$$u = -\frac{25}{6} = -4.167\ cm$$

The closest distance at which the person should can keep the book is, $u = -4.167$ cm

To find farthest object distance, lens equation is, $\frac{1}{v'} - \frac{1}{u'} = \frac{1}{f'}$

Rewriting for farthest object distance, $\frac{1}{u'} = \frac{1}{v'} - \frac{1}{f'}$

Substituting, $\frac{1}{u'} = \frac{1}{\infty} - \frac{1}{5}; u = -5\ cm$

The farthest distance at which the person can keep the book is, $u' = -5\ cm$

(b) To find magnification in near point focusing, $m = 1 + \frac{D}{f} = 1 + \frac{25}{5} = 6$

To find magnification in normal focusing,

$$m = \frac{D}{f} = \frac{25}{5} = 5$$

6.13.1.3. Resolving power of microscope

The diagram related to the calculation of resolution of microscope is illustrated in Figure 6.85. A microscope is used to see the details of the object under observation. The ability of microscope depends not only in magnifying the object but also in resolving two points on the object separated by a small distance d_{min}. Smaller the value of d_{min} better will be the resolving power of the microscope.

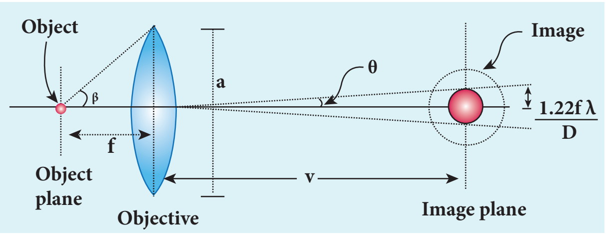

Figure 6.85 Resolving power of microscope

The radius of central maxima is already derived as equation 6.159,

$$r_o = \frac{1.22\lambda v}{a} \qquad (6.179)$$

In the place of focal length f we have the image distance v. If the difference between the two points on the object to be resolved is d_{min}, then the magnification m is,

$$m = \frac{r_o}{d_{min}} \qquad (6.180)$$

$$d_{min} = \frac{r_o}{m} = \frac{1.22\lambda v}{am} = \frac{1.22\lambda v}{a(v/u)} = \frac{1.22\lambda u}{a} \quad [\because m = v/u]$$

$$d_{min} = \frac{1.22\lambda f}{a} \quad [\because u \approx f] \qquad (6.181)$$

On the object side,

$$2\tan\beta \approx 2\sin\beta = \frac{a}{f} \therefore \quad [a = f 2\sin\beta] \quad (6.182)$$

$$d_{min} = \frac{1.22\lambda}{2\sin\beta} \qquad (6.183)$$

To further reduce the value of d_{min} the optical path of the light is increased by immersing the objective of the microscope in to a bath containing oil of refractive index n.

$$d_{min} = \frac{1.22\lambda}{2n\sin\beta} \qquad (6.184)$$

Such an objective is called oil immersed objective. **The term $n\sin\beta$ is called *numerical aperture NA*.**

$$d_{min} = \frac{1.22\lambda}{2(NA)} \qquad (6.185)$$

6.13.2 Compound microscope

The diagram of a compound microscope is shown in Figure 6.86. The lens near the object, called the ***objective***, forms a real, inverted, magnified image of the object. This serves as the object for the second lens which is the ***eyepiece***. Eyepiece serves as a simple microscope that produces finally an enlarged and virtual image. The first inverted image formed by the objective is to be adjusted close to, but within the focal plane of the eyepiece so that the final image is formed nearly at infinity or at the near point. The final image is inverted with respect to the original object. We can obtain the magnification for a compound microscope.

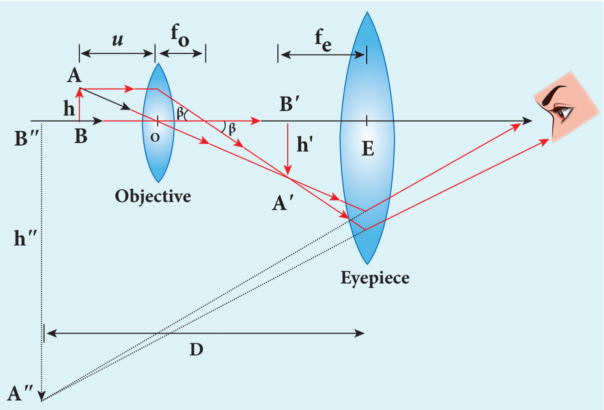

Figure 6.86 Compound microscope

6.13.2.1. Magnification of compound microscope

From the ray diagram, the linear magnification due to the objective is,

$$m_o = \frac{h'}{h} \qquad (6.186)$$

From the Figure 6.85, $\tan\beta = \frac{h}{f_o} = \frac{h'}{L}$, then

$$\frac{h'}{h} = \frac{L}{f_o} \qquad (6.187)$$

$$m_o = \frac{L}{f_o} \qquad (6.188)$$

Here, the distance L is between the first focal point of the eyepiece to the second focal point of the objective. This is called the tube length L of the microscope as f_o and f_e are comparatively smaller than L.

If the final image is formed at P (near point focussing), the magnification m_e of the eyepiece is,

$$m_e = 1 + \frac{D}{f_e} \qquad (6.189)$$

The total magnification m in near point focusing is,

$$m = m_o m_e = \left(\frac{L}{f_o}\right)\left(1 + \frac{D}{f_e}\right) \qquad (6.190)$$

If the final image is formed at infinity (normal focusing), the magnification m_e of the eyepiece is,

$$m_e = \frac{D}{f_e} \qquad (6.191)$$

The total magnification m in normal focusing is,

$$m = m_o m_e = \left(\frac{L}{f_o}\right)\left(\frac{D}{f_e}\right) \qquad (6.192)$$

EXAMPLE 6.42

A microscope has an objective and eyepiece of focal lengths 5 cm and 50 cm respectively with tube length 30 cm. Find the magnification of the microscope in the (i) near point and (ii) normal focusing.

Solution

$f_0 = 5cm = 5\times10^{-2}m;\ \ f_e = 50cm = 50\times10^{-2}m;$

$L = 30cm = 30\times10^{-2}m;\ D = 25cm = 25\times10^{-2}m$

(i) The total magnification m in near point focusing is, $m = m_o m_e = \left(\frac{L}{f_o}\right)\left(1 + \frac{D}{f_e}\right)$

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Substituting,

$$m = m_o m_e = \left(\frac{30\times10^{-2}}{5\times10^{-2}}\right)\left(1+\frac{25\times10^{-2}}{50\times10^{-2}}\right)$$
$$=(6)(1.5)=9$$

(ii) The total magnification m in normal focusing is, $m = m_o m_e = \left(\frac{L}{f_o}\right)\left(\frac{D}{f_e}\right)$

Substituting,

$$m = m_o m_e = \left(\frac{30\times10^{-2}}{5\times10^{-2}}\right)\left(\frac{25\times10^{-2}}{50\times10^{-2}}\right)$$
$$=(6)(0.5)=3$$

6.13.3 Astronomical telescope

An astronomical telescope is used to get the magnification of distant astronomical objects like stars, planets, moon etc. The image formed by astronomical telescope will be inverted. It has an objective of long focal length and a much larger aperture than the eyepiece as shown in Figure 6.87. Light from a distant object enters the objective and a real image is formed in the tube at its second focal point. The eyepiece magnifies this image producing a final inverted image.

