${}^{12}\text{C}$, ${}^{16}\text{O}$ (even-even nuclei i.e. nuclei having even number of protons and even number of neutrons) indicate that these nuclei are more stable than their immediate neighbours.

The binding energy per nucleon curve is very useful in explaining the phenomena of nuclear fission and nucleon fusion.

Example 26.3 : Mass of a Boron (${}^10\text{B}$) atom is 10.811 u. Calculate its mass in kg.

Solution : Since $u = 1.660565 \times 10^{-27}\text{kg}$,

$$10.811\text{u} = 10.811 \times 1.660565 \times 10^{-27}\text{kg}$$

$$= 17.952368 \times 10^{-27}\text{kg}$$



INTEXT QUESTIONS 26.2

1. The mass of the nucleus of ${}^7\text{Li}$ atom is 6.01513 u. Calculate mass defect and binding energy per nucleon. Take, $m_p = 1.00727\text{ u.}$; $m_n = 1.00865\text{ u}$ and $1\text{ u} = 931\text{ MeV}$.

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2. Calculate the radius of the nucleus of ${}^8_4\text{Be}$ atom.

[Use $R = r_0 A^{1/3}$; $r_0 = 1.2 \times 10^{-15}\text{m}$]

26.2 HOW DO NUCLEONS CLING TOGETHER : NUCLEAR FORCE

Once physicists accepted the neutron-proton hypothesis of nucleus, an important question arose : How do nucleons cling together? In other words : What is the nature of force that binds nucleons? Since gravitation and electromagnetic interactions explain most of the observed facts, you may be tempted to identify one of these forces as the likely force. However, the extremely small size of the nucleus, where protons and neutrons are closely packed, suggests that forces should be strong, short range and attractive. These attractive forces can not have electrostatic origin because electrostatic forces between protons are repulsive. And if only these were operative, the nucleons would fly away, which is contrary to experience. Moreover, the forces between nucleons are responsible for the large binding energy per nucleon (nearly 8 MeV). Let us consider the gravitational force. No doubt, it is a force of attraction between every pair of nucleons. However, it is far too weak to account for the powerful attractive forces between nucleons. If the magnitude of nucleon-nucleon force is taken to be unity, the gravitational force would be of the order of 10^{-39} . We may, therefore, conclude that the purely attractive forces between nucleons are of a new type with no analogy whatever with the forces known in the realm of classical physics. This new attractive force is called **nuclear force**.

26.2.1 Characteristic Properties

You may recall that the gravitational as well as electrostatic forces obey inverse square law. However, the nucleons are very densely packed and the nuclear force that holds the nucleons together in a nucleus must exist between the neighbouring nucleons. Therefore, nuclear force should be a short range force operating over very short distances ($\sim 10^{-15}\text{m}$).

These nuclear forces must account for the attractive force between :

- a proton and a neutron;
- two protons; and
- two neutrons.

Since binding energy per nucleon, B is the same, irrespective of the mix of neutrons and protons in the nucleus, we are quite justified in considering the force between them as equivalent. That is, nuclear force is *charge independent*.

The nuclear force shows the property of *saturation*, which means that nucleons show only limited attraction. That is, each nucleon in a nucleus interacts with only neighbouring nucleons instead of all nucleons from one end of the nucleus to the other.

If nuclear forces had only attractive character, nucleons should have coalesced under their influence. But we all know that the average separation between nucleons is constant, resulting in a nuclear volume proportional to the total number of nucleons. The possible explanation is that nuclear forces exhibit attractive character only so long as nucleons are separated through a certain critical distance. For distances less than this critical value, the character of nuclear forces changes abruptly; attraction should change to repulsion. (You should not confuse this repulsion with electrostatic repulsion.) These qualitative aspects of nuclear forces are shown in Fig. 26.3

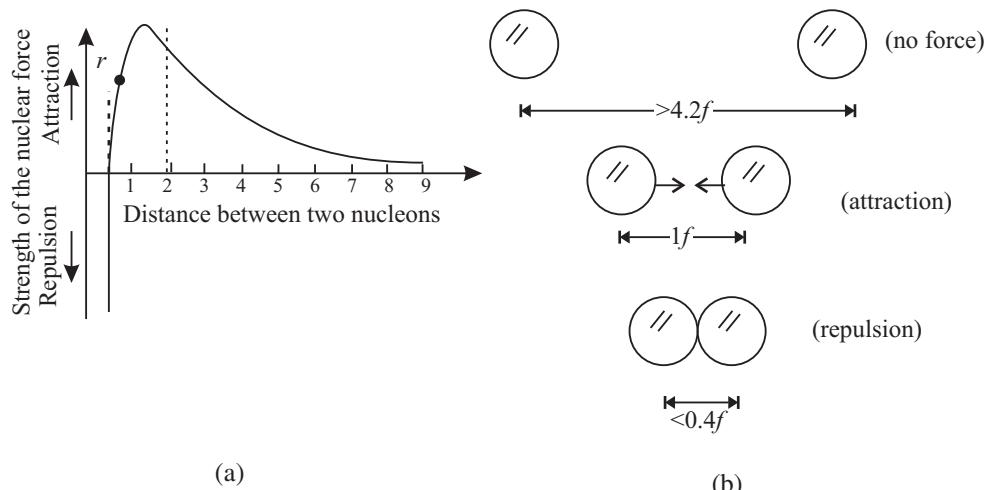


Fig. 26.3 : a) Typical variation of nuclear forces with distance, and **b)** effect of inter-nuclear distance on the force between nucleons.

26.3 RADIOACTIVITY

What is the age of our earth? How do geologists estimate the age of rocks and fossils found during excavations? What is radio-therapy which is used to treat malignant cells? The answers to all these interesting and useful questions are inherent in the study of radioactivity; a natural phenomenon in which atoms emit radiations to attain stability. Though it was discovered by chance, it opened flood gates for new physics. It finds wide use in industry, agriculture and medical care. Let us learn about it now.

26.3.1 Discovery

The story of discovery of radioactivity is very interesting. In 1896, French physicist A.H. Becquerel was working on the phenomenon of fluorescence (in which some

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substances emit visible light when they are exposed to ultra-violet radiations). In one of the drawers of his desk, he had kept a collection of various minerals, besides several unopened boxes of photographic plates. Somehow, the collection of minerals remained untouched for a considerable period of time. One day Becquerel used one of the boxes of photographic plates to photograph something. When he developed the plates, he was disappointed to find that they were badly fogged as if previously exposed to light. He tried the other boxes of photographic plates and found them also in the same poor condition. He could not understand as to why plates were fogged because all the boxes were sealed and the plates inside were wrapped with thick black paper.

Becquerel was puzzled and investigated the situation further. He found that uranium placed in his drawer had done the damage and concluded that there must be some new type of penetrating radiation originating from the uranium salt. This radiation was named *Becquerel rays* and the phenomenon of emission of this radiation was named *radioactivity*. The elements exhibiting this phenomenon were called *radioactive elements*.

Soon after this discovery, and based on an exhaustive study, Madame Marie Curie alongwith her husband Pierre Curie, isolated an element from uranium ore by a painstaking method known as chemical fractionating. This new element, which was a million times richer in the mysterious rays than uranium, was given the name radium. Another radioactive element discovered by Madam Curie was named polonium in honour of her native country-Poland.

26.3.2 Nature of Radiations

In 1899, Lord Rutherford, a British physicist, analysed the Becquerel rays emitted by radioactive elements. He established the existence of two distinct components : α -particles and β -rays. The existence of third radiation – gamma rays – was established by P. Villard.

We know that nuclei of all atoms contain positively charged protons, which repel each other strongly due to electrostatic repulsion. To overcome this repulsion, neutrons in the nuclei act as glue. But in case of heavier nuclei, this electrostatic repulsion is so strong that even the addition of neutrons is not able to keep the nuclei stable. To achieve stability, such nuclei disintegrate spontaneously by emitting α and β particles along with γ -rays as shown in Fig 26.4. So, we can say that in natural radioactivity, α , β and γ -rays are emitted.

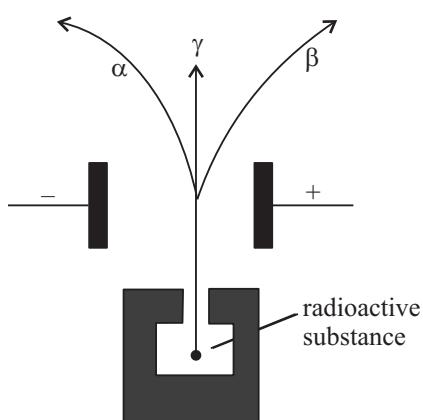


Fig. 26.4 : Emission of α , β and γ radiations

The emitted radiation is called the **radioactive radiation** and the process of disintegration (break-up) of atomic nuclei (by emitting α , β and γ -rays) is called **radioactive decay**. Sometimes, the break-up can be induced by bombarding stable nuclei with other light particles (like neutron and protons). It is then called *artificial radio-activity*.

The characteristic features of this phenomenon are that it is spontaneous and in the case of α or β emission, a new nucleus belonging to a new element is formed. That is, one element gets converted into another element. This is thus a nuclear disintegration phenomenon and suggests the possibility of mutation of new nuclei. Let us first study the characteristic properties of α , β , and γ radiations.

(i) α -particles

Alpha particles are helium nuclei (${}^4_2\text{He}$) and consist of two protons and two neutrons. Detailed studies of these particles revealed the following properties :

- Being charged particles, they get deflected in electric and magnetic fields.
- They produce fluorescence in substances like zinc sulphide and barium platinocyanide, affect a photographic plate, can induce radioactivity in certain elements and produce nuclear reactions.
- They have great ionizing power. A single particle in its journey through a gas can ionize thousands of gas atoms before being absorbed.
- They have little penetration power through solid substances, and get scattered by thin foils of metals. They can be stopped by 0.02 mm thick aluminum sheet.
- The energies of α particles emitted from a radioactive substance is a characteristic of the emitting nucleus. This corresponds to a variation in their velocity from $1.4 \times 10^7 \text{ m s}^{-1}$ to $2.05 \times 10^7 \text{ m s}^{-1}$.

(ii) β -particles

β -Particles can be both positively and negatively charged. They originate in the nucleus in the process of conversion of a neutron into a proton, and vice versa. Further studies of β -particles have revealed the following properties.

- Being charged particles, they get deflected by electric and magnetic fields.
- They produce fluorescence in materials like zinc-sulphide and barium platinocyanide; and affect photographic plates.
- They can ionize gas atoms but to a much smaller extent than the α -particles.
- Negatively charged β -particles can pass through a few mm of aluminium sheets. They are about 100 times more penetrating than α -particles.



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- Average energies of negative β -particles vary between 2 MeV and 3 MeV. Due to their small mass, their velocities vary in range from $0.33c$ to $0.988c$, where c is velocity of light.

(iii) γ -rays

γ -rays are electromagnetic waves of high frequency, and as such highly energetic. They are characterized with the following properties :

- They do not get deflected by electric or magnetic fields. They travel with velocity of light in free space.
- Their penetration power is more than that of α and β -particles; γ -rays can penetrate through several centimeters of iron and lead sheets.
- They have ionizing power that is smaller compared to that of α and β -particles.
- They can produce fluorescence in materials and affect a photographic plate.
- They knock out electrons from the metal surfaces on which they fall and heat up the surface. Hard γ -rays (i.e. high energy γ -rays) are used in radio therapy of malignant cells.

Marie Curie (1867–1934)

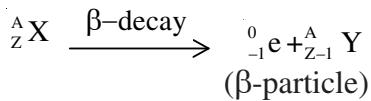
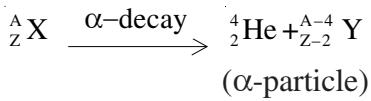


Marie Curie shared the 1903 Nobel prize in physics with A. Henri Becquerel and her husband Pierre Curie for her studies in the field of radioactivity. She was the first person in the world to receive two Nobel prizes; the other Nobel prize she received was in chemistry in 1911. Later her daughter Joliot also won the Nobel prize in chemistry for her discovery of artificial radioactivity.

26.3.3 Radioactive Decay

In any radioactive decay, spontaneous emission consists of either a α -particle or a β -particle. The emission of an α -particle from a radioactive nucleus (called *parent nucleus*) changes it into a new nucleus (new element is called *daughter nucleus*) with its atomic number decreased by two and its mass number decreased by four. Similarly, emission of a β particle changes the parent nucleus into a daughter nucleus with its atomic number increased by unity (if it is β^- emission) but its mass number remains unchanged. The emission of γ -rays does not change the atomic number or the mass number of the parent nucleus and hence no new nucleus is formed.

Note that in any nuclear disintegration, the charge number (Z) and the mass number (A) are always conserved. Thus for any radioactive nucleus, denoted by X , the nuclear transformations may be written as :



The asterisk over the symbol of element implies that it is in an excited state.

26.3.4 Law of Radioactive Decay

We now know that if we have a given amount of radioisotope, it will gradually decrease with time due to disintegrations. The law describing radioactive decay is very simple. The rate of radioactive disintegration is independent of external factors such temperature, pressure etc. and depends only on the law of chance. It states that *the number of radioactive atoms disintegrating per second is proportional to the number of radioactive atoms present at that instant of time*. This is called *law of radioactive decay*.

Let N_0 be the number of radioactive atoms, at $t = 0$, and $N(t)$ be the number of radioactive atoms at time t . If dN denotes the number of atoms that decay in time dt , then $(N - dN)$ signifies the number of radioactive atoms at time $(t + dt)$. Hence, rate of decay

$$\frac{dN(t)}{dt} \propto N,$$

or

$$\frac{dN(t)}{dt} = -\lambda N(t) \quad (26.4)$$

where λ denotes decay constant, which is characteristic of the radioactive substance undergoing decay. The negative sign signifies that the number of nuclei decreases with time. This relation can be rearranged as

$$\lambda = -\frac{1}{N(t)} \frac{dN(t)}{dt} \quad (26.5)$$

Thus, decay constant (λ) may be defined as the ratio of the instantaneous rate of disintegration to the number of radioactive atoms present at that instant.

The law of decay is sometimes also expressed in exponential form and is also called the *law of exponential decay*. To obtain the exponential form, we integrate Eq. (26.4) with respect to time :

$$N(t) = N_0 \exp(-\lambda t) \quad (26.6)$$

The most important conclusion from this law is that N will become zero only when $t = \infty$.



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We rewrite Eqn. (26.4) as

$$\frac{dN(t)}{N(t)} = -\lambda dt$$

On integration, we get
 $\ln N(t) = -\lambda t + k$.

At $t = 0$, $N(t) = N_0$
 $\therefore k = \ln N_0$

Hence

$$\ln N(t) - \ln N_0 = -\lambda t$$

$$\text{or } \ln \left(\frac{N(t)}{N_0} \right) = -\lambda t$$

On taking antilog, we obtain the required result:

$$N(t) = N_0 \exp(-\lambda t)$$

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Thus, no radioactive element will disappear completely even after a very long time.

The radioactive decay law clearly shows that even if the number of atoms N_0 for different radioactive elements is same initially, at a later time they will have different values of $N(t)$ due to different values of their decay constants (λ). They will thus show different rates of disintegration. This is determined by their half-life ($T_{1/2}$) and average lives (T_a).

Units of Disintegration

The decay constant is measured in units of per second. The **activity** of a radioactive substance at any instant of time is measured by its rate of disintegration. Its SI unit has been named becquerel :

$$1 \text{ becquerel} = 1 \text{ disintegration per second.}$$

Another unit of the decay constant is **curie**.

$$1 \text{ curie} = 3.7 \times 10^{10} \text{ disintegrations per second.}$$

which is the rate of disintegration of radium (Ra) measured per second per gram.

Yet another unit is ‘rutherford’ (rd) :

$$1 \text{ rd} = 10^6 \text{ disintegrations per second.}$$

26.3.5 Half Life ($T_{1/2}$)

The half life ($T_{1/2}$) of any radioactive element is defined as the time in which the number of parent radioactive atoms decreases to half of the initial number.

By definition, at $t = T_{1/2}$, $N = N_0/2$. Therefore, using Eqn. (26.6), we can write

$$N_0/2 = N_0 \exp(-\lambda T_{1/2})$$

$$\text{or } \lambda T_{1/2} = \log_e 2$$

$$\begin{aligned} \text{or } T_{1/2} &= \frac{\log_e 2}{\lambda} \\ &= \frac{2.303 \times \log_{10} 2}{\lambda} \\ &= \frac{2.303 \times 0.3010}{\lambda} \\ &= \frac{0.693}{\lambda} \end{aligned}$$

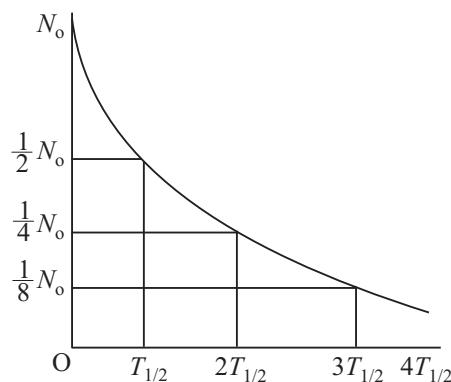


Fig. 26.5 : Radioactive decay curve

Thus, half-life of any radioactive substance is inversely proportional to its decay constant and is a characteristic property of the radioactive nucleus. The half-life of ^{14}C (radioactive carbon) is 5730 years. This means that one gram of ^{14}C will be reduced to 0.5 g in 5730 years. This number will be further reduced to $\frac{0.5}{2} = 0.25$ g in another 5730 years. i.e. in a total period of 11460 years. Refer to Fig. 26.5 to see how a radioactive sample decays with time.



Notes

Example 26.4 : An animal fossil obtained in the Mohanjodaro – excavation shows an activity of 9 decays per minute per gram of carbon. Estimate the age of the Indus Valley Civilisation. Given the activity of ^{14}C in a living specimen of similar animal is 15 decays per minute per gram, and half life of ^{14}C is 5730 years.

Solution : ^{14}C is radioactive isotope of carbon. It remains in fixed percentage in the living species. However, on death, the percentage of ^{14}C starts decreasing due to radioactive decay. Using radioactive decay law, we can write

$$N(t) = N_0 \exp(-\lambda t)$$

so that

$$N/N_0 = \exp(-\lambda t)$$

or

$$9/15 = \exp(-\lambda t)$$

or

$$\log_e(9/15) = -\lambda t$$

or

$$\log_e\left(\frac{15}{9}\right) = \lambda t$$

which gives

$$t = 1/\lambda [\log_e(15/9)]$$

Here $T_{1/2} = 0.693/\lambda = 5730$ years. Therefore,

$$t = 2.303 \times (5730/0.693) [\log_{10}15 - \log_{10}9]$$

Hence

$$t = 4224.47 \text{ years.}$$

Thus, the specimen containing carbon –14 existed 4224.47 years ago. Hence the estimated age of Indus valley civilisation is 4225 years.



INTEXT QUESTIONS 26.3

- How can you say that radioactivity is a nuclear disintegration phenomenon?
- Compare the ionizing and penetration powers of α , β and γ - radiations.

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Atoms and Nuclei



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Nuclei and Radioactivity

3. Apply the law of conservation of charge and mass numbers to determine the values of a and b in the following decay - equations :
 - (i) ${}_{Z}^{A}X = {}_{2}^{4}\text{He} + {}_{a}^{b}\text{Y} + \gamma$
 - (ii) ${}_{Z}^{A}X = {}_{-1}^{0}\text{e} + {}_{a}^{b}\text{Y} + \gamma$
4. The half-life of a radioactive substance is 5 years. In how much time, 10g of this substance will reduce to 2.5g?

Applications of Radioactivity

Radioactivity finds many applications in our every day life. Some of these are given below.

- (i) **In medicine** : In the treatment of cancer (radiotherapy), a radio-active cobalt source which emits x-rays is used to destroy cancerous cells. The decay of a single radioactive atom can be registered by an instrument placed at a remote location outside a container wall. This high sensitivity is utilized in *tracer technique* as an important tool in medical diagnostics, like the detection of ulcer in any part of the body. A few radioactive atoms of some harmless element (${}_{11}^{24}\text{Na}$) are injected into the body of a patient. Their movement can then be recorded. The affected part absorbs the radioactive atoms whose flow is, therefore, stopped and the diseased part of the body is easily located.
- (ii) **In agriculture** : By exposing the seeds to controlled γ radiation, we are able to improve the quality and yield of crops, fruits and vegetables. Radiating these before their storage helps in saving from decay.
- (iii) **In geology** : In estimating the age of old fossils. The normal activity of living carbon containing matter is found to be about 15 decays per minute for every gram of carbon. This activity arises from the small proportion of radioactive carbon –14 present in the atmosphere with the ordinary carbon –12. This isotope (${}^{14}\text{C}$) is taken by plants from the atmosphere and is present in animals that eat plants. Thus, about one part in 10^8 radioactive carbon is present in all living beings (all animals and plants). When the organism is dead, its interaction with the atmosphere (i.e. absorption, which maintains the above equilibrium) ceases and its activity begins to fall. From this, the age of the specimen can be approximately estimated. This is called *carbon-dating* and is the principle of determining the age of old fossils by archeologists.

The same technique has been used in estimating the age of earth from the measurements of relative amounts of ^{238}U and ^{206}Pb in geological specimens containing uranium ore. Assume that the specimen of ore contained only uranium and no lead at the time of birth of the earth. With the passage of time, uranium decayed into lead. The amount of lead present in any specimen will therefore indicate its age. The present age of the earth, using this method, has been estimated to be about 4 billion years.

- (iv) **In industry :** γ -radiations are used to find the flaws (or imperfections) in the inner structure of heavy machinery. For example, if there is an air bubble inside, the penetration of γ -rays will be more at that point.



Notes



WHAT YOU HAVE LEARNT

- The nucleus in an atom contains positively charged protons and uncharged neutrons.
- The number of protons inside the nucleus of an atom of any element gives the atomic number of the element.
- The sum of the number of protons and neutrons in the nucleus of an atom is called its mass number.
- The atoms having same atomic number but different mass numbers are called isotopes.
- The atoms with same mass number but different atomic numbers are called isobars.
- The atoms with same number of neutrons are called isotones.
- The nucleons inside the nucleus of every atom are bound together by strong attractive nuclear forces which are short-range and charge-independent.
- The mass of a nucleus is found to be less than the sum of the masses of its nucleons. This difference in mass is called mass-defect. It is a measure of the binding energy.
- The size (volume) of the nucleus depends on its mass number.
- The spontaneous emission of α -particle or β -particle followed by γ -emission from any nucleus is called radioactivity.

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Atoms and Nuclei



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- The α -particles have been identified as helium nuclei, while β -particles have been identified as fast moving electrons. The γ -rays are electromagnetic waves of extremely short wavelength.
- According to the law of radioactive decay, the number of radioactive atoms disintegrating per second is proportional to the number of radioactive atoms present at that instant.
- The half life of a radioactive substance is the time during which the number of radioactive atoms reduce to half of its original number.
- The law of exponential decay is $N(t) = N_0 \exp(-\lambda t)$.



TERMINAL EXERCISE

1. When does a radioactive sample disintegrate?
2. Differentiate between isotopes and isobars.
3. Explain the characteristics of binding energy per nucleon versus mass number curve.
4. What is the nature of nuclear force? Give its characteristics.
5. Explain how decay constant is related to half-life of a radioactive substance.
6. Define the following terms:
 - (i) Atomic number; (ii) Mass number; (iii) Mass defect;
 - (iv) Binding energy of nucleons; (v) Half-life; (vi) Average life;
 - (vii) Decay constant.
7. State the law of radioactive decay.
8. What is carbon dating? What is its importance?
9. Calculate the number of neutrons, protons and electrons in the following atoms.
 - (i) $^{23}_{11}\text{Na}$; (ii) ^2_1H ; (iii) $^{238}_{92}\text{U}$; (iv) $^{35}_{17}\text{Cl}$;
10. Calculate the mass defect and binding energy of nucleons for the following nuclei.
 - (i) ^4_2He ; (ii) ^7_3Li ; (iii) $^{14}_7\text{N}$;

Given, $1 \text{ u} = 1.660566 \times 10^{-27} \text{ kg} = 931 \text{ MeV}$, Mass of a proton = 1.007276 u .
 Mass of a neutron = 1.008665 u , Mass of ${}_2^4\text{He}$ atom = 4.00260 u , Mass of
 ${}_3^7\text{Li}$ atom = 7.01601 u , Mass of ${}_7^{14}\text{N}$ atom = 14.00307 u .

11. Using the present day abundance of the two main uranium isotopes and assuming that the abundance ratio could never have been greater than unity, estimate the maximum possible age of the earth's crust. Given that the present day ratio of ${}^{238}\text{U}$ and ${}^{235}\text{U}$ is $137.8 : 1$; Half life of ${}^{238}\text{U}$ is $= 4.5 \times 10^9$ year; and that of ${}^{235}\text{U}$ is 7.13×10^8 years.
12. If the activity of a radioactive sample drops to $\frac{1}{16}$ th of its initial value in 1 hour and 20 minutes, Calculate the half-life.



Notes



ANSWERS TO INTEXT QUESTIONS

26.1

1.

| Isotopes | Isobars | Isotones |
|--|---|---|
| ${}_6^{12}\text{C}$ and ${}_6^{14}\text{C}$ | ${}_32^{76}\text{Ge}$ & ${}_34^{76}\text{Se}$ | ${}_1^2\text{H}$ & ${}_2^3\text{He}$ |
| ${}_1^1\text{H}$ and ${}_1^2\text{H}$ & ${}_1^3\text{H}$ | ${}_18^{40}\text{Ar}$ & ${}_20^{40}\text{Ca}$ | ${}_6^{14}\text{C}$ & ${}_8^{18}\text{O}$ |
| ${}_8^{16}\text{O}$ & ${}_8^{18}\text{O}$ | ${}_32^{76}\text{Ge}$ & ${}_34^{76}\text{Se}$ | ${}_11^{23}\text{Na}$ & ${}_12^{24}\text{Mg}$ |
| ${}_17^{35}\text{Cl}$ & ${}_17^{37}\text{Cl}$ | ${}_1^3\text{H}$ & ${}_2^3\text{He}$ | ${}_13^{27}\text{Al}$ & ${}_14^{28}\text{Si}$ |
| ${}_82^{206}\text{Pb}$ & ${}_82^{207}\text{Pb}$ | ${}_3^7\text{Li}$ & ${}_4^7\text{Be}$ | ${}_13^{27}\text{Al}$ & ${}_14^{28}\text{Si}$ |
| ${}_92^{238}\text{U}$ & ${}_92^{239}\text{U}$ | | |

2. (i) heavier; (ii) mass; (iii) nucleons; (iv) 14; (v) 14 (vi) atomic.

3. Atomic number.

26.2

1. $\Delta m = 1.041358 \text{ u}; 969.5 \text{ MeV}$. 2. $2.4 \times 10^{-15} \text{ m}$.

26.3

1. Nuclear disintegration usually involves α or β emission which results in change of atomic and mass numbers of the parent element. With the emission of α

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and β particles, the heavier nuclei shed some of their mass resulting in comparatively lighter nuclei. Hence, it is a nuclear disintegration phenomenon.

- ## 2. Ionizing power of

$$\alpha > \beta > \gamma$$

Penetration power of

$$\alpha < \beta < \gamma$$

3. i) $a = Z - 2$ and $b = A - 4$
ii) $a = Z + 1$ and $b = A$.
 4. Two half life times are required – one for reduction from 10 to 5 grams and the other from 5 to 2.5 grams, i.e.. 10 years.

Answers to Problems in Terminal Exercise

9. (i) 12, 11, 11 (ii) 1, 1, 1 (iii) 146, 92, 921 (iv) 18, 17, 17

10. (i) 0.034, 28MeV (ii) 0.044, 37.86 MeV (iii) 0.10854,
101MeV

11. 6×10^9 years

12. 20 min



NUCLEAR FISSION AND FUSION

We all know that the sun supports life on the earth by continuously providing energy. It has been doing so for the last several billion years and will continue to do so for billions of years to come. What is the source of this huge amount of energy emitted by the sun? This question fascinated human mind always. But now we reliably know that the energy in the core of sun is produced by fusion of hydrogen nucli into helium at very high temperatures. This is also true of other stars. Imitation of these conditions in a fusion reactor is being highlighted as the ultimate source of all our energy requirements in coming years.

Similarly, you must have read about energy security and the role of nuclear energy to produce electricity in our nuclear reactors at Tarapore, Kota, Kaiga, Narora, Kalpakkam and Kakrapara. Similarly, you may have read in newspapers that on August 6, 1945, an atom bomb dropped over Hiroshima, a large city of Japan, destroyed the entire city almost completely in a span of a few seconds and lacs of lives were lost. It released an energy equivalent to that released by the explosion of a 20,000 ton TNT (tri-nitro toluene) bomb and was completely new in human history. Since then, more powerful (atomic, hydrogen and neutron) bombs have been made whose destructive power is equivalent to several Mega tons of TNT. The super powers are said to have stockpiled a large number of such bombs. The destructive power of their stock is so enormous that they can destroy the entire earth several times over. The physical process responsible for such colossal amount of energy is nuclear fission. You will now learn about these processes.



OBJECTIVES

After studying this lesson, you should be able to

- *state conservation laws for nuclear reactions;*

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Notes

Nuclear Fission and Fusion

- explain the terms nuclear chain reaction, controlled and uncontrolled fission chain reactions;
- describe working of a nuclear reactor; and
- explain the mechanism of production of energy in stars.

27.1 CHEMICAL AND NUCLEAR REACTIONS

27.1.1 Chemical Reaction

We know that all substances are made up of atoms. In lesson 26, you learnt that electrons in the outermost orbit govern the chemical properties of an element. That is, atoms combine with other atoms or molecules (a group of atoms) and rearrange their valence electrons. This is accompanied by reduction in their potential energy.

The formation of a new compound molecule due to rearrangement of valence electrons in interacting atoms and molecules with the release or absorption of energy is called a chemical reaction. In this process, the nucleus is not affected at all. Even the electrons in the inner orbits remain unaffected.

An example of a chemical reaction is the interaction of carbon atoms with oxygen molecules to produce carbon dioxide :



In this chemical reaction, 4.08 eV energy is released for each reacting carbon atom. It is called the binding energy (B.E) of CO₂ molecule. Reactions which result in release of energy are said to be *exothermic*. Chemical reactions which require energy to be supplied to be initiated are *endothermic*. For example, if 4.08 eV of energy is given to a CO₂ molecule under suitable conditions, it will break up into its constituents:



As shown in Eq. (27.1), 4.08 eV energy leaves the system to form CO₂ gas. Therefore, the mass of CO₂ molecule will be less than the total mass of C and O₂ by a mass equivalent of 4.08 eV. The loss of mass Δm can be calculated using the relation $E = mc^2$:

$$\Delta m = \frac{4.08 \times 1.602 \times 10^{-19}}{9 \times 10^{16}} = 7.26 \times 10^{-36} \text{ kg} \quad (27.3)$$

Such a small change in mass cannot be detected and we say that the mass is conserved in chemical reactions, though slight change of mass does occur.

The important points to be noted in chemical reactions are

- Energies of the order of 10 eV are involved.
- Change of mass is of the order of 10^{-35} kg, which is extremely small and we say that the mass is conserved.
- The total number of atoms of each type on the right hand side of the chemical equation is always equal to the total number of atoms of each type on the left hand side.



Notes

27.1.2 Nuclear Reactions

In nuclear reactions, the nuclei, not electrons, of the reactants interact with each other. They result in the formation of new elements. This process is also called transmutation of nuclei. From the previous lesson, you may recall that in nuclear reactions energies of the order of MeV are involved.

We know that the entire positive charge of an atom is concentrated in its nucleus, whose size is of the order of 10^{-15} m. The nucleus is surrounded by electrons revolving in certain specified orbits. These create a strong electrostatic potential barrier (also called the Coulomb barrier) as shown in Fig. 27.1. The Coulomb barrier is about 3 MeV for carbon nuclei and 20 MeV for lead nuclei. It means that a charged projectile aimed at a nucleus will experience strong repulsion by the Coulomb barrier of the target nucleus. If the kinetic energy of projectile is not large enough to penetrate the barrier, it will come back without producing any nuclear reaction. For a proton to enter a carbon nucleus and produce transmutation, its energy should be more than 3 MeV or so. It is because of the large amounts of energy involved in nuclear reactions that we do not observe these reactions in everyday life at ordinary temperatures and pressures.

The phenomenon of nuclear **transmutation** or nuclear reaction was discovered by Lord Rutherford in the year 1919. He bombarded nitrogen gas with high energy α -particles of energy 7.7 M eV obtained from a polonium source. He observed

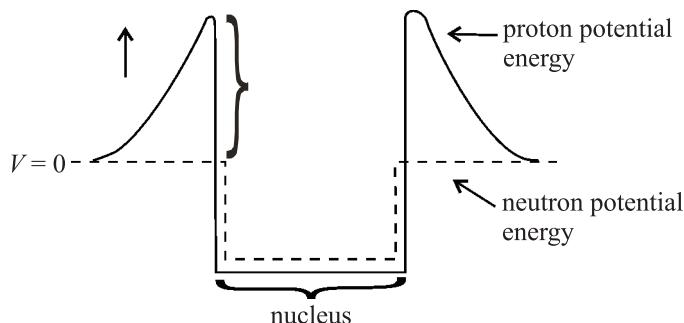


Fig. 27.1 : Proton and neutron potential energies near a nucleus

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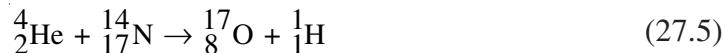
Atoms and Nuclei



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that nitrogen transformed into oxygen. This change was accompanied by high energy protons :



The oxygen nuclei and protons carry away 6.5 MeV. Clearly this reaction can occur if

1.2 MeV energy is supplied from outside. Therefore, it is an endothermic nuclear reaction. When aluminium is bombarded by 7.7 M eV alpha particles from polonium, the following nuclear reaction takes place and 10.7 MeV energy is released:



Here we see that more energy is released than the input energy; it is an exothermic reaction. Note that there is a gain of nearly 3 MeV energy per reaction, which is approximately 700,000 times the energy released in burning of one carbon atom. *But this reaction can't be used for production of energy because out of 125,000 incident alpha particles only one succeeds in producing the reaction.* Hence on the whole, there is much more energy spent than produced.

Nuclear reactions can also be produced by protons, deuterons, neutrons and other light nuclei. Of these, *neutrons are the best projectiles for producing nuclear reactions; being neutral particles, they do not experience Coulomb repulsion..* Thus even thermal neutrons (i.e. neutrons having energy 0.0253 eV) can penetrate the target nucleus and produce a nuclear reaction.

Some typical examples of nuclear reactions produced by protons, deuterons and neutrons are:



Like chemical reactions, nuclear reactions also follow conservation laws. We state these now.