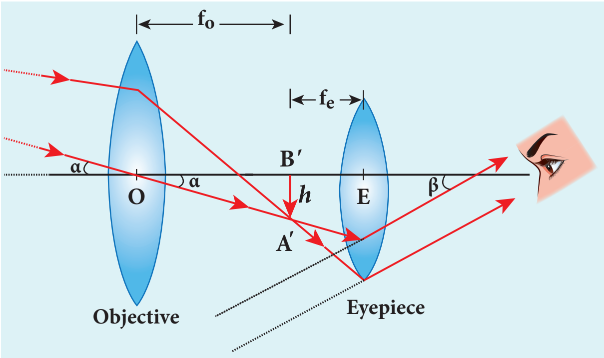


Figure 6.87 Astronomical telescope

6.13.3.1 Magnification of astronomical telescope

The magnification m is the ratio of the angle β subtended at the eye by the final image to the angle α which the object subtends at the lens or the eye.

$$m = \frac{\beta}{\alpha} \qquad (6.193)$$

From the diagram, $m = \frac{h/f_e}{h/f_o}$ (6.194)

$$m = \frac{f_o}{f_e} \qquad (6.195)$$

The length of the telescope is approximately, $L = f_o + f_e$

EXAMPLE 6.43

A small telescope has an objective lens of focal length 125 cm and an eyepiece of focal length 2 cm. What is the magnification of the telescope? What is the separation between the objective and the eyepiece? Two stars separated by 1′ will appear at what separation when viewed through the telescope?

Solution

f_o = 125 cm; f_e = 2 cm; m = ?; L = ?; θ_i = ?

Equation for magnification of telescope, $m = \frac{f_o}{f_e}$

Substituting, $m = \frac{125}{2} = 62.5$

Equation for approximate length of telescope, $L = f_o + f_e$

Substituting, L = 125+2 = 127 cm = 1.27 m

Equation for angular magnification, $m = \frac{\theta_i}{\theta_0}$

Rewriting, $\theta_i = m \times \theta_0$

Substituting,

$$\theta_i = 62.5\times1' = 62.5' = \frac{62.5}{60} = 1.04^\circ$$

6.13.4 Terrestrial telescope

A terrestrial telescope is used to see object at long distance on the surface of earth. Hence, image should be erect. A terrestrial telescope has an additional erecting lens to make the final image erect as shown in Figure 6.88.

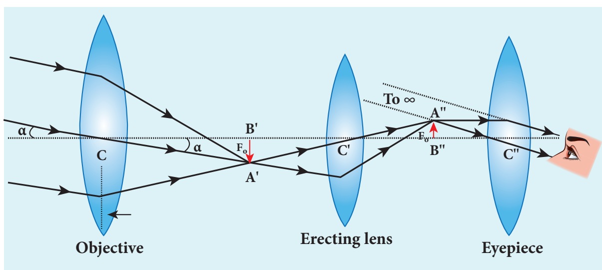


Figure 6.88 Terrestrial telescope

6.13.5 Reflecting telescope

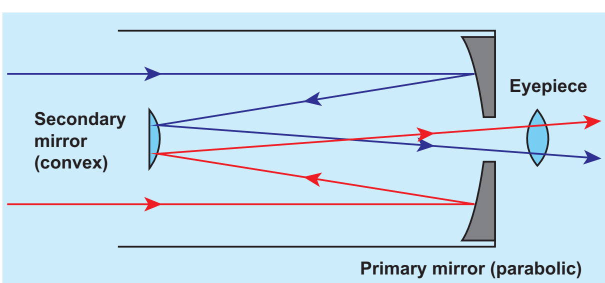


Figure 6.89 Reflecting telescope

Modern telescopes use a concave mirror rather than a lens for the objective. It is rather difficult and expensive to make lenses of large size which form images that are free from any optical defect. **Telescopes with mirror objectives are called *reflecting* telescopes.** They have several advantages. Only one surface it to be polished and maintained. Support can be given from the entire back of the mirror rather than only at the rim for lens. Mirrors weigh much less compared to lens. But the one obvious problem with a reflecting telescope is that the objective mirror would focus the light inside the telescope tube. One must have an eye piece inside obstructing some light. This problem could also be overcome by introducing a secondary mirror which would take the light outside the tube for view as shown in the Figure 6.89.

6.13.6 Spectrometer

The spectrometer is an optical instrument used to study the spectra of different sources of light and to measure the refractive indices of materials. It is shown in Figure 6.90. It consists of basically three parts. They are (i) collimator (ii) prism table and (iii) Telescope.

Figure 6.90 Spectrometer

(i) Collimator

The collimator is an arrangement to produce a parallel beam of light. It consists of a long cylindrical tube with a convex lens at the inner end and a vertical slit at the outer end of the tube. The distance between the slit and the lens can be adjusted such that the slit is at the focus of the lens. The slit is kept facing the source of light. The width of the slit can be adjusted. The collimator is rigidly fixed to the base of the instrument.

(ii) Prism table

The prism table is used for mounting the prism, grating etc. It consists of two circular metal discs provided with three levelling screws. It can be rotated about a vertical axis passing through its centre and its position can be read with verniers V_1 and V_2 . The prism table can be raised or lowered and can be fixed at any desired height.

(iii) Telescope

The telescope is an astronomical type. It consists of an eyepiece provided with cross wires at one end and an objective lens at its other end. The distance between the objective lens and the eyepiece can be adjusted so that the telescope forms a clear image at the cross wires, when a parallel beam from the collimator is incident on it.

The telescope is attached to an arm which is capable of rotation about the same vertical axis as the prism table. A circular scale graduated in half degree is attached to it. Both the telescope and prism table are provided with radial screws for fixing them in a desired position and tangential screws for fine adjustments.

Adjustments of the spectrometer

The following adjustments must be made before doing the experiment using spectrometer.

(a) ***Adjustment of the eyepiece*** The telescope is turned towards an illuminated surface and the eyepiece is moved to and fro until the cross wires are clearly seen.

(b) ***Adjustment of the telescope*** The telescope is adjusted to receive parallel rays by turning it towards a distant object and adjusting the distance between the objective lens and the eyepiece to get a clear image on the cross wire.

(c) ***Adjustment of the collimator*** The telescope is brought along the axial line with the collimator. The slit of the collimator is illuminated by a source of light. The distance between the slit and the lens of the collimator is adjusted until a clear image of the slit is seen at the cross wire of the telescope. Since the telescope is already adjusted for parallel rays, a well-defined image of the slit can be formed, only when the light rays emerging from the collimator are parallel.

(d) ***Levelling the prism table*** The prism table is adjusted or levelled to be in horizontal position by means of levelling screws and a spirit level.

6.13.6.1 Determination of refractive index of material of the prism

The preliminary adjustments of the telescope, collimator and the prism table of the spectrometer are made. The refractive index of the prism can be determined by knowing the angle of the prism and the angle of minimum deviation.

(i) Angle of the prism (A)

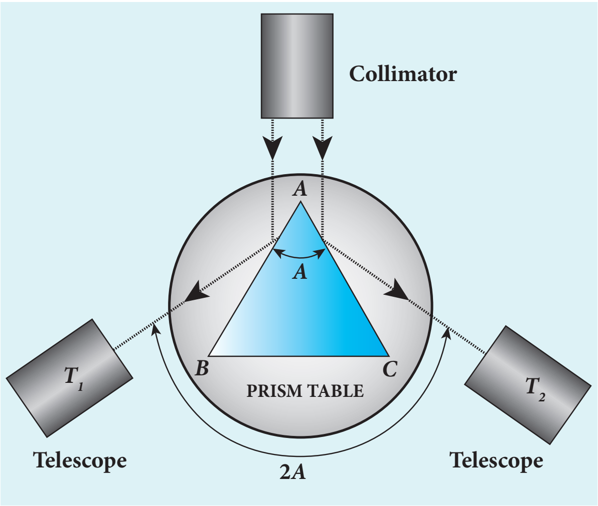


Figure 6.91 Angle of prism

The prism is placed on the prism table with its refracting edge facing the collimator as shown in Figure 6.91. The slit is illuminated by a sodium light (monochromotic light). The parallel rays coming from the collimator fall on the two faces AB and AC. The telescope is rotated to the position T_1 until the image of the slit formed by the reflection at the face AB is made to coincide with the vertical cross wire of the telescope. The readings of the verniers are noted. The telescope is then rotated to the position T_2 where the image of the slit formed by the reflection at the face AC coincides withthe vertical cross wire. The readings are again noted.

The difference between these two readings gives the angle rotated by the telescope, which is twice the angle of the prism. Half of this value gives the angle of the prism A.

(ii) Angle of minimum deviation (D)

The prism is placed on the prism table so that the light from the collimator falls on a refracting face, and the refracted image is observed through the telescope as shown in Figure 6.92. The prism table is now rotated so that the angle of deviation decreases. A stage comes when the image stops for a moment and if we rotate the prism table further in the same direction, the image is seen to recede and the angle of deviation increases. The vertical cross wire of the telescope is made to coincide with the image of the slit where it turns back. This gives the minimum deviation position.

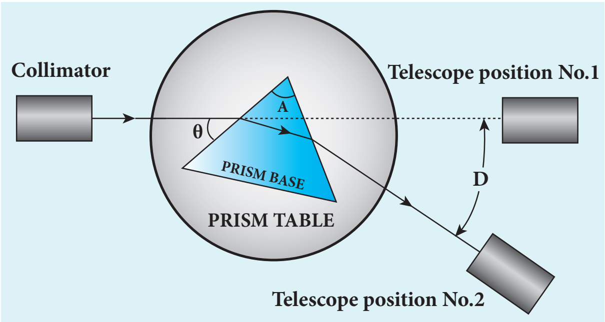


Figure 6.92 Angle of minimum deviation

The readings of the verniers are noted. Now, the prism is removed and the telescope is turned to receive the direct ray and the vertical cross wire is made to coincide with the image. The readings of the verniers are noted. The difference between the two readings gives the angle of minimum deviation D. The refractive index of the material of the prism n is calculated using the formula,

$$n = \frac{\sin\left(\frac{A+D}{2}\right)}{\sin\left(\frac{A}{2}\right)} \qquad (6.103)$$

The refractive index of a liquid may be determined in the same way using a hollow glass prism filled with the given liquid. Spectrometer experiments are discussed in the Practicals given in Volume 1 of this book.

6.13.7 The eye

Eye is a natural optical instrument given by God to the human beings. The internal structure and the Physics aspect of the functioning of different parts of human eye are discussed already in (X Physics Unit-2). As the eye lens is flexible, its focal length can be changed to some extent. When the eye is fully relaxed, the focal length is maximum and when it is strained the focal length is minimum. The image must be formed on the retina for a clear vision. The diameter of eye for a normal adult is about 2.5 cm. Hence, the image-distance, in other words, the distance between eye lens and retina is fixed always at 2.5 cm for a normal eye. We can just discuss the optical functioning of eye without giving importance to the refractive indices of the two liquids, aqueous humor and virtuous humor present in the eye A person with normal vision can see objects kept at infinity in the relaxed condition with maximum focal length f_{max} of the eye as shown in Figure 6.93(a). Also at a distance of 25 cm in the strained condition with minimum focal length f_{min} of the eye as shown in Figure 6.93(b).

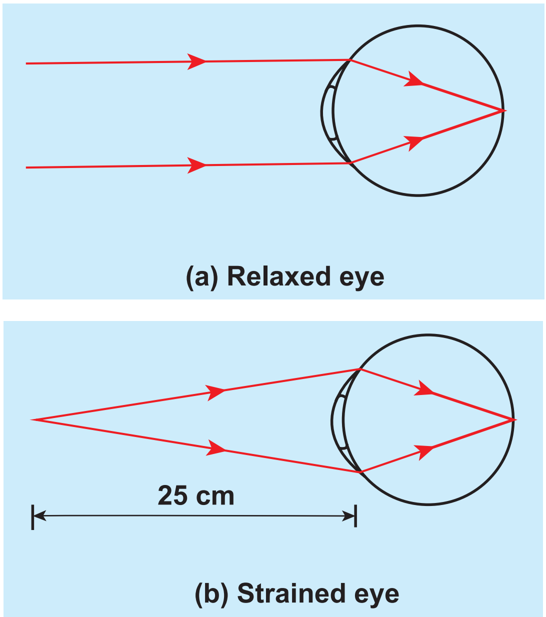


Figure 6.93 Focusing of normal eye

Let us find f_{max} and f_{min} of human eye from the lens equation given below.

$$\frac{1}{f}=\frac{1}{v}-\frac{1}{u} \qquad (6.69)$$

When the object is at infinity, $u = -\infty$, and $v = 2.5$ cm (distance between eye lens and retina), the eye can see the object in relaxed condition with f_{max}. Substituting these values in the lens equation gives,

$$\frac{1}{f_{\max}}=\frac{1}{2.5\ cm}-\frac{1}{-\infty}$$

$$f_{\max}=2.5\ cm$$

When the object is at near point, $u = -25$ cm, and $v = 2.5$ cm, the eye can see the object in strained condition with f_{min}. Substituting these values in the lens equation gives,

$$\frac{1}{f_{\min}}=\frac{1}{2.5\ cm}-\frac{1}{-25\ cm}$$

$$f_{\min}=2.27\ cm$$

See, the small variation of $f_{\max}-f_{\min}=0.23\ cm$ of the focal length of eye lens makes objects visible from infinity to near point for a normal person. Now, we can discuss some common defects of vision in the eye.

6.13.7.1 Nearsightedness (*myopia*)

A person suffering from nearsightedness or *myopia* cannot see distant objects clearly. This may result because the lens has too short focal length due to thickening of the lens or larger diameter of the eyeball than usual. These people have difficulty in relaxing their eye more than what is needed to overcome this difficulty. Thus, they need correcting lens.

The parallel rays coming from the distant object get focused before reaching the retina as shown in Figure 6.94(a). But, these persons can see objects which are nearer. Let x be the maximum distance up to which a person with nearsightedness can see as shown in Figure 6.94(b). To overcome this difficulty, the virtual image of the object at infinity should be formed at a distance x from the eye using a correcting lens as shown in Figure 6.93(c).