27.1.3 Conservation Laws for Nuclear Reactions

- *The sum of the mass numbers of the reactants is equal to the sum of mass numbers of the products. In Eqn. (27.7), mass number 7 = 3 + 4 = 6 + 1 is conserved.*

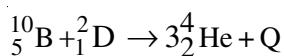
- The sum of atomic numbers of the reactants is equal to the sum of atomic numbers of the products. In Eqn. (27.7), atomic number $4 = 3 + 1 = 2 + 2$ is conserved.
- Nuclear reactions follow the law of conservation of energy. We know that mass is concentrated form of energy. Therefore the sum of input kinetic energy plus the mass of the reactants is equal to the output kinetic energy plus the mass of the products.
- Nuclear reactions follow the law of conservation of momentum, which results in distribution of kinetic energy among various product nuclei.

Now, answer the following questions.



INTEXT QUESTIONS 27.1

- Complete the following equations of nuclear reaction.
 - ${}_{9}^{19}\text{F} + {}_{1}^1\text{H} \rightarrow {}_{8}^{16}\text{O} + ?$
 - ${}_{13}^{27}\text{Al} + {}_{0}^1\text{n} \rightarrow ? + {}_{2}^4\text{He}$
 - ${}_{90}^{234}\text{Th} \rightarrow {}_{91}^{234}\text{Pa} + ?$
 - ${}_{29}^{63}\text{Cu} + {}_{1}^2\text{D} \rightarrow {}_{30}^{64}\text{Zn} + ?$
- Calculate the energy released in the nuclear reaction given below



Given that $m({}^{10}\text{B}) = 10.01294$ u; $m({}_{1}^2\text{D}) = 2.014103$ u, and $m({}_{2}^4\text{He}) = 4.002604$ u.

- ${}_{7}^{14}\text{N}$ nucleus, on bombarding with alpha particles, produces ${}_{8}^{17}\text{O}$. Write down the reaction equation and calculate the energy released.

Given that: $m({}_{7}^{14}\text{N}) = 14.003014$ u; $m({}_{8}^{17}\text{O}) = 16.999138$ u; $m({}_{2}^4\text{He}) = 4.002604$ u; $m({}_{1}^1\text{H}) = 1.007825$ u and energy of α particle = 7.7 MeV.

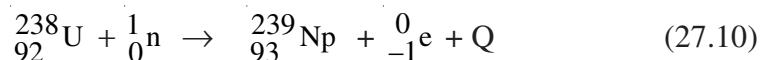


Notes



27.2 NUCLEAR FISSION

The story of discovery of fission is very fascinating. In the year 1938, Enrico Fermi, Otto Hahn and others irradiated uranium nuclei with slow neutrons to produce transuranic elements (having Z greater than 92), which do not occur in nature. When incident neutrons were captured by the uranium nuclei, the neutron-proton ratio increased. In reducing this ratio, it was expected that uranium would become β -active. That is a neutron would essentially behave as if it has changed into a proton resulting in the release of a β -particle and some energy according to the equation:



In this process, a new transuranic element having atomic number 93 was expected to be produced. In fact, Fermi and his co-researchers observed β -activities with half-lives different from any of the known values for heavy elements in the vicinity of uranium. From those observations, they concluded that transuranic elements had been produced. And to identify the element, they carried out chemical analysis but failed.

In the same year, Otto Hahn and Fritz Strausmann carried out a series of experiments and established that barium, an element of intermediate mass number, rather than a transuranic element, was one of the products of the reaction and it was accompanied by release of nearly 200 MeV of energy. This result – the product of slow neutron bombardment of uranium was barium – was completely unexpected and defied all knowledge of nuclear physics of that time. These findings were reported in *Nature* in Dec. 1938.

Initially, Lise Meitner and Otto Frisch explained these results on the basis of liquid drop model of nucleus and named this process *nuclear fission* using the analogy with biological cell division. Later on, Bohr and Wheeler calculated the amount of energy released in the process, confirming the physical basis of this model.



**Enrico Fermi
(1901 – 1954)**

Enrico Fermi, the Italy born physicist, was responsible for peaceful uses of nuclear energy for mankind. He demonstrated that nuclear transformations may occur in any element exposed to stream of neutrons. He achieved self-sustained nuclear fission chain reaction in 1942.

Fermi was only 25 years old when he formulated the Fermi–Dirac statistics, applicable to particles having half integral spin values (called fermions). At the time of his premature death, he was engrossed in theoretical studies of cosmic radiations.

27.2.1 Mechanism of Nuclear Fission

In the year 1939, Bohr and Wheeler developed the theory of fission using the analogy between nuclear forces and the forces which bind molecules in a liquid. They predicted that $^{235}_{92}\text{U}$ was more fissile than $^{235}_{92}\text{U}$. Refer to Fig. 27.2. It shows the schematics of nuclear fission of $^{235}_{92}\text{U}$ by thermal neutrons according to the equation.

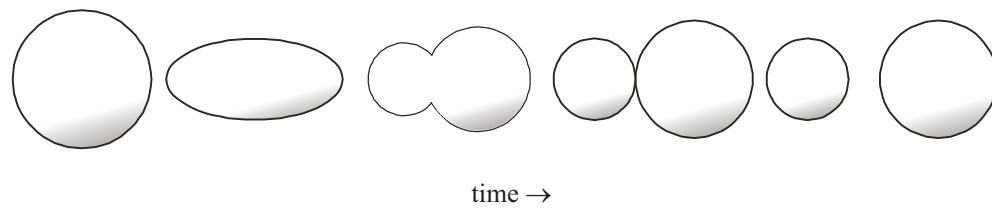
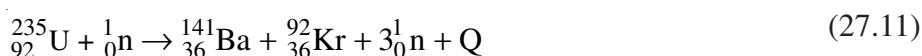


Fig. 27.2 : Nuclear-fission of a nucleus according to the liquid drop model

The emitted neutrons have energy of the order of a few MeV, and $Q \approx 200\text{MeV}$.

Note that a fission event occurs within 10^{-17}s of neutron capture and fission neutrons are emitted within about 10^{-14}s of the event. Moreover, the fission fragments are of unequal mass; one being 1.5 to 2 times heavier than the other. Also, Eqn. (27.11) gives only one of the more than 40 different modes in which a $^{235}_{92}\text{U}$ nucleus can fission. It means that about 80 different nuclei of intermediate masses are produced in the fission of $^{235}_{92}\text{U}$. The heavier fragments lie in the mass range 125–150 with a maximum around 140, whereas the lighter fragments lie in the range 80 – 110 with a maximum around 95. The number of neutrons emitted is either two or three and the average number of neutrons produced per fission of ^{235}U is 2.54

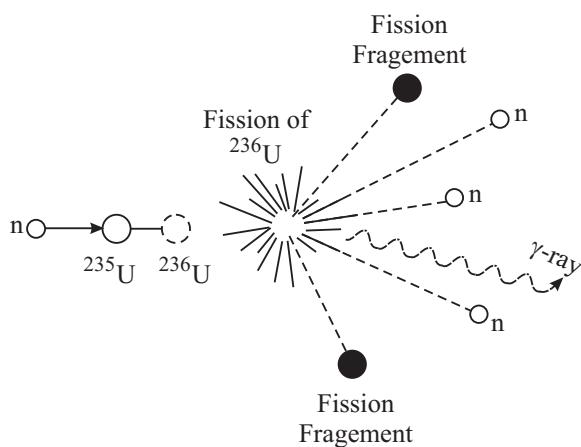


Fig. 27.3 : Nuclear fission

Notes



MODULE - 7

Atoms and Nuclei



Notes

Nuclear Fission and Fusion

Bohr and Wheeler treated the nucleus as a charged spherically symmetric liquid drop in its equilibrium (lowest energy) state. According to them, when a nucleus captures a thermal neutron, the binding energy (BE) of this neutron, which is 6.8 MeV per atomic mass unit for ^{235}U , is released. This energy excites the nucleus and distorts its shape. While the force of surface tension tries to restore the original shape, the Coulomb force tends to distort it further. As a result, it oscillates between spherical and dumb bell shapes, as shown in Fig.27.2, depending on the energy of excitation. When the energy gained by the nucleus is large, the amplitude of these oscillations pushes the nucleus into dumb bell shape. When the distance between the two charge centres exceeds a critical value, electrostatic repulsion between them overcomes nuclear surface tension and pushes the nucleus into two parts resulting in fission.

A substance like ^{235}U which undergoes fission by thermal neutrons is called a fissile material. Other fissile materials are ^{233}Th , ^{233}U and ^{239}Pu . You may note that all these nuclei have odd mass number and even atomic number.

We can estimate the amount of energy released in the fission of ^{235}U by calculating the mass defect as follows:

Table 27.1 Energy Generated in a Nuclear Reaction

| Reactants | Mass | Products | Mass |
|-------------------|--|------------------------|--------------|
| ^{235}U | 235.0439 u | $^{141}_{56}\text{Ba}$ | 140.9139 u |
| ^1n | 1.008665 u | $^{92}_{36}\text{Kr}$ | 91.8973 u |
| | | $3 \times Vn$ | 3.025995 u |
| Total mass | 236.052565 u | Total mass | 235.837195 u |
| Mass defect | 0.21537u | | |
| Energy released | $0.21537 \times 931 \approx 200 \text{ MeV}$ | | |

27.2.2 Nuclear Chain Reaction

You have now learnt that when a neutron is captured by ^{235}U , it splits into two fragments and 2-3 neutrons are emitted. These are capable of causing further fissions. This immediately presented the exciting possibility of maintaining a fission chain reaction in which each fission event removes one neutron and replaces that by more than two. When

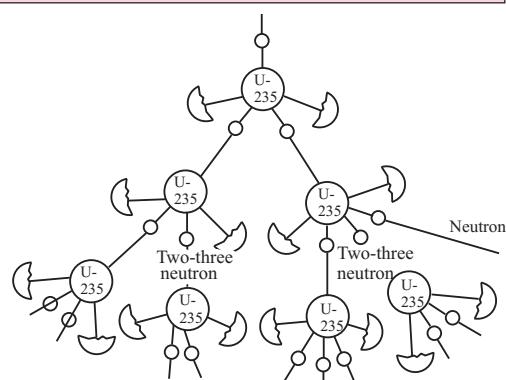


Fig. 27.4 : Nuclear Chain reaction



Notes

the rate of production of neutrons equals the rate of loss of neutrons, the reaction is said to be *self-sustained*. The device designed to maintain a self-sustained and controlled chain reaction is called a *nuclear reactor*.

Nuclear reactors are usually classified according to the purpose for which they are used. So a nuclear power reactor is used to produce electricity and a research reactor is used to produce radioisotopes for medical purposes, carrying out experiments for refinements or applied research. We also categorise nuclear reactors as fast and thermal, depending on the energy of neutrons causing fission. In India, we have thermal power reactors at Tarapore, Narora, Kota, Kaiga, etc. At Kalpakkam, we are developing a fast breeder research reactor.

You will now learn about a nuclear reactor in brief.

27.3 NUCLEAR REACTOR

Ever since the first nuclear reactor was constructed by Fermi and his co-workers at the university of Chicago USA, a large number of reactors have been built the world over primarily to meet demand for energy. Some countries generate as much as 70% of their total energy from nuclear reactors. In India, the contributions of nuclear energy is only about 2%, but efforts are on to increase this share. In absolute terms, we are generating about 20,000 MWe from nuclear reactors.

Nuclear reactors have huge complex structures and great care has to be exercised in designing them. The basic principle of a nuclear power plant is very simple and analogous to any power plant. The heat liberated in fission is used to produce steam at high pressure and high temperature by circulating a coolant, say water, around the fuel. (In a coal fired station, coal is burnt to produce steam. Since one fission event generates about 7×10^5 times more energy than that produced in burning one atom of carbon, we can cut down on emission of greenhouse gases substantially by switching over to nuclear energy. However, there are some complex social and political issues with global dimensions that will ultimately decide our ultimate nuclear energy options.)

The steam runs a turbine–generator system to produce electricity. (In research reactors, the heat is discharged into a river or sea. You many have heard about Bhahha Atomic Research Centre at Trombay, Mumbai or Indira Gandhi Atomic Research Centre at Kalpakkam. The heat generated by the research reactors at these centres is discharged into the Arabian sea and the Bay of Bengal, respectively.)

The general features of a reactor are illustrated in Fig. 27.5. All nuclear reactors consist of:

- A *reactor core*, where fission takes place resulting in release of energy. It has fuel rods (embedded in a moderator in a thermal reactors), and *control rods* to maintain the chain reaction at the desired level. *Coolant* is circulated to remove the heat generated in fission. Usually, heavy water or ordinary water are used as coolants and cadmium or boron are used for control rods.

MODULE - 7

Atoms and Nuclei



Notes

Nuclear Fission and Fusion

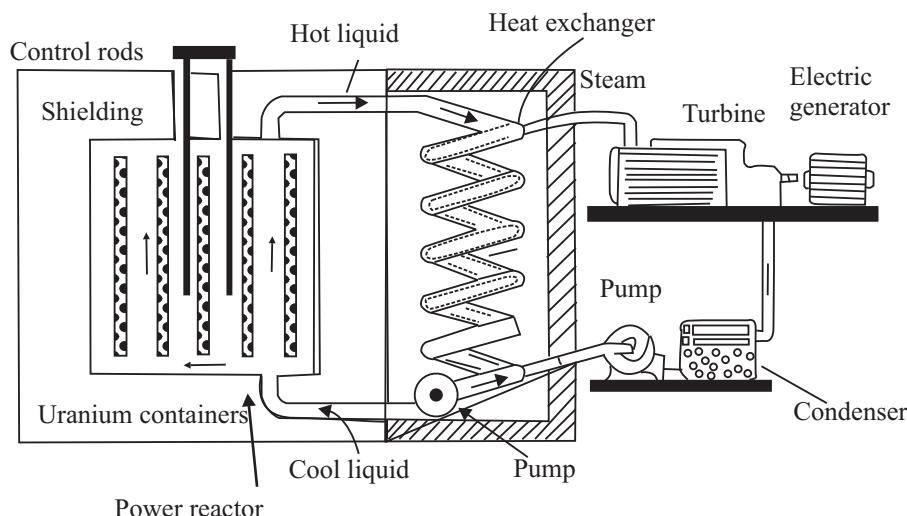


Fig. 27.5 : Schematic diagram of a nuclear reactor

- A *reflector* is put next to the core to stop neutron leakage from the core.
- The whole assembly is placed inside a vessel, called *pressure vessel*. Usually, a few inches thick stainless steel is used for this purpose.
- A thick *shield* is provided to protect the scientists and other personnel working around the reactor from radiations coming from the reactor core. It is usually in the form of a thick concrete wall.
- The entire structure is placed inside a *reactor building*. It is air tight and is maintained at a pressure slightly less than the atmospheric pressure so that no air leaks out of the building.

The heat generated inside the reactor core of a reactor due to fission is removed by circulating a *coolant*. The heated coolant is made to give up its heat to a secondary fluid, usually water in a heat exchanger. This generates steam, which is used to drive turbine-generator system to produce electricity in a power plant and discharged into a river/lake/sea in a research reactor.

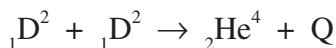


INTEXT QUESTIONS 27.2

1. Why does a $^{238}_{92}\text{U}$ nucleus become β -active after absorbing a neutron?
2. Out of $^{238}_{92}\text{U}$, ^{141}Ba , ^{239}Pu , and $^{12}_6\text{C}$, which nucleus is fissile?
3. How much energy is released when $^{235}_{92}\text{U}$ undergoes nuclear fission?

27.4 NUCLEAR FUSION

You now know that uranium nucleus can be made to split into lighter nuclei resulting in release of huge amount of energy. You may now ask: Can we combine lighter nuclei to produce energy? To discover answer to this question, refer to the binding energy per nucleon (BE/A) curve (Fig.26.2). You will note that binding energy per nucleon increases as we go from hydrogen to helium. It means that helium is more stable than hydrogen. Consider the following reaction:



You can easily calculate the B.E of reactants and products:

$$\text{Total B.E of reactants, } BE_1 = 2 \times 2.22 = 4.44 \text{ MeV}$$

$$\text{Total B.E of products, } BE_2 = 28.295 \text{ MeV}$$

$$Q = (BE_2 - BE_1) \approx 24 \text{ MeV}$$

∴ Note that the energy released per nucleon in this reaction is $24/4=6$ MeV, which is nearly seven times the energy released per nucleon ($200/238 = 0.83$ MeV) in a nuclear fission event.

The process in which two light nuclei combine to form a heavier nucleus is called nuclear fusion.

Fusion process presents itself as a more viable energy option. However, the process of fusion is more difficult to achieve than nuclear fission because both the deuterons are positively charged. When we try to bring them together to fuse into one nucleus, they repel each other very strongly and the reaction is ordinarily impossible.

To achieve this reaction, the deuterons have to be heated to nearly 10 million kelvin so that they acquire sufficient kinetic energy to overcome repulsion before they collide to fuse into helium nucleus. But the problems associated with maintaining such high temperatures continuously and containing the reactants together has not yet been solved fully. The controlled thermonuclear reaction necessary for harnessing this source of energy is however not far now.

Almost inexhaustible amount of deuterium (heavy hydrogen) is present in the ocean. Once we begin to harness this source, our energy problem should be solved for ever. We will get an endless supply of cheap electricity without any pollution. This is because one gram of deuterium (heavy hydrogen) yields about 100,000 kW h of energy.

27.4.1 Energy in the Sun and Stars

The stars like our sun are very massive objects. They have been continuously emitting tremendous amount of energy for the last billions of years.



Notes

Table 27.2 : Binding Energy per nucleon (BE/A) of some light nuclei

| Nucleon | BE/A(in MeV) |
|-----------------|--------------|
| ${}^2\text{D}$ | 1.11 |
| ${}^3\text{T}$ | 2.827 |
| ${}^3\text{He}$ | 2.573 |
| ${}^4\text{He}$ | 7.074 |
| ${}^6\text{Li}$ | 5.332 |
| ${}^7\text{Li}$ | 6.541 |

MODULE - 7

Atoms and Nuclei



Notes

Nuclear Fission and Fusion

Such a huge amount of energy cannot be obtained by burning conventional fuels like coal. Nuclear fission can also not be the source of this energy, because heavy elements do not exist in the sun in large quantity. The sun mainly consists of hydrogen and helium gases. Then you may like to know: What is the source of energy in the sun? This question has engaged human intellect for long. As a child, you must have gazed the sky when you learnt the rhyme: Twinkle twinkle little star, How I wonder what you are!

You may know that the huge mass of the sun produces extremely strong gravitational field, which compresses its constituent gases by enormous pressure resulting in the rise of temperature to millions of kelvin at its centre. It has been estimated that the temperature at the centre of the sun is 20 million kelvin. At such high temperatures and pressures, gas molecules travel at high speeds and collide setting in thermonuclear reaction and resulting in the release of large amount of energy.

Bethe proposed that fusion of hydrogen into helium is responsible for the energy produced in stars:

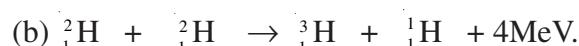


The overall result here is: four hydrogen nuclei fuse into a helium nucleus with the release of two positrons (electron-like microscope particles of the same mass but positive charge) and 26.8 MeV energy. ***The tremendous amount of energy released in a thermo-nuclear reaction is the source of energy in stars.*** The quantity of hydrogen in the sun is sufficient to keep it shining for nearly 8 billion years more.



INTEXT QUESTIONS 27.3

1. 200 MeV energy is released in fission of one ${}_{92}^{235}\text{U}$ nucleus and 26.8 MeV energy is released in fusion of 4 protons. Which process releases more energy per unit mass?
2. Consider the following reactions:



Calculate Q in the first reaction and mass of tritium in the second reaction.

Given $m({}_{1}^2\text{H}) = 2.014103\text{u}$, $m({}_{2}^4\text{He}) = 4.002604\text{u}$, $m({}_{1}^1\text{H}) = 1.007825\text{u}$ and

$m({}_{3}^7\text{Li}) = 7.015982\text{u}$.

27.5 NUCLEAR ENERGY

We need energy for all economic activities in life. The amount of energy consumed per capita is a measure of advancement of a nation. According to a recent UNESCO report (2007), we are consuming about 40% more than what mother earth can generate in the form of food, water and energy. In fact, the human society has been continuously striving for energy security and looking for newer sources of energy. Due to over use, conventional sources of energy are depleting very fast and may exhaust completely in the next one hundred years. The nuclear energy is perhaps an important option for meeting our future energy needs through peaceful applications. Let us discuss these now.

27.5.1 Peaceful Applications

The most important peaceful application of nuclear energy is in the *generation of electricity*. One of the main advantages of nuclear power plant is that the fuel is not required to be fed into it continuously like the gas or coal in a thermal power plant. Further, it does not *pollute the environment to the extent discharge of smoke or ash from fossil fuel/power plants do*. The fuel once loaded in a reactor runs for nearly 6 months at a stretch. Because of this nuclear power plants have been used to power huge ships and submarines.

However, spent fuel of a reactor is highly radioactive because a large number of radio-isotopes are present in it. India has developed its own facility to treat spent fuel and extract it from those *radio-isotopes which find uses in agriculture, medicine, industry and research*. To avoid the spread of radioactive radiations from the radioactive wastes, the radioactive wastes are generally embedded deep inside salt mines in heavy steel cases. Yet, it has evoked considerable controversy due to its destructive potential which was displayed on August 6, 1945, when an atom bomb was dropped on Hiroshima (Japan) killed hundred thousand people in a very short time. Subsequently, even more powerful hydrogen and nitrogen bombs have been developed. These can destroy this beautiful planet many times over.



Notes

Nuclear Power in India

The possibility of harnessing nuclear power for civil use was recognised by Dr H.J. Bhabha soon after India got independence. He outlined a three stage development plan for meeting country's nuclear power needs. These are :

- Employ pressurised Heavy Water Reaction (PHWR) fuelled by natural uranium to generate electricity and produce plutonium as a by-product.
- Set up fast breeder reactors burning the plutonium to breed U-233 from thorium.
- Develop the second stage and produce a surplus of fissile material.

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Atoms and Nuclei



Notes

Nuclear Fission and Fusion

Nuclear power has been produced in India through 14 small and one mid-sized nuclear power reactors in commercial operation, eight under construction and more planned. As of now, nuclear power contributes nearly 2×10^{10} kW h of electricity – 3% of total power capacity available.

Government policy is to have 20 GWe of nuclear capacity operating by 2020 and 25% nuclear contribution is foreseen by 2050.



INTEXT QUESTIONS 27.4

1. What type of reactors are used in India for power generation?
2. How much $^{235}_{92}\text{U}$ undergoes fission in an atomic bomb which releases energy equivalent to 20,000 tons of TNT. (Given that 1 g of TNT gives out 1000 calorie of heat).

27.5.2 Hazards of Nuclear Radiations and Safety Measures

The living and non-living things around us constitute our environment. In this environment, a delicate balance has existed for millions of years between the flora, fauna, aquatic and human life. This balance is now being threatened. One of the factors disturbing this balance is the ever increasing pollution in our environment. Out of the various types of pollutants present in our environment, the one which has very serious long term biological effects are the ‘nuclear radiations’. Earlier these were present only because of natural sources like the radioactive minerals and cosmic rays, but now their presence is increasing day by day due to man-made sources. The major present day man-made sources of nuclear radiations are the nuclear tests, nuclear installations like the nuclear research facilities, nuclear reactors, and radio isotopes in treating diseases.

Nuclear radiations dissociate complex molecules of living tissues through ionisation and kill the cells. They induce cancerous growth, cause sterility, severe skin burns, and lower the body resistance against diseases. They disrupt the genetic process, mainly in the unborn child, and show their effects even upto five generations. Nuclear radiations affect us not only directly, but also indirectly by affecting the flora, fauna and the aquatic life around us. They kill vegetation, fishes and animals.

The damage caused by nuclear radiations depends on the exposed part of the body, as well as on the energy, intensity and the nature of the radiation. Different parts of human body show different sensitivities to radiation. The α -particles are, as a rule, quite harmful because of their high ionising power. The damaging effects of different radiations are generally compared in terms of their ‘relative biological effectiveness’, called the RBE factors. These factors for different particles/rays are given in Table 27.3.

Table 27.3: RBE factors of different radiations

| Particles/rays | RBE factors |
|---|-------------|
| X-rays, γ -rays, β -particles | 1 |
| Thermal neutrons | 2 to 5 |
| Fast neutrons | 10 |
| α -particles, high energy ions of O, N, etc. | 10 to 20 |



Notes



WHAT YOU HAVE LEARNT

- Valence electrons take part in chemical reactions and the energy involved in such reactions is of the order of 1eV.
- In a nuclear reaction, the atomic nuclei interact to form a new element.
- Energy involved in nuclear reaction is of the order of MeV.
- In a nuclear reaction, atomic number, mass number and charge are conserved.
- When a heavy nucleus like uranium is bombarded by slow neutrons, it splits into two fragments with release of 2-3 neutrons and 200 MeV energy. This process is known as nuclear fission.
- Substances that undergo fission are called fissile substances. ^{233}Th , ^{233}U , ^{235}U , and ^{239}Pu are fissile materials.
- Chain reaction occurs when more than one emitted neutron induces further fission for each primary fission.
- Nuclear reactor is a device to sustain controlled chain reaction.
- In nuclear fusion two light nuclei are fused into one.
- For producing nuclear fusion, the reacting nuclei must be heated to nearly 20 million kelvin to gain sufficient kinetic energy to overcome the Coulombian potential barrier.
- In stars energy is produced by nuclear fusion reaction.
- Amount of hydrogen consumed in the sun is nearly 400×10^6 ton per second.
- Radio-isotopes find diverse applications in agriculture, medicine and industry.



TERMINAL EXERCISE

1. How does a nuclear reaction differ from a chemical reaction?
2. What is the use of moderator and absorber in a fission reactor?

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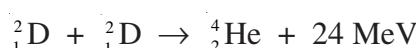
Atoms and Nuclei



Notes

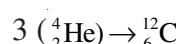
Nuclear Fission and Fusion

3. On the basis of B.E per nucleon versus mass number curve, explain nuclear fusion.
4. What is a nuclear reaction? State the conservation laws obeyed in nuclear reactions. Give three examples of nuclear reactions.
5. What is nuclear fission? Give an example to illustrate your answer.
6. Calculate the mass of ^{235}U consumed to generate 100 mega watts of power for 30 days.
7. Heavy hydrogen undergoes the following fusion reaction



Calculate the amount of heavy hydrogen used in producing the same energy as above. Compare the two results.

8. What is nuclear fusion? Write an equation of nuclear fusion to support your answer.
9. What is the source of energy in the sun? How is it generated? Illustrate with an example.
10. Describe the construction of an atomic reactor.
11. Calculate the energy released in a fusion reaction



Given, the mass of one α -particle = 4.00263u.



ANSWERS TO INTEXT QUESTIONS

27.1

1. a. $^{19}_9\text{F} + ^1_1\text{H} \rightarrow ^{16}_8\text{O} + ^4_2\text{He};$
b. $^{27}_{13}\text{Al} + ^1_0\text{n} \rightarrow ^{24}_{11}\text{Na} + ^4_2\text{He};$
c. $^{234}_{90}\text{Th} \rightarrow ^{234}_{90}\text{Pa} + ^0_{-1}\text{e};$
d. $^{63}_{29}\text{Cu} + ^2_1\text{D} \rightarrow ^{64}_{30}\text{Zn} + ^1_0\text{n}$
2. 17.9MeV
3. $^{14}_7\text{N} + ^4_2\text{He} \rightarrow ^{17}_8\text{O} + ^1_1\text{H} + 6.5\text{MeV.}$

27.2

1. Due to increase of n/p ratio above the natural ratio, its stability decreases. To decrease the ratio to attain more stability, it emits a β -particle.
2. ^{239}Pu
3. 200 MeV.

27.3

1. (1) In fission the energy released is 0.84 MeV/u where as in fusion. It is 6.7 MeV/u. Thus energy released per unit mass is more in the later case.
2. (a) 17.3 MeV, (b) 2.69 MeV.

27.4

1. Pressurized Heavy Water Reactor
2. nearly 1 kg.

Answers to Problems in Terminal Exercise

6. 30.6 kg
7. 146.6 g
11. 7.35 MeV

**Notes**

SENIOR SECONDARY COURSE

ATOMS AND NUCLEI

STUDENT'S ASSIGNMENT – 7

Maximum Marks: 50

Time : 1½ Hours

INSTRUCTIONS

- Answer All the questions on a separate sheet of paper
- Give the following information on your answer sheet:
 - Name
 - Enrolment Number
 - Subject
 - Assignment Number
 - Address
- Get your assignment checked by the subject teacher at your study centre so that you get positive feedback about your performance.

Do not send your assignment to NIOS

1. What is the ratio of the energies of first and second orbits of hydrogen atom? (1)
2. Express 1 eV in J. (1)
3. Express 1 u in kg. (1)
4. Why is the wave nature of matter not apparent to our daily observations? (1)
5. What happens to the average life of the radioactive sample when its mass decreases? (1)
6. What is the use of moderator in a nuclear fission reaction? (1)
7. How does the velocity of photoelectrons change when the velocity wavelength of incident radiation is increased? (1)
8. Distinguish between isotopes and isobars. (1)
9. Obtain the de Broglie wavelength associated with an electron accelerated through a potential difference of 200 V. (2)
10. With the help of examples distinguish between chemical and nuclear reactions. (2)
11. Write the postulates of Bohr's theory of hydrogen atom. (2)
12. Show that the density of a nucleus is independent of its mass number. (2)
13. Distinguish between half life and average life of a radioactive substance. Find the half life of a radioactive element which reduces to $\frac{1}{8}$ th of its initial mass in 16 days. (4)
14. "Specific charge (e/m) is more important physical quantity as compared to charge (e) or mass (m) of a fundamental particle." – Give examples in support of this statement. (4)

15. Write nuclear equations for :

- (a) the alpha decay of $^{286}_{88}\text{Ra}$
 - (b) the β^- – decay of $^{15}_{32}\text{P}$
 - (c) the β^+ – decay of $^{11}_{6}\text{C}$
 - (d) the γ – decay of $^{60}_{27}\text{Co}$
- (4)

16. Obtain the binding energy per nucleon of $^{11}_7\text{N}$ nucleus. Give

$$m_p = 1.00783 \text{ u}$$

$$m_n = 1.00867 \text{ u}$$

$$m_N = 14.00307 \text{ u.}$$

17. For scattering of α -particles by an atom of atomic number π , the relation between impact parameter b and the scattering angle θ is given by

$$b = \frac{Ze^2 \cot \theta / 2}{4\pi\epsilon_0(mv^2 / 2)}.$$

- (a) What is the value of scattering angle for which $b = 0$?
 - (b) Why is it that the mass of the nucleus does not enter the formula but the charge does?
 - (c) For a given value of b , does the angle of deflection increase or decrease with increasing energy?
 - (d) For a given energy of the α -particle, does the scattering angle increase or decrease with decrease in impact parameter?
- (4)

18. The total energy of an electron in the ground state of the hydrogen atom is about (- 13.eV). Find:

- (a) What is the kinetic energy of the electron in its first excited state?
- (b) What is the potential energy of the electron in its first excited state?
- (c) What is the total energy of the electron in its first excited state?
- (d) Which of the above answers would change if the choice of zero potential energy is altered?

19. In an experimental study of photoelectric effect the values of stopping potentials for various wavelengths were obtained as under :
(5)

| | | | | | |
|--------------------|------|------|------|------|------|
| $\lambda(A^\circ)$ | 3650 | 4358 | 5461 | 6907 | 8400 |
| $V_s (\text{V})$ | 1.28 | 0.95 | 0.74 | 0.16 | 0 |

(a) Determine the value of planck's constant h .

(b) Estimate the threshold frequency and work function for the material.

20. Derive a relation for the de Broglie wavelength associated with an electron accelerated in a potential field of V volts.

Which part of em spectrum you find radiations having wavelengths comparable to the de Broglie wavelength associated with a 100 eV electron.
(5)

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SEMICONDUCTORS AND SEMICONDUCTING DEVICES

Ever since man moved out of the cave and settled into a civil society, his quest for comfort has increased continuously. The invention of fire and wheel proved turning points in human history. Probably, the next big development was the grey revolution, which transformed the way of communication, transportation and living. Sitting in our living rooms, we can connect to our loved ones face-to-face across oceans and continents using computer mediated video-conferencing.

To make all this possible solid state electronic devices have played a significant role. Electronics is a branch of science and technology in which electrons are manipulated to do some specific tasks. Scientists have studied the electrical nature of materials and developed a concept of **energy bands** in terms of which solids can be classified as conductors, semiconductors and insulators. Semiconductors are mostly used for developing electronic devices. Silicon and germanium are the most familiar semiconductor materials. Normally, the conductivity of a semiconductor lies in-between the conductivities of metals and insulators. However, at absolute zero, the semiconductor also acts like a perfect insulator. The conductivity of a semiconductor is influenced by adding some impurity element called *dopant*.

In this lesson you will learn about various types of semiconductors, their behaviour and how they are combined to form useful devices such as Zener diode, solar cell, photodiode, light emitting diode and transistor, etc. You will also learn to draw the I-V characteristics of Zener diode, LED, photo diode and solar cell.



OBJECTIVES

After studying this lesson, you should be able to :

- *explain what energy bands are and how they are used to classify materials as conductors, insulators and semiconductors;*

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- differentiate between (i) intrinsic and extrinsic and (ii) n-type and p-type semiconductors;
- explain formation of depletion region and barrier potential in a p-n junction diode;
- describe I-V characteristics of a p-n junction diode in the forward and reverse biases;
- describe different type of diodes, viz. zener, LED, photo diode and solar cell and their I-V characteristics;
- explain the action of a transistor;
- describe the effect of doping, size and function of different regions in a transistor;
- list the differences between p-n-p and n-p-n transistors;
- list different configurations in which a transistor can be connected and describe their input and output characteristics; and
- compare different configurations of a transistor in terms of their input/output resistance, gain and applications.

28.1 ENERGY BANDS IN SOLIDS

While studying the structure of atoms you have learnt that electrons in an isolated atom stay in certain discrete, well defined energy states. When two atoms come closer to form a stable structure, such that the separation between them tends to be lesser than their diameter (d), the energy states tend to overlap, which is forbidden by Pauli's exclusion principle. Hence, they get modified and corresponding to each of the interaction energy states, two energy states are created: one slightly lower than the normal state which is called the bonding state and the other slightly higher than the normal state called the antibonding state.

In solids, very large number of atoms (typically 10^{23} atoms per cm^3) come together to form crystals. If N atoms interact corresponding to each of the energy states, $2N$ energy states are created. All these energy states are so close to each other (typically $\Delta E \sim 10^{-23}$ eV) that we cannot practically discriminate between them. This quasi continuous distribution of energy states, which are though separate but practically indiscriminable, is called **energy band**.