The focal length of the correcting lens for a myopic eye can be calculated using the lens equation.

$$\frac{1}{f}=\frac{1}{v}-\frac{1}{u} \qquad (6.69)$$

Here, $u = -\infty$, $v = -x$. Substituting these values in the lens equation gives,

$$\frac{1}{f}=\frac{1}{-x}-\frac{1}{-\infty}$$

Focal length f of the correcting lens is,

$$f=-x \qquad (6.196)$$

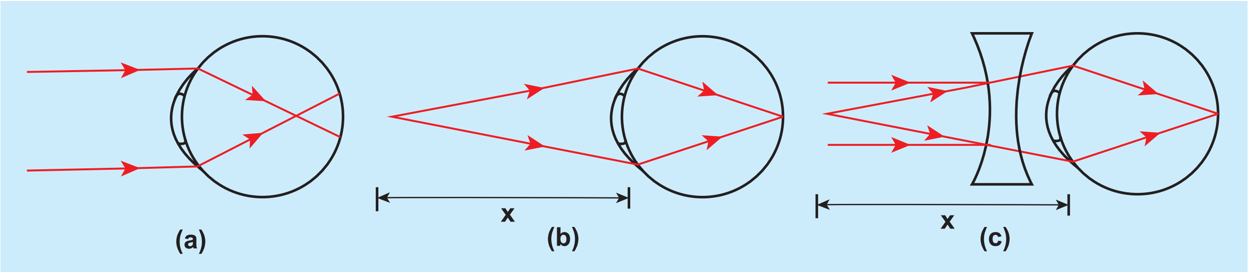


Figure 6.94 Myopic eye and correction

The negative sign in the above result suggests that the lens should be a concave lens. Basically, the concave lens slightly diverges the parallel rays from infinity and makes them focus now at the retina which got earlier focused before reaching retina in the unaided condition.

6.13.7.2 Farsightedness (*hypermetropia*)

A person suffering from farsightedness or *hypermetropia* or *hyperopia* cannot clearly see objects close to the eye. It occurs when the eye lens has too long focal length due to thining of eye lens or shortening of the eyeball than normal. The least distance for clear vision for these people is appreciably more than 25 cm and the person has to keep the object inconveniently away from the eye. Thus, reading or viewing smaller things held in the hands is difficult for them. This kind of farsightedness arising due to aging is called *presbyopia* as the aged people cannot strain their eye more to reduce the focal length of the eye lens.

The rays coming from the object at near point get focused beyond the retina as shown in Figure 6.95(a). But, these persons can see objects which are far say, more than 25 cm. Let y be the minimum distance from the eye beyond which a person with farsightedness can see as shown in Figure 6.95(b). To overcome this difficulty, the virtual image of the object at y should be formed at a distance of 25 cm (near point) from the eye using a correcting lens as shown in Figure 6.95(c).

The focal length of the correcting lens for a hypermetropic eye can be calculated using the lens equation.

$$\frac{1}{f}=\frac{1}{v}-\frac{1}{u} \tag{6.69}$$

Here, $u = -y$, $v = -25$ cm. Substituting these values in the lens equation gives,

$$\frac{1}{f}=\frac{1}{-y}-\frac{1}{-25\,cm}$$

Simplifying the above equation gives,

$$\frac{1}{f}=\frac{1}{25\,cm}-\frac{1}{y}=\frac{y-25\,cm}{y\times 25\,cm}$$

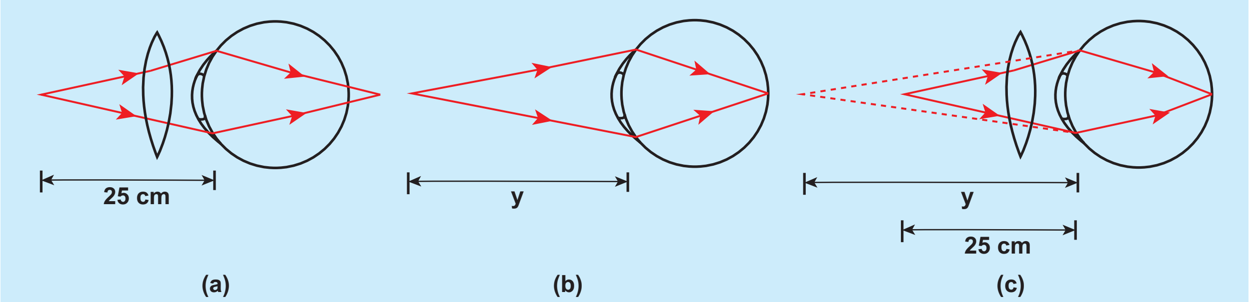


Figure 6.95 Hypermetropic eye and correction

$$f = \frac{y \times 25\ cm}{y - 25\ cm} \qquad (6.197)$$

The focal length calculated using above formula will be positive as y is always greater than 25 cm. The positive sign of the focal length suggests that the lens should be a convex lens. In principle, the convex lens slightly converges the rays coming from beyond y and makes them focus now at the retina which got earlier focused beyond retina for the unaided eye.

6.13.6.3 Astigmatism

Astigmatism is the defect arising due to different curvatures along different planes in the *eye* lens. Astigmatic person cannot see all the directions equally well. The defect due to astigmatism is more serious than myopia and hyperopia. The remedy to astigmatism is using of lenses with different curvatures in different planes to rectify the defect. In general, these specially made glasses with different curvature for different planes are called as cylindrical lenses.

Due to aging people may develop combination of more than one defect. If it is the combination of nearsightedness and farsightedness then, such persons may need a converging glass for reading purpose and a diverging glass for seeing at a distance. Bifocal lenses and progressive lenses provide solution for these problems.

EXAMPLE 6.44

Calculate the power of the lens of the spectacles necessary to rectify the defect of nearsightedness for a person who could see clearly only up to a distance of 1.8 m.

Solution

The maximum distance the person could see is, x = 1.8 m.

The lens should have a focal length of, $f = -x$ m = −1.8 m.

It is a concave or diverging lens.

The power of the lens is,

$$P = -\frac{1}{1.8\ m} = -0.56\ \text{diopter}$$

EXAMPLE 6.45

A person has farsightedness with the minimum distance he could see clearly is 75 cm. Calculate the power of the lens of the spectacles necessary to rectify the defect.

Solution

The minimum distance the person could see clearly is, y = 75 cm.

The lens should have a focal length of,

$$f = \frac{y \times 25\ cm}{y - 25\ cm}$$

$$f = \frac{75\ cm \times 25\ cm}{75\ cm - 25\ cm} = 37.5\ cm$$

It is a convex or converging lens.

The power of the lens is,

$$P = \frac{1}{0.375\ m} = 2.67\ \text{diopter}$$

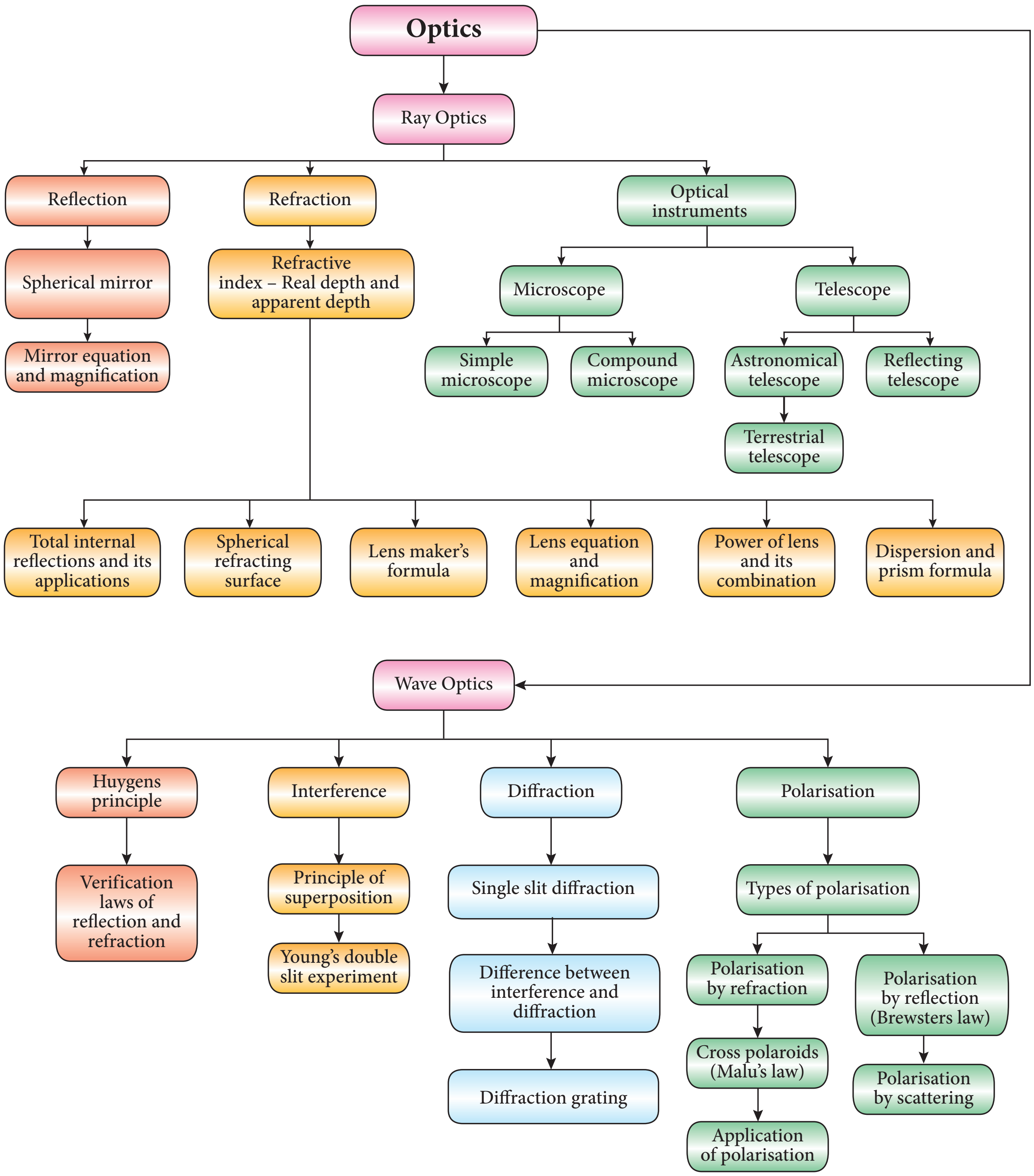
SUMMARY

- In ray optics, light is treated as a ray in the direction of light.
- Light undergoes reflection at polished surfaces and it is governed by laws of reflection.
- In general, plane mirrors form virtual and laterally inverted images at equal distance inside the mirror.
- The height of plane mirror needed to see a person fully in a mirror is half of the height of person.
- Spherical mirrors form a part of a sphere.
- Paraxial rays are the rays travelling close to the principal axis of the mirror and make small angles withit.
- There is a relation between f and R in spherical mirrors for paraxial rays.
- Image formation in spherical mirrors is based on mirror equation.
- There is a set of Cartesian sign conventions to be followed to trace image formed by spherical mirrors.
- Light travels with lesser velocity in optically denser medium.
- Optical path is the equivalent path travelled in vacuum in the same time light travels through a optically denser medium.
- The phenomenon of refraction is governed by laws of refraction (Snell's law).
- The apparent depth is always lesser than actual depth.
- Refraction takes place in atmosphere due to different layers of air with varying refractive indices.
- Total internal reflection takes place when light travels from denser to rarer medium with the angle of incidence greater than critical angle.
- There are several applications of total internal reflection.
- A glass slab produces lateral displacement or shift of ray entering into it.
- Thin lenses are formed by two spherical refracting surfaces.
- The image tracing in thin lenses is done with the Cartesian sign conventions and with the help of lens equation.
- Power and focal length are inverse to each other.
- There is effective focal length for lenses in contact and out of contact.
- Prism produces deviation on the incident ray.
- Angle of deviation depends on angle of prism, angle of incidence and refractive index of material of prism.
- The refractive index of prism depends on angle of prism and angle of minimum deviation.
- When white light travels through a medium, different colours travel with different speeds leading to dispersion of light.
- Dispersive power is the measure of ability of the medium to disperse white light.
- Rainbow is formed by dispersion of light by droplets of water.

- Light can be scattered by the particles present in atmosphere.
- The scattering of light by particles of size less than wavelength of light is called Rayleigh scattering which is inversely proportional to fourth power of wavelength.
- If the scattering is by suspended dust particles whose size is greater than wavelength of light, the scattering is independent of wavelength.
- There are four theories on light each explaining few aspects of light.
- Light has wave and particle nature.
- In wave optics we treat light propagating as a wavefront.
- Huygens' principle explains the propagation of light as wavefront.
- Laws of reflection and refraction are proved by Huygens' principle.
- In interference, two light waves are added to get varying intensities at different points.
- Coherent sources produce monochromatic light waves in phase or with constant phase difference.
- Coherent sources are obtained by intensity division, wavefront division and real and virtual images of light source.
- Young's double slit uses wavefront division to obtain coherent sources.
- Interference with polychromatic (white) light produces coloured interference fringes.
- Thin films appear coloured due to interference of white light.
- Bending of light around sharp edges is called diffraction.
- There are two types of diffractions called Fresnel and Fraunhofer diffractions
- Diffraction takes place at single slit which has a width comparable to the wavelength of light.
- Fresnel's distance is the distance up to which ray optics is obeyed.
- Diffraction can also happen in grating which has multiple slits of thickness comparable to wavelength of light used.
- Using diffraction grating and spectrometer wavelength of monochromatic light and also different colours of polychromatic light can be determined.
- Resolution is the quality of image which is decided by diffraction effect and Rayleigh criterion.
- Resolution is measured by the smallest distance which could be seen clearly without the blur due to diffraction.
- Polarisation is restricting electric or magnetic field vibrations to one plane.
- Polarisation is obtained by selective absorption, reflection, double refraction and scattering.
- Malus' law gives the intensity of emerging light when a polarised light enters two polaroids kept at an angle.
- Brewster's law relates angle of polarisation and refractive index of the medium.
- Optically active crystals can be classified as uniaxial and biaxial crystals.
- When light enters in to optically active crystals, double refraction takes place.