The process of interaction of energy states (and thereby energy band formation) starts from outer unfilled energy states and then proceeds to valence level. The band formed of unfilled energy levels is called conduction band and the one formed of filled valence levels is called valence band. The relative position of



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28.1.1 Classification of Solids as Conductors, Semiconductors and Insulators on the basis of Energy Bands

If in a solid at equilibrium separation, the conduction band (CB) and valence bands (VB) overlap as it happens in case of metals the material is conductor [Fig. 28.1(a)].

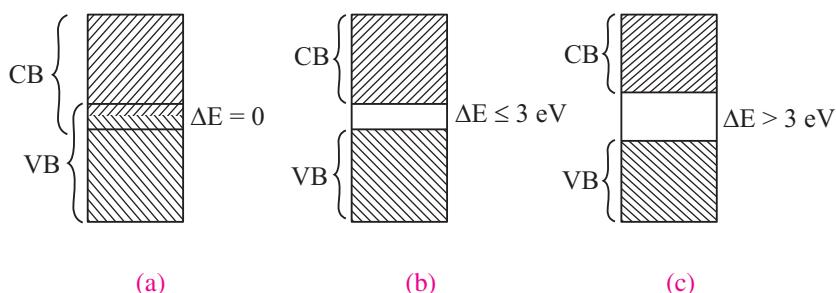


Fig. 28.1 Energy band in (a) Conductors (b) Semiconductors (c) Insulators

If at equilibrium separation the conduction band is completely empty, valence band is completely full and there is a small band gap ($\Delta E \leq 3$ eV) between the highest level of valence band and lowest level of conduction band, called a forbidden energy gap, the solid is a semiconductor. [Fig. 28.1(b)].

If at equilibrium separation, the CB is completely empty, VB is completely filled and there is a large band gap ($\Delta E > 3$ eV) the solid is an insulator.

28.2 INTRINSIC AND EXTRINSIC SEMICONDUCTORS

Semiconductors are classified on the basis of their purity as intrinsic (pure) and extrinsic (impure) semiconductors. Let us now learn about these.

28.2.1 An Intrinsic Semiconductor

Pure silicon and germanium are intrinsic semiconductors as the electrons in these elements are all tightly held in their crystalline structure, i.e., they do not have free electrons. When energy is added to pure silicon in the form of heat, say, it can cause a few electrons to break free of their bonds, leaving behind a hole in each case. (The absence of electrons is treated as positively charged particles having the same amount of positive charge as the negative charge on an electron.) These electrons move randomly in the crystal. These electrons and holes are called *free carriers*, and move to create electrical current. However, there are so few of them in pure silicon that they are not very useful.

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Note that in an intrinsic semiconductor, electrons and holes are always generated in pairs and the negative charge of free electrons is exactly balanced by the positive charge of holes. However, a hole only shifts its position due to the motion of an electron from one place to another. So we can say that when a free electron moves in a crystal because of thermal energy; its path deviates whenever it collides with a nucleus or other free electrons. This gives rise to a zig-zag or random motion, which is similar to that of a molecule in a gas.

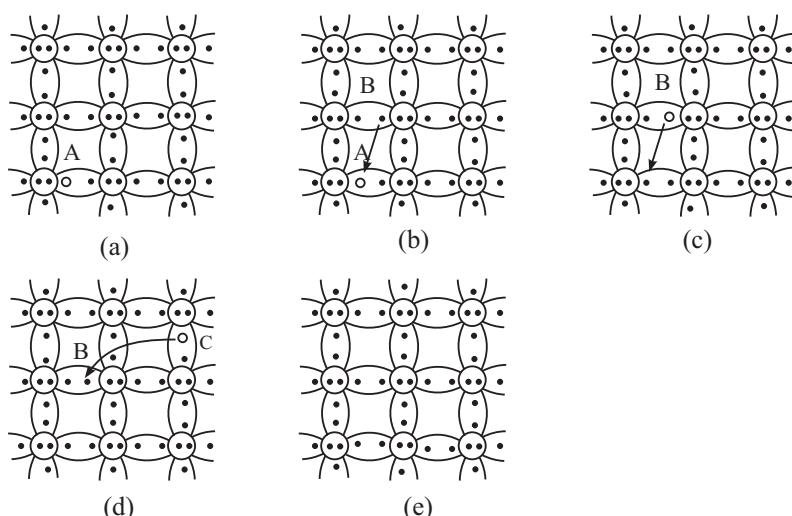


Fig. 28.2 : Movement of electrons and holes in a semiconductor

Now refer to Fig. 28.2(a) and consider the electron-hole pair generated at point A. The free electron drifts in the crystal leaving behind a hole. The broken bond now has only one electron and this unpaired electron has tendency to acquire an electron and complete its pair by forming a covalent bond. Due to thermal energy, the electron from neighbouring bond, say at point B, may get excited to break its own bond and jump into the hole at A. As a result, the hole at A vanishes and a new hole appears at B (Fig. 28.2(c)). Thus motion of electron from point B to point A causes the hole to move from A to B.

You may now like to ask: What will happen when hole at B attracts and captures a valence electron from neighbouring bond at C? The movement of electron from C to B causes movement of hole from B to C [see Fig. 28.2(d) and (e)]. Conventionally, the flow of electric current through the semiconductor is taken in the same direction in which holes move.

At absolute zero temperature, all valence electrons are tightly bound to their parent atoms and intrinsic semiconductor behaves as an insulator. At room temperature, the thermal energy makes a valence electron in an atom to move away from the influence of its nucleus. Therefore, a covalent bond is broken and electron becomes free to move in the crystal, resulting in the formation of a vacancy, called hole. **Thus, due to thermal energy, some electron-hole pairs are generated and semiconductor exhibits small conductivity.** For example, at



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room temperature (300 K), Ge has intrinsic carrier concentration of about $2.5 \times 10^{19} m^{-3}$ electron-hole pair. As temperature increases, more electron-hole pairs are generated and conductivity increases. Alternatively, we can say that resistivity decreases as temperature increases. It means that semiconductors have negative temperature coefficient of resistance.

28.2.2 An Extrinsic Semiconductor

You now know that intrinsic semiconductors have high resistivity. Also their conductivity shows little flexibility. For these reasons, *intrinsic (pure) semiconductors are of little use; at best these can be used as a heat or light sensitive resistance*. These limitations are overcome by adding a small and measured quantity of another material to intrinsic (pure) semiconductor, which either increases the number of holes or electrons.

Note that the word impurity is being used here because we are adding atoms of some other element to a pure material.

The process in which some atoms of a pure or intrinsic semiconductor are replaced by impurity atoms from their lattice-sites is called **doping** and the impurity so added is called **dopant**. Such doped semiconductors are called **extrinsic** semiconductors.

The doped semi-conductor normally belongs to group IV of periodic table and the dopants are generally taken from either **group III** (having three valence electrons) or **group V** (having five valence electrons) of the Periodic Table. Fig. 28.3 shows a small portion of the Periodic Table. Here groups III and V have been highlighted to indicate the types of materials generally used for doping.

| | III | IV | V | VI |
|----|-----|----|----|----|
| II | Al | Si | P | S |
| Zn | Ga | Ge | As | Se |
| Cd | In | Sn | Sb | Te |
| Hg | | | | |

Fig. 28.3 : A part of the Periodic Table. Group III and V elements are used for doping an intrinsic semiconductor of group IV.

Normally we add a very small amount of impurity atoms to the pure simiconductor. It is of the order of one atom per 10^8 atoms of intrinsic semiconductor. These atoms change the balance of charge carriers; either they add free electrons or create holes. Either of these additions makes the material more conducting. Thus, most of the charge carriers in extrinsic semiconductors originate from the impurity atoms.

28.2.3 *n*-and *p*-type Semiconductors

From the electronic configuration of Si ($1s^2, 2s^2, 2p^6, 3s^2, 3p^2$), you will recall that ten electrons are tightly bound to the nucleus and four electrons revolve around the nucleus in the outermost orbit. In an intrinsic silicon semiconductor, the Si

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atom attains stability by sharing one electron each with four neighbouring Si atoms. (This is called *covalent bonding*). The same holds true for germanium; its electronic configuration is $1s^2, 2s^2, 2p^6, 3s^2, 3p^6, 3d^{10}, 4s^2, 4p^2$. When silicon (or germanium) is doped with a pentavalent (five electrons in the outermost orbit) atom like phosphorus, arsenic or antimony, four electrons form covalent bonds with the four neighbouring silicon atoms, but the fifth (valence) electron remains unbound and is available for conduction, as shown in Fig. 28.4. Thus, when a silicon (or germanium) crystal is doped with a pentavalent element, it develops excess free electrons and is said to be an *n-type* semiconductor. Such impurities are known as *donor impurities*.

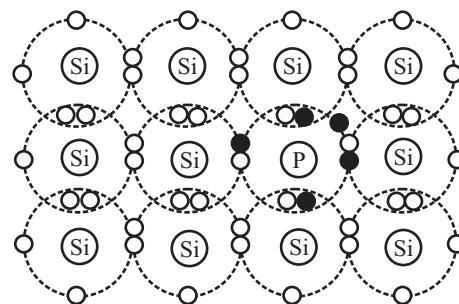


Fig. 28.4 : Covalent bonding in a *n*-type semiconductor

If silicon (or germanium) is doped with a trivalent (three electrons in the outermost shell) atom like boron, aluminium, gallium or indium, three valence electrons form covalent bonds with three silicon atoms and deficiency of one electron is created. This deficiency of electron is referred to as *hole*. It is shown in Fig. 28.5. Such a semiconductor is said to be a *p-type* semiconductor and the impurities are known as *acceptor impurities*.

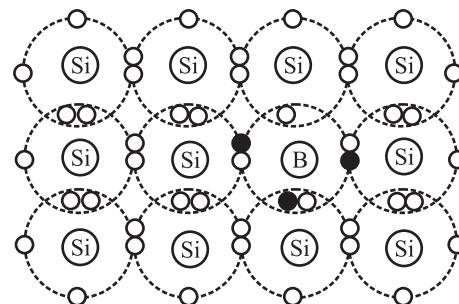


Fig. 28.5 : Covalent bonding in a *p*-type semiconductor

You may now like to ask: Is a *n*-type semiconductor negatively charged or a *p*-type semiconductor positively charged? The answer to this question is not in affirmative.



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INTEXT QUESTIONS 28.1

1. At 300 K, pure silicon has intrinsic carrier concentration of $1.5 \times 10^{16} \text{ m}^{-3}$. What is the concentration of holes and electrons?
2. The *n*-type semiconductor is obtained by doping with
 - (i) trivalent impurity
 - (ii) pentavalent impurity
 - (iii) tetravalent impurity
 - (iv) trivalent as well as tetravalent
3. An intrinsic semiconductor can be converted into an extrinsic semiconductor by addition of This process in called
4. Electrons in *n*-type semiconductor and holes in *p*-type semiconductor are the carriers.
5. An extrinsic semiconductor has resistivity as compared to an intrinsic semiconductor.

28.3 A *p-n* JUNCTION

You now know that *n*-type and *p*-type semiconductors respectively have electrons and holes as majority charge carriers. What do you think will happen if a *n*-type material is placed in contact with a *p*-type material? Shall we obtain some useful device? If so, how? To answer such questions, let us study formation and working of a *p-n* junction.

28.3.1 Formation of a *p-n* Junction

To form a *p-n* junction, the most convenient way is to introduce donor impurities on one side and acceptor impurities into the other side of a single semiconducting crystal, as shown in Fig. 28.6.

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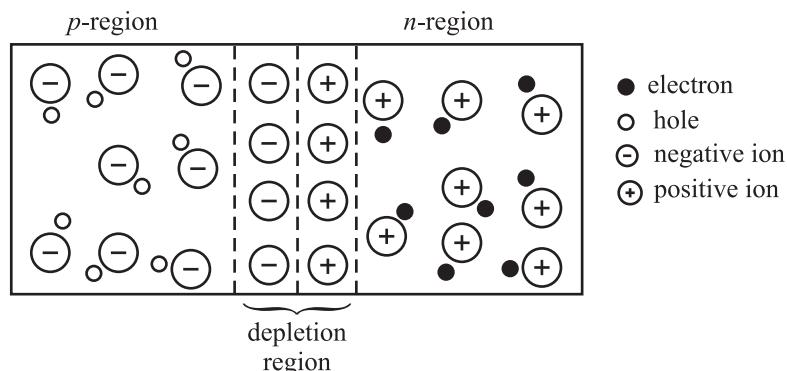


Fig. 28.6 : A p-n junction with depletion region

We now know that there is greater concentration of electrons in the *n*-region of the crystal and of holes in the *p*-region. Because of this, electrons tend to diffuse to the *p*-region and holes to the *n*-region and recombine. Each recombination eliminates a hole and a free electron. This results in creation of positively and negatively charged ions near the junction in *n* and *p* regions, respectively. As these charges accumulate, they tend to act as shield preventing further movement of electrons and holes across the junction. Thus, after a few recombinations, a narrow region near the junction is depleted in mobile charge carriers. It is about $0.5 \mu\text{m}$ thick and is called the *depletion region* or *space-charge region*.

Due to accumulation of charges near the junction, an electric field is established. This gives rise to electrostatic potential, known as *barrier potential*. This barrier has polarities, as shown in Fig. 28.7. When there is no external electric field, this barrier prevents diffusion of charge carriers across the junction.

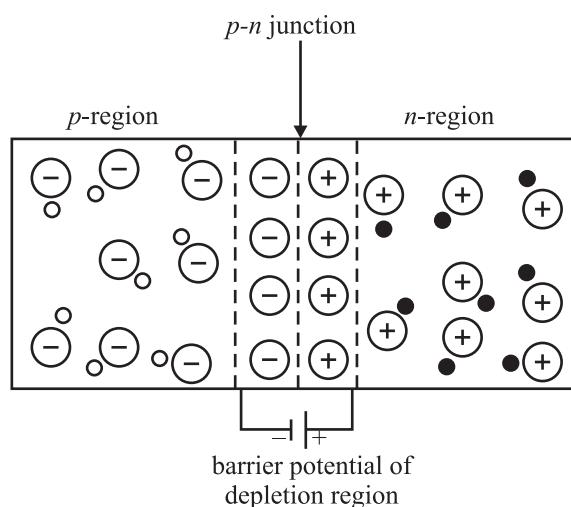


Fig.28.7 : Barrier potential due to depletion region

The barrier potential is characteristic of the semiconductor material. It is about 0.3 V for Ge and about 0.7 V for Si. The junction acts as a diode. It is symbolically represented as shown in Fig. 28.8(a). Here A corresponds to *p*-region and acts as

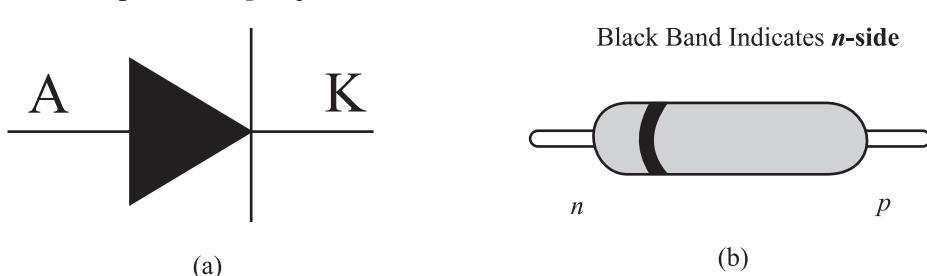


Fig. 28.8: (a) Symbol of a *p-n* junction (diode). The arrow gives the direction of conventional current. It is from *p* to *n* region (b) A *p-n* junction diode available in the market.

You may have noted that semiconductor diodes are designated by two letters followed by a serial number. The first letter indicates the material: *A* is used for material with a band gap of 0.6 eV to 1.0eV such as germanium. *B* is used for material with a band gap of 1.0eV to 1.3eV, such as silicon. The second letter indicates the main application: *A* signifies detection diode, *B* denotes a variable capacitance diode, *E* for tunnel diode, *Y* for rectifying diode and *Z* denotes Zener diode. The serial numbers specify power rating, peak reverse voltage, maximum current rating, etc. (We have to refer to manufacturer's catalogue to know exact details.) For example, *BY127* denotes a silicon rectifier diode and *BZ148* represents a silicon Zener diode.

To make visual identification of anode and cathode, the manufacturers employ one of the following ways :

- the symbol is painted on the body of the diode;
- red and blue marks are used on the body of the diode. Red mark denotes anode, whereas blue indicates the cathode;
- a small ring is printed at one end of the body of the diode that corresponds to the cathode. The band in Fig. 28.8(b) indicates the *n*-side of the *p-n* junction.

Note that we have to work within the specified ranges of diode ratings to avoid damage to the device.



INTEXT QUESTIONS 28.2

1. Fill in the blanks:
 - (a) When a *p-n* junction is formed, the diffuse across the junction.
 - (b) The region containing uncompensated acceptor and donor ions is called region.

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- (c) The barrier potential in silicon is V and in germanium, it is V.
- (d) In a *p-n* junction with no applied electric field, the electrons diffuse from *n*-region to *p*-type region as there is concentration of in *n*-region as compared to *p*- region.
2. Choose the correct option:
- (a) The potential barrier at the *p-n* junction is due to the charges on the either side of the junction. These charges are
- (i) majority carriers
 - (ii) minority carriers
 - (iii) fixed donor and acceptor ions.
 - (iv) none of above
- (b) In a *p-n* junction without any external voltage, the junction current at equilibrium is
- (i) due to diffusion of minority carriers only
 - (ii) due to diffusion of majority carriers only
 - (iii) zero, as no charges are crossing the junction
 - (iv) zero, as equal and opposite charges are crossing the junction
- (c) In a semiconductor diode, the barrier potential repels
- (i) minority carriers in both the regions
 - (ii) majority carriers in both the regions
 - (iii) both the majority and the minority carriers
 - (iv) none of the above
3. Why is depletion region named so? What is depletion region made of?

28.4 FORWARD AND REVERSE BIASED *p-n* JUNCTION

Biassing means application of voltage. To make a *p-n* junction to conduct, we have to make electrons move from the *n*-type region to the *p*-type region and holes moving in the reverse direction. To do so, we have to overcome the potential barrier across the junction by connecting a battery to the two ends of the *p-n* junction diode. The battery can be connected to the *p-n* junction in two ways:

- Positive terminal of the battery connected to the *p*-side and negative terminal of the battery connected to the *n*-side. This is called **forward bias** [Fig. 28.9(a)].
- Positive terminal of the battery connected to the *n*-side and negative terminal of the battery connected to the *p*-side. This is called **reverse bias** [Fig. 28.9(b)].

When a junction is forward biased and the bias exceeds barrier potential, holes are compelled to move towards the junction and cross it from the *p*-region to the



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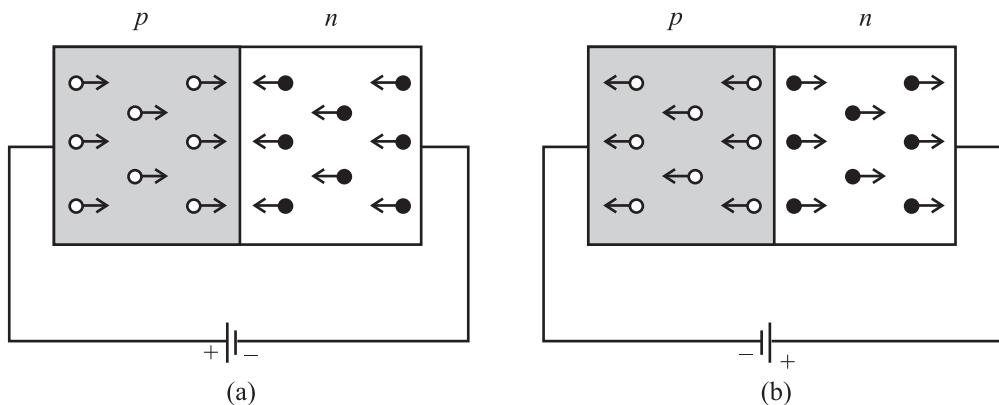


Fig. 28.9 : (a) Forward biased, and (b) reverse biased p - n junction

When the p - n junction is **reverse biased**, holes in the p -region and electrons in the n -region move away from the junction. Does it mean that no current shall flow in the circuit? No, a small current does flow even now because of the fewer number of electron-hole pairs generated due to thermal excitations. This small current caused by minority carriers is called **reverse saturation current** or **leakage current**. In most of the commercially available diodes, the reverse current is almost constant and independent of the applied reverse bias. Its magnitude is of the order of a few microamperes for Ge diodes and nanoamperes in Si diodes.

A p - n junction offers low resistance when forward biased, and high resistance when reverse biased. This property of p - n junction is used for ac rectification.

When the reverse bias voltage is of the order of a few hundred volt, the current through the p - n junction increases rapidly and damages it due to excessive power dissipation. The voltage at which a diode breaks down is termed as **breakdown voltage**. Physically, it can be explained as follows: When a reverse bias is applied, a large electric field is established across the junction. This field (i) accelerates the available minority carriers, which, in turn, collide with the atoms of the semiconductor material and eject more electrons through energy transfer (avalanche effect), and (ii) breaks covalent bonds by exerting large force on electrons bound by the bonds. This results in creation of additional electron-hole pairs in the junction region (Zener effect). Both these processes give rise to large reverse current even for a small increment in reverse bias voltage. This process is termed as *Zener breakdown*.

**INTEXT QUESTION 28.3**

1. Define forward bias.
2. Define reverse bias.
3. Fill in the blanks:
 - (a) When forward bias is applied on a *p-n* junction diode, the width of the depletion region
 - (b) When a *p-n* junction diode is reverse biased, the width of depletion region
 - (c) When the reverse bias voltage is made too high, the current through the *p-n* junction abruptly. This voltage is called.
4. Choose the correct option:
 - (a) In a forward biased junction
 - (i) the holes in the *n*-region move towards the *p*-region
 - (ii) there is movement of minority carriers
 - (iii) charge carriers do not move
 - (iv) majority carriers in both the regions (*n* and *p*-regions) move into other regions.
 - (b) In a reverse biased junction
 - (i) there is no of potential barrier
 - (ii) there is movement of majority carriers only
 - (iii) there is movement of minority carriers only
 - (iv) none of the above
5. State two types of reverse breakdowns which can occur in a *p-n* junction diode and differentiate between them.

28.5 CHARACTERISTICS OF *p-n* JUNCTION DIODES

The practical application of a semiconductor device in electronic circuits depends on the current and voltage (*I-V*) relationship, as it gives vital information to a circuit designer as well as a technician. Therefore, with the help of *I-V* characteristics, we can know how much current flows through the junction diode at a particular voltage.

28.5.1 Forward Bias Characteristics

Refer to Fig. 28.10(a). You will note that to draw forward bias characteristic of a *p-n* junction diode, the positive terminal of a battery (*B*) is connected to *p*-side of



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the diode through the rheostat. (Alternative by we can use a variable battery.) The voltage applied to the diode can be varied with the help of the rheostat. The milliammeter (mA) measures the current in the circuit and voltmeter (V) measures the voltage across the diode. The direction of conventional current is the same as the direction of the diode arrow. Since current experiences little opposition to its flow through a forward biased diode and it increases rapidly as the voltage is increased, a resistance (R) is added in the circuit to limit the value of current. If this resistance is not included, the diode may get permanently damaged due to flow of excessive current through it.

The $I-V$ characteristic curve of a $p-n$ junction in forward bias is shown in Fig. 28.10(b).

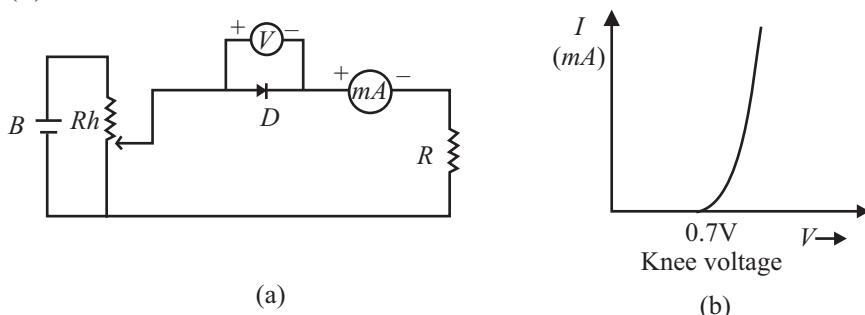


Fig. 28.10 : (a) Circuit diagram $I-V$ characteristics of a $p-n$ junction diode in forward bias, and (b) typical characteristics curve.

Note that the characteristic curve does not pass through origin; instead it meets the V -axis at around 0.7V. It means that the $p-n$ junction does not conduct until a definite external voltage is applied to overcome the barrier potential. The forward voltage required to get the junction in conduction mode is called **knee voltage**. It is about 0.7 V for Si and 0.3 V for Ge $p-n$ junction.

This voltage is needed to start the hole-electron combination process at the junction. As the applied voltage is increased beyond knee voltage, the current through the diode increases linearly. For voltage of around 1V, the current may attain a value of 30-80 mA.

28.5.2 Reverse Bias Characteristics

To draw reverse bias characteristics of a $p-n$ junction, we use the circuit diagram shown in Fig. 28.11 (a). If you compare it with Fig. 28.10(a) for forward $I-V$ characteristics, you will note two changes:

- (i) The terminals of the junction are reversed.
- (ii) Instead of milliammeter, microammeter (μA) is used.

A typical $I-V$ characteristic curve of a $p-n$ junction in reverse bias is shown in Fig 28.11(b).

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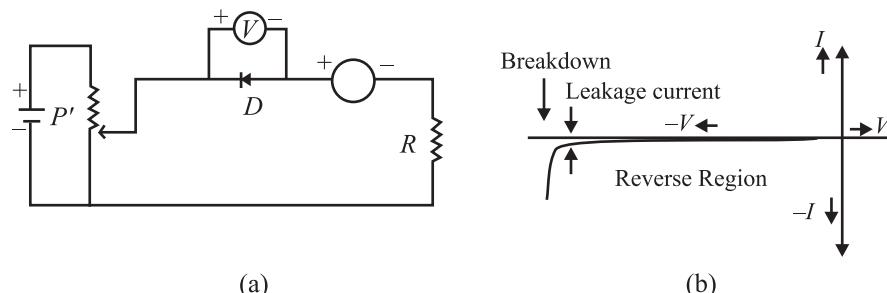


Fig.28.11 : a) Circuit diagram to obtain I - V characteristics of a p - n junction in reverse bias, and b) reverse bias characteristic curve

Note that the junction current is comparatively much less in reverse bias for all voltages below the breakdown voltage. And at breakdown voltage, the current increases rapidly for a small increase in voltage. Moreover, comparison of Fig. 28.10(b) and 28.11(b) reveals that a p - n junction diode offers low resistance when it is forward biased and high resistance when reverse biased. At the breakdown voltage in reverse biased p - n junction diode, the sharp increase in reverse current is due to sudden decrease in resistance offered by the junction.

From this we may conclude that a p - n junction diode conducts in only one direction, i.e. has unidirectional conduction of current, with electrons flowing from the n -type region to p -type end in forward bias.

You may have seen turnstiles at a metro subway station that let people go through in only one direction. A diode is a one-way turnstile for electrons.

p - n junction diodes find wide applications. These include :

1. The unidirectional conducting property of a diode is used to convert ac voltage into dc voltage as a *rectifier*. Diodes are also used in adaptors to recharge batteries of cell phones, CDplayers, laptops, etc. You will study about it in detail in the next lesson.
2. A device that uses batteries often contains a diode as it simply blocks any current from leaving the battery, if it is reverse biased. This protects the sensitive electronics in the device.



INTEXT QUESTIONS 28.4

1. Explain the concept of knee voltage.
2. (a) The knee voltage in case of silicon diode is whereas in germanium diode it is
 (b) In a p - n junction diode, the current flows only in direction.
 (c) The reverse saturation current is of the order of for germanium diodes.



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3. Choose the correct option :

- (a) The $I-V$ characteristics of a $p-n$ junction diode in forward bias show
 - (i) a non-linear curve
 - (ii) linear curve
 - (iii) linear as well as non-linear portions
 - (iv) none of above
- (b) When a $p-n$ junction is forward biased and the voltage is increased, the rapid increase in current for relatively small increase in voltage occurs
 - (i) almost immediately
 - (ii) only when the forward bias exceeds the potential barrier
 - (iii) when there is breakdown of the junction
 - (iv) none of the above

28.6 TYPES OF DIODES

By adjusting the levels of doping, doping material and the geometry (size, area etc.) of a $p-n$ junction diode, we can modify its electrical and optical behaviour. In this section, we have listed diodes whose properties have been deliberately modified to obtain specific capabilities. Each of these diodes has its own schematic symbol and reflects its nature and functions.

You can use the following table to make a comparison between different diodes:

| Name | Symbol | Construction | Principle mechanism | Main | Main use function |
|-------------|--------|---|--|--|-------------------------------------|
| Zener diode | | $p-n$ junction diode with heavily doped p - & n - regions. Very narrow depletion layer (< 10 nm). | Zener breakdown mechanism | Provides continuous current in reverse breakdown voltage region without being damaged. | Voltage stabilization or regulation |
| Photo-diode | | $p-n$ junction diode. Uses light (or photo) emitting semiconductor materials, with very thin p -region, whose thickness is determined by wavelength of radiation to be detected | Photovoltaic effect into electrical current in | Converts an optical input controls in VCR & TV reverse bias. | Receivers for remote |

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| | | | | | |
|------------|--|---|---------------------|--|---|
| LED | | <i>p-n</i> junction diode with materials having band energies corresponding to near infrared region or visible light region (GaAsP or InP) | Electroluminescent | Changes an electrical input to a light output in forward bias. | Used in multimeters, digital watches, instrument displays, calculators, switch boards, burglar alarm and remote control devices |
| Solar cell | | <i>p-n</i> junction diode in which either <i>p</i> or <i>n</i> region is made very thin to avoid significant absorption of light before reaching the junction | Photovoltaic effect | Conversion of solar energy into electrical energy | 1. In satellites to power systems. 2. To charge batteries. 3. Calculators |

28.6.1 I-V Characteristics of Zener diode

Zener diode is fabricated by heavily doping both the *p*- and *n*-sides of the junction. Hence depletion layer formed is very thin ($< 10^{-6}$ m). And the electric field across the depletion layer is extremely high ($\sim 5 \times 10^6$ N C $^{-1}$) even for a small reverse bias voltage of 5 V. The I-V characteristics of a Zener diode is shown in Fig. 28.12. It is seen that when the applied reverse voltage (*V*) reaches the breakdown voltage (*V_z*) of the Zener diode, there is a large change in the current. After the breakdown voltage *v_z*, a large change in the current can be produced by almost insignificant change in the reverse bias voltage. Zener voltage remains constant, even though the current through the Zener diode varies over a wide range.

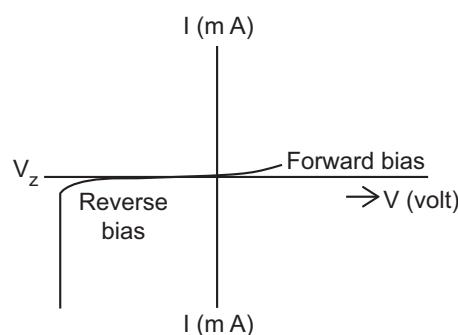
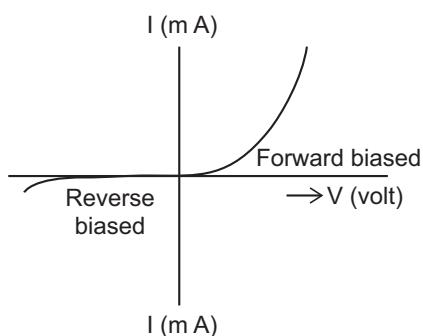


Fig. 28.12

**Notes****28.6.2 I-V Characteristics of light-emitting diode**

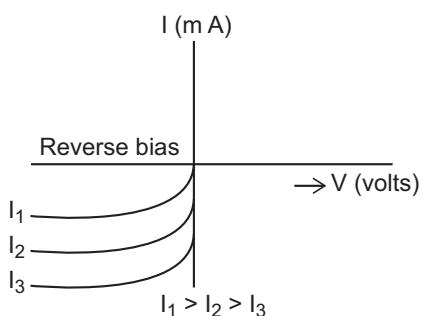
In light-emitting diode (LED) when the forward current of diode is small the intensity of light emitted is small. As the forward current increases, the intensity of the emitted light increases and reaches a maximum value. Further increase in the forward current results in a decrease of light intensity. LEDs are biased such that the light emitting efficiency is maximum.

The I-V characteristics of a LED is similar to that of a Si junction diode as shown in Fig. 28.13. But the threshold voltages are much higher and slightly different for each colour. The reverse breakdown voltage of LEDs are very low, typically around 5V.

**Fig. 28.13****28.6.3 I-V Characteristics of Photo diode**

The photo diode is fabricated such that the generation of electron – hole pairs takes place in or near the depletion region in the diode. Due to the electric field of the junction, the electrons and holes are separated before they recombine. The direction of the electric field is such that the electrons reach the *n*-side and the holes reach the *p*-side. The electrons are collected on the *n*-side and the holes are collected on the *p*-side giving rise to an emf. When an external load is connected, the current flows. The magnitude of the photocurrent depends on the intensity of the incident light.

I-V Characteristics of a photo diode are shown in Fig. 28.14.

**Fig. 28.14**

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28.6.4 I-V Characteristics of Solar Cell

The generation of emf, when the light falls on a solar cell is due to the following three basic processes: generation, separation and collection. Generation of electron – hole pairs is due to the light (with $h\nu > E_g$) close to the junction. Separation of electrons and holes is due to the electric field of the depletion layer. Electrons are swept to the *n*-side and the holes to the *p*-side.

The electrons reaching the *n*-side are collected by the front contact and the holes reaching the *n*-side are collected by the back contact. Thus the *p*-side becomes positive and the *n*-side negative giving rise to photo voltage.

The I-V characteristic curves of a solar cell are shown in Fig. 28.15. They are drawn in the fourth quadrant of the coordinate axis. This is because a solar cell does not draw current but supplies the same to the load.

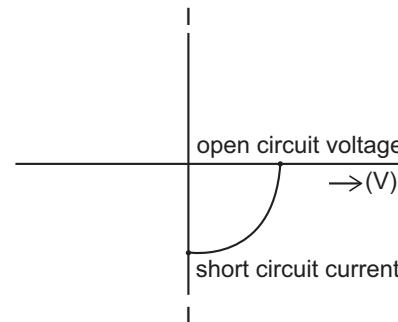


Fig. 28.15



INTEXT QUESTIONS 28.5

1. Choose the correct option
 - (a) A zener diode is operated in
 - (i) Forward bias
 - (ii) Reverse bias
 - (iii) Both of the above
 - (iv) None of the above
 - (b) Zener diode is
 - (i) A highly doped *p-n* junction diode
 - (ii) A lowly doped *p-n* junction diode
 - (iii) A moderately doped *p-n* junction diode
 - (iv) Another name of normal *p-n* junction diode
 - (c) A zener diode is used as a
 - (i) amplifier
 - (ii) rectifier
 - (iii) constant current device
 - (iv) constant voltage device

**Notes**

2. Fill in the blanks
 - a) The zener diode is based on the breakdown mechanism.
 - b) A photodiode is operated in bias.
 - c) In a photodiode, the $p-n$ junction is made from semiconductor material.
 - d) LED's are made up of the conductor material from of the periodic table.
 - e) The light emitting diodes operate in bias.
 - f) The arrow in the symbol of LED symbolizes of light.
 - g) In an LED light is emitted due to of electrons and holes.
 - h) LED is based on the principle of
 - i) Solar cells are based on effect.
 - j) When sunlight having energy than the band gap energy falls on the solar cell, it is and frees electron-hole pairs.
3. How does the separation of electrons and holes take place in a solar cell?

28.7 TRANSISTORS – pnp AND npn

In the preceding sections, you have learnt about a $p-n$ junction diode, which permits current to flow in only one direction. This limits its applications to rectification and detection. A more useful semiconductor device is a bipolar junction transistor.

The invention of transistor by John Bardeen, Walter Brattain and William Shockley in 1948 at Bell laboratory in USA revolutionised the electronic industry. The transistors find many varied uses in our daily life ranging from gas lighter to toys to amplifiers, radio sets and television. In the form of switching device, these can be used to regulate vehicular traffic on the roads. They form key elements in computers, space vehicles, power systems in satellites and communication.

A transistor is basically a silicon or germanium crystal containing three alternate regions of p and n -type semiconductors as shown in Fig. 28.16. These three regions are called *emitter*(E), *base*(B) and *collector*(C). The middle region is the base and the outer two regions are emitter and collector. Note that the emitter and collector are of the same type (p or n) and collector is the largest of the three regions.

The base terminal controls the current flowing between the emitter and the collector. This control action gives the transistor an added advantage over the diode, which has no possibility of controlling the current flow. Depending on the type of doping, the transistors are classified as $n-p-n$ or $p-n-p$. In general, the level of doping decreases from emitter to collector to base.

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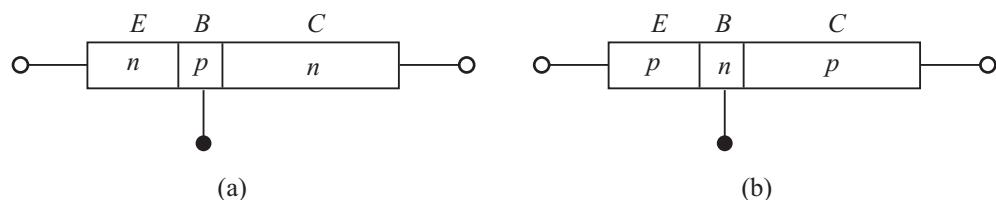
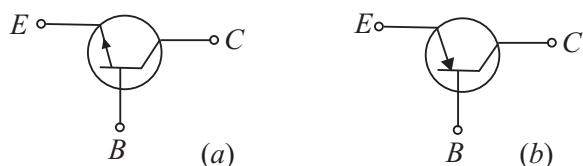


Fig. 28.16 : (a) n-p-n, and (b) p-n-p transistor

The names of the terminals of a transistor give clear indication of their functions. In case of a *n-p-n* transistor, the majority carriers (electrons) from the emitter are injected into base region. Since base is a very lightly doped thin layer, it allows most of the electrons injected by the emitter to pass into the collector. Being the largest of three regions, the collector dissipates more heat compared to the other two regions.



Figs 28.17 : Symbols of a) n-p-n, and b) p-n-p transistors

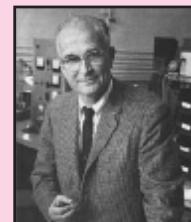
The symbolic representations of *n-p-n* and *p-n-p* transistors are shown in Fig. 28.17. The arrow head indicates the direction of flow of conventional current.

You may now like to ask : Why does the arrow head point outward in case of *n-p-n* transistor and inward in case of *p-n-p* transistor?

In a *n-p-n*, transistor, the emitter current is due to flow of electrons from emitter to base, and the conventional current flows from base to emitter and hence the arrow head points out from the base. In case of *p-n-p* transistor, the emitter current comprises flow of holes from emitter to base. Thus the conventional current flows from emitter to base.

Since transistors are bipolar devices, their operation depends on both the majority and minority carriers.

William Bradford Shockley (1910 – 1989)



England born, American physicist W.B. Shockley was one of the three scientists who received 1956 Nobel Prize in physics for the discovery of transistor. Basically a solid state physicist, shockley contributed significantly to the development of theoretical understanding of bands in semiconductors, order and disorder in alloys; theory of vacuum tubes, theory of dislocations and theory of ferromagnetic domains. He is truely one of the pioneers of electronic revolution.



Notes

28.7.1 Working Principle

You are familiar with the working of a *p-n* junction. We now discuss the working principle of a transistor and consider an *n-p-n* transistor first because it is more commonly used.

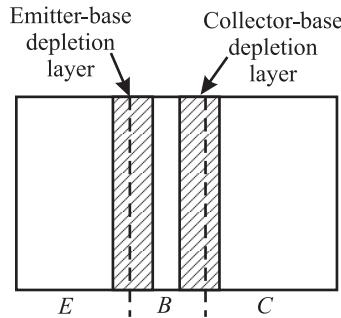
When no voltage is applied across the transistor, diffusion of free electrons across the junctions produces two depletion layers, as shown in Fig. 28.18. For each depletion layer, the barrier potential is about 0.7V at 25°C for a silicon transistor and 0.3V for a germanium transistor. As you may be aware, silicon transistors are more widely used than germanium transistors because of higher voltage rating, greater current ratings, and low temperature sensitivity. For our discussion, we refer to silicon transistors, unless otherwise indicated.

Since the three regions in a transistor have different doping levels, the depletion layers have different widths. If a region is heavily doped, the concentration of ions near the junction will be more, resulting in thin depletion layer and vice versa. Since the base is lightly doped as compared to emitter and collector, the depletion layers extend well into it, whereas penetration in emitter/collector regions is to a lesser extent (Fig. 28.18). Moreover, the emitter depletion layer is narrower compared to collector depletion layer.

In order to make a transistor function properly, it is necessary to apply suitable voltages to its terminals. This is called **biasing** of the transistor.

A *n-p-n* Transistor

A typical biasing scheme of a *n-p-n* transistor is shown in Fig. 28.19(a). Note that *the emitter-base junction is forward biased while the collector-base junction is reverse biased*. We therefore expect a large emitter current and low collector current. But in practice, we observe that the collector current is almost as large as the emitter current. Let us understand the reason. When forward bias is applied to the emitter, free electrons in the emitter have to overcome the barrier potential to enter the base region [see Fig. 28.19(b)]. When V_{BE} exceeds barrier potential (0.6 to 0.7V for silicon transistor), these electrons enter the base region, as shown in Fig. 28.19(c). Once inside the base, these electrons can flow either through the thin base into the external base lead or across the collector junction into the collector region. The downward component of base current is called *recombination current*. It is small because the base is lightly doped and only a few holes are available. Since the base region is very thin and it receives a large number of



Figs 28.18: Depletion layers in a transistor when no voltage is applied

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electrons, for $V_{BE} > 0.7V$, most of these electrons diffuse into the collector depletion layer. The free electrons in this layer are pushed (by the depletion layer field) into the collector region [(Fig. 28.19(d))] and flow into the external collector lead. So, we can say that a steady stream of electrons leaves the negative source terminal and enters the emitter

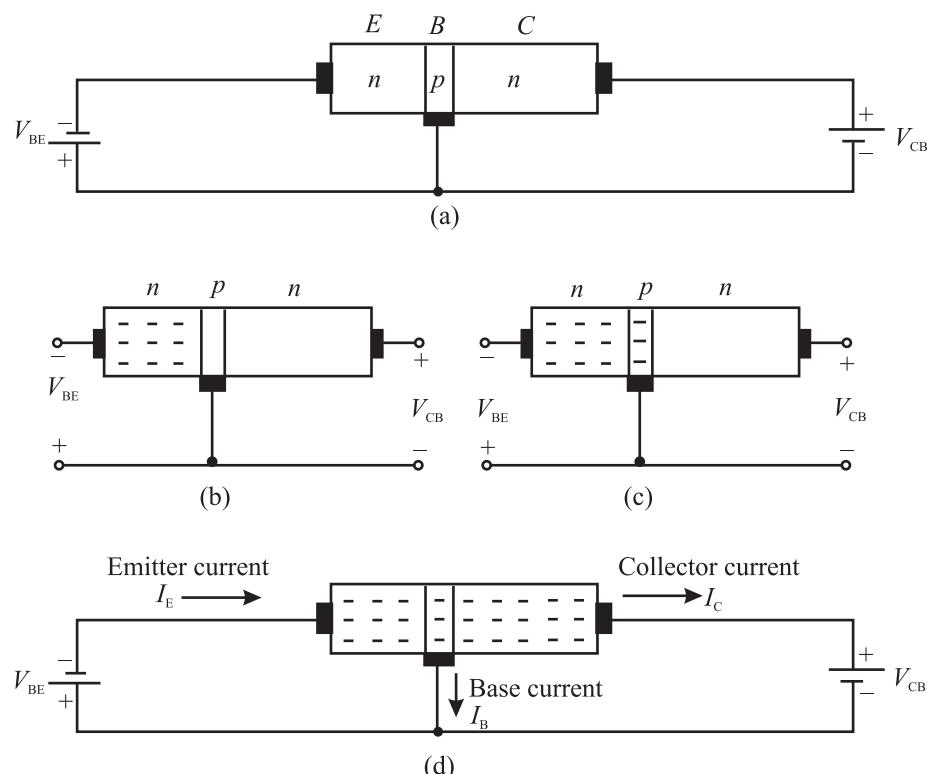


Fig. 28.19 : A *n-p-n* transistor when (a) emitter is forward-biased and collector is reverse-biased, (b) free electrons in an emitter, (c) free electrons injected into base; and (d) free electrons pass through the base to the collector.

region. The forward bias forces these electrons to enter the base region. Almost all these electrons diffuse into the collector depletion layer through the base. The depletion layer field then pushes a steady stream of electrons into the collector region. In most transistors, more than 95 percent emitter-injected electrons flow to the collector; less than 5 percent flow to the external base lead.

From this you should not conclude that you can connect two discrete diodes back to back to get a transistor. This is because in such a circuit, each diode has two doped regions and the overall circuit would have four doped regions and the base region would not be the same as in a transistor. *The key to transistor action, therefore, is the lightly doped thin base between the heavily doped emitter and the intermediately doped collector.* Free electrons passing through the base stay in base for a short time and reach the collector.



Notes

The relation between collector current (I_C) and emitter current (I_E) is expressed in terms of signal current gain, α , of a transistor. It is defined as

$$\alpha = \frac{I_C}{I_E}. \quad (28.1)$$

You should note that the value of α is nearly equal to but always less than one.

Similarly, we can relate the collector current to the base current in a transistor. It is denoted by greek letter beta:

$$\beta = \frac{I_C}{I_B} \quad (28.2)$$

Beta signifies the current gain of the transistor in common-emitter configuration. The value of β is significantly greater than one.

Since emitter current equals the sum of collector current and base current, we can write

$$I_E = I_C + I_B$$

On dividing throughout by I_C , we get

$$\frac{I_E}{I_C} = 1 + \frac{I_B}{I_C}. \quad (28.3)$$

In terms of α and β , we can rewrite it as

$$\frac{1}{\alpha} = 1 + \frac{1}{\beta}$$

or

$$\beta = \frac{\alpha}{1 - \alpha} \quad (28.4)$$

Let us now consider how a *p-n-p* transistor differs from a *n-p-n* transistor in its details.

A *p-n-p* Transistor

A *p-n-p* transistor biased for operation in the active region is shown in Fig 28.20. Note that we reverse the battery terminals when *n-p-n* transistor is substituted by *p-n-p* transistor.

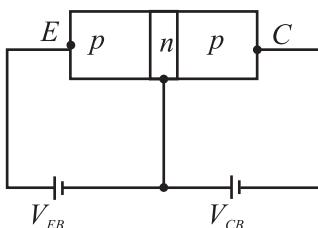


Fig. 28.20 : A *p-n-p* transistor biased for active operation

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As before, the emitter - base junction is forward biased by battery of voltage V_{EB} and the collector base junction is reverse biased by a battery of voltage V_{CB} . The resistance of the emitter-base junction is very small due to its forward bias as compared to the collector-base junction (which is reverse biased). Therefore, we apply small forward bias voltage (0.6V) to the emitter-base junction, whereas the reverse bias voltage applied to the collector-base junction is of much higher value (9V).

The forward bias of emitter-base junction makes the majority carriers, that is the holes, in emitter (*p*-region), to diffuse to the base (*n*-region), on being repelled by the positive terminal of the battery. As width of the base is extremely thin and it is lightly doped, very few (two to five percent) of total holes that enter the base recombine with electrons and 95% to 98% reach the collector region. Due to reverse bias of the collector- base region, the holes reaching this region are attracted by the negative potential applied to the collector, thereby increasing the collector current (I_C). Therefore, increase in emitter current (I_E) increases collector current. And Eqns. (28.1) – (28.4) hold in this case as well.



INTEXT QUESTION 28.6

1. Choose the correct option:
 - (a) The arrow head in the symbol of a transistor points in the direction of
 - (i) hole flow in the emitter region
 - (ii) electron flow in emitter region
 - (iii) majority carriers flow in the above region
 - (iv) none of the above
 - (b) The emitter current in a transistor in normal bias is
 - (i) less than the collector current
 - (ii) equal to sum of base current and collector current
 - (iii) euqlal to base current
 - (iv) none of the above
2. Fill in the blanks
 - (a) A ransistor has regions and junctions.
 - (b) In a transistor, has the least thickness.
 - (c) The emitter region is doped, whereas region has the least doping.
 - (d) The collector of the transistor has size and doping.

- (e) The transistor is said to be in active region when junction is forward biased and junction is reverse biased.
- (f) The two types of transistors are and

You now know the working principle of a transistor. Let us learn the various ways in which a transistor is biased.



Notes

28.7.2 Transistor Configurations

A transistor is a two-port device; it can take an input and deliver an output. For both input and output, two terminals are needed. This can be done in a transistor by making one of the three terminals common. The configurations of a transistor in which one of the terminals is common to both input and output are shown in Fig. 28.21.

- When emitter is common to both input and output circuits, we obtain common emitter (*CE*) configuration (Fig. 28.21a);
- When base is common to both input and output circuits, we obtain common base (*CB*) configuration (Fig. 28.21b); and
- When collector is common to both input and output circuits, we have common collector (*CC*) configuration (Fig. 28.21c).

In each of these configurations, the transistor characteristics are unique. The *CE* configuration is used most widely because it provides voltage, current and power gains. In the *CB* configuration, the transistor can be used as a constant current source while the *CC* configuration is usually used for impedance matching.

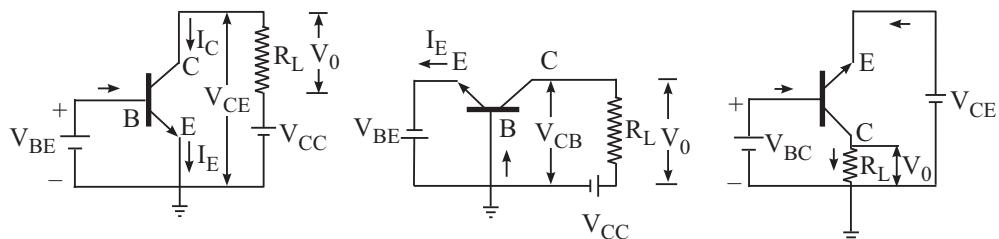


Fig. 28.21: Transistor configuration: a) *CE*, b) *CB*, and c) *CC*

For each configuration, we can plot three different characteristics: a) input characteristics, b) output characteristics, and c) transfer characteristics, depending on the nature of quantities involved.

Table 28.2 gives various quantities related to each of these characteristics in all the three configurations and the transistor constants of interest.

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Table 28.2: Physical quantities of interest in different characteristics of a transistor

| Configuration | Input Characteristic | Output characteristic | Transfer characteristic | Important transistor constant |
|---------------|---|--|-------------------------|-------------------------------------|
| CE | V_{BE} and I_B with V_{CE} as parameter | V_{CE} and I_C with I_B as parameter | I_B and I_C | Current gain, β |
| CB | V_{BE} and I_E with V_{CB} as parameter | V_{CB} and I_C with I_E as parameter | I_E and I_C | Large signal current gain, α |
| CC | V_{CB} and I_B with V_{CE} as parameter | V_{CE} and I_E with I_B as parameter | I_B and I_E | |

To work with a transistor, you will be required to identify its base, emitter and collector leads. To do so, you can follow the following steps.

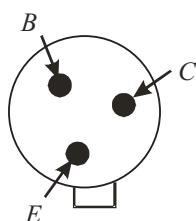


Fig. 28.22 : Identifying transistor leads.

Look for the a small notch provided on the metallic cap. The terminal close to the notch is emitter. To identify other two terminals, turn the transistor up-side-down. You can easily identify the base and the collector as shown in Fig. 28.22.

Like a *p-n* junction diode, transistors are also designated with two letters followed by a serial number. The first letter gives an indication of the material. *A* is for germanium and *B* is for silicon. The second letter indicates the main application: *C* is used for audio frequency transistors, *D* for power transistors and *F* for radio-frequency transistors. The serial number consists of digits assigned by the manufacturer for identification. For example, *AC 125* represents germanium transistor for *AF* applications.

28.8 TRANSISTOR CHARACTERISTICS

As mentioned earlier, operation of a transistor can be studied with input and output *I-V* characteristics. The nature of these characteristics is unique and depends on the configuration used. Let us first study *CE* configuration.

28.8.1 Common Emitter (CE) Configuration of a *n-p-n* Transistor

Common emitter characteristics of a transistor relate voltage and current when emitter is common to both input and output circuits. The circuit diagram for *CE*



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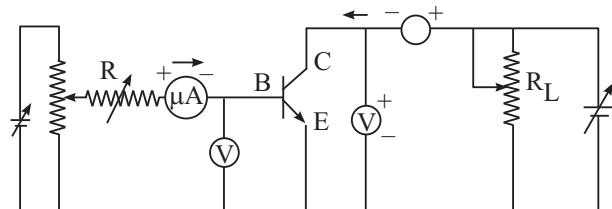


Fig. 28.23 : Circuit diagram for input and output characteristics of a transistor in *CE* configuration.

Input characteristics

In *CE* configuration, the input characteristics show the variation of I_B with V_{BE} when V_{CE} is held constant. To draw this characteristic, V_{CE} is kept at a suitable value with the help of R and R_1 . Then V_{BE} is changed in steps and corresponding values of I_B are measured with the help of microammeter, connected to base. Fig. 28.24. shows typical input characteristics of a *n-p-n* transistor in *CE* configuration.

Note that for a given value of V_{CE} , the curve is as obtained for forward biased *p-n* junction diode. For $V_{BE} < 0.5V$, there is no measurable base current ($I_B = 0$). However, I_B rises steeply for $V_{BE} > 0.6V$.

From the reciprocal of the slope of input characteristic, we get input resistance of the transistor defined as the ratio of small change in base - emitter voltage to the small change produced in the base current at constant collector - emitter voltage:

$$R_{ie} = \left. \frac{\Delta V_{BE}}{\Delta I_B} \right|_{V_{CE}} \quad (28.5)$$

Usually, the value of R_{ie} is in the range $20-100\Omega$. You should note that since the curve is not linear, the value of input resistance varies with the point of measurement. As V_{CE} increases, the curve tends to become more vertical and the value of R_{ie} decreases.

Output characteristics

The output characteristic curves depict the variation of collector current I_C with V_{CE} , when base current I_B is kept constant. To draw output characteristics, I_B is

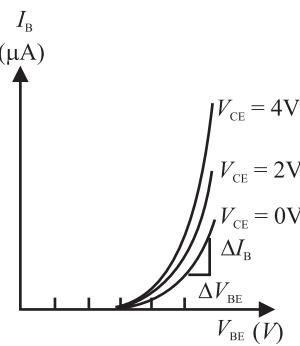


Fig. 28.24 : Input characteristics of a typical *npn* transistor in *CE* configuration

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fixed, say at $10 \mu\text{A}$, by adjusting R_1 and R . V_{CE} is then increased from 0 to 10 V in steps of 0.5V by varying R_2 and the corresponding value of I_C is noted. Similarly, the output characteristics can be obtained at $I_B = 40\mu\text{A}, 60\mu\text{A}, 80\mu\text{A}$. However, in no case, the maximum base current rating of the transistor should be exceeded.

The output characteristics of this configuration are shown in Fig. 28.25.

From the output characteristics, you will note that I_C changes with increase in V_{CE} for a given value of I_B and I_C increases with I_B for a given V_{CE} . From these characteristics, we can calculate output admittance (h_{oe}):

$$h_{oe} = \frac{\Delta I_C}{\Delta V_{CE}} \quad (28.6)$$

where Δ denotes a small change.

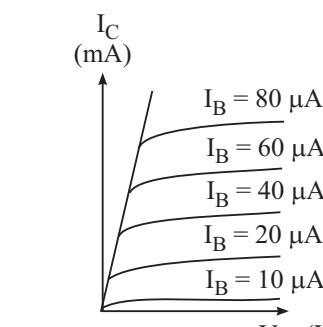


Fig. 28.25 : Output characteristics of a typical *npn* transistor in *CE* configuration

28.8.2 Common Emitter (CE) Configuration of a *pnp* Transistor

In the preceding section, you learnt to draw input and output characteristics of a *n-p-n* transistor in common emitter configuration. Now we will consider a *p-n-p* transistor. Fig. 28.26 shows the circuit diagram for *CE* characteristics of a *p-n-p* transistor. The transistor is biased to operate in the active region. The microammeter and voltmeter are used in the base-emitter circuit to measure the base current (I_B) and the voltage between base and emitter. Similarly, milliammeter and voltmeter are connected in collector-emitter circuit to measure the collector current (I_C) and voltage between collector and emitter (V_{CE}).

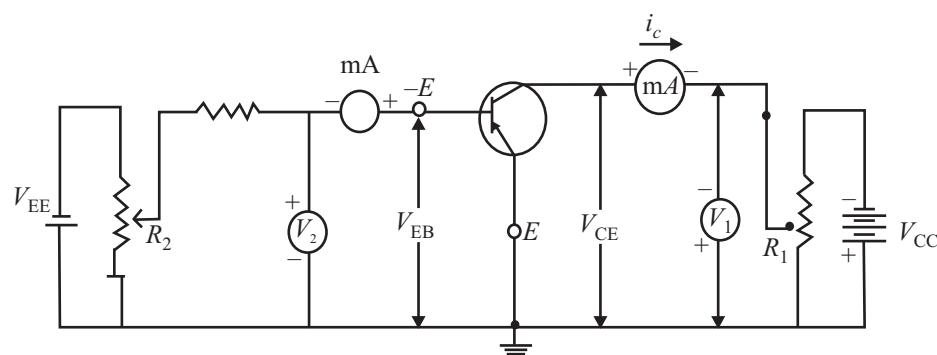


Fig. 28.26 : Circuit diagram for obtaining input and output characteristics of a *p-n-p* transistor in *CE* configuration

Input Characteristics

Input characteristics are graphs between V_{BE} and I_B at different constant values of V_{CE} .

To plot input characteristics, the potentiometer R_1 in the emitter-collector circuit is adjusted till the voltmeter shows constant value. Then potentiometer in the emitter-base circuit is adjusted in such a way that base-emitter voltage is zero. For this value, base current is also observed to be zero. Keeping the V_{CE} constant, V_{BE} is increased gradually and change in base current is noted with the help of microammeter. To plot input characteristics at $V_{CE} = -2V$, say, the potentiometer in emitter-collector circuit is adjusted till the voltmeter in the same circuit reads 2V. Then potentiometer in the emitter-base circuit is adjusted to make V_{BE} zero. Then V_{BE} is increased gradually, keeping V_{CE} constant. Similarly the input characteristics of the transistor in the CE configuration can be drawn for different values of $V_{CE} = -6V, 1V$ and so on. Fig. 28.27 shows typical input characteristics of CE configuration. As may be noted, the nature of *input characteristics* is similar to the forward characteristics of *p-n junction diode*. The base current remains zero as long as the base voltage is less than the barrier voltage (for silicon transistor, it is $\sim 0.7V$). As the base voltage exceeds barrier voltage, current begins to increase slowly and then rises abruptly.

You may also recall that these curves are similar to the ones obtained for the CE configuration for *n-p-n* transistor.

From the reciprocal of the slope of the curve of input characteristic, the a.c input resistance of the transistor can be calculated.

- **a.c input resistance (R_{in})** of the transistor in CE configuration is expressed as:

$$R_{in} = \left. \frac{\Delta V_{BC}}{\Delta I_B} \right|_{V_{CE}} = \text{constant} \quad (28.7)$$

In this configuration R_{in} is typically of the order of one $k\Omega$.

Notes

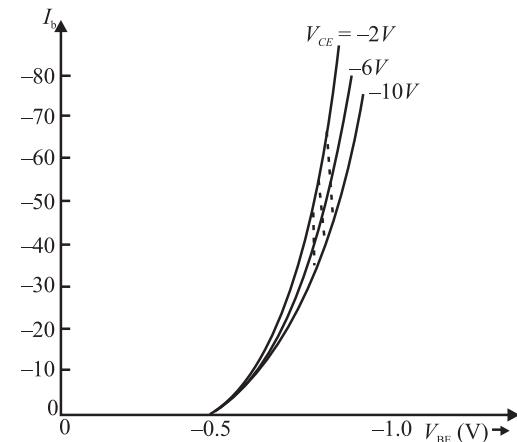


Fig. 28.27 : Input characteristics of a typical *p-n-p* transistor in *CE* configuration.

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Output Characteristics

These are graphs between collector-emitter voltage (V_{CE}) and the collector current (I_C) at different constant values of base current (I_B).

To draw these characteristics, V_{CE} is made zero and V_{BE} is adjusted till the microammeter in the base-emitter circuit is set to read a constant value. Thus V_{CE} is adjusted to make I_B constant at a particular value.

Now keeping I_B constant, V_{CE} is increased from zero in a number of steps and the corresponding collector current I_C is noted with the help of milliammeter connected in series with collector.

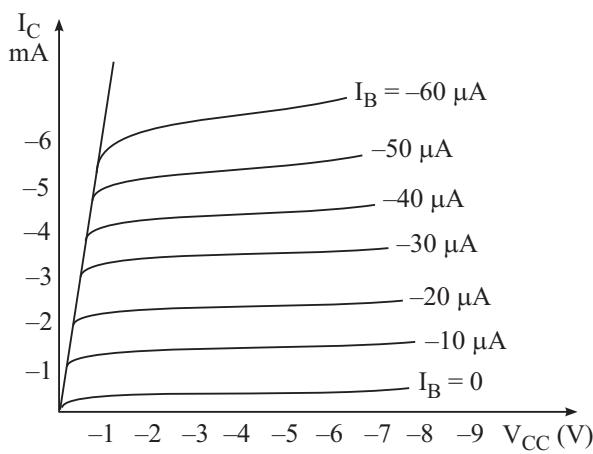


Fig 28.28 : Output characteristics of a typical *pnp* transistor in *CB* configuration

How can we plot the output characteristics at $I_B = 50 \mu A$? To do so, V_{BE} is adjusted till milliammeter reads $50 \mu A$. Increase V_{CE} gradually and note corresponding values of I_C . The graph between V_{CE} and I_C gives the output characteristics at $I_B = 50 \mu A$. Similarly, the output characteristics can be obtained at $I_B = 100 \mu A, 200 \mu A$ and so on. Fig. 28.28 shows output characteristics of *p-n-p* transistor for *CE* configuration.

Example 28.1 : Calculate the current gain β of a transistor if the current gain $\alpha = 0.98$

$$\text{Solution: } \beta = \frac{\alpha}{1-\alpha} = \frac{0.98}{1-0.98} = 49$$

Example 28.2 : In a transistor, 1 mA change in emitter current changes collector current by 0.99 mA. Determine the a.c current gain.

Solution: Given $\Delta I_e = 1 \text{ mA} = 1 \times 10^{-3} \text{ A}$ and $\Delta I_c = 0.99 \text{ mA} = 0.99 \times 10^{-3} \text{ A}$

$$\text{Therefore, a.c current gain of the transistor } \alpha = \frac{\Delta I_c}{\Delta I_e} = \frac{0.99 \times 10^{-3}}{1 \times 10^{-3}} \text{ A} = 0.99$$



INTEXT QUESTIONS 28.7

1. Fill in the blanks
 - (a) The curve relates the input current with input voltage, for a given output voltage.



Notes



WHAT YOU HAVE LEARNT

- Semiconductors are materials like silicon (Si) and germanium (Ge), which have conductivities midway between insulators and conductors.
- Semiconductors are of two types : Intrinsic (pure) and extrinsic (doped).
- Extrinsic semiconductors can be *p*-type (doped with 3rd group impurities) or *n*-type (doped with 5th group impurities).
- A *p-n* junction diode consists of a *n*-type region and a *p*-type region, with terminals on each end.
- When a *p-n* junction is formed, diffusion of holes and electrons across the junction results in a depletion region which has no mobile charges.
- The ions in the region adjacent to the depletion region generate a potential difference across the junction.
- A forward biased *p-n* junction offers low resistance to flow of electrons.
- A reverse biased *p-n* junction diode offers high resistance to flow of current.
- A *p-n* junction allows current to flow in only one direction.
- There are various types of diode e.g. photo diode light emitting diode, Zener diode and solar cell.
- A photo diode is always connected in reverse bias.
- A transistor consists of three separate regions (emitter, base and collector) and two junctions. Emitter is most heavily doped and base is the least doped. While collector has the largest size, base is the thinnest.
- Transistor can either be *n-p-n* type or *p-n-p* type.
- A transistor can be connected in any of the three configurations: common collector (*CE*), common base (*CB*) or common emitter (*CE*).
- The characteristics of a transistor vary according to the configuration of the transistor.
- *CE* configuration is preferred over other configurations as it provides high current gain and voltage gain.

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TERMINAL EXERCISE

1. Describe the most important characteristic of a $p-n$ junction diodes.
2. Explain the formation of depletion region in a $p-n$ junction diode.
3. Which charge carriers conduct forward current in a $p-n$ junction diode?
4. Differentiate between
 - (i) Forward bias and reverse bias
 - (ii) Avalanche and zener breakdown
5. Explain the working of $p-n-p$ and $n-p-n$ transistors.
6. Define current gains α and β of a transistor.
7. For $\alpha = 0.998$, calculate change in I_C if change in I_E is 4 mA.
8. What are energy bands in solids? How are they formed? How do we classify solids as conductor semiconductors and insulators on the basis of energy bands?



ANSWERS TO INTEXT QUESTIONS

28.1

- | | |
|------------------------------|-------------|
| 1. 1.5×10^{15} each | 2. (ii) |
| 3. impurity, doping | 4. majority |
| | 5. lower |

28.2

- | | | |
|--------------------------|-----------------------|----------|
| 1. (a) majority carriers | (b) depletion region | |
| (c) 0.7, 0.3 | (d) higher, electrons | |
| 2. (a) (iii), | (b) (iii), | (c) (ii) |

28.3

- | | | |
|------------------|---------------|----------------------------------|
| 3. (a) decreases | (b) increases | (c) increases, breakdown voltage |
| 4. (a) (iv); | (b) (iii) | |

28.4

- | | | |
|------------------------|---------|------------------|
| 2. (a) 0.7 V, 0.3 V; | (b) one | (c) micro ampere |
| 3. (a) (iii); (b) (ii) | | |

28.5

1. (ii), (i), (iv)
2. (a) Zener (b) reverse (c) light sensitive
 (d) group III-V (e) forward (f) emission
 (g) recombination (h) electroluminiscence
 (i) photovoltaic (j) more, absorbed
3. Separation of electrons and holes takes place due to the electric field of the depletion layer.

28.6

1. (a) (i); (b) (ii)
2. (a) Three, two; (b) Base (c) Most heavily, base
 (d) largest size, moderate (e) Emitter-base, collector-base
 (f) *npn, pnp*

28.7

1. (a) input characteristic (b) output characteristics
 (c) collector, emitter (d) base and emitter, base and collector

Answers to Problems in Terminal Exercise

7. 3.992 mA

**Notes**

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APPLICATIONS OF SEMICONDUCTOR DEVICES

In the last lesson, you learnt the working principle of semiconductor devices like *p-n* junction diode, Zener diode, LED, solar cells and transistors. Due to their miniature size and special electrical properties, these devices find applications in almost every household appliances and gadgets like gas lighter, security alarm, radio, TV, telephone, tape recorder, CD player, computer, fan regulator, emergency lights etc. All control mechanisms in big industries and flight control equipments in an aeroplane and power systems in satellites use semiconductor devices. In a way, it is now difficult to imagine life without these.

In this lesson you will learn some simple applications of diodes and transistors. This discussion is followed by an introduction to elements of **digital electronics**. This branch of electronics handles special types of signals/waveforms, which can assume only two values, 0 and 1. Digital electronics is based on the concept of **logic gates**. These gates accept input in digital form and give output according to the logic operation it is supposed to perform. You will learn about logic gates, their symbols and circuit implementation in this lesson.



OBJECTIVES

After studying this lesson, you should be able to:

- *explain the use of diode as a half-wave and a full-wave rectifier;*
- *explain the use of Zener diode as voltage regulator;*
- *describe the uses of a transistor as an amplifier, a switch and an oscillator;*
- *explain the logic gates with their Truth Tables; and*
- *realize logic gates using simple circuit elements.*

29.1 APPLICATIONS OF *p-n* JUNCTION DIODES

You now know that a *p-n* junction exhibits asymmetric electrical conduction, i.e., its resistance in forward bias is different from that in reverse bias. This property of a diode is used in rectification, i.e., conversion of an ac signal into a dc signal (of constant magnitude). In every day life, we may need it to charge a cell phone, laptop etc. Let us now learn about it.



Notes

29.1.1 *p-n* Junction Diode as a Rectifier

You have learnt in Lessons of Module 5 that the electricity supply in our homes provides us ac voltage. It is a sinusoidal signal of frequency 50 Hz. It means that voltage (or current) becomes zero twice in one cycle, i.e., the waveform has one positive and other negative half cycle varying symmetrically around zero voltage level. The average voltage of such a wave is zero. Let us now learn the mechanism to convert an ac into dc.