- In double refraction, ordinary ray obeys laws of refraction and extraordinary ray does not obey laws of refraction.
- Nicol prism separates ordinary and extraordinary rays by transparent cement called Canada balsam.
- Light scattered by molecules at perpendicular direction to the incident light is found to be plane polarised.
- Single convex lens can act as a simple microscope when object is within the focal length.
- Two focusing namely, near point focusing and normal focusing are possible.
- In near point focusing, the image is formed at 25 cm which is the distance of distinct vision for normal eye. Whereas, in normal focusing the image is formed at infinity.
- To find magnification in near point focusing we use lateral magnification and for normal focusing we use angular magnification.
- Resolution of microscope could be improved by using oil immersed eye piece.
- Compound microscope has improved magnification.
- Astronomical telescope has an eye piece of long focal length and eye piece of short focal length.
- Terrestrial telescopes have one addition lens for producing erect image.
- Reflecting telescopes have advantages as well as disadvantages.
- Eye has three problems (i) nearsightedness, (ii) farsightedness, (iii) astigmatism.

CONCEPT MAP



Optics
Ray Optics
Reflection
Refraction
Optical instruments
Spherical mirror
Refractive index – Real depth and apparent depth
Microscope
Telescope
Mirror equation and magnification
Simple microscope
Compound microscope
Astronomical telescope
Reflecting telescope
Terrestrial telescope
Total internal reflections and its applications
Spherical refracting surface
Lens maker's formula
Lens equation and magnification
Power of lens and its combination
Dispersion and prism formula
Wave Optics
Huygens principle
Interference
Diffraction
Polarisation
Verification laws of reflection and refraction
Principle of superposition
Young's double slit experiment
Single slit diffraction
Difference between interference and diffraction
Diffraction grating
Types of polarisation
Polarisation by refraction
Polarisation by reflection (Brewsters law)
Cross polaroids (Malu's law)
Polarisation by scattering
Application of polarisation

Multiple choice questions

1. The speed of light in an isotropic medium depends on,

(a) its intensity

(b)its wavelength

(c) the nature of propagation

(d) the motion of the source w.r.t medium

D7F8Y5

2. A rod of length 10 cm lies along the principal axis of a concave mirror of focal length 10 cm in such a way that its end closer to the pole is 20 cm away from the mirror. The length of the image is, (AIPMT Main 2012)

(a) 2.5 cm (b) 5cm

(c) 10 cm (d) 15cm

3. An object is placed in front of a convex mirror of focal length of f and the maximum and minimum distance of an object from the mirror such that the image formed is real and magnified. (IEE Main 2009)]

(a) $2f$ and c (b) c and ∞

(c) f and O (d) None of these

4. For light incident from air on a slab of refractive index 2, the maximum possible angle of refraction is,

(a) 30^{o} (b) 45^{o}

(c) 60^{o} (d) 90^{o}

5. If the velocity and wavelength of light in air is V_a and λ_a and that in water is V_w and λ_w, then the refractive index of water is,

(a) $\frac{V_w}{V_a}$ (b) $\frac{V_a}{V_w}$

(c) $\frac{\lambda_w}{\lambda_a}$ (d) $\frac{V_a \lambda_a}{V_w \lambda_w}$

6. Stars twinkle due to,

(a) reflection

(b) total internal reflection

(c) refraction

(d) polarisation

7. When a biconvex lens of glass having refractive index 1.47 is dipped in a liquid, it acts as a plane sheet of glass. This implies that the liquid must have refractive index,

(a) less than one

(b) less than that of glass

(c) greater than that of glass

(d) equal to that of glass

8. The radius of curvature of curved surface at a thin planoconvex lens is 10 cm and the refractive index is 1.5. If the plane surface is silvered, then the focal length will be,

(a) 5 cm (b) 10 cm

(c) 15 cm (d) 20 cm

9. An air bubble in glass slab of refractive index 1.5 (near normal incidence) is 5 cm deep when viewed from one surface and 3 cm deep when viewed from the opposite face. The thickness of the slab is,

(a) 8 cm (b) 10 cm

(c) 12 cm (d) 16 cm

10. A ray of light travelling in a transparent medium of refractive index n falls, on a surface separating the medium from air at an angle of incidents of 45^{o}. The ray can undergo total internal reflection for the following n,

(a) $n = 1.25$ (b) $n = 1.33$

(c) $n = 1.4$ (d) $n = 1.5$

11. A plane glass is placed over a various coloured letters (violet, green, yellow, red) The letter which appears to be raised more is,

(a) red (b) yellow

(c) green (d) violet

12. Two point white dots are 1 mm apart on a black paper. They are viewed by eye of pupil diameter 3 mm approximately. The maximum distance at which these dots can be resolved by the eye is, [take wavelength of light, λ = 500 nm]

(a) 1 m (b) 5 m

(c) 3 m (d) 6m

13. In a Young's double-slit experiment, the slit separation is doubled. To maintain the same fringe spacing on the screen, the screen-to-slit distance D must be changed to,

(a) $2D$ (b) $\frac{D}{2}$

(c) $\sqrt{2}D$ (d) $\frac{D}{\sqrt{2}}$

14. Two coherent monochromatic light beams of intensities I and $4I$ are superposed. The maximum and minimum possible intensities in the resulting beam are [IIT-JEE 1988]

(a) $5I$ and I (b) $5I$ and $3I$

(c) $9I$ and I (d) $9I$ and $3I$

15. When light is incident on a soap film of thickness 5×10^{-5} cm, the wavelength of light reflected maximum in the visible region is 5320 Å. Refractive index of the film will be,

(a) 1.22 (b) 1.33

(c) 1.51 (d) 1.83.

16. First diffraction minimum due to a single slit of width 1.0×10^{-5} cm is at 30°. Then wavelength of light used is,

(a) 400 Å (b) 500 Å

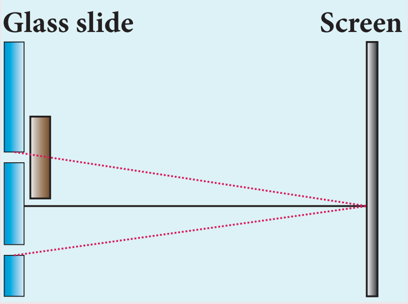
(c) 600 Å (d) 700 Å

17. A ray of light strikes a glass plate at an angle 60°. If the reflected and refracted rays are perpendicular to each other, the refractive index of the glass is,

(a) $\sqrt{3}$ (b) $\frac{3}{2}$

(c) $\sqrt{\frac{3}{2}}$ (d) 2

18. One of the of Young's double slits is covered with a glass plate as shown in figure. The position of central maximum will,



(a) get shifted downwards

(b) get shifted upwards

(c) will remain the same

(d) data insufficient to conclude

19. Light transmitted by Nicol prism is,

(a) partially polarised

(b) unpolarised

(c) plane polarised

(d) elliptically polarised

20. The transverse nature of light is shown in,

(a) interference

(b) diffraction

(c) scattering

(d) polarisation

Answers

1) b	2) b	3) d	4) a	5) b
6) c	7) d	8) b	9) c	10) d
11) d	12) b	13) a	14) c	15) b
16) b	17) a	18) b	19) c	20) d