(a) Half-Wave Rectification

Refer to Fig. 29.1. The signal from ac mains is fed into a step down transformer *T* which makes it available at the terminals *X* and *Y*. The load resistance R_L is connected to these terminals through a *p-n* junction diode *D*. You may now like to ask : Why have we used a step down transformer? This is done due to the fact that most devices require voltage levels lower than 220V. The stepped down ac signal is obtained at the output of stepdown transformer. The potential at terminal *X* with respect to *Y* will vary as a sine function with time, as shown in Fig. 29.2(a). In the positive half cycle, during the time interval 0 to $T/2$, diode *D* will be forward biased and conduct, i.e., current flows through R_L from *A* to *B*. However, during the negative half cycle, i.e., in the interval $T/2$ to T , *D* is reverse biased and the junction will not conduct, i.e. no current flows through R_L . This is shown in Fig. 29.2(b). Since the *p-n* junction conducts only in one-half cycle of the sine wave, it acts as a half-wave rectifier.

During the non-conducting half cycle, the maximum reverse voltage appearing across the diode is equal to the peak ac voltage V_m . The maximum reverse voltage that a diode can oppose without breakdown is called its **Peak Inverse Voltage(PIV)**. For rectification, we must choose a diode having PIV greater than the peak ac voltage to be rectified by it; otherwise it will get damaged. The dc voltage, V_{dc} across R_L , as measured by voltmeter in case of half-wave rectifier, is given by

$$V_{dc} = V_m/\pi \quad (29.1)$$

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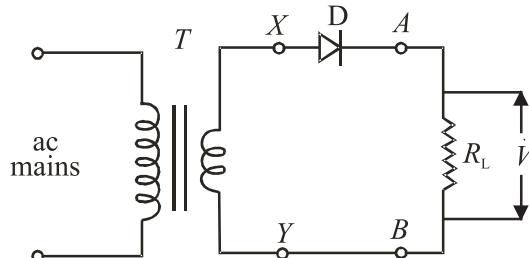


Fig. 29.1: Half wave rectifier circuit

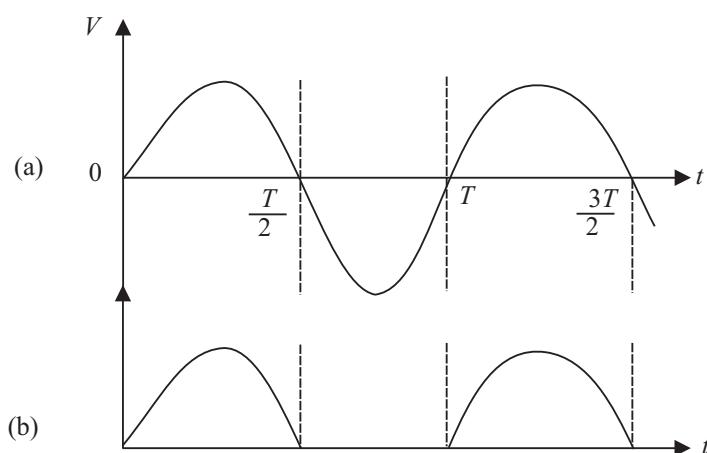


Fig. 29.2: (a) Input ac voltage, and (b) half-wave rectified output

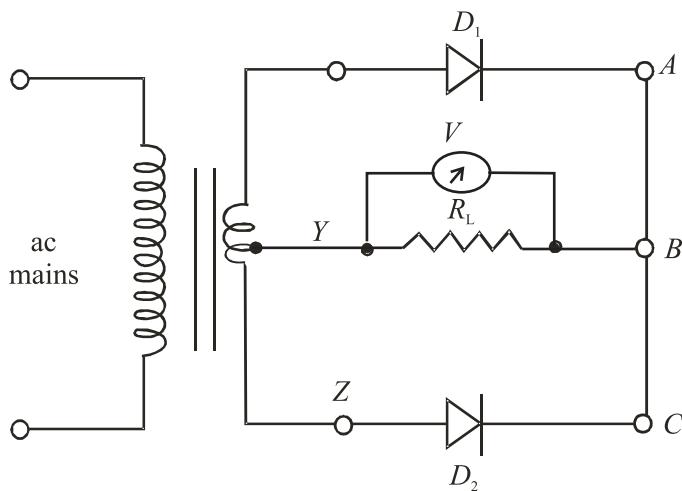
where V_m is the peak ac voltage. The dc current I_{dc} through the load resistance R_L is given by

$$I_{dc} = \frac{V_{dc}}{R_L} = \frac{V_m}{\pi R_L} \quad (29.2)$$

Note that in this case, we are utilizing only half of the input power and obviously it is not an efficient way of obtaining dc. You may logically think that instead of one, we should use two diodes in such a way that they conduct in alternate cycles. This is known as full-wave rectification. Let us learn about it now.

(b) Full-Wave Rectification

For full-wave rectification, we feed the input signal in a centre tapped step down transformer. (It has two identical secondary windings connected in series.) D_1 and D_2 are two $p-n$ junction diodes, as shown in Fig. 29.3. One end of the load resistance R_L is connected to the central point Y of the secondary windings and the other end is connected to the cathode terminals of the diodes D_1 and D_2 . The anodes of these diodes are connected respectively to the ends X and Z of the secondary windings. The potentials at the ends X and Z are in opposite phase with respect to Y , i.e., when potential of X is positive, Z will be negative and vice versa. It is shown graphically in Fig. 29.4 (a) and (b).



Notes

Fig. 29.3 : A full-wave rectifier circuit using two diodes

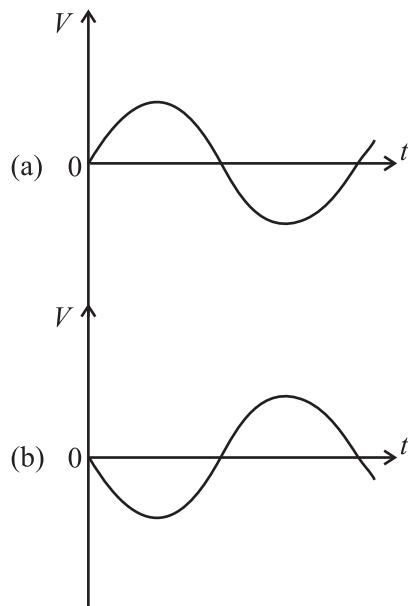


Fig. 29.4 : (a) Potential at point X is positive with respect to Y, and (b) potential of point Z is negative with respect to Y

Suppose that to start with, terminal X is positive and Z is negative with respect to Y. In this condition, diode D_1 will conduct but D_2 will not conduct. The current will flow through the load from B to Y and the output voltage across R_L is as shown in Fig 29.5(a). During the next half cycle, terminal X will be negative and Z will be positive. Under this condition, diode D_2 conducts and current will again pass through the load resistance in the same direction, that is from B to Y . The corresponding waveform is shown in Fig. 29.5(b).And the net output across R_L is pulsating , as shown in Fig. 29.5(c).

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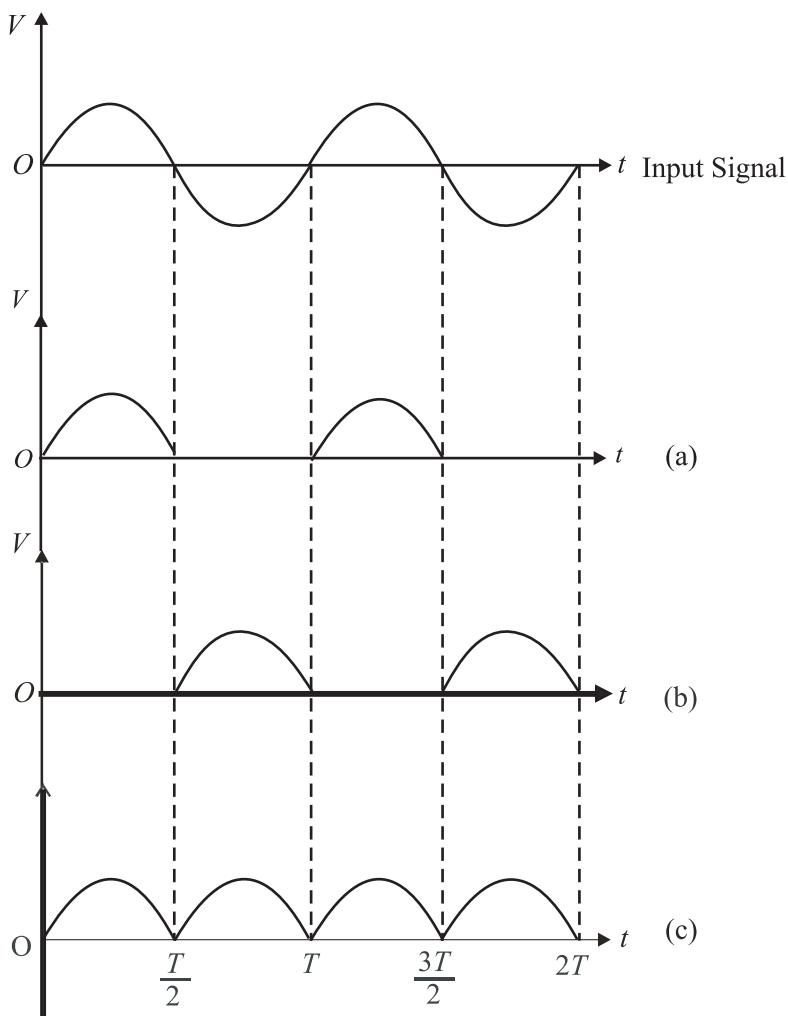


Fig. 29.5 : Voltage across R_L a) when D_1 conducts, b) D_2 conducts, c) net output of full wave rectifier

Since current through the load now flows over the entire cycle of the sine wave, this is called full-wave rectification. The dc voltage V_{dc} and dc current I_{dc} are given by

$$V_{dc} = 2 \times V_m / \pi \quad (29.3)$$

and
$$I_{dc} = \frac{V_{dc}}{R_L} = \frac{2V_m}{\pi R_L} \quad (29.4)$$

Note that the unidirectional current flowing through the load resistance after full-wave rectification pulsates from maximum to minimum (zero) and is not useful for any practical application. To reduce the fluctuating component and obtain more steady current, we filter the pulsating part. You may be eager to know as to how do we achieve this. Let us now discover answer to this important question.

Filtering

We recall that impedance offered by a capacitor to the flow of ac depends on its frequency. Therefore, a capacitor C connected across the load resistance, as shown in Fig. 29.6, filters out high frequency component.

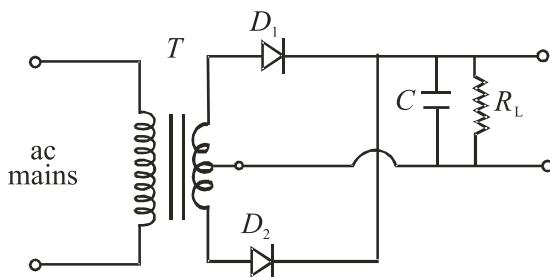


Fig. 29.6 : Circuit diagram for *capacitor-filter* in full-wave rectification

The capacitor gets charged to nearly maximum potential V_m when diode D_1 conducts for period $t = T/4$. When the current tends to decrease for $T/4 < t < T/2$, the capacitor discharges itself and tries to maintain current through the load, reducing fluctuations considerably, as shown in Fig. 29.7. The larger the value of capacitor and the load resistance, the lower will be the fluctuations in the rectified dc. The capacitor C connected across the load to reduce fluctuations is called a *filter capacitor*. In a power supply, we use LC and $C-L-C$ (or π) filters to reduce the rippling effect. You will learn about these in detail in your higher classes.

In high quality power supplies combination of inductors and capacitor $L - C - L$ or $C - L - C$ is used. Depending on the way, these components are connected these filters are called 'T' or 'Π'.

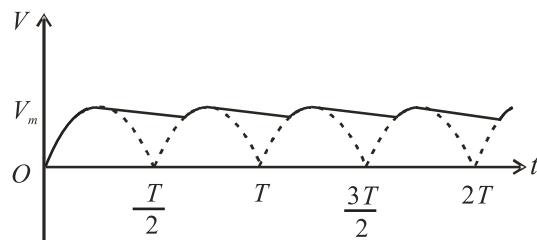


Fig 29.7: Output voltage when capacitor is used to filter ac

Special p-n junction, called Zener diode, acts as voltage regulator in reverse bias. You will now study about it.

29.1.2 Zener Diode as a Voltage Regulator

The half-and full-wave rectifiers with filters are the simplest type of power supplies. These provide almost pure dc but have one deficiency. When load current is increased by decreasing resistance, the output voltage drops. This is because, when large current is drawn, the filter capacitor gets discharged more and its voltage across the load resistor reduces. Similarly, if the ac input changes, the dc output voltage also varies. Obviously, a supply with varying output voltage affects the performance of different devices being operated with it. For example, if we operate an amplifier, the quality of sound reproduced by it will get deteriorated.

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To remove this deficiency, a Zener diode is used with simple power supplies which gives constant dc voltage. Such a circuit is called regulated power supply.

The Zener regulated voltage supply circuit is shown in Fig. 29.8. It consists of a Zener diode with breakdown voltage V_z . This will be equal to the stabilized output voltage V_o . A suitable series resistance R_s is included to control circuit current and dissipate excess voltage. The anode of Zener diode is connected to the negative terminal of input supply, and the cathode is connected in series with R_s to positive terminal of input supply, that is, the Zener is connected in reverse bias condition. The load resistance R_L is connected across the Zener diode. The Zener regulator will only operate if the input supply voltage to the regulator, V_i is greater than V_z . After breakdown, the voltage across it remains nearly constant and is independent of the current passing through it. The current I_s flowing through R_s is given by the equation

$$I_s = (V_i - V_z)/R_s \quad (29.5)$$

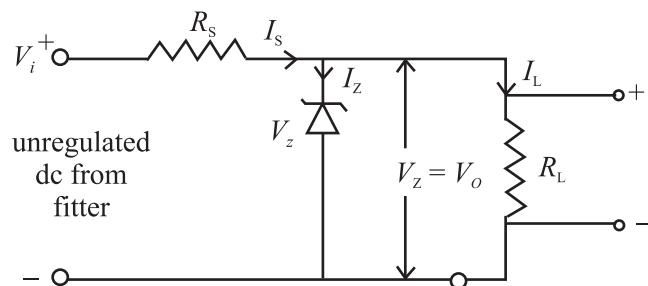


Fig. 29.8 : Zener diode as a stabilizer

This current divides in two parts: the Zener current I_z and load current I_L . Applying Kirchoff's law, we can write

$$\begin{aligned} I_s &= I_z + I_L \\ \text{or } I_z &= I_s - I_L \end{aligned} \quad (29.6)$$

For Zener diode to operate, some current $I_{z_{\min}}$ should always flow through it. Therefore, the load current I_L should always be less than the main current I_s . Typical value of $I_{z_{\min}}$ may range from 5 mA to 20 mA.

If load current is zero, the entire I_s will pass through Zener diode and output voltage V_o will be equal to V_z . When some load current is drawn, say I_L , the Zener current will decrease by the same amount but the output voltage will remain V_z . Similarly, if the ac main voltage increases or decreases, the input voltage, V_i will increase or decrease accordingly. It will result in change of I_s given by Eqn.(29.5). Due to change in I_s , the change in V_i will appear as a drop across the series resistance R_s . The Zener voltage V_z and hence V_o will remain unchanged. Thus we see that the output voltage has been stabilized against the variations in the current and the input voltage.



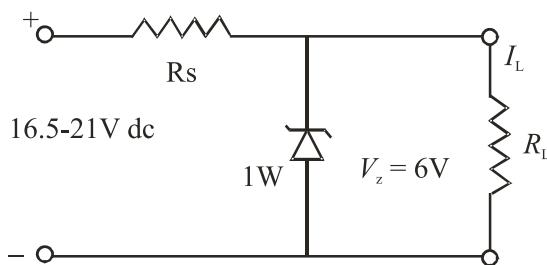
Notes

The power dissipation in Zener diode is given by the relation

$$P_d = V_z \times I_z \quad (29.7)$$

This dissipation should not exceed the maximum power dissipation rating recommended by the manufacturer for Zener diode. Let us now understand the design of a Zener regulated power supply with one example.

Example 29.1: The load current varies from 0 to 100 mA and input supply voltage varies from 16.5 V to 21 V in a circuit. Design a circuit for stabilized dc supply of 6 V.



Solution: We choose a Zener diode of 6 V. Let $I_{z_{\min}}$ be 5 mA. The maximum current will flow through the Zener when there is no load current. Its magnitude will be $(100+5)$ mA = 0.105 A.

The value of R_s is determined by the minimum input voltage and maximum required current:

$$R_s = \frac{V_{z_{\min}} - V_z}{I_{\max}} = \frac{16.5V - 6V}{105mA} = 100 \Omega$$

The current through the Zener diode will be maximum when the input voltage is maximum, that is 21 V and $I_L = 0$. Therefore, the maximum Zener current $I_{\max} = (21V - 6V)/100 \Omega \approx 0.15$ A.

The maximum power dissipation in the diode is $6V \times 0.15A = 0.9W$.

It means that we should use a Zener diode of 6 V, 1 W and resistance R_s of 100Ω . It should be connected in the circuit as shown above. It will give a stable output of 6 V for the specified ranges of load and input variation.



INTEXT QUESTIONS 29.1

1. Draw a circuit of full-wave rectifier with a filter capacitor.
2. What will be the output voltage, if you connect a Zener diode in forward bias instead of reverse bias in the regulator circuit of Example 29.1?

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29.2 TRANSISTOR APPLICATIONS

You learnt the working principle of transistor in detail in the last lesson. Normally, the collector is reverse biased and no current flows in collector-emitter circuit. If we pass a very small current in the base circuit, a very large current starts flowing in the collector circuit. This property has made a transistor indispensable for vast electronic applications. But here we have discussed its applications as an amplifier, as a switch, and as an oscillator (frequency generator).

29.2.1 Transistor as an Amplifier

An electrical signal is voltage or current, which is coded with some useful information. For example, when we speak in front of a microphone, its diaphragm vibrates and induces a very small voltage in its coil, depending on the intensity of sound. This induced voltage appears as a weak signal and can not operate a loudspeaker to reproduce sound. To make it intelligible, it is fed into a device called amplifier. The amplifier increases the level of input signal and gives out magnified output. If V_i is the input signal voltage fed to the amplifier and V_o denotes the amplified output, their ratio is called *voltage gain*.

$$i.e., \quad A_v = \frac{V_o}{V_i} \quad (29.8)$$

Similarly, we can define the current gain and power gain as

$$A_I = \frac{i_o}{i_i} \quad (29.10)$$

$$A_p = \frac{P_o}{P_i} \quad (29.11)$$

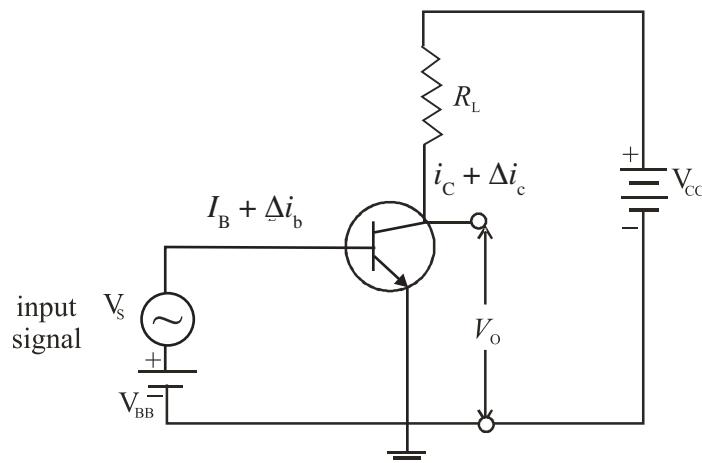


Fig. 29.9: Basic amplifier circuit using a *n-p-n* transistor in *CE* mode



Notes

The circuit for transistor as an amplifier is shown in Fig. 29.9. Here an *n-p-n* transistor is used in CE mode. Its collector is reverse biased through the load resistance R_L by the battery V_{CE} . When a base current I_B flows, some collector current I_C will start flowing. On decreasing I_B , a stage will be reached when I_C becomes almost zero. This is the lower limit of variation of I_B . Similarly, on increasing I_B again, a stage of saturation is reached and I_C stops increasing. This corresponds to the upper limit of variation of I_B . For faithful amplification of input signal, a base current equal to the mean of these two limiting values of I_B is passed through the base by forward biasing it with battery V_{BB} . We can choose the operating point in the centre of linear operating range of the transistor. This is called biasing of the base. A signal source providing an input signal v_s is connected in series with v_{BB} .

Due to addition of oscillating signal voltage v_s to v_{BB} , the base current changes by an amount Δi_b around the dc biasing current I_B . The signal voltage is kept low so that the signal current Δi_b if added and subtracted from I_B does not cross the upper and lower limits of the base current variation. Otherwise, the transistor will go into cut off or saturation region and the amplified output will be highly distorted and noisy. Note that signal current

$$\Delta i_b = v_s / r_i \quad (29.12)$$

where r_i is the input impedance. This change in base current Δi_b results in a large change in collector current, say Δi_c given by

$$\Delta i_c = \beta \Delta i_b = \beta v_s / r_i \quad (29.13)$$

where β is the ac current amplification factor, equal to $\Delta i_c / \Delta i_b$. From (Eqn. 29.13) we get

$$v_s = \Delta i_c \times r_i / \beta \quad (29.14)$$

By applying Kirchhoff's law to the output circuit in Fig. 29.9, we have

$$V_{CC} = V_{CE} + I_C R_L \quad (29.15)$$

On differentiating Eqn. (29.15), we get

$$dV_{CC} = dV_{CE} + dI_C \times R_L \quad (29.16)$$

Since V_{CC} is constant, $dV_{CC} = 0$. Therefore, we get

$$dV_{CE} = - dI_C \times R_L$$

But dV_{CE} is the change in output Δv_o and dI_C in i_c . Therefore,

$$\Delta v_o = - \Delta i_c \times R_L$$

The voltage gain A_v of the amplifier is given by

$$\begin{aligned} A_v &= v_o / v_s = - (\Delta i_c \times R_L) / (\Delta i_c \times r_i / \beta) \\ &= -\beta \times R_L / r_i \end{aligned} \quad (29.17)$$

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The ratio β/r_i is called **transconductance** of transistor and is denoted by g_m . Hence Eqn. (29.17) can be written as

$$A_v = -g_m \times R_L \quad (29.18)$$

The negative sign indicates that input and output are in opposite phase, i.e. they differ in phase by 180° . The power gain is given by

$$A_p = A_i \times A_v = \beta \times A_v \quad (29.19)$$

Note that power gain does not mean that the law of conservation of energy is violated in an amplifier. The ac power output of the amplifier is more than the ac input signal power but this gain is achieved at the cost of dc power supplied by the voltage source.

John Bardeen (1908 – 1991)



John Bardeen is the only researcher in history of science who received two Nobel Prizes in Physics. He was born in Madison, Wisconsin USA, in a highly educated family. He was so bright a kid that his parents moved him from third grade to Junior high school. He did his graduation in Electrical Engineering. But, he also had to struggle for his career. After spending three years as geophysicist with Gulf Oil Company, he went to Princeton for his Ph.D. in Mathematical Physics. After a brief stint at Harvard and Minnesota and in Naval Ordnance Labs, he joined William Shockley's research group at Bell Laboratories. With Walter Brattain, he developed the first transistor for which Bardeen, Brattain and Shockley were conferred the 1956 Nobel Prize in Physics.

Bardeen shared his second Nobel in 1972 with Leon C Cooper and R Schieffer for their theoretical work on superconductivity.



INTEXT QUESTION 29.2

1. For a CE mode amplifier, v_i is 20 mV and v_o is one volt. Calculate voltage gain.
2. The P_o of an amplifier is 200 times that P_i . Calculate the power gain.
3. For a CE amplifier, $R_L = 2000 \Omega$, $r_i = 500 \Omega$ and $\beta = 50$. Calculate voltage gain and power gain.

29.2.2 Transistor as a switch

In day-to-day life, we use electrical switches to put the gadgets like lamps, fans, machines on or off manually. Note that the switch has two distinct states, viz on



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and off. In electronics, we come across situations where we need to apply an input to some device in the form of two distinct voltage levels. This is as if we were operating a switch. When switch is on, one voltage level is applied but when switch is off, the other one is applied. Typically, such voltage levels are used in computers, where digital signals are employed. This is done by using a transistor in the non-linear region of its operation. In the transistor characteristics shown in Fig 29.10, we see two extreme regions: *cut-off region and saturation region*. The (jagged) region below the zero base ($I_B = 0$) signifies the *cut off* regions. The transistor does not conduct and entire supply voltage V_{CC} appears across the transistor between the collector and the emitter (V_{CE}). That is, the output voltage at the collector is V_{CC} .

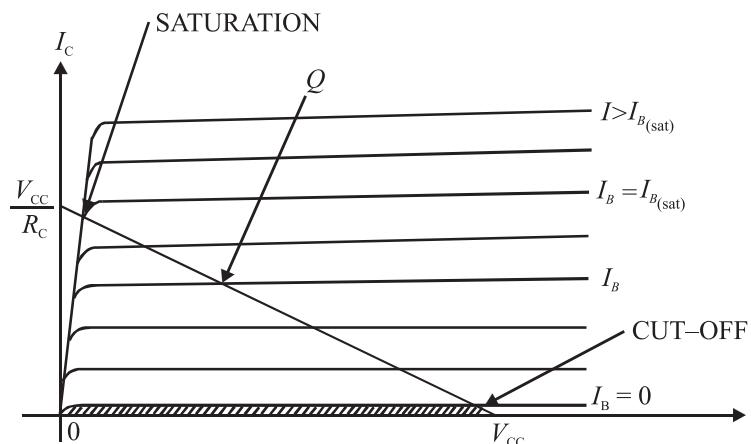


Fig. 29.10 : Transistor output characteristics

When the base current I_B is greater than its saturation value, the transistor conductor fully and collector-emitter voltage V_{CE} is almost zero. In such a case, the output voltage obtained between collector and ground is zero and entire voltage drop

appears across R_L . That is , the collector current $I_C = \frac{V_C}{R_L}$.

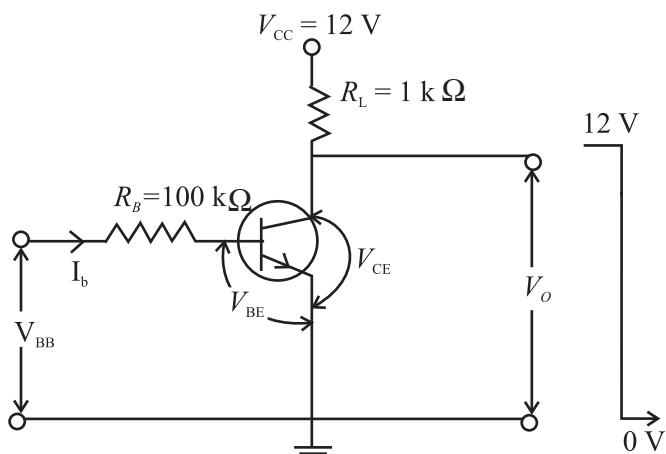


Fig. 29.11: Transistor as a switch

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Fig 29.11 shows a typical circuit of transistor as a switch. The control signal for switching the transistor on or off is given in the form of V_{BB} . For the input loop, we can write

$$I_B R_B + V_{BE} - V_{BB} = 0$$

When $V_{BB} = 0$, we get

$$I_B = -\frac{V_{BE}}{R_B} \quad (29.20)$$

Since I_B is less than zero, the transistor is cut off, and

$$V_0 = V_{CC} \quad (29.21)$$

If $V_{BB} = 5V$, and $V_{BE} = 0.7 V$ for the chosen transistor, from Eqn. (29.20) we get

$$I_B (100 \text{ k}\Omega) + 0.7V - 5 \text{ V} = 0.$$

$$\therefore I_B = \frac{5\text{V} - 0.7\text{V}}{100\text{k}\Omega} = 43 \mu\text{A}$$

For normal transistors, this value of base current is enough to drive the transistor to full saturation. In this case, $V_O = V_{CE_{sat}} = 0$ and the collector current

$$I_C = \frac{V_{CC}}{R_L} = \frac{12\text{V}}{1\text{K}\Omega} = 12\text{mA}.$$

This kind of switch can also be used as an indicator in displays. For example, if we connect an LED in series with the collector resistor, as shown in Fig 29.12, the collector current drives the LED on for high (+5V) input, and it lights up. Whenever input is zero, the LED is off because no collector current flows through the circuit.

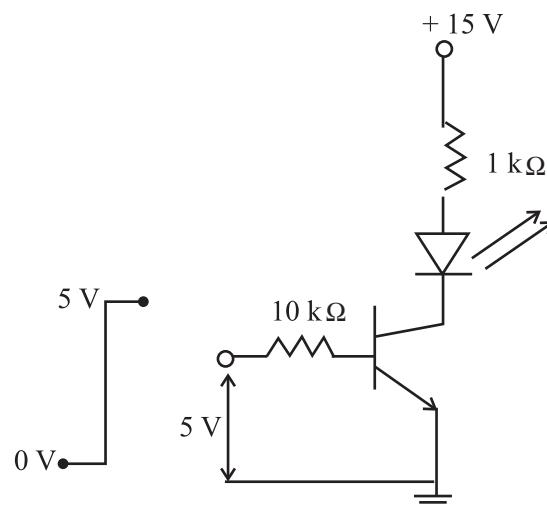


Fig. 29.12: LED indicator using transistor switch



Notes

29.2.3 Transistor as an Oscillator

An electronic oscillator is a device which generates continuous electrical oscillations. In a simple oscillator circuit, a parallel LC circuit is used as resonant circuit and an amplifier is used to feed energy to the resonant circuit. It can generate frequencies from audio to radio range depending on the choice of L and C .

We know that when a charged capacitor is connected across an inductor, the charge oscillates. But due to loss of energy by radiation and heating of wires, the energy is lost and the amplitude of oscillations decays with time. To build a sinusoidal oscillator, where the oscillations are sustained (i.e. they do not decay), we need an amplifier with positive feedback. The basic idea is to feed a part of output signal in input signal. By adjusting the gain of the circuit and the phase of the feedback signal, energy dissipated in each cycle is replenished to get sustained oscillations of desired frequency.

Schematically, we can depict an oscillator to be made up of two main blocks: an amplifier with gain A , and a feedback circuit with feedback factor β , as shown in Fig 29.13.

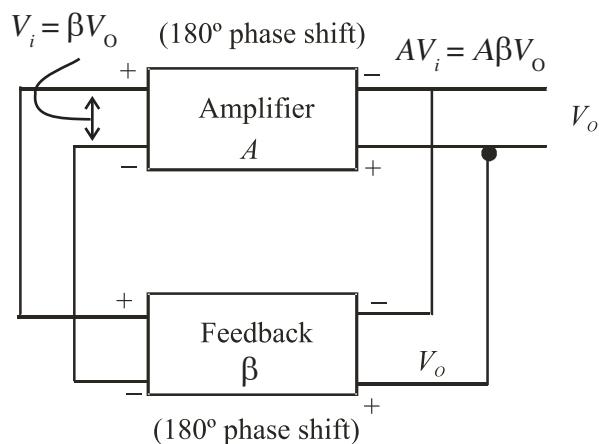


Fig. 29.13: Schematic diagram of an oscillator

In case, $A \beta < 1$, V_o decreases continuously. On the other hand, if $A \beta > 1$, V_o increases gradually. But if $A\beta = 1$, we get constant value of V_o leading to sustained oscillations.

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Now, we consider a *CE* amplifier, like the one discussed in Sec. 29.2.1. It has 180° phase difference between the input and output, i.e. it has negative gain ($-A$). To keep the total feedback gain $A\beta = 1$, we require that β is also negative; equal to $-A^{-1}$. That is, it is necessary to introduce a phase shift of 180° in the feedback circuit as well.

In Fig. 29.14, we have shown a circuit diagram of an oscillator using *LC* tank circuit and a transistor amplifier in *CE* mode. This is called Colpitt's Oscillator.

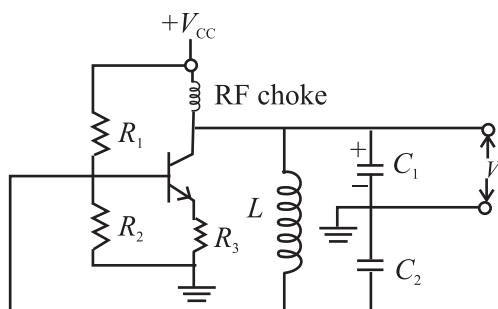


Fig. 29.14 : Colpitt's Oscillator

In this circuit C_1 , C_2 and L form the tank circuit. The oscillating current is generated in this circuit, which is at its resonant frequency. The output is obtained across C_1 , the feedback is provided across C_2 connected to the base of the transistor amplifier in *CE* mode. In this case 180° is introduced by the amplifier and another 180° phase shift is provided by the capacitor C_2 which is connected between ground and other end of the inductor coil. Hence, the total loop gain is positive. When the gain of transistor amplifier is sufficiently large at the resonant frequency, we obtain sustained oscillations at the output.

29.3 LOGIC GATES

In electronics, we come across mainly two types of waveforms. The information carried by these waveforms is called signal. When the signal takes any value within a range of amplitude at any instant of time, it is called a continuous signal. When the signal takes the value only at certain times, it is called a discrete signal. When the signal takes only particular finite number of amplitude values, it is called a **digital signal** (Fig. 29.15).

The digital signal varies in steps and typically has only two widely separated values '0' and '1'. These are called bits. Normally 0V corresponds to bit '0' and 5 V corresponds to bit '1'. Since the levels are so widely separated, any noise riding on the signal within the range of almost 2V, [(0V + 2V) for level '0' and (5V - 2V) for level '1', does not affect the signal value. Hence these signals are immune to noise. The signals used in a computer are digital. The information is

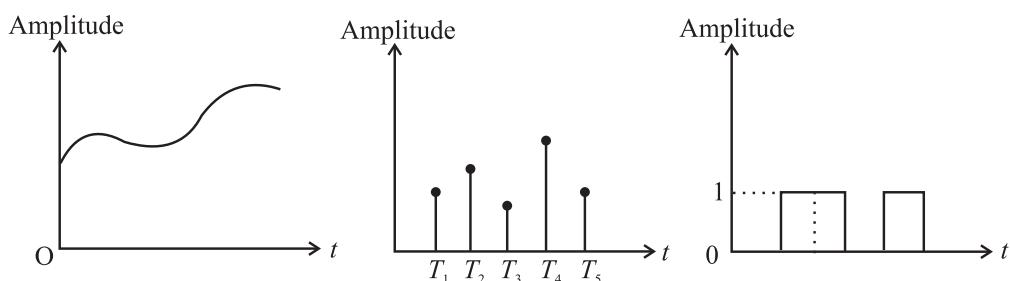


Fig. 29.15: a) continuous signal, b) discrete signal, and c) digital signal

Different mathematical operations can be performed on the digital signal. The mathematics governing these operations is called Boolean algebra.

In Boolean algebra, the basic operations are addition and multiplication. If it is a digital data that takes value 0 or 1, the following identities hold:

$$A \times 0 = 0 \quad (29.22)$$

$$A + 1 = 1 \quad (29.23)$$

The circuits which perform these operations are called *logic gates*. Let us now learn about basic logic gates.

29.3.1 Basic Logic Gates

Logic gates are devices which have one or more inputs and one output. They give different output when the input bits differ in their arrangement. The output produced by these gates follows the laws of Boolean logic. There are three basic types of logic gates :

1. AND Gate, 2. OR Gate, 3. NOT Gate

These gates perform multiplication, addition and inversion (negation) operations, respectively. Let us now learn the working of these logic gates.

1. AND Gate

An AND gate can have two or more inputs but only one output. The logic symbol of a two input AND gate is given Fig 29.16(a). We can understand the behaviour of an AND gate by considering a number of electrical switches connected in series. For examples, switches A and B are two inputs of the gate and the bulb gives the output Y. The ON switch stands for logic input '1' and OFF switch stands for logic input '0'. In this case, the bulb will glow only if it is connected to the supply voltage. This will happen only if both A and B switches are simultaneously ON (or '1'). The behaviour of output Y at various values of A and B is shown in Table in Fig. 29.16(c). This table is called *Truth Table*.



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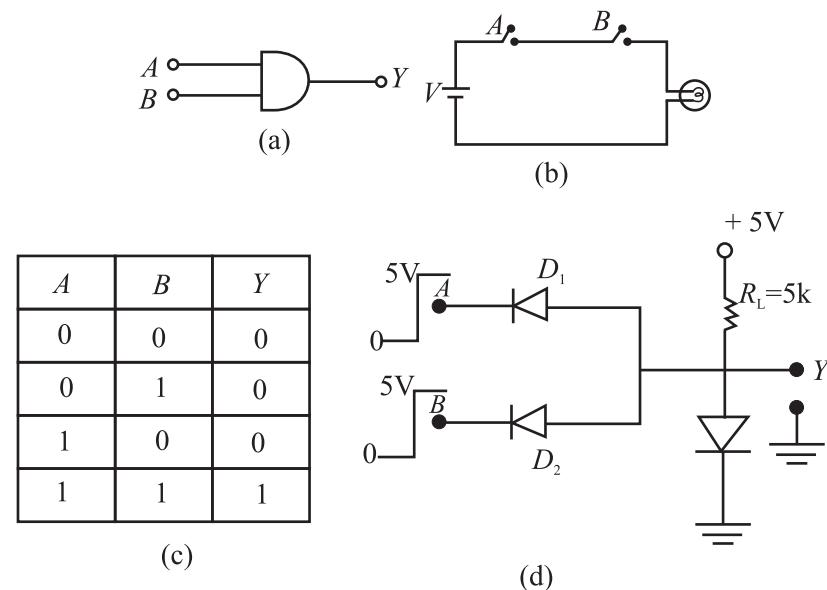


Fig 29.16: a) Symbol of AND gate, b) switch implementation of AND gate, c) Truth Table of AND Gate, and d) diode implementation of AND gate.

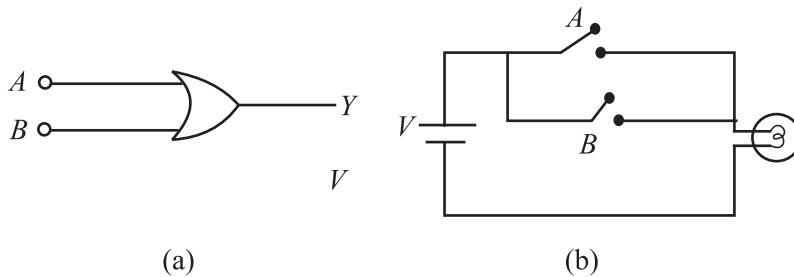
The Boolean expression for the AND operation is represented as

$$Y = A \cdot B = AB = A \times B \text{ and read as } A \text{ AND } B.$$

Realization of AND Gate : The logic gate realized by using diodes is called a DDL Gate (Diode-Diode Logic Gate). The diode implementation of a two-input AND gate is shown in Fig.29.16 (d). The anodes of two diodes D_1 and D_2 connected in parallel are forward biased by a 5 V battery through a $5\text{k}\Omega$ resistance. The output is taken from the anode. Cathode wires A and B serve as input terminals. When either A or B or both the terminals are grounded, the respective diode will conduct and a potential drop will develop across the resistance and output will be 0.7 V, i.e. logic '0'. When both the terminals are connected to 5V (i.e. for input 1, 1), neither of the diodes will conduct and output will be 5 V, i.e. logic '1'

2 OR Gate

The OR gate can have two or more inputs and only one output. The logic symbol of a two input OR gate is given in Fig 29.17(a). We can explain the behaviour of an OR gate with the help of a number of electrical switches connected in parallel. For a two input OR gate, two switches are connected, as shown in Fig.29.17(b). The switch A and B are the two inputs of the gate and the bulb gives output Y. The ON switch stands for logic input '1' and OFF switch stands for logic input '0'. The glowing bulb stands for logic output '1' and the non-glowing bulb for logic output '0'. In this case, when either A OR B or both the switches are ON, the supply voltage reaches the output and the bulb glows. The input-output correlation for an OR gate is shown in the Truth Table given in Fig. 29.17(c).



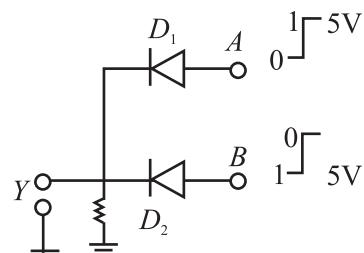
(a)

(b)

 Notes

| A | B | Y |
|---|---|---|
| 0 | 0 | 0 |
| 0 | 1 | 1 |
| 1 | 0 | 1 |
| 1 | 1 | 1 |

(c)



(d)

Fig 29.17: a) Symbol of OR gate, b) switch implementation of OR gate, c) Truth Table of OR gate, and d) diode implementation of OR gate

The Boolean expression for an OR operation is represented as

$$Y = A + B \text{ and read as } A \text{ or } B.$$

Realization of OR Gate: The diode implementation of a two-input OR gate is shown in Fig. 29.17 (d). The cathodes of diodes D₁ and D₂ connected in parallel are grounded through a 5 kΩ resistance. The output is taken from the cathode and the two anode wires A and B serve as input terminals. When either A or B or both the terminals are connected to the positive terminal of the 5 V battery, the respective diode/diodes will conduct and potential at the output will be about 5V i.e. logic '1'. When both the switches are open, output will be 0 V i.e. logic '0'.

3 NOT Gate

Another important gate used in digital signal handling is the **NOT** gate, which inverts the signal, i.e., if input is '1' then output of NOT gate is '0' and for '0' input, the output is '1'.

The symbol for NOT gate is shown in Fig. 29.18(a). The Truth Table of NOT gate is shown in fig. 29.18(b).

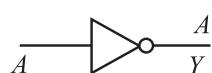
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(a)

| A | $Y = A$ |
|---|---------|
| 0 | 1 |
| 1 | 0 |

(b)

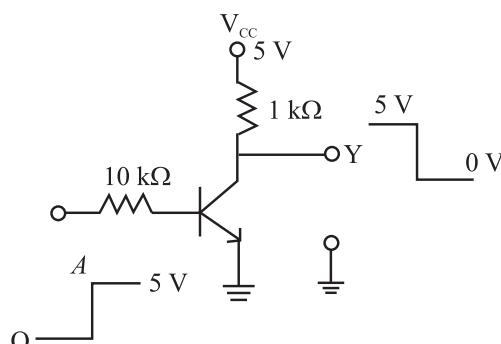


Fig. 29.18: (a) Symbol of NOT gate, (b) Truth Table of NOT gate, and (c) circuit implementation of NOT gate

The circuit to implement a NOT gate is identical to that used for a transistor as a switch. This is shown in Fig. 29.18(c). When input A is at '0' level, transistor is off and the entire V_{CC} voltage (5V) appears at the output Y . When input A is '1' (5V), the transistor conducts and output voltage Y is '0'.

The inversion operation is indicated by a bar on the top of the symbol of the input e.g. in the Truth Table we can write, $Y = \text{NOT}(A) = \bar{A}$

So far we have discussed basic logic gates. You may now ask: Can we combine these to develop other logic gates? You will discover answer to this question in the following section.

29.3.2 Combination Logic Gates

Two most important gates formed by combination of logic gates are (1) NAND [NOT+AND] and (2) NOR [NOT+OR] gates. In digital electronics, a NAND gate or a NOR gate serves as a building block because use of multiple number of either of these gates allows us to obtain OR, AND and NOT gates. For this reason, these are called universal gates. Let us now learn about combination logic gates.

1. NAND Gate

The NAND Gate is obtained by combining AND gate and NOT gate, as shown in Fig. 29.19 (a). Here the output Y of AND gate is inverted by the NOT gate to get the final output Y . The logic symbol of a NAND gate is shown in Fig. 29.19(b). The Truth Table of a NAND gate is given in Fig. 29.19(c). It can be obtained by inverting the output of an AND gate. The truth table of a NAND gate shows that it gives output '1' when at least one of the inputs is '0'. The Boolean expression of a NAND operation is represented as

$$Y = A \cdot B = A \times B = AB$$



Notes

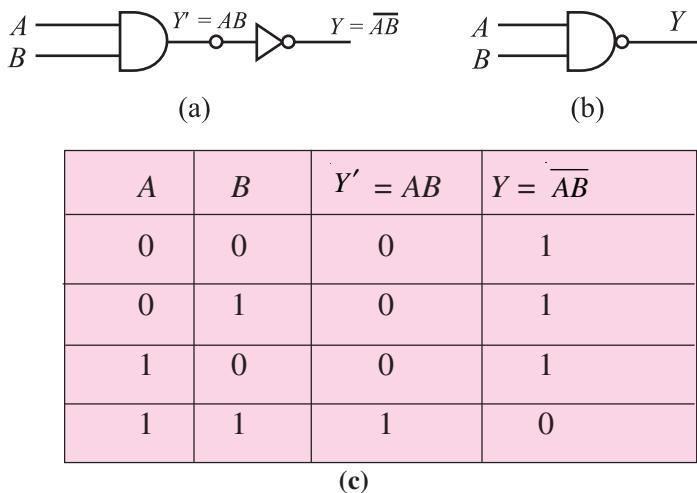


Fig. 29.19 : a) NAND as combination logic gate, b) symbol of NAND GATE, and c) Truth Table of a NAND gate

2. NOR Gate

The NOR gate, obtained by combining an OR gate and NOT gate, is shown in Fig. 29.20(a). Here the output of OR gate, Y' , is inverted by the NOT gate to get the final output Y . The logic symbol of a NOR gate is given in Fig. 29.20(b). The Truth Table of a NOR gate given in Fig. 29.20(c), can be arrived at by inverting the output of an OR gate. The Truth Table of a NOR gate shows that it gives output ‘1’ only when both the inputs are ‘0’

The Boolean expression for a NOR operation is represented as $Y = \overline{A + B}$.

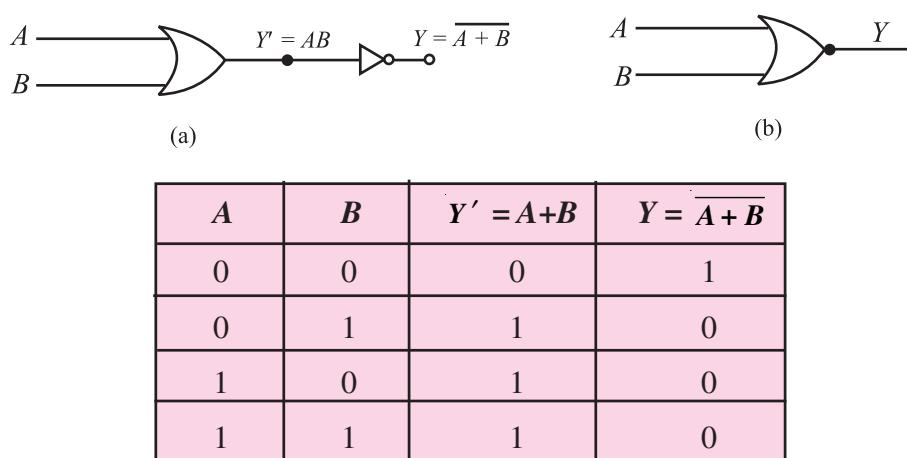


Fig. 29.20 : a) NOR as combination logic gate, b) symbol of NOR gate, and c) Truth Table of NOR gate

As mentioned earlier, the NAND and NOR gates are basic building blocks of all the logic gates. Let us now see, how we can obtain the three basic gates AND, OR and NOT by using NAND gates.

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29.3.3 Realization of Basic Gates from NAND Gate

The NAND gate is considered to be the universal gate because all other gates can be realized by using this gate.

(a) Realization of a NOT gate : If two input leads of a NAND gate are shorted together, as shown in Fig. 29.21, the resulting gate is a NOT gate. You can convince yourself about this by writing its truth table.

Here we have $A = B$



Fig. 29.21 : NAND gate as NOT gate

(b) Realization of an AND gate : The AND gate can be realized by using two NAND gates. The output of one NAND gate is inverted by the second NAND gate used as NOT gate as shown in Fig 29.22(a). The combination acts as an AND gate, as is clear from the Truth Table given in Fig. 29.22(b).

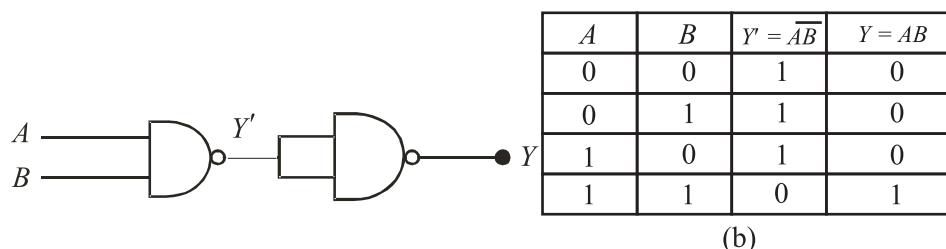


Fig. 29. 22: a) NAND gates connected to implement AND gate and
b) Truth Table of AND gate using NAND gate

c) Realization of an OR gate : The OR gate can be realized by using three NAND gates. Two NAND gates are connected as inverters and their outputs are fed to the two inputs of a NAND gate, as shown in Fig. 29.23. The combination acts as an OR gate.

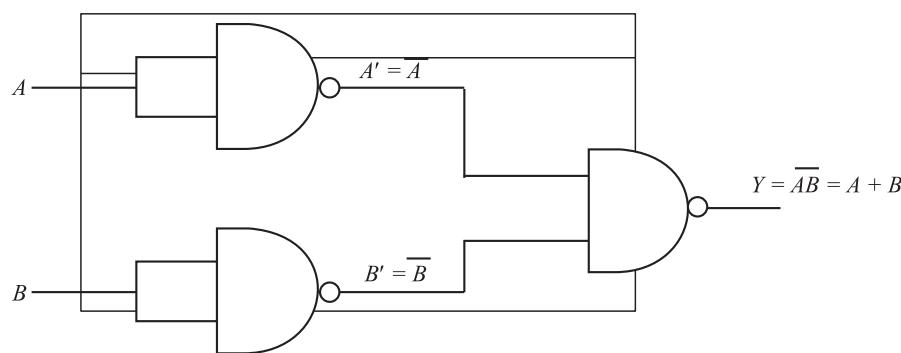


Fig. 29.23 : Three NAND gates connected as OR gate



INTEXT QUESTIONS 29.3

Complete the following table from Fig. 29.23 to prove that it is an OR gate.

| A | B | A' | B' | Y |
|---|---|------|------|---|
| 0 | 0 | — | — | — |
| 0 | 1 | — | — | — |
| 1 | 0 | — | — | — |
| 1 | 1 | — | — | — |



Notes



WHAT YOU HAVE LEARNT

- A $p-n$ junction diode can be used as a rectifier to convert ac into dc.
- A half-wave rectified dc contains more ac component than the full-wave rectified dc.
- A Zener diode stabilizes the output of a power supply.
- In a stabilizer, the Zener diode dissipates more power when the current taken by the load is less.
- For amplification, a transistor needs input current.
- Transistor can be used as a switch by biasing it into saturation and cut-off regions.
- There are three basic logic gates: AND, OR and NOT.
- NAND gate is a universal gate because it can be used to implement other gates easily.



TERMINAL QUESTION

1. Why the Peak Inverse Voltage (PIV) of a $p-n$ junction diode in half-wave rectifier with filter capacitor is double of that without the capacitor?
2. Explain how a Zener diode helps to stabilize dc against load variation.
3. What should be the range of variation of amplitude of input signal for proper working of an amplifier?
4. Draw a circuit using diodes and transistors to implement a NOR gate.

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ANSWERS TO INTEXT QUESTIONS

29.1

1. See Fig.29.6
2. In case of full wave rectifier, both diodes D_1 and D_2 charge C to maximum voltage of V_{\max} in alternate half cycles. Hence, the PIV of the diodes should be $2 \times V_{\max}$.
3. $R_z = 100\Omega$, $R_s = 100\Omega$ and $R = R_z + R_s = 200\Omega$
Hence,

$$I = \frac{21}{200} = 0.105A$$

$$\text{and } V = IR = 0.105 \times 100 \\ = 10.5V$$

29.2

$$1. |A_v| = \frac{V_o}{V_i} = \frac{1V}{20\text{ mV}} = 50$$

$$2. A_p = \frac{P_o}{P_i} = 200$$

$$3. |A_v| = \frac{\beta \times R_L}{r_i} = \frac{50 \times 2000\Omega}{500\Omega} = 200$$

$$A_p = \beta A_v = 50 \times 200 = 10000.$$

29.3

| A | B | A' | B' | Y |
|-----|-----|------|------|-----|
| 0 | 0 | 0 | 0 | 0 |
| 0 | 1 | 1 | 1 | 1 |
| 1 | 0 | 1 | 0 | 1 |
| 1 | 1 | 1 | 1 | 1 |

30**Notes**

COMMUNICATION SYSTEMS

Communication is a basic characteristic of all living beings. Communication entails transmitting and receiving information from one individual/place to another. In the world of animals, communication is made by mechanical, audio and chemical signals. You may have observed how sparrows begin to chirp loudly on seeing an intruder, who can put their life in danger. However, human beings are blessed with very strong means of communication – speech. We can express what we see, think and feel about whatever is happening around us. That is to say, we use sound (an audible range, 20Hz - 20kHz) and light (in visible range, 4000 Å – 7000 Å), apart from mechanical (clapping, tapping) and opto-mechanical signals (nodding, gesturing), for communication. You must realise that language plays a very significant role in making sense out of spoken or written words. It comes naturally to us. Prior to the written alphabet, the mode of communication was oral. The second era of communication began with the invention of printing press. Invention of the telegraph in the early nineteenth century marked the beginning of the third stage. Revolutionary technological developments enabled as rapid, efficient and faithful transfer of information. Using tools and techniques such as telegraph, fax, telephone, radio, mobiles, satellites and computers, it is possible to communicate over long distances. The oceans and mountain ranges no longer pose any problem and the constraints of time and distance seem to be non-existent. On-line learning, (education), publishing (research), banking (business) which were topics in science fiction not too long ago, are now routine activities. In fact, combination of computers with electronic communication techniques has opened a very powerful and fertile field of information and communication technologies (ICT_s).

Have you ever thought about the technology that has made all this development possible? You will discover answers to this question in this lesson.

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OBJECTIVES

After studying this lesson, you will be able to:

- list the elements of a long distance communication system;
- explain the terms analogue and digital signals;
- describe how electromagnetic waves act as carriers of information;
- Specify the bandwidth of signals (speech, TV and digital data);
- list various transmission media and state bandwidths specific to them;
- explain importance of ground, sky and space wave propagation;
- state need for modulation; and
- explain production and detection of amplitude modulation wave.

30.1 A MODEL COMMUNICATION SYSTEM

Communication systems endeavour to transmit information from

- one to one, i.e., point-to-point communication;
- one to many, i.e., broadcast communication; and
- many to many, i.e., telephone conference call or a chat room.

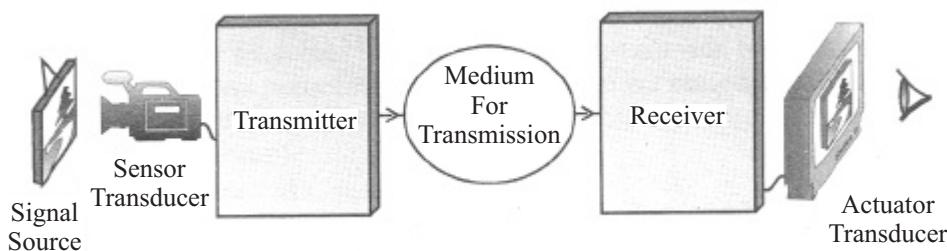


Fig. 30.1 : A schematic arrangement for the communication system.

In a typical modern day communication system, the information is in the form of electrical signals (voltage or current), spread over a range of frequencies called the **signal bandwidth**. (Some **noise** gets added to the signal and tries to obscure the desired information.) For scientific analysis of any system, we model the system into its basic components. You will now learn about these.

30.1.1 Elements of a Communication System

Refer to Fig. 30.1. It shows building blocks of a typical communication system. As may be noted, the essential elements of a communication system are:



Notes

30.2 TYPES OF SIGNALS – ANALOGUE AND DIGITAL

You now know that communication of information involves use of signals, which are classified on the basis of their origin and nature. Accordingly we have

- continuous time (analog) and discrete time (digital) signals;
- coded and uncoded signals;
- periodic and aperiodic signals;
- energy and power signals; and
- deterministic and random signal.

Of these, we will consider only analog and digital systems. The sound produced by human being in conversation/interaction or photograph are converted into continuously varying electrical analog signal [Fig. 30.2(a)]. But in modern electronic communication systems, these are converted into discrete form, which has finite values at different instances of time and zero otherwise [Fig. 30.2 (b), (c)] from Fig. 30.2, you will note that the waveforms used to represent correspond to a particular frequency and are periodic; while one of these is sinusoidal, the another is pulsed. In fact these may be viewed as a sub-class of sine and square waveforms.

Information can be packaged in both analog (or continuous) and digital (or discrete) forms. Speech, for example, is an analog signal which varies continuously with time. In contrast, computer files consist of a symbolic “discrete-time” digital signal.

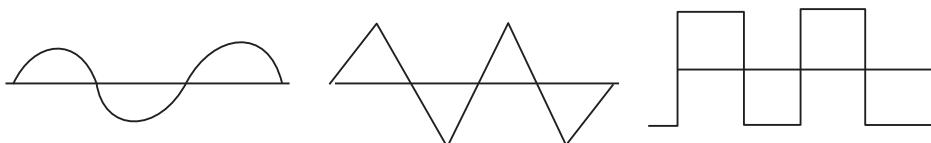


Fig. 30. 2 : Examples of (a) continuous (sinusoidal) and (b) discrete signals.

In the digital format, signals are in the form of a string of **bits** (abbreviated from **binary digits**), each bit being either ‘ON’ or ‘OFF’ (1 or 0). The **binary** system

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refers to a number system which uses only two digits, 1 and 0 (as compared to the **decimal** system which uses ten digits from 0 to 9). We can convert all information-bearing signals into discrete-time, amplitude-quantised digital signals. In a compact disc (CD), the audio is stored in the form of digital signals, just as a digital video disc (DVD) stores the video digitally.

Communication systems can be either fundamentally analog, such as the amplitude modulation (AM) radio, or digital, such as computer networks. Analog systems are in general, less expensive than digital systems for the same application, but digital systems are more efficient give better performance (less error and noise), and greater flexibility.

30.3 BAND WIDTH OF SIGNALS

The most crucial parameter in communication systems is the signal bandwidth, which refers to the frequency range in which the signal varies. However, it has different meaning in analog and digital signals. While analog bandwidth measures the range of spectrum each signal occupies, digital bandwidth gives the quantity of information contained in a digital signal. For this reason, analog bandwidth is expressed in terms of frequency, i.e. Hz, the digital bandwidth is expressed in terms of bits per second (bps). The frequency range of some audio signals and their bandwidths are given in Table 30.1. Note that human speech has bandwidth of nearly four kilo hertz. The bandwidth is about 10kHz in amplitude modulated (AM) radio transmission and 15kHz in frequency modulated (FM) transmission. However, the quality of signal received from FM broadcast is significantly better than that from AM. The bandwidth of a video signal is about 4.2MHz and television broadcast channel has bandwidth of 6MHz. The bandwidth of a typical modem, a device used for communication of digital signals over analog telephone lines, are 32kbps, 64 kbps or 128 kbps.

Table 30.1: Typical audiobandwidths

| Source | Frequency range(H_E) | Bandwidth (kHz) |
|----------------------------------|--------------------------|-----------------|
| Guitar | 82–880 | ... 0.8 |
| Violin | 196–2794 | ... 2.6 |
| Vowels (a,e,i,o,u) consonants | 250–5000 | ... 4.0 |
| Telephone signal | 200–3200 | ... 3.0 |



Notes

30.3.1 Electromagnetic Waves in Communication

In communication, we use different ways to transport the electrical signal from the transmitter to the receiver. From Modules on electricity and magnetics, you may recall that current passes through a metal conductor in the form of current signal or voltage drop, through air in the form of electromagnetic radiation or converted into light signal and sent through an optical fibre. Irrespective of the mode transmission of signal is governed by the classical theory of electromagnetic wave propagation, given by Maxwell.

As the name suggests, e.m. waves consist of electric and magnetic fields, which are inseparable. An electric field varying in time produces a space-time varying magnetic field, which, in turn, produces electric field. This mutually supporting role results in propagation of electromagnetic waves according to e.m. laws. The pictorial representation of a plane e.m. wave is shown in Fig. 30.3.

Mathematically we can express these as $E = E_0 \sin(kz - \omega t)$ and $H = H_0 \sin(kz - \omega t)$. The direct experimental evidence for the existence of e.m. waves came in 1888 through a series of brilliant experiments by Hertz. He found that he could detect the effect of e.m. induction at considerable distances from his apparatus.

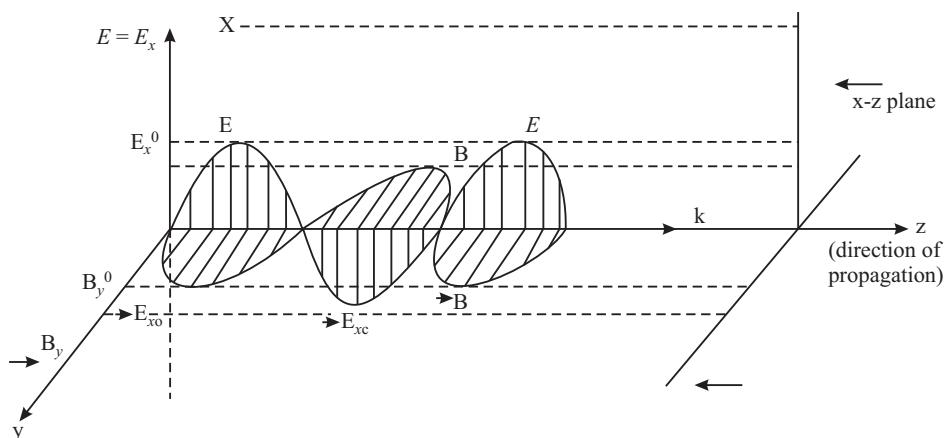


Fig. 30. 3: Propagation of electromagnetic waves

By measuring the wavelength and frequency of e.m. waves, he calculated their speed, which was equal to the speed of light. He also showed that e.m. waves exhibited phenomena similar to those of light. The range of wavelengths, as we now know is very wide from radio waves (λ is 1m to 10m) to visible light (400nm) as shown in Fig. 30.4. This generated a lot of interest and activity. In 1895 Indian physicist Jagadis Chandra Bose produced waves of wavelength in the range 25mm to 5m and demonstrated the possibility of radio transmission. This work was put to practical use by Guglielmo Marconi who, succeeded in transmitting e.m. waves across the Atlantic Ocean. This marked the beginning of the era of communication using e.m. waves. Marconi along with Carl Ferdinand Braun, received the 1909 Nobel Prize in physics for their work on wireless telegraphy.

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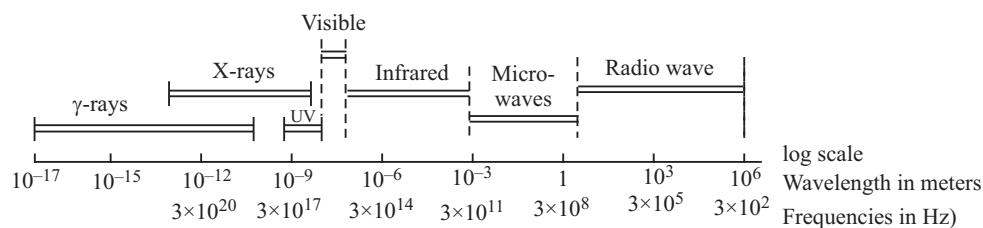


Fig. 30.4: The electromagnetic spectrum: The wave values of length correspond to vacuum (or air) The boundaries between successive regions of the spectrum are not sharply defined.

In a communication system, a transmitter radiates electromagnetic waves with the help of an antenna. These waves propagate in the space and captured by the receiver. At the receiver, another antenna extracts energy (information) from the electromagnetic waves. Now we use radio waves for different purposes television (TV) broadcasts, AM (amplitude modulated) and FM (frequency modulated) radio broadcasts, police and fire radios, satellite TV transmissions, cell phone conversations, and so on. Each such signal uses a different frequency, and that is how they are all separated.

You will learn the details of the mechanism of these transmissions and working of some common communication devices in the following two lessons. In Table 30.2, we have listed internationally accepted electromagnetic spectrum relevant for radio and TV broadcast, popular band names, and their application.

(Frequencies ν in Hz are related to wavelengths λ in m in vacuum through the relationship $c = \nu\lambda$, where $c = 3 \times 10^8$ m/s is the speed of electromagnetic waves in vacuum.)

Table 30.2: Radio frequency bands

| Band | Frequency Range | Wavelength Range | Application |
|-------------------------------|-----------------|------------------|--|
| Extremely Low Frequency (ELF) | < 3 kHz | > 100 km | Mains electricity |
| Very Low Frequency (VLF) | 3 - 30 kHz | 100 – 10 km | SONAR |
| Low Frequency (LF) | 30 - 300 kHz | 10 – 1 km | Marine navigator |
| Medium Frequency (MF) | 300 kHz - 3 MHz | 1 km – 100 m | Medium wave radio |
| High Frequency (HF) | 3 - 30 MHz | 100 – 1 m | short wave radio |
| Very High Frequency (VHF) | 30 – 300 MHz | 10 – 1 m | FM radio |
| Ultra High Frequency (UHF) | 300 MHz – 3 GHz | 1 m – 10 cm | commercial, TV, Radio, Radar |
| Super High Frequency (SHF) | 3 – 30 GHz | 10 – 1 cm | Satellite communiction, cellular mobile, commercial TV |



Notes

AM radio is broadcast on bands, popularly known as the Long wave: 144 - 351 kHz (in the LF), the Medium wave: 530 - 1,700 kHz (in the MF), and the Short wave: 3 – 30 MHz (HF). **Medium wave** has been most commonly used for commercial AM radio broadcasting. **Long wave** is used everywhere except in North and South Americas, where this band is reserved for aeronautical navigation. For long- and medium-wave bands, the wavelength is long enough that the wave diffracts around the curve of the earth by ground wave propagation, giving AM radio a long range, particularly at night. **Short wave** is used by radio services intended to be heard at great distances away from the transmitting station; the far range of short wave broadcasts comes at the expense of lower audio fidelity. The mode of propagation for short wave is ionospheric.

Table 30. 3 : Frequency ranges for commercial FM-radio and TV broadcast

| Frequency Band | Nature of Broadcast |
|----------------|--|
| 41 – 68 MHz | VHF TV |
| 88 – 104 MHz | FM Radio |
| 104 – 174 MHz | S Band (Sond-erkanal meaning Special Channel) for cable TV networks |
| 174 – 230 MHz | VHF TV |
| 230 – 470 MHz | H (Hyper) Band for cable TV networks |
| 470 – 960 MHz | UHF TV |

Frequencies between the broadcast bands are used for other forms of radio communication, such as walkie talkies, cordless telephones, radio control, amateur radio, etc.

You must have read about Internet enabled mobile phones and Internet Protocol Television. Have you ever thought as to which technology is enabling such empowerment? Is it fibre optic communication? Does laser play any role? You will learn answers to all such questions in the next unit.



INTEXT QUESTIONS 30.1

- What is an electromagnetic wave?
- Calculate the wavelength of a radio wave of frequency of 30 MHz propagating in space.
- What is the frequency range of (i) visible light, (ii) radio waves?

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Jagadis Chandra Bose (1858 – 1937)

Jagadis Chandra Bose, after completing his school education in India, went to England in 1880 to study medicine at the University of London. Within a year, he took up a scholarship in Cambridge to study Natural Science at Christ's College – one of his lecturers at Cambridge, Professor Rayleigh had a profound influence on him. In 1884 Bose was awarded B.A degree by Cambridge university and B.Sc degree by London University. Bose then returned to India and took teaching assignment as officiating professor of physics at the Presidency College in Calcutta (now Kolkata). Many of his students at the Presidency College were destined to become famous in their own right. Satyendra Nath Bose who became well known for his pioneering work on Bose-Einstein statistics and M.N. Saha who gave revolutionary theory of thermal ionisation, which enabled physicists to classify the stars into a few groups.



In 1894, J.C. Bose converted a small enclosure adjoining a bathroom in the Presidency College into a laboratory. He carried out experiments involving refraction, diffraction and polarization. To receive the radiation, he used a variety of junctions connected to a highly sensitive galvanometer. He developed the use of *galena* crystals for making receivers, both for short wavelength radio waves and for white and ultraviolet light. In 1895, Bose gave his first public demonstration of radio transmission, using these electromagnetic waves to ring a bell remotely and to explode some gunpowder. He invited by Lord Rayleigh, to give a lecture in 1897. Bose reported on his microwave (2.5 cm to 5 mm) experiments to the Royal Institution and other societies in England. But Nobel prize alluded him probably for want of vivid practical application of this work by him. By the end of the 19th century, the interests of Bose turned to response phenomena in plants. He retired from the Presidency College in 1915, and was appointed Professor Emeritus. Two years later the Bose Institute was founded in Kolkata. Bose was elected a Fellow of the Royal Society in 1920.