Short Answer Questions

1. State the laws of reflection.
2. What is angle of deviation due to reflection?
3. Give the characteristics of image formed by a plane mirror.
4. Derive the relation between f and R for a spherical mirror.
5. What are the Cartesian sign conventions for a spherical mirror?
6. What is optical path? Obtain the equation for optical path of a medium of thickness d and refractive index n.
7. State the laws of refraction.
8. What is angle of deviation due to refraction?
9. What is principle of reversibility?
10. What is relative refractive index?
11. Obtain the equation for apparent depth.
12. Why do stars twinkle?
13. What is critical angle and total internal reflection?
14. Obtain the equation for critical angle.
15. Explain the reason for glittering of diamond.
16. What are mirage and looming?
17. Write a short notes on the prisms making use of total internal reflection.
18. What is Snell's window?
19. Write a note on optical fibre.
20. Explain the working of an endoscope.
21. What are primary focus and secondary focus of convex lens?
22. What are the sign conventions followed for lenses?
23. Arrive at lens equation from lens maker's formula.
24. Obtain the equation for lateral magnification for thin lens.
25. What is power of a lens?
26. Derive the equation for effective focal length for lenses in contact.
27. What is angle of minimum deviation?
28. What is dispersion?
29. How are rainbows formed?
30. What is Rayleigh's scattering?
31. Why does sky appear blue?
32. What is the reason for reddish appearance of sky during sunset and sunrise?
33. Why do clouds appear white?
34. What are the salient features of corpuscular theory of light?
35. What is wave theory of light?
36. What is electromagnetic wave theory of light?
37. Write a short note on quantum theory of light.
38. What is a wavefront?
39. What is Huygens' principle?
40. What is interference of light?
41. What is phase of a wave?
42. Obtain the relation between phase difference and path difference.
43. What are coherent sources?
44. What is intensity division?

45. How does wavefront division provide coherent sources?
46. How do source and images behave as coherent sources?
47. What is bandwidth of interference pattern?
48. What is diffraction?
49. Differentiate between Fresnel and Fraunhofer diffraction.
50. Discuss the special cases on first minimum in Fraunhofer diffraction.
51. What is Fresnel's distance? Obtain the equation for Fresnel's distance.
52. Mention the differences between interference and diffraction.
53. What is a diffraction grating?
54. What are resolution and resolving power?
55. What is Rayleigh's criterion?
56. What is polarisation?
57. Differentiate between polarised and unpolarised light
58. Discuss polarisation by selective absorption.
59. What are polariser and analyser?
60. What are plane polarised, unpolarised and partially polarised light?
61. State and obtain Malus' law.
62. List the uses of polaroids.
63. State Brewster's law.
64. What is angle of polarisation and obtain the equation for angle of polarisation.
65. Discuss about pile of plates.
66. What is double refraction?
67. Mention the types of optically active crystals with example.
68. Discuss about Nicol prism.
69. How is polarisation of light obtained by scattering of light?
70. Discuss about simple microscope and obtain the equations for magnification for near point focusing and normal focusing.
71. What are near point and normal focusing?
72. Why is oil immersed objective preferred in a microscope?
73. What are the advantages and disadvantages of using a reflecting telescope?
74. What is the use of an erecting lens in a terrestrial telescope?
75. What is the use of collimator?
76. What are the uses of spectrometer?
77. What is myopia? What is its remedy?
78. What is hypermetropia? What is its remedy?
79. What is presbyopia?
80. What is astigmatism?

Long Answer Questions

1. Derive the mirror equation and the equation for lateral magnification.
2. Describe the Fizeau's method to determine speed of light.
3. Obtain the equation for radius of illumination (or) Snell's window.
4. Derive the equation for acceptance angle and numerical aperture, of optical fiber.
5. Obtain the equation for lateral displacement of light passing through a glass slab.
6. Derive the equation for refraction at single spherical surface.

7. Obtain lens maker's formula and mention its significance.
8. Derive the equation for thin lens and obtain its magnification.
9. Derive the equation for effective focal length for lenses in out of contact.
10. Derive the equation for angle of deviation produced by a prism and thus obtain the equation for refractive index of material of the prism.
11. What is dispersion? Obtain the equation for dispersive power of a medium.
12. Prove laws of reflection using Huygens' principle.
13. Prove laws of refraction using Huygens' principle.
14. Obtain the equation for resultant intensity due to interference of light.
15. Explain the Young's double slit experimental setup and obtain the equation for path difference.
16. Obtain the equation for bandwidth in Young's double slit experiment.
17. Obtain the equations for constructive and destructive interference for transmitted and reflected waves in thin films.
18. Discuss diffraction at single slit and obtain the condition for n^{th} minimum.
19. Discuss the diffraction at a grating and obtain the condition for the m^{th} maximum.
20. Discuss the experiment to determine the wavelength of monochromatic light using diffraction grating.
21. Discuss the experiment to determine the wavelength of different colours using diffraction grating.
22. Obtain the equation for resolving power of optical instrument.
23. Discuss about simple microscope and obtain the equations for magnification for near point focusing and normal focusing.
24. Explain about compound microscope and obtain the equation for magnification.
25. Obtain the equation for resolving power of microscope.
26. Discuss about astronomical telescope.
27. Mention different parts of spectrometer and explain the preliminary adjustments.
28. Explain the experimental determination of material of the prism using spectrometer.

Conceptual Questions

1. Why are dish antennas curved?
2. What type of lens is formed by a bubble inside water?
3. It is possible for two lenses to produce zero power?
4. Why does sky look blue and clouds look white?
5. Why is yellow light preferred to during fog?
6. Two independent monochromatic sources cannot act as coherent sources, why?
7. Does diffraction take place at the Young's double slit?
8. Is there any difference between coloured light obtained from prism and colours of soap bubble?

9. A small disc is placed in the path of the light from distance source. Will the center of the shadow be bright or dark?

10. When a wave undergoes reflection at a denser medium, what happens to its phase?

Numerical Problems

1. An object is placed at a certain distance from a convex lens of focal length 20 cm. Find the distance of the object if the image obtained is magnified 4 times.

[Ans: –15 cm.]

2. A compound microscope has a magnification of 30. The focal length of eye piece is 5 cm. Assuming the final image to be at least distance of distinct vision, find the magnification produced by the objective.

[Ans: 5]

3. An object is placed in front of a concave mirror of focal length 20 cm. The image formed is three times the size of the object. Calculate two possible distances of the object from the mirror.

[Ans: –40/3 cm and –80/3 cm]

4. A small bulb is placed at the bottom of a tank containing water to a depth of 80 cm.What is the area of the surface of water through which light from the bulb can emerge out? Refractive index of water is 1.33. (Consider the bulb to be a point source.)

[Ans: 2.6 m^2]

5. A thin converging glass lens made of glass with refractive index 1.5 has a power of + 5.0 *D*. When this lens is immersed in a liquid of refractive index *n*, it acts as a divergent lens of focal length 100 cm. What must be the value of *n*?

[Ans: 5/3]

6. If the distance *D* between an object and screen is greater than 4 times the focal length of a convex lens, then there are two positions of the lens for which images are formed on the screen. This method is called conjugate foci method. If *d* is the distance between the two positions of the lens, obtain the equation for focal length of the convex lens.

[Ans: $\left(f = \dfrac{D^2 - d^2}{4D}\right)$

7. A beam of light of wavelength 600 nm from a distant source falls on a single slit 1 mm wide and the resulting diffraction pattern is observed on a screen 2 m away. What is the distance between the first dark fringe on either side of the central bright fringe?

[Ans: 2.4 mm]

8. In Young's double slit experiment, the slits are 2 mm apart and are illuminated with a mixture of two wavelength $\lambda_0 = 750$ nm and $\lambda = 900$ nm. What is the minimum distance from the common central bright fringe on a screen 2 m from the slits where a bright fringe from one interference pattern coincides with a bright fringe from the other?