30.4 COMMUNICATION MEDIA

There are two types of communication channels: wireline (using guided media) or wireless (using unguided media). *Wireline* channels physically connect the transmitter to the receiver with a “wire,” which could be a twisted pair of transmission lines, a coaxial cable or an optical fibre. Consequently, wireline channels are more private and comparatively less prone to interference than wireless channels. Simple wireline channels connect a single transmitter to a single



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receiver, i.e., it is a point-to-point connection. This is most commonly observed in the telephone network, where a guided medium in the form of cable carry the signal from the telephone exchange to our telephone set. Some wireline channels operate in the broadcast mode, i.e., one or more transmitters are connected to several receivers, as in the cable television network.

Wireless channels are much more public, with a transmitter antenna radiating a signal that can be received by any antenna tuned close by. In radio transmission, the wireless or unguided propagation of radio waves from the transmitter to the receiver depends on the frequency of the electromagnetic waves. As you will learn in this lesson, the waves are transmitted as ground (or surface) waves, sky waves, or space waves by direct line-of-sight using tall towers, or by beaming to artificial satellites and broadcasting from there. Wireless transmission is flexible endowed with the advantage that a receiver can take in transmission from any source. As a result, desired signals can be selected by the tuner of the receiver electronics, and avoid unwanted signals. The only disadvantage is that the interference and noise are more prevalent in this case.

For transmitting em signals, we use microwave frequencies, you may recall that it varies from 1GHz to 300GHz. This frequency range is further divided into various bands. Indian satellite INSAT – 4C operates in the C band (4 – 8 GHz), whereas Edusat operates in Ku bond (12–18 GHz).

30.4.1 Transmission lines

For guided signal transmission, a transmission line – a material medium forms a path. As such, the construction of a transmission line determines the frequency range of the signal that can be passed through it. Fig. 30.5 shows some typical transmission lines. The simplest form of transmission line is a pair of parallel conductors separated by air or any dielectric medium. These are used in telephony. However, such lines tend to radiate, if the separation between the conductors is nearly half of the frequency corresponding to the operating frequency. This may lead to noise susceptibility, particularly at high frequencies, and limit their utility. To overcome this problem, we use a *twisted pair of wires*. These are used in computer networking.

At high signal frequencies ($\leq 3\text{GHz}$) we minimise radiation losses by using *coaxial cables*, where one conductor is hollow and the second conductor is placed inside it at its centre throughout the length of the cable. These conductors are separated by dielectric spacer layers of polythylene and the electric field is confined in the annular space in between the conductors. These cables are used for carrying cable TV signals. It is important to note that ideally the dielectrics should have infinite resistance. But in practice, their resistance is finite and that too decreases

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with frequency. As a result, even coaxial cables are useful in a limited range (upto a maximum of 40GHz when special dielectric materials are used). Beyond 40GHz, we use *waveguides*. However, for frequencies greater than 300GHz, their dimensions become too small (is 4mm or so) and it presents practical problems. Above this frequency, we use optical fibres for guided wave transmission.

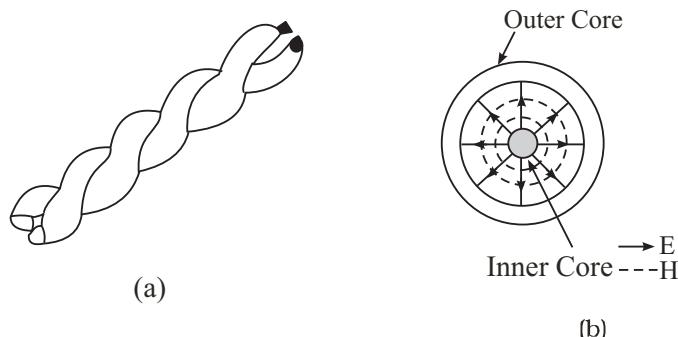


Fig. 30.5: (a) A twisted pair (b) A coaxial cable

30.4.2 Optical Fibre

The 1960 invention of the **laser** (acronym for Light Amplification by Stimulated Emission of Radiation) completely revolutionized communication technology. The laser, which is a highly coherent source of light waves, can be used as an enormously high capacity carrier wave for information carrying signals (voice, data or video) transmitted through an **optical waveguide**, such as an **optical fibre**. The basic principle involved in all long distance communication systems is **multiplexing**, i.e., simultaneous transmission of different messages over the same pathways. To illustrate it, let us consider transmission of an individual human voice. The frequency band required for transmitting human voice extends from $v_1 = 200\text{Hz}$ to $v_2 = 4000\text{ Hz}$, i.e., the information contained in this frequency band can be transmitted in any band whose width is $v_1 - v_2 = 3800\text{ Hz}$, regardless of the region of the spectrum in which it is located. Higher frequency regions have far more room for communication channels, and hence, have a much greater potential capacity than the lower frequencies. The frequency corresponding to the visible optical region at 600 nm is $5 \times 10^{14}\text{ Hz}$, while that at a wavelength of 6 cm is $5 \times 10^9\text{ Hz}$. Thus, the communication capacity of visible light in an optical fibre is about 100,000 times greater than that of a typical microwave in a metallic conductor.

The most extensively used optical waveguide is the step-index optical fibre that has a cylindrical central glass or plastic core (of refractive index n_1) and a cladding of the same material but slightly (about 1%) lower refractive index (n_2). There is usually an outer coating of a plastic material to protect the fibre from the physical environment (Fig. 30.6).

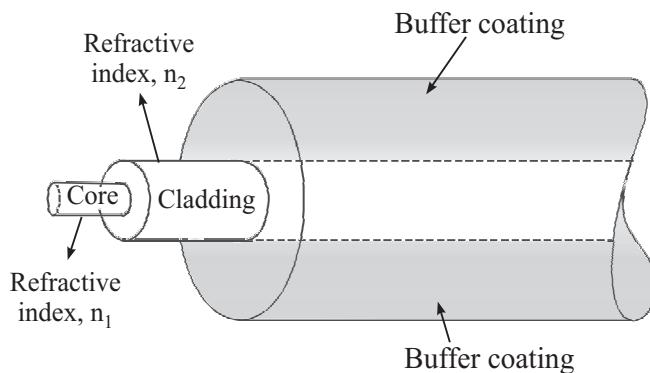


Fig. 30.6: A typical optical fibre with a doped silica core and a pure silica cladding.

When light from the core (n_1) is incident on the interface of the cladding ($n_2 < n_1$), the *critical angle* of incidence for *total internal reflection* is given by $\theta_c = \sin^{-1}(n_2/n_1)$. Thus in an optical fibre, the light ray is made to enter the core such that it hits the core-cladding interface at an angle $\theta_i > \theta_c$. The ray then gets guided through the core by repeated total internal reflections at the upper and lower core-cladding interfaces. You may recall from wave optics that when a plane wave undergoes total internal reflection, a wave propagates in the cladding (rarer medium) along the interface, with its amplitude decreasing exponentially away from the interface. The entire energy of the wave in the core is reflected back, but there is a power flow along the interface in the cladding. Such a wave is called an *evanescent wave*, and is extensively used in integrated optics for the coupling the energy of a laser beam into a thin film waveguide (Fig. 30.7)

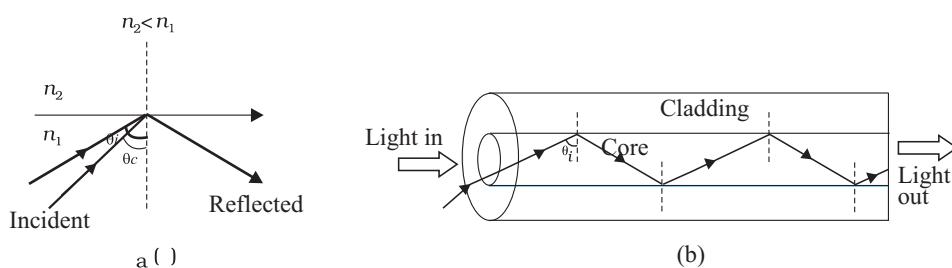


Fig. 30.7: (a) Total internal reflection (b) Ray confinement in actual optical fibre



INTEXT QUESTIONS 30.2

1. What is a coaxial cable? Write down its frequency range of operation.
2. State the basic principle used for guiding light in an optical fibre.

30.5 UNGUIDED MEDIA

The wireless communication between a transmitting and a receiving station utilising the space around the earth, i.e. atmosphere is called *space communication*. The

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earth's atmosphere plays a very interesting role the propagation of e.m. waves from one place to another due to change in air temperature, air density, electrical conductivity and absorption characteristics with height. For example, most of the radiations in infrared region are absorbed by the atmosphere. The ultraviolet radiations are absorbed by the ozone layer.

Five layers are considered to play main role in communication:

- *C layer* at about 60km above the surface of earth reflects e.m. waves in the frequency range 3kHz – 300kHz. It is therefore used for direct long range communication.
- *D layer* at a height of about 80km reflects e.m. waves in the low frequency range (3kHz – 300kHz) but absorbs waves in the medium frequency range (300 kHz – 3MHz) and high frequency range (3 – 30MHz).
- *E layer* at a height of about 110km helps in propagation of waves in the medium frequency range but reflects waves in the high frequency range in the day time.
- *F₁ layer* at a height of about 180 km lets most of the high frequency waves to pass through.
- *F₂ layer* (at a height of 300 km in daytime and 350 km at night) reflects e.m. waves upto 30MHz and allows waves of higher frequencies to pass through.

You may recall from your earlier classes that, based on the variation of temperature, air density and electrical conductivity with altitude, the atmosphere is thought to be made up of several layers. The atmospheric layer close to the earth called the *troposphere* extends up to about 12 km above the sea level. The temperature in troposphere vary between 290K (at the equator) to 220K (at tropopause). The air density is maximum but electrical conductivity is the least compared to other layers. The next layer up to about 50 km is called the *stratosphere*. An ozone layer is in the lower stratosphere extends from about 15 km to about 30 km. The layer above the stratosphere and up to about 90 km is called the *mesosphere*. The minimum temperature in mesosphere is about 180K. Beyond mesosphere upto 350km, there is a zone of ionised molecules and electrons called the *ionosphere*. In ionosphere, temperature increases with height to about 1000K. The ionosphere affects the propagation of radio waves. It is divided into D, E, F and F₂ regions based on the number density of electrons, which increases with height from about 10^9 m^{-3} in D region to 10^{11} m^{-3} in E region and 10^{12} m^{-3} in F₂ layer¹. These variations in temperature, density and conductivity arise due to different absorption of solar radiations at different heights and changes in composition etc.

The essential feature of space communication is that a signal emitted from an antenna of the transmitter has to reach the antenna of the receiver. Depending on the frequency of radio wave, it can occur as *ground wave*, *space wave*, *sky wave* and via satellite communication. Let us now learn about these.



Notes

30.5.1 Ground Wave Propagation

In ground *wave* propagation, the electromagnetic waves travel along the surface of the earth. These can bend around the corners of the objects but are affected by terrain. A vertical antenna is used to transmit electromagnetic waves. If electric field E is vertical, and the magnetic field B is horizontal, the direction of propagation k is horizontal but perpendicular to both E and B vectors. The material properties of the ground, such as its conductivity refractive index and dielectric constant, are seen to control propagation of such waves. That is why ground waves propagation is much better over sea than desert. In practice, ground waves are rapidly attenuated due to scattering by the curved surface of the earth. A larger wavelength results in smaller attenuation. That is, ground waves are more useful as lower frequencies & constitute the only way to communicate into the ocean with submarines. Moreover, this mode of propagation is suitable for short range communication. For these reasons, ground wave propagation is used for radio wave (300kHz – 3MHz) transmission.

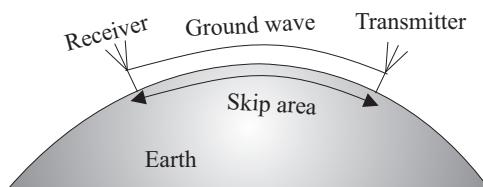


Fig. 30.8: Ground wave propagation

30.5.2 Sky Wave or Ionospheric Propagation

In sky *wave* or **ionospheric** propagation, the electromagnetic waves of frequencies between 3MHz – 30MHz launched by a transmitting antenna travel upwards, get reflected by the ionosphere and return to distant locations. In this mode, the reflecting ability of the ionosphere controls the propagation characteristics of the sky wave. The ionosphere acts as an invisible electromagnetic “mirror” surrounding the earth – at optical frequencies it is transparent, but at radio frequencies it reflects the electromagnetic radiation back to earth.

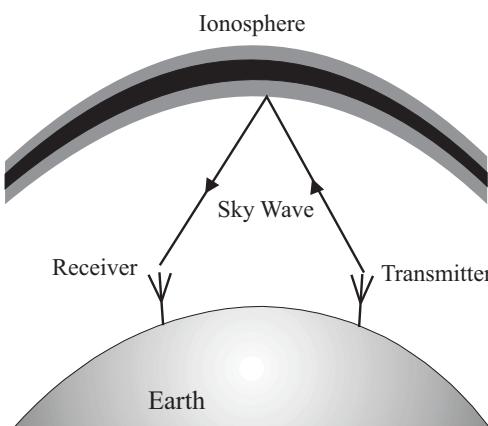


Fig. 30.9: Skywave propagation

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The maximum distance along the surface of the earth that can be reached by a single ionospheric reflection ranges between 2010 and 3000 km depending on the altitude of the reflecting layer. The communication delay encountered with a single reflection ranges between 6.8 and 10 ms, a small time interval. This mode of propagation is used for long-distance (short wave) communication in the frequency range approximately between 5 and 10 MHz. Above 10 MHz, the waves pass through the ionosphere and do not reflect back to the earth. It is, however, subject to erratic daily and seasonal changes due to variations in the number density and height of the ionized layers in the ionosphere. The composition of the ionosphere at night is different than during the day because of the presence or absence of the sun. That is why international broadcast is done at night because the reflection characteristics of the ionosphere are better at that time.

30.5.3 Space Wave Propagation

You may have seen very high antennas at radio station. These are used for broadcasting. In space wave propagation, some of the VHF radio waves (30 MHz – 300MHz) radiated by an antenna can reach the receiver travelling either directly through space or after reflection by the curvature of the earth. (Note that earth reflected waves are different from ground waves.)

In practice, direct wave mode is more dominant. However, it is limited to the so-called *line-of-sight* transmission distances and curvature of earth as well as height of antenna restrict the extent of coverage.

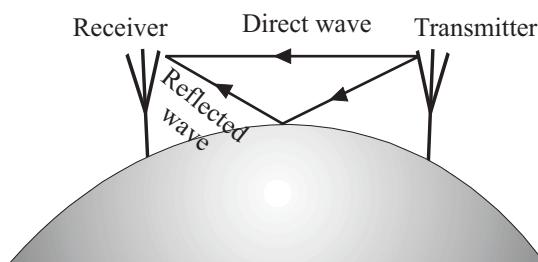


Fig. 30.10: Space-wave propagation

So far you have learnt that ground waves suffer conduction losses, space waves have limitations due to line of sight and sky waves penetrate the ionisation beyond a certain frequency. Some of these difficulties were circumvented with the launch of communication satellites in the 1950s. Satellite communication has brought about revolutionary changes in the form and format of transmission and communication. We can now talk in real time at a distance. Let us now learn about it.

30.5.4 Satellite Communication

The basic principle of satellite communication is shown in Fig. 30.11. The modulated carrier waves are beamed by a transmitter directly towards the satellite.



Notes

The satellite receiver amplifies the received signal and retransmits it to earth at a different frequency to avoid interference.

These stages are called uplinking and down-linking.

As we have seen already in connection with communication with light waves, the capacity of a communication channel can be increased by increasing the frequency of communication. How high up can we go in frequency? You now know that the ionosphere does not reflect waves of frequencies above 10 MHz, and for such high frequencies we prefer space wave propagation with direct transmission from tall towers. But this line-of-sight transmission also has a limited range or reach. Hence for long-range wireless communication with frequencies above 30 MHz, such as for TV transmission in the range of 50-1000 MHz, communication through a satellite is used.

The gravitational force between the earth and the satellite serves as the centripetal force needed to make the satellite circle the earth in a *freefall* motion at a height of about 36,000 km. An orbit in which the time of one revolution about the equator exactly matches the earth's rotation time of one day is called a *geostationary* orbit, i.e., the satellite appears to be stationary relative to the earth. Ground stations transmit to orbiting satellites that amplify the signal and retransmit it back to the earth. If the satellites were not in geostationary orbits, their motion across the sky would have required us to adjust receiver antenna continually. Two other orbits are also currently being used for communication satellites: (i) *polar circular orbit* at a height of about 1000 km almost passing over the poles (i.e., with an inclination of 90°), and (ii) *highly elliptical inclined orbit* (with an inclination of 63°) for communications in regions of high altitudes.

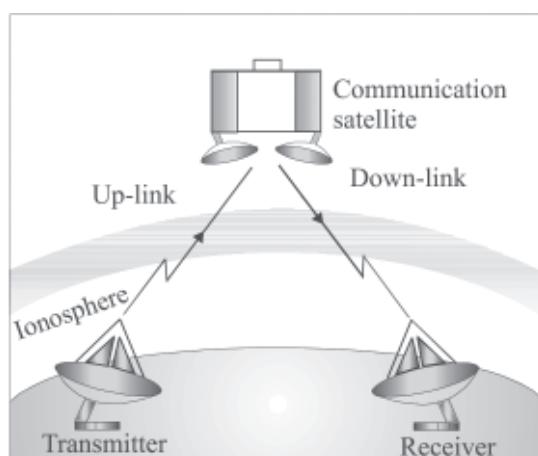


Fig. 30.11: Satellite communication.

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INTEXT QUESTIONS 30.3

1. Why do you hear some radio stations better at night than in the day?
2. Choose the correct option in each case:
 - (a) Frequencies in the UHF range normally propagate by means of
 - (i) Ground Waves
 - (ii) Sky Waves
 - (iii) Surface Waves
 - (iv) Space Waves.
 - (b) Satellites are used for communication
 - (i) With low (< 30 MHz) frequencies and for a small range
 - (ii) With low (< 30 MHz) frequencies and for a long range
 - (iii) With high (> 30 MHz) frequencies and for a small range
 - (iv) With high (> 30 MHz) frequencies and for a long range.

EDUSAT

The Indian Space Research Organisation (ISRO), Department of Space, Government of India, launched an exclusive education satellite EDUSAT in Sept. 2004. The satellite has its footprints all over the country and operates in KU band. It is designed to provide services for seven years. This satellite has capability for radio and TV broadcast, Internet-based education, data broadcasting, talk-back option, audio-video interaction, voice chat on Internet and video conferencing. It has opened up numerous possibilities: a teacher of a leading educational institution in a city may video-conference with students of a remote school, or school drop-outs in villages may receive Internet-based education support and get back into mainstream education system. EDUSAT has the capability of telecasting 72 channels. A large number of networks have been created by state governments and national institutions including NIOS. Such networks are being successfully used to impart education even in regional languages.

30.6 MODULATION – ANALOGUE AM AND FM, DIGITAL (PCM)

The process of processing a signal to make it suitable for transmission is called *modulation*. Most of the information-bearing signals in day-to-day communication are audio signals of frequency less than 20 kHz. For small distances, we can form direct link. But it is not practical to transmit such signals to long distances. This is because of the following two reasons:



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- The signal should have an antenna or aerial of size comparable to the wavelength of the signal so that the time variation of the signal is properly sensed by the antenna. It means that for low-frequency or long-wavelength signals, the antenna size has to be very large.
- The power carried by low frequency signals is small and can not go far. It is because of continuous decline or attenuation due to absorption/radiation loss. It means that for long distance transmission high frequencies should be used. But these can not carry useful information. We are therefore confronted with a situation analogous to the following:

On a front port, Indian army spots advancing enemy forces. To minimise loss of life and save the post from falling to enemy, they need reinforcement from the base camp. But by the time an army jawan goes, conveys the message and the reinforcement reaches, the port would have fallen. Therefore, it wants a carrier, say a horse, which can run fast. But the horse can not deliver the message. The way out is: Put the jawan on the horseback; let the horse run and jawan convey the message.

For signal transmission, audio signal acts as jawan and high (radio) frequency acts as the horse (carrier). So we can say that by superimposing a low frequency signal on a high frequency carrier wave, we process a signal and make it suitable for transmission. We convert the original signal into an electrical signal, called the *base band signal* using a signal generator. Next we superimpose the base band signal over carrier waves in the modulator. The change produced in the carrier wave is known as modulation of the carrier wave and the message signal used for modulation is known as *modulating signal*. The carrier wave can be continuous or pulsed. Since a sinusoidal wave, is characterised by amplitude, frequency and phase it is possible to modulate (i.e. modify) either of these physical parameter.

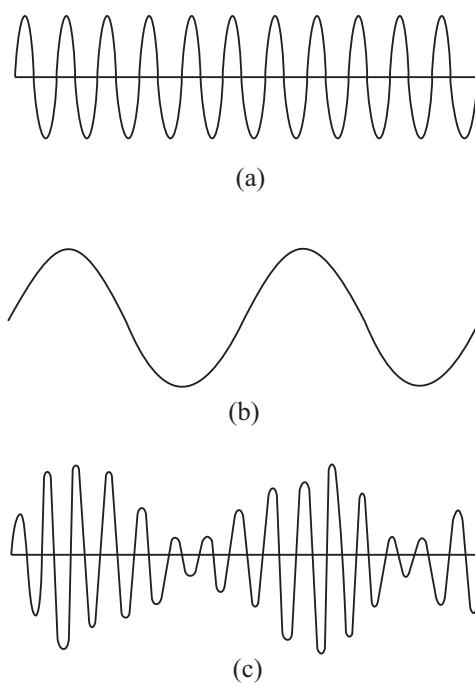


Fig. 30.12: Modulation of a carrier wave by a modulating signal: (a) a sinusoidal carrier wave of high frequency, (b) a modulating signal (message or information signal) of low frequency, (c) amplitude modulated carrier wave.

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This is known as analog modulation. There are different types of analog modulation: **Amplitude Modulation (AM)**; **Frequency Modulation (FM)**; and **Phase Modulation (PM)**, respectively. For pulsed carrier waves, **Pulse Code Modulation (PCM)** is the preferred scheme.

In **Amplitude modulation**, the amplitude of a high-frequency carrier wave (Fig. 30.12a) is modified in accordance with the strength of a low-frequency audio or video modulating signal (Fig. 30.12.b). When the amplitude of the modulating wave increases, the amplitude of the modulated carrier also increases and vice-versa — the envelope of the modulated wave takes the form depending on the amplitude and frequency of modulating signal (Fig. 30.12.c).

To understand this, we write expressions for instantaneous amplitudes of audio signal and carrier wave:

$$v_a(t) = v_{ao} \sin \omega_a t \quad (30.1a)$$

$$\text{and} \quad v_c(t) = v_{co} \sin \omega_c t \quad (30.1b)$$

where ω_a and ω_c are the angular frequencies and v_{ao} and v_{co} denote of audio and carrier waves, respectively. denote the amplitudes. In amplitude modulation the modulating (audio) signal is superimposed on the carrier wave, so that the amplitude of the resultant modulated wave can be expressed as

$$\begin{aligned} A(t) &= v_{co} + v_a(t) = v_{co} + v_{ao} \sin \omega_a t \\ &= v_{co} \left[1 + \frac{v_{ao}}{v_{co}} \sin \omega_a t \right] \end{aligned} \quad (30.2)$$

Hence the modulated wave can be expressed as

$$v_c^{\text{mod}}(t) = A \sin \omega_c t = v_{co} \left[1 + \frac{v_{ao}}{v_{co}} \sin \omega_a t \right] \sin \omega_c t \quad (30.3)$$

From Eqn. (30.3) we note that the instantaneous amplitude of the modulated wave is determined by the amplitude and frequency of the analog audio signal. The ratio v_{ao}/v_{co} gives us a measure of the extent to which carrier amplitude is varied by the analog modulating signal and is known as amplitude modulation index. We will denote it by m_a . In terms of modulation index, we can rewrite Eqn. (30.3) as

$$\begin{aligned} v_c^{\text{mod}} &= v_{co} (1 + m_a \sin \omega_a t) \sin \omega_c t \\ &= v_{co} \sin \omega_c t + v_{co} m_a \sin \omega_a t \sin \omega_c t \\ &= v_{co} \sin \omega_c t + \frac{v_{co} m_a}{2} \cos(\omega_c - \omega_a)t - \frac{v_{co} m_a}{2} \cos(\omega_c + \omega_a)t \end{aligned} \quad (30.4)$$



Notes

From Eqn. (30.4) we note that

- the modulated wave shown in Fig. 30.12(c) has three components. The first term represents carrier wave the second term whose frequency is lower than that of the carrier wave, constitutes the lower side band, and the third term with frequency higher than the carrier wave is the upper side band; and
- the frequency of the modulating signal is not directly contained in the amplitude modulated wave.

If the modulating signal in an AM system is given by

$v_a = 4\sin 6283t$ and frequency of the lower side band is 3.5×10^5 Hz, the angular frequency of the carrier wave is given by

$$\begin{aligned}\omega_c &= \omega_a + 2\pi \times (3.5) \times 10^5 \\ &= 6283 + 22 \times 10^5 \\ &= (2200 + 6.283) \times 10^3 \text{ rad} \\ &= 2.206 \times 10^6 \text{ rad}\end{aligned}$$

It is important to appreciate that the most efficient information transfer takes place when maximum power transmitted by the communication system is contained in the side bands.

The block diagram of a basic analog AM transmitter is shown in Fig. 30.13 (a). The oscillator provides a fixed frequency and the power amplifier modulates the signal.

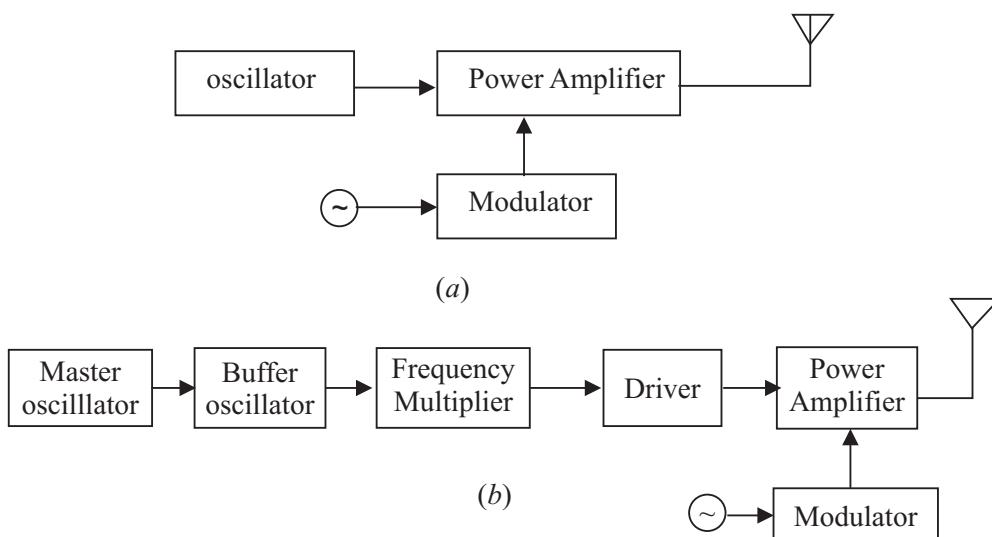


Fig. 30.13 Block diagram of (a) a basic and (b) practical AM transmitter

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For any broadcast, the maximum power that can be radiated is controlled by the GOI. It is in the range 500W to 50kW for radio transmitters. Every broadcaster is allocated a definite frequency, which has to be observed strictly to avoid interference with other signals. To ensure this, undesirable frequencies are filtered out by using coupling circuits. We will not go into these details further.

The most popular form of radio communication in India over the past 50 years had been medium wave (520 – 1700kHz) and short wave (4.39 – 5.18MHz; 5.72 – 6.33MHz) analog AM broadcast. It continues to have the widest spread, though analog FM broadcast is now being preferred because of better quality. Moreover, radio waves are now comparatively free and private broadcasters are also entering the field in a big way. FM radio stations are also being created by educational institutions for education as well as empowerment of rural youth and homemakers. In TV transmission, audio is frequency modulated whereas the video (picture) is amplitude modulated.

In **frequency modulation**, the amplitude of the carrier wave remains constant, but its frequency is continuously varied in accordance with the instantaneous amplitude of the audio or video signal. When the amplitude of the modulating signal voltage is large, the carrier frequency goes up, and when the amplitude of the modulating signal is low, the carrier frequency goes down, i.e., the frequency of the FM wave will vary from a minimum to a maximum, corresponding to the minimum and maximum values of the modulating signal (Fig. 30.14).

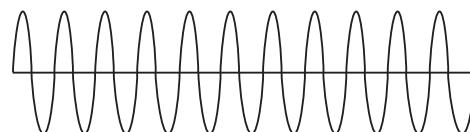


Fig. 30.14: Frequency modulated carrier wave

An FM Transmitter essentially contains an oscillator, whose frequency of the carrier is varied depending on the input audio signal. (It is usually accomplished by varying capacitance in an LC oscillator or by changing the charging current applied to a capacitor, for example, by the use of a reverse biased diode, since the capacitance of such a diode varies with applied voltage.) After enhancing the power of the modulated signal, it is fed to the transmission antenna. Low-frequency radio broadcast stations use amplitude modulation, since it is a simple, robust method.

Phase modulation involves changing the phase angle of the carrier signal in accordance with the modulating frequency. Analog pulse modulation is either amplitude modulated or time modulated. Similarly, digital pulse modulation is of two types: pulse code modulation and pulse delta modulation.



Notes

30.7 DEMODULATION

The modulated signal carrying the information, once radiated by the antenna, travels in space. Since there are so many transmitting stations, thousands of signals reach our antenna.

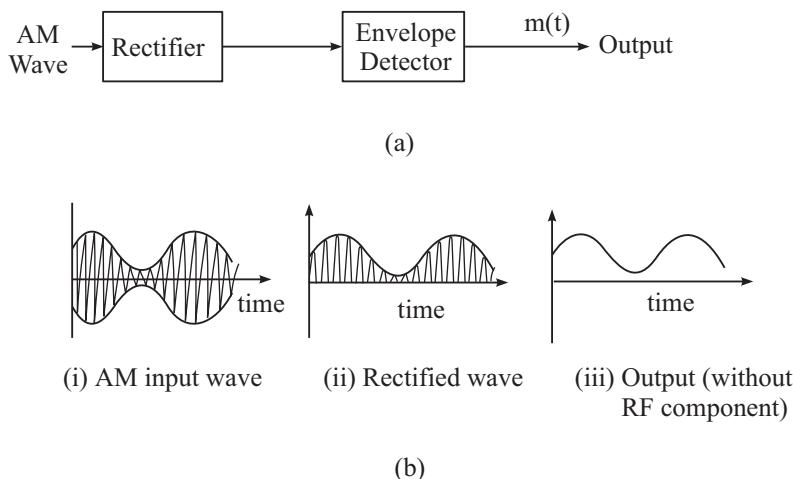


Fig. 30.15: (a) Block diagram of a detector for AM signal. (b) (i) AM modulated input wave (ii) Rectified wave (iii) output demodulated wave

We have to choose the desired signal and decouple the carrier wave and the modulating signal. This process is known as *demodulation*. The modulated signal of the form shown in Fig. 30.15(a) is passed through 30.15(a) a rectifier to produce the output shown in Fig. 30.15(b). This envelop of the signal (b) is the message signal. In order to reveal the message signal, the signal is passed through a envelope detectors (demodulator) which may consists of a simple RC circuit.



INTEXT QUESTIONS 30.4

- Choose the correct option in each case:
 - Modulation is used to
 - reduce the bandwidth used

MODULE - 8

Semiconductors Devices
and Communication



Notes

Communication Systems

- (ii) separate the transmissions of different users
(iii) ensure that information may be transmitted to long distances
(iv) allow the use of practical antennas.
- (b) AM is used for broadcasting because
- it is more noise immune than other modulation systems
 - it requires less transmitting power compared to other systems
 - it avoids receiver complexity
 - no other modulation system can provide the necessary bandwidth for faithful transmission.
2. What is the optimum size of a radio antenna.

COMMUNICATION APPLICATIONS

In recent years, the world of communication has advanced rapidly from printed texts to the telegraph, the telephone, the radio, the television, mobiles, Internet and computer conferencing (Audio and video). Countries all over the world are striving to achieve high standards of national and international communications. Radio and TV broadcasting through communication satellites is routinely achieved to reach out to the majority of the population even in remote corners of the globe. The domestic system of automatic telephone exchanges is usually connected by modern networks of fibre-optic cable, coaxial cable, microwave radio relay, and a satellite system.

Cellular or mobile telephone services are now widely available and include roaming service, even to many foreign countries. The cellular system works as a radio network of base stations and antennas. (The area of a city covered by one base station is called a cell, whose size ranges from 1 km to 50 km in radius.) A cell phone contains both a low-power transmitter and a receiver. It can use both of them simultaneously, understand different frequencies, and can automatically switch between frequencies. The base stations also transmit at low power. Each base station uses carefully chosen frequencies to reduce interference with neighbouring cells.

In a situation where multiple personal computers are used, as perhaps in your local study centre, it helps to get all the computers connected in a network so that they can “talk” to each other, and we can

- share a single printer between computers;
- share a single Internet connection among all the computers;
- access shared files and documents on any computer;
- play games that allow multiple users at different computers; and
- send the output of a device like a DVD player to other computer(s).



Notes

To install such a network of personal computers, there are three steps:

- Choose the technology for the network. The main technologies to choose between are standard Ethernet, phone-line-based, power-line-based and wireless.
- Buy and install the hardware.
- Configure the system and get everything talking together correctly.

The Internet is a vast network of computers throughout the world. It combines many different forms of communications. As the technology advances it could replace all other forms of communication by combining them into one. Magazines and newspapers are already being put online along with libraries, art, and research. Unlike most forms of communication, it facilitates access to vast store of information through the World Wide Web (WWW). The World Wide Web is the multimedia part of the Internet and combines text with sound, photos, drawings, charts, graphs, animation, and even video. New innovations such as Java, a web-based programming language, allow simple tasks to be performed inside the document. The more widespread the Internet becomes, the more important and powerful type of communication it will become. In India, several hundred thousand schools are being provided access to computers and Internet to impure the quality of education. The MHRD is developing a one stop portal –Sakshat– which can be accessed by you. The National Institute of open schooling is also contributing for it.

You may be aware of variouis complex communication systes in use around us; like: radio, TV, fax machine, internet etc.

A cathode-ray tube in a television set, has a “cathode” which emits a ray of electrons in a vacuum created inside a glass “tube”. The stream of electrons is focused and accelerated by anodes and hits a screen at the other end of the tube. The inside of the screen is coated with phosphor, which glows when struck by the beam. The cathods beam carries the vedio signals from the object and forms the image an the screen accordingly.

In a fax machine, the document to be transmitted is scanned by a photo sensor to generate a signal code before it is transmitted through a telephone line.

A modem (modulator/demodulator) can convert a digital bit stream into an analog signal (in the modulator) and vice-versa (in the demodulator). It is used as a transmitter to interface a digital source to an analogue communication channel, and also as a receiver to interface a communication channel to a digital receiver

MODULE - 8

Semiconductors Devices
and Communication



Notes

Communication Systems



WHAT YOU HAVE LEARNT

- In a typical modern-day long distance communication system, the information is in the form of electrical signals (voltage or current).
- The essential elements of a communication system are (i) a transmitter (ii) a medium or mechanism to carry the signal over long distances, and (iii) a receiver to intercept the signal and retrieve the information.
- An antenna or aerial is essentially a system of conductors, which is an effective radiator and absorber of electromagnetic waves in the desired radio frequency region.
- Analog signals are physical signals that vary continuously with time while digital signals have the form of discrete pulses.
- Digital communication systems are more efficient, give better performance, and greater flexibility than their analog counterparts.
- AM radio is broadcast on three bands, the Long wave at 144 – 351 kHz (in the LF), the Medium wave at 530 – 1,700 kHz (in the MF), and the Short wave at 3 – 30 MHz (HF). FM radio is broadcast on carriers at 88 – 104 MHz (in the VHF). Commercial TV transmission is in the VHF-UHF range.
- Electrical communications channels are wireline (using guided media) or wireless (using unguided media).
- Multiplexing refers to the process of simultaneous transmission of different messages (each with some frequency bandwidth) over the same path way. The higher the frequency of the carrier, the higher is its message-carrying capacity.
- Comparing the different wireline channels, the communication capacity of visible light (of frequency of about 10^{14} Hz) in an optical fibre is thus much larger than that of typical microwave (of frequency of about 10^9 Hz) in a metallic conductor.
- An optical fibre guides a light beam (from a laser) from its one end to the other by the process of total internal reflection at the interface of the inner core (of refractive index n_1) and the cladding (of refractive index $n_2 > n_1$).
- In the wireless radio transmission, a system of conductors called antenna or aerial launches the carrier radio waves in space and also detects them at the receiver location. The propagation of radio waves in the atmosphere depends on the frequency of the waves. Low and medium frequency radio waves up to about 1 MHz are used in ground (or surface) wave communication. Medium frequency (MF) waves of 300 kHz – 3 MHz are largely absorbed by the ionosphere. The high-frequency (HF) waves of 3 – 30 MHz are, however, reflected back by the ionosphere. VHF and UHF waves are transmitted either



Notes

- by direct line-of-sight using tall towers (space wave or tropospheric propagation), or by beaming to artificial satellites and broadcasting from there.
- The cellular or mobile telephone system works as a radio network in which a city is divided into ‘cells’ of 1 km to 50 km in radius, and each cell is covered by one base station. A cellular phone contains a low-power transmitter and a low-power receiver.
- An analogue signal is completely described by its samples, taken at equal time intervals T_s , if and only if the sampling frequency $f_s = 1/T_s$ is at least twice the maximum frequency component of the analogue signal.
- Low frequencies can not be transmitted to long distances using aerials or antennas of practical dimensions. Low-frequency messages are loaded on a high frequency carrier signal by a process called modulation. In amplitude modulation (AM), the amplitude of a high-frequency carrier wave are modified in accordance with the strength of a low-frequency information signal. In frequency modulation (FM), the amplitude of the carrier wave remains constant, but its frequency is continuously varied in accordance with the instantaneous amplitude of the information signal, i.e., the frequency of the modulated carrier wave varies from a minimum to a maximum corresponding to the minimum and maximum values of the modulating signal.
- In the digital pulse code modulation (PCM) technique, the modulating signal is first sampled, the magnitude (with respect to a fixed reference) of each sample is quantised, and then the binary code is usually transmitted modulating an analogue current in a landline.



TERMINAL EXERCISE

- What are the essential elements of a communication system?
- What is an antenna?
- What are the important characteristics of a receiver in a communication system?
- Distinguish between the terms analogue and digital signals. Define a ‘bit’.
- The VHF band covers the radio frequency range of 30 – 300 MHz. Using the known relationship of speed to frequency and wavelength of an electromagnetic wave, determine the VHF wavelength range in vacuum. Take the speed of light in vacuum to be $3 \times 10^8 \text{ ms}^{-1}$
- Long distance radio broadcasts use shortwave bands. Explain.
- Satellites are used for long distance TV transmission. Justify.
- The core of an optical fibre is made of glass with a refractive index of 1.51 and the cladding has a refractive index of 1.49. Calculate the critical angle for total internal reflection.

MODULE - 8

Semiconductors Devices
and Communication



Notes

Communication Systems

9. List some advantages of creating a local network of personal computers.
10. What do you understand by modulation. Explain its need.
11. Explain the role of modulation and demodulation in communication system.



ANSWERS TO INTEXT QUESTIONS

30.1

1. E.M. waves are time varying electrical and magnetic field travelling in space with a speed of $3 \times 10^8 \text{ ms}^{-1}$ at right angles to each other.
2. $\lambda = \frac{C}{v} = \frac{3 \times 10^8 \text{ ms}^{-1}}{30 \times 10^6 \text{ s}^{-1}} = 10 \text{ m}$
3. (i) frequency range of visible light is 10^{14} Hz – 10^{15} Hz
(ii) frequency range of radio waves is 30 kHz–300 MHz

30.2

1. Co-axial cable is a pair of point to point connector where in one conductor is in the form of hollow cylinder and the other a solid wire at the axis of the first conductor the two being separated by an insulator. The one used for frequency range 3.0 GHz–40 GHz.
2. The basic principle used in optical fibre is total internal reflection due to which light beam may move along an optical fibre without any loss in energy.

30.3

1. Sky wave communication is normally better during night, because in absence of sun ionosphere's composition is settled so that it acts as a better reflector.
2. (a) iv, (b) (iii)

30.4

1. (a) (iii), (b) (iii)
2. The minimum size of a transmitting antenna is comparable to wavelength of signal to be transmitted. For maximum power transmission size of antenna should be at least $\lambda/4$.

Answer to Problems in Terminal Exercise

5. 10 m- 1 m
8. $\sin^{-1} n_2/n_1 = 80.66^\circ$

SENIOR SECONDARY COURSE

SEMICONDUCTORS AND SEMICONDUCTOR DEVICES

STUDENT'S ASSIGNMENT – 8

Maximum Marks: 50

Time : 1½ Hours

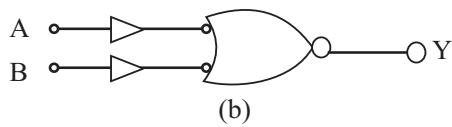
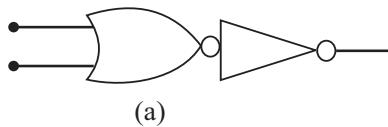
INSTRUCTIONS

- Answer All the questions on a separate sheet of paper
- Give the following information on your answer sheet:
 - Name
 - Enrolment Number
 - Subject
 - Assignment Number
 - Address
- Get your assignment checked by the subject teacher at your study centre so that you get positive feedback about your performance.

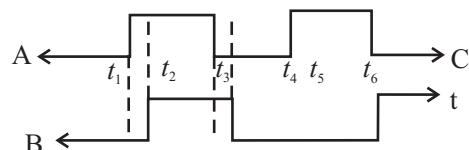
Do not send your assignment to NIOS

- | | | |
|-----|--|---|
| 1. | Name the majority charge carriers in n-type semiconductor? | 1 |
| 2. | Draw symbol for a n-p-n transistor. | 1 |
| 3. | Explain the meaning of the term doping in semiconductors. | 1 |
| 4. | What is the effect of forward biasing a p-n junction on the width of depletion region around it? | 1 |
| 5. | How do you identify collector and emitter in a transistor. | 1 |
| 6. | Draw the logic symbol of a NOR gate. | 1 |
| 7. | Out of silicon and germanium which has more free charge carrier density at room temperature. Why? | 1 |
| 8. | In common base configuration current gain is less than 1 but still can have a voltage gain. How? | 1 |
| 9. | Distinguish between a LED and a solar cell. Draw diagram of each. | 2 |
| 10. | Draw the characteristics of a pn junction diode in (i) forward bias (ii) reverse bias. | 2 |
| 11. | In a half wave rectifier input frequency is 50 Hz. What is its output frequency? What is the output frequency of a full wave rectifier for the same input frequency. | 2 |

12. Two amplifiers are connected one after the other in series. The first amplifier has a voltage gain of 10 and the second has a voltage gain of 20. If the input signal is 0.01v, calculate the output ac signal. 2
13. How can you realize an AND gate with the help of *p-n* junction diodes? Draw the circuit and explain to truth table. 4
14. For a common emitter amplifier, the audio signal voltage across a $5\text{ k}\Omega$ collector resistance is 5v. Suppose the current amplification factor of the transfer is 100, find the input signal voltage and base current, if the base resistance is $1\text{ k}\Omega$. 4
15. Define current gain in common base configuration and common emitter configuration. Establish a relation between the two. 4
16. With the help of suitable diagrams explain
 (a) how does a capacitor convert fluctuating *ac* steady *dc*.
 (b) how a zener diode stabilizes dc output against load variations. 4
17. Explain :
 (i) Why a transistor has to be biased for using it as an amplifier,
 (ii) How the range of variation of amplitude of input signal is decided for the proper working of a transistor,
 (iii) Why the voltage gain of an amplifier can not be increased beyond a limit by increasing load resistance. 4
18. Identify the logic gates indicated by circuits given below.



Corresponding to the input signal at A and B as shown below draw output waveform for each ats.



19. With the help of a circuit diagram explain how a transistor can be used as an amplifier? 5
20. Draw a circuit diagram for studying the characteristics. Draw the input and output characteristics and explain the current gain obtained. 5

QUESTIONS PAPER DESIGN

Subject: Physics

Paper Marks: 80

Class: Senior Secondary

Duration: 03 Hrs.

1. Weightage by Objectives

| Objective | Marks | % of the Total Marks |
|-----------------------|-----------|----------------------|
| Knowledge | 20 | 25 |
| Understanding | 40 | 50 |
| Application and Skill | 20 | 25 |
| Total | 80 | 100 |

2. Weightage by Types of Question

| Type of Questions | Marks × No. of Questions | Marks Allotted |
|---------------------------------|--------------------------|-----------------|
| Essay (E) | 6×4 | 24 |
| Short Answers I (SA1) | 4×7 | 28 |
| Short Answers II (SA2) | 2×9 | 18 |
| Multiple Choice Questions (MCQ) | 1×10 | 10 |
| Total | 30 Questions | 80 Marks |

3. Weightage as per the Content

| Sr. No. | Module | Marks |
|---------|---|-----------|
| 1 | Motion, Force and Energy | 14 |
| 2 | Mechanics of Solids and Fluids | 06 |
| 3 | Thermal Physics | 06 |
| 4 | Oscillations and Waves | 06 |
| 5 | Electricity and Magnetism | 16 |
| 6 | Optics and Optical Instruments | 14 |
| 7 | Atoms and Nuclei | 08 |
| 8 | Semiconductor Devices and Communication | 10 |
| | Total Marks | 80 |

4. Difficulty Level

| | Easy | Average | Difficult | Total |
|-------------------|------|---------|-----------|-------|
| Percent Weightage | 25 % | 50 % | 25 % | 100 % |
| Marks Allotted | 20 | 40 | 20 | |
| No. of Questions | 08 | 15 | 07 | |

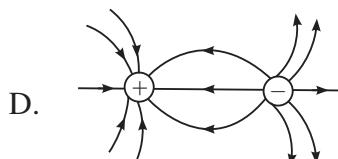
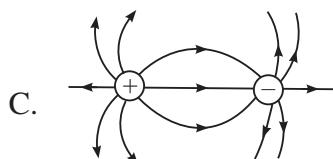
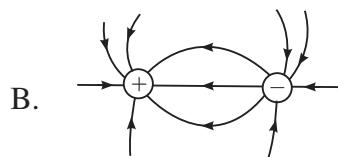
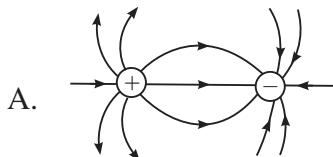
Note : Some internal choices are given in application questions.

5. Time Management

| Type of Questions | Total Time 180 minutes |
|---------------------------------|------------------------|
| Essay (E) | 60 |
| Short Answers I (SA1) | 50 |
| Short Answers II (SA2) | 35 |
| Multiple Choice Questions (MCQ) | 15 |
| Reading and Revision | 20 |

SAMPLE QUESTIONS PAPER

1. Which of the following diagrams best represents electric field lines between two equal and opposite charges?



1

2. A rectangular current carrying loop of wire is placed near a long current carrying straight conductor with one of its sides parallel to the conductor as shown in figure. The loops will be



- A. attracted towards the conductor whatever be the direction of current in the loop.
- B. repelled by the conductor whatever be the direction of current in the loop.
- C. attracted towards the conductor when the current in the loop flows anticlockwise.
- D. repelled by the conductor when the current in the loop flows anticlockwise.

1

3. The energy required to cause a transition of an electron from $n = 1$ to $n = 2$ orbit of hydrogen atom is

- | | |
|-------------|------------|
| A. 10.2 eV | B. 13.6 eV |
| C. 12.09 eV | D. 1.51 eV |

4. The number of electrons emitted by cathodes in a photo tube is 3.0×10^4 when it is illuminated by a small bright source placed 50 cm away. If the source is carried to a distance of 100 cm, the number of electrons emitted by photocathode would be

- | | |
|----------------------|----------------------|
| A. 3.0×10^4 | B. 7.5×10^3 |
| C. 1.5×10^4 | D. 6.0×10^4 |

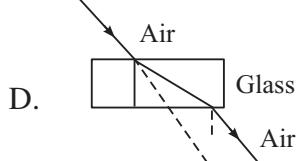
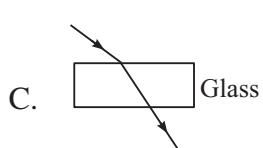
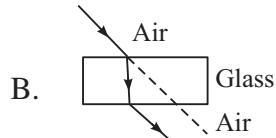
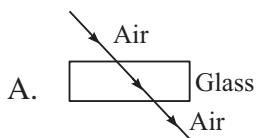
1

5. In a forward biased p-n junction diode

- A. charge carriers do not move at all.
- B. only the holes in the p-region move towards the n-region

Sample Questions Paper

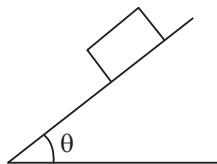
- C. only the minority charge carriers move due to thermal energies
D. majority charge carrier diffuse from either side towards the junction 1
6. A n-type extrinsic semiconducting material is
- A. negatively charged
 - B. positively charged
 - C. electrically neutral
 - D. negatively or positively charged depending on the dopant 1
7. For a CE mode amplifier, v_i is 10 mV and v_o is one volt. The voltage gain of the amplifier would be
- A. 50
 - B. 20
 - C. 100
 - D. 10 1
8. In a n-type semiconductor
- A. electrons are majority charge carriers and dopants are trivalent atom.
 - B. electrons are minority carriers and dopants are pentavalent atoms
 - C. holes are the minority carrier and dopants are pentavalent atoms
 - D. holes are the majority charge carriers and dopants are trivalent atoms 1
9. A secondary rainbow occurs when a ray of light undergoes
- A. one refraction and one internal reflection in a rain drop
 - B. two refractions and two internal reflections in a rain drop
 - C. two refractions and one internal reflection in a rain drop
 - D. one refraction and two internal reflections in a rain drop 1
10. Which of the following represents the refraction of a ray of light through a glass slab correctly



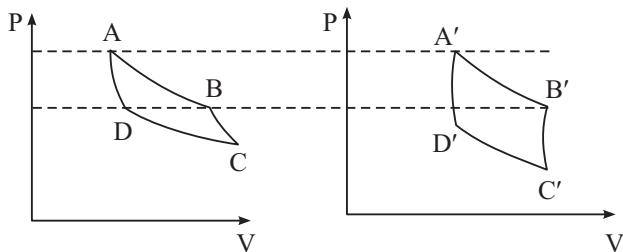
11. State second law of motion. Show that for an object of constant mass, acceleration of the object is directly proportional to the force applied on it. 2

Sample Questions Paper

12. A body of mass m is placed on a rough plane inclined at an angle θ . Draw a diagram showing forces acting on the body and write expression for frictional force if the block stays stationary on the inclined plane. 2

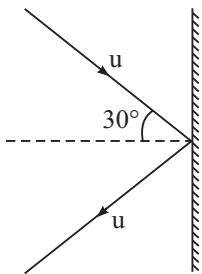


13. State stoke's law for the viscous force acting on a sphere of radius r , moving in a fluid of viscosity η with velocity v . Obtain the unit of coefficient of viscosity in terms of kg, metre and second. 2
14. Figures below show the p-v diagrams of two carnot engines. Which of the two engines is more efficient. (It is given that both engines draw equal heat from the source). 2



15. A spectral line in the light from a star shows a shift towards the red end of the spectrum. If the shift is 0.03%, calculate the velocity of recession of the star. 2
16. How will you convert a galvanometer into (i) an ammeter (ii) a voltmeter? 2
17. Draw a diagram showing refraction of a ray of light through a glass prism. Mark angle of deviation and angle of emergence in the diagram. 2
18. In fission of a $^{236}_{92}\text{U}$ nucleus 200 MeV energy is released whereas in fusion of 4 protons 26.8 Mev energy is released. Which of these processes gives more energy per unit atomic mass? Explain. 2
19. The axis of a 100 turn circular coil (area of cross-section $3.85 \times 10^{-3} \text{ m}^2$) is parallel to a uniform magnetic field. The magnitude of the field changes at a constant rate from 25 mT to 50 mT. in 250 ms. Calculate the magnitude of induced emf across the coil. 2
20. Explain (i) Isothermal (ii) Adiabatic (iii) isobaric and (iv) isochoric processes. 4
21. A ball of mass 50g strikes a rigid wall at an angle 30° with a speed of 10 ms^{-1} and gets reflected without any change in speed as shown in figure. Find the magnitude and direction of the impulse imparted to the ball by the wall. 4

Sample Questions Paper



22. Explain why a small drop of liquid always spherical. Obtain an expression for pressure difference between inside and outside a spherical liquid drop. 4

23. A transverse harmonic wave on a string is given by

$$y(x, t) = 3.0 \sin(36t + 0.018x + \pi/4)$$

where x and y are in cm and t in seconds. The positive direction of x is from left to right.

- A. What is the speed and direction of propagation of the wave?
- B. What are its amplitude and initial phase?
- C. What is the least distance between its two successive crests? 4

24. An alternating voltage $E = E_0 \sin \omega t$ is applied to a circuit comprising of an inductor (L), a capacitor (C) and a resistor (R) in series. Obtain an expression for (i) impedance of the circuit and (ii) phase angle between voltage and current. 4

25. Distinguish between angular dispersion and dispersive power. for a prism of angle $A = 60^\circ$, the angle of minimum deviation is also A. Calculate its refractive index. 4

26. What is mass defect? The mass of the nucleus of $^{14}_7\text{N}$ atom is 14.00307 u. Calculate mass defect and binding energy per nucleon. Take $m_p = 1.00727$ u, $m_n = 1.00865$ u and $1\text{u} = 931$ meV. 4

or

An electron, a proton and an alpha particle have the same kinetic energy, which of these particles has the shortest de Broglie wave and which have longest wavelength.

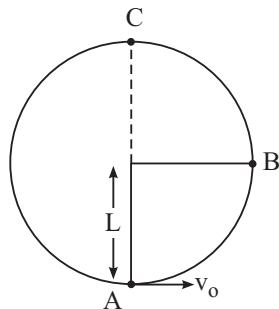
27. Define resistance. State the factors on which the resistance of a conductor depends. How are resistances connected in (i) series (ii) parallel. Find the expressions for equivalent resistance of two different resistors in these two arrangements. 6

28. What is meant by rectification? State the principle of a rectifier. With the help of a circuit diagram, explain the working of a half wave rectifier. Draw the input and output waveforms for a half wave rectifier. 6

29. What is meant by interference of light? Describe Young's double slit experiment to obtain interference pattern on a screen. Show that the intensity of the resultant wave in Young's experiment is proportional to the square of the amplitude of the incident wave. Discuss the conditions for constructive and destructive interference. 6

Sample Questions Paper

30. A bob of mass m , tied to a string of length L is rotated in a vertical circle in such a manner that the string has zero tension at point C, as shown in figure. Find the ratio of the kinetic energies at points B and C. 6



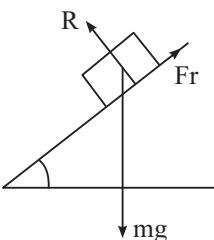
or

A body of mass 20 kg is initially moving with a speed of 5 ms^{-1} . A force of 40 N is applied on the body for 3 seconds.

- (i) find the final speed of the body after 3 seconds.
- (ii) What is the distance covered during this period?
- (iii) How much work has been done during this period?
- (iv) Find the initial K.E. of the body.
- (v) Find the final K.E. of the body.
- (vi) Show that the work done is equal to the change in K.E. for the body.

MARKING SCHEME

1. A
 2. D
 3. A
 4. B
 5. D
 6. C
 7. C
 8. C
 9. B
 10. B
11. Statement of the law (module 1 p 65) 1
 Direction according to the law (module 1 p 65) 1

12.  1

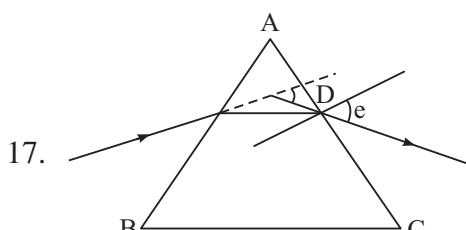
- $$F_r = mg \sin\theta$$
- or $F_r \leq \mu mg \cos\theta$ 1
13. According to stoke's law 1
- $$F = 6\pi\eta rv$$

$$\eta = \frac{F}{6\pi rv} = \frac{\text{kgms}^{-2}}{\text{m} \cdot \text{ms}^{-1}} = \text{kg m}^{-1}\text{s}^{-1}$$

14. Engine (ii) is more efficient $\frac{1}{2}$

$$\therefore \eta = \frac{W}{Q}$$

Marking Scheme

- Q is same for both $\frac{1}{2}$
 and w = area enclosed in p-v diagram which is more in case of II $\frac{1}{2}$
15.
$$\frac{\Delta\lambda}{\lambda} = \frac{v}{c}$$
 1
 $v = c \frac{\Delta\lambda}{\lambda} = 3 \times 10^8 \times \frac{0.03}{100} = 9 \times 10^4 \text{ ms}^{-1}$ 1
16. (a) By connecting a suitable low resistance in parallel to the galvanometer coil 1
 (b) By connecting a suitable high resistance in series with the galvanometer coil. 1
17.  1+1
18. More energy per unit amu is released in fusion. 1
 In fission it is less than 1 MeV/amu $\frac{1}{2}$
 In fusion it is more than 6.7 MeV/amu $\frac{1}{2}$
19. $N = 100, A = 3.85 \times 10^{-3} \text{ m}^2, B = (50 - 25) = 25 \text{ mT}$
 $|e| = NA \left(\frac{B_2 - B_1}{t} \right)$ 1
 $= 100 \times 3.85 \times 10^{-3} \times \frac{20 \times 10^{-1}}{250 \times 10^{-3}}$ $\frac{1}{2}$
 $= 3.85 \times 10^{-2} \text{ V}$ $\frac{1}{2}$
 $= 38.5 \text{ mV}$ $\frac{1}{2}$
20. Explanation of
 (i) Isothermal process (p302) 1
 (ii) Adiabatic process (p302) 1
 (iii) Isobaric process (p303) 1
 (iv) Isochoric process (p303) 1

Marking Scheme

21. $p_i = 0.05 \times 10 = 0.5 \text{ kg ms}^{-1} = p_f$

$$p_i^x = 0.5 \cos 30 \quad p_i^y = 0.5 \sin 30 \quad 1$$

$$p_f^x = 0.5 \cos 30 \quad p_f^y = 0.5 \sin 30$$

$$\text{Impulse} = p_f - p_i = (p_f^x i + p_f^y j) - (p_i^x i + p_i^y j) \quad 1$$

$$= (p_f^x - p_i^x) i \quad 1$$

$$= -0.866 \text{ kg ms}^{-1} \quad 1$$

$$\text{Impulse} = 0.866 \text{ kg ms}^{-1} \quad \frac{1}{2}$$

22. Explanation 1

Derivation 3

23. (a) $v = \frac{w}{k} = \frac{36}{0.018} = 2000 \text{ cm s}^{-1} = 20 \text{ ms}^{-1}$ 1

The wave is travelling towards negative
x direction 1

(b) $a = 3.0 \text{ cm}$ $\phi_0 = TA$ $\frac{1}{2} + \frac{1}{2}$

(c) Least distance between two successive crests $\frac{1}{2}$

$$= \lambda = \frac{2\pi}{0.018} = 3.5 \text{ m} \quad \frac{1}{2}$$

24. Connection diagram $\frac{1}{2}$

Phase diagram $\frac{1}{2}$

Derivation of expression for z 2

Derivation of expression for $\tan \theta$ 1

25. The difference between the angles of deviation for any two wavelengths (colours)
is known as the angular dispersion for those wavelength 1

Marking Scheme

The ratio of the angular dispersion to the mean deviation is known as the dispersive power (w) of the material of the prism

$\frac{1}{2}$

$$\begin{aligned}
 w &= \frac{s_v - \delta_r}{s_y} \\
 &= \frac{\sin\left(\frac{A + \sin}{2}\right)}{\sin\frac{A}{2}} \\
 &= \frac{\sin 60^\circ}{\sin 30^\circ} \\
 &= \sqrt{3}
 \end{aligned}$$

$\frac{1}{2}$

1

$\frac{1}{2}$

$\frac{1}{2}$

26. The mass of the nucleus of an atom of any element is always found to be less than the sum of the masses of its constituent nucleon. The difference between the sum of masses of nucleon and the mass of the nucleus is called mass defect.

1

$$\begin{aligned}
 \Delta m &= 7m_p + 7m_n - m(^{14}_7N) \\
 &= 7 \times 1.00727 + 7 \times 1.00865 - 14.00307 \\
 &= 7.05089 + 7.06055 - 14.00307 \\
 &= 14.11144 - 14.00307 \\
 &= 0.10837 \text{ u}
 \end{aligned}$$

$\frac{1}{2}$

1

$\frac{1}{2}$

$$\begin{aligned}
 BE &= \Delta m \times 931 \\
 &= 100.89247 \text{ Mev}
 \end{aligned}$$

$\frac{1}{2}$

$\frac{1}{2}$

$$\begin{aligned}
 BE/\text{amu} &= \frac{\Delta m \times 931}{A} \\
 &= \frac{100.89247}{14} = 7.206 \text{ MeV}
 \end{aligned}$$

$\frac{1}{2}$

or

26. $\lambda = \frac{h}{P}$

$\frac{1}{2}$

$$\text{K.E.}; E = \frac{1}{2}mv^2 = \frac{m^2v^2}{2m} = \frac{p^2}{2m}$$

$\frac{1}{2}$

Marking Scheme

$$\Rightarrow \lambda = \frac{h}{\sqrt{2mE}} \quad \frac{1}{2}$$

$$\Rightarrow \lambda \propto \frac{1}{\sqrt{m}} \quad \frac{1}{2}$$

of the 3 particles given α -particle is the heaviest and electron is the lightest particle. 1

$\therefore \lambda_e$ will be greatest and λ_α will be shortest. $\frac{1}{2} + \frac{1}{2}$

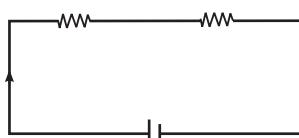
27. Resistance of a conductor is the ratio of P.D applied across it and the current flowing through it

$$\text{i.e. } R = \frac{V}{I} \quad \frac{1}{2}$$

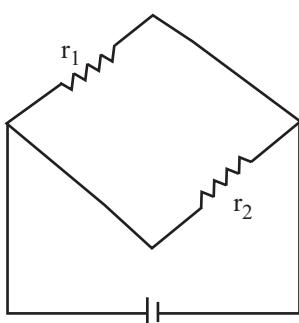
Resistance of a conductor depends on

Nature of material, length, area of cross section and temperature of the conductor 1

- (i) resistors connected end to end such that same current flows through each of them $\frac{1}{2}$



- (ii) resistors connected in such a way that one end of all the conductors is connected to positive terminal of the battery and the other end of negative terminal. In this case same P.D. is applied across all the resistors. 1



In series combination

$$\begin{aligned} V &= V_1 + V_2 \\ &= Ir_1 + Ir_2 \end{aligned}$$

Marking Scheme

$$= I(r_1 + r_2)$$

$$\frac{V}{I} = R = r_1 + r_2 \quad 1$$

In parallel combination $I = I_1 + I_2$

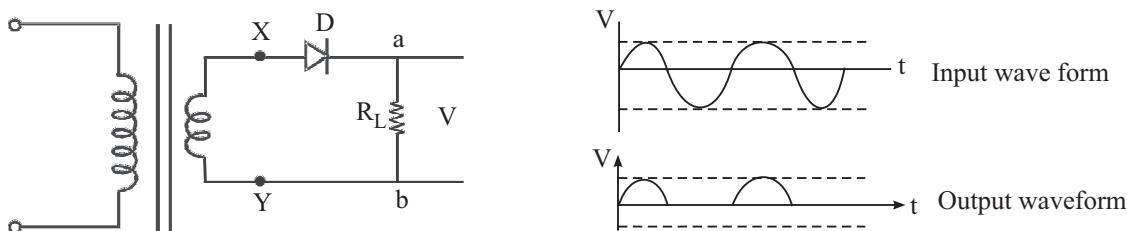
$$= \frac{V}{r_1} + \frac{V}{r_2} = V \left(\frac{r_1 + r_2}{r_1 r_2} \right) \quad 2$$

$$R = \frac{V}{I} = \frac{r_1 r_2}{r_1 + r_2}$$

28. The process of conversion of ac into dc is called rectification 1

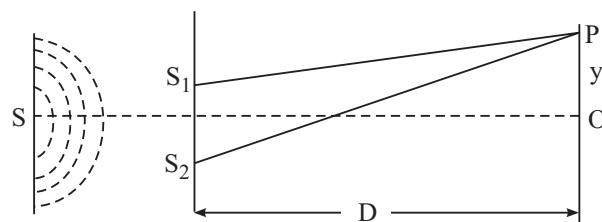
A pn junction diode conducts when it is forward biased and does not conduct in reverse bias 1

for the input wave form shown the diode is forward biased during the first half cycle ($0 - T/2$) and hence conducts and current flows through R_C from a to b. During the next half cycle ($T/2 - T$) D is reverse biased and hence no current flows through R_C during ($T - 3T/2$) again current flows through R_C from



2 + 2

29. The phenomenon of redistribution of energy in space due to superposition of light waves from two coherent sources. 1



Description of experimental set up of Young's double slit experiment

$$\text{Derivation of } A = 2a \cos\left(\frac{\delta}{2}\right) \quad 1\frac{1}{2}$$

Marking Scheme

$$I \propto A^2$$

$$\propto 4A^2 \cos^2\left(\frac{\delta}{2}\right)$$

$\frac{1}{2}$

Constructive interference

$$I_{\max} = 4a^2$$

when $\cos^2\delta/2 = 1$

$$\cos \delta/2 = 1$$

$$\delta = 0, 2\pi, 4\pi \dots 2n\pi$$

$\frac{1}{2}$

Destructive interference

$$I_{\min} = 0$$

when $\cos^2 \delta/2 = 0$

$$\delta = \pi, 3\pi, 5\pi \dots (2n + 1)\pi$$

$\frac{1}{2}$

$$30. T - mg = \frac{mv_o^2}{L}$$

$\frac{1}{2}$

$$0 - mg = \frac{mv_c^2}{L} \Rightarrow v_c = \sqrt{gL}$$

1

$$T_1 - 0 = \frac{mv_B^2}{L}$$

$\frac{1}{2}$

$$\frac{1}{2}mv_o^2 - \frac{1}{2}mv_B^2 = mgL$$

1

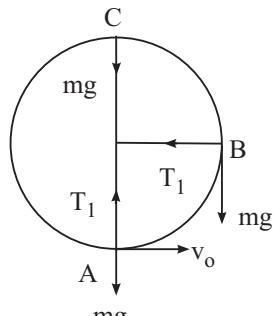
$$\Rightarrow v_B^2 = 3gL$$

$\frac{1}{2}$

$$\frac{k\varepsilon \text{ at } B}{k\varepsilon \text{ at } c} = \frac{\frac{1}{2}mv_B^2}{\frac{1}{2}mv_c^2} = \frac{v_B^2}{v_c^2} = \frac{3gL}{gL} = \frac{3}{1}$$

$1\frac{1}{2}$

$$= 3 : 1$$



Marking Scheme

or

$$(i) \quad a = \frac{F}{m} = \frac{40}{20} = 2 \text{ ms}^{-2} \quad \frac{1}{2}$$

$$v = u + at$$

$$= 5 + 2 \times 3 = 11 \text{ ms}^{-1} \quad \frac{1}{2}$$

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

$$= 5 \times 3 + \frac{1}{2} \times 2 \times (3)^2$$

$$= 15 + 9 = 24 \quad 1$$

$$(iii) \quad W = FS$$

$$= 40 \times 24$$

$$= 960 \text{ J} \quad 1$$

$$(iv) \quad \text{Initial K.E.} = \frac{1}{2}mu^2$$

$$= \frac{1}{2} \times 20 \times (5)^2$$

$$= 250 \text{ J} \quad 1$$

$$(v) \quad \text{Final K.E.} = \frac{1}{2}mv^2$$

$$= \frac{1}{2} \times 20 \times (11)^2$$

$$= 1210 \text{ J} \quad 1$$

$$(vi) \quad \text{Change in K.E.} = 1210 - 250 = 960 \text{ J}$$

$$\text{Also work done} = 960 \text{ J}$$

$$\text{So work done} = \text{change in K.E. of the body} \quad 1$$

Complete and Post the feedback form today

| Feed back on Lessons | | | | | | | | | | |
|-----------------------------|-------------|-----------|-------------|-----------|----------|---------|---------------|------------|----------------------|-------------|
| Lesson No. | Lesson Name | Content | | | Language | | Illustrations | | What You Have Learnt | |
| | | Difficult | Interesting | Confusing | Simple | Complex | Useful | Not useful | Very helpful | Not helpful |
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--Final fold and seal--

--Fourth fold--

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You must have enjoyed going through your course books. It was our endeavor to make the study material relevant, interactive and interesting. Production of material is a two way process. Your feedback would help us improve the study material. Do take a few minutes of your time and fill-up the feedback form so that an interesting and useful study material can be made.

Thank you
Coordinator
(Physics)

Feed back on Questions

| Lesson No. | Lesson Name | Intext Questions | | Terminal Questions | | |
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--Second Fold--



Yours suggestion

Did you consult any other book to study Physics?
If Yes, give reason for consulting it

Yes/No

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