[Ans: 4.5 mm]

9. In Young's double slit experiment, 62 fringes are seen in visible region for sodium light of wavelength 5893 Å. If violet light of wavelength 4359 Å is used in place of sodium light, then

what is the number of fringes seen? [Ans: 84]

10. A compound microscope has a magnifying power of 100 when the image is formed at infinity. The objective has a focal length of 0.5 cm and the tube length is 6.5 cm. What is the focal length of the eyepiece.

[Ans: 2 cm]

BOOKS FOR REFERENCE

1. Frances A. Jenkins and Harvey E. White, Fundamentals of Optics, 4th Edition, McGraw Hill Book Company, (2011).
2. David Halliday, Robert Resnick and Jearl Walker,Fundamentals of Physics, 6th Edition,John Wiley & Sons Inc., (2004).
3. H.C. Verma,Concepts of Physics [Part-1], 1st Edition, Bharathi Bhawan Publishers & Distributers Pvt. Ltd., (2008).
4. Roger A. Freedman, Hugh D. Young, Sears and Zemansky's University Physics, 12th Edition, Pearson, (2011).

ICT CORNER

Optics

In this activity you will be able to explore the behaviour of a Young's double slit experiment by adjusting the slit separation, the distance to the screen, and the wave-length of the light.

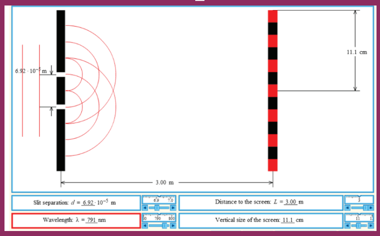
Topic: Young's double slit experiment.

STEPS:

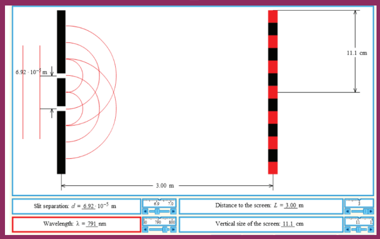
- Open the browser and type
- 'tutor-homework.com/Physics_Help/double_slit_experiment.html' in the address bar.
- Change the slit separation (distance between two sources) andobservehow the pattern of bright and dark fringes changes.
- What happens to the fringe width if distance between the source and screen decreases?
- Observe how does the fringe width in interference pattern vary with the wavelength of incident light?

Observe the pattern of bright and dark fringes by clicking the Run button.

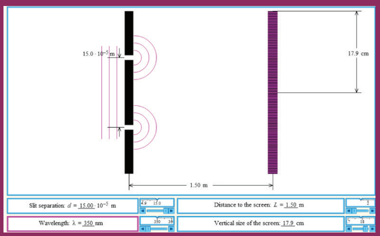
Step1



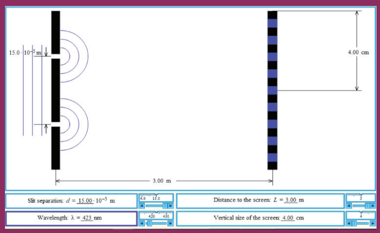
Step2



Step3



Step4



Note:

Use flash enabled browser or install flash player in your system.

URL:

http://tutor-homework.com/Physics_Help/double_slit_experiment.html

* Pictures are indicative only.

* If browser requires, allow **Flash Player** or **Java Script** to load the page.

B263_12_PHYSICS_EM

UNIT 8

ATOMIC AND NUCLEAR PHYSICS

All of physics is either impossible or trivial. It is impossible until you understand it, and then it becomes trivial

– Ernest Rutherford

LEARNING OBJECTIVES

In this unit, the students are exposed to

- electric discharge through the gases
- determination of specific charge by J.J. Thomson experiment
- determination of electronic charge by Millikan's oil drop experiment
- atom models – J.J. Thomson and Rutherford
- Bohr atom model and hydrogen atom
- atomic spectrum and hydrogen spectrum
- structure and properties of nucleus
- various classification of nuclei based on atomic and mass number
- mass defect and binding energy
- relation between stability and binding energy curve
- alpha, beta and gamma decay
- law of radioactive decay
- nuclear fission and fusion
- elementary ideas of nuclear reactors
- qualitative idea of elementary particles

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8.1 INTRODUCTION

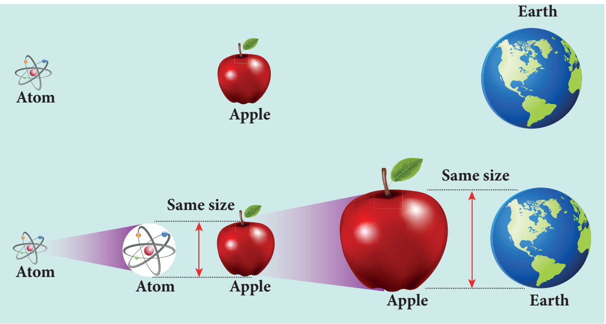


Figure 8.1 Comparision of size of an atom with that of an apple and comparision of size of an apple with that of the Earth

In earlier classes, we have studied that anything which occupies space is called matter. Matter can be classified into solids, liquids and gases. In our daily life, we use water for drinking, petrol for vehicles, we inhale oxygen, stainless steel vessels for cooking, etc. Experiences tell us that behaviour of one material is not same as another, this means that the physical and chemical properties are different for different materials. In order to understand this, we need to know the fundamental constituents of materials.

When an object is divided repeatedly, the process of division could not be done beyond a certain stage in a similar way and we end up with a small speck. This small speck was defined as an atom. The word atom in Greek means 'without division or indivisible'. The size of an atom is very very small. For an example, the size of hydrogen atom (simplest among other atoms) is arround 10^{-10} m. An American Physicist Richard P. Feynman said that if the atom becomes the size of an apple, then the apple becomes the size of the earth as shown in Figure 8.1. Such a small entity is an atom.

In this unit, we first discuss the theoretical models of atom to understand its structure. The Bohr atom model is more successful than J. J. Thomson and Rutherford atom models. It explained many unsolved issues in those days and also gave better understanding of chemistry.

Later, scientists observed that even the atom is not the fundamental entity. It consists of electrons and nucleus. Around 1930, scientists discovered that nucleus is also made of proton and neutron. Further research discovered that even the proton and neutron are made up of fundamental entities known as quarks.

In this context, the remaining part of this unit is written to understand the structure and basic properties of nucleus. Further how the nuclear energy is produced and utilized are discussed.

8.2 ELECTRIC DISCHARGE THROUGH GASES

Gases at normal atmospheric pressure are poor conductors of electricity because they do not have free electrons for conduction. But by special arrangement, one can make a gas to conduct electricity.

A simple and convenient device used to study the conduction of electricity through gases is known as gas discharge tube. The arrangement of discharge tube is shown in Figure 8.2. It consists of a long closed glass tube (of length nearly 50 cm and diameter of 4 cm) inside of which the gas in pure form is filled usually. The small opening in the tube is connected to a high vacuum pump and a low-pressure gauge. This tube is fitted with two metallic plates known as electrodes which are connected to secondary of an induction coil. The electrode connected to positive of secondary is known as anode and the electrode to the negative of the secondary is cathode. The potential of secondary is maintained about 50 kV.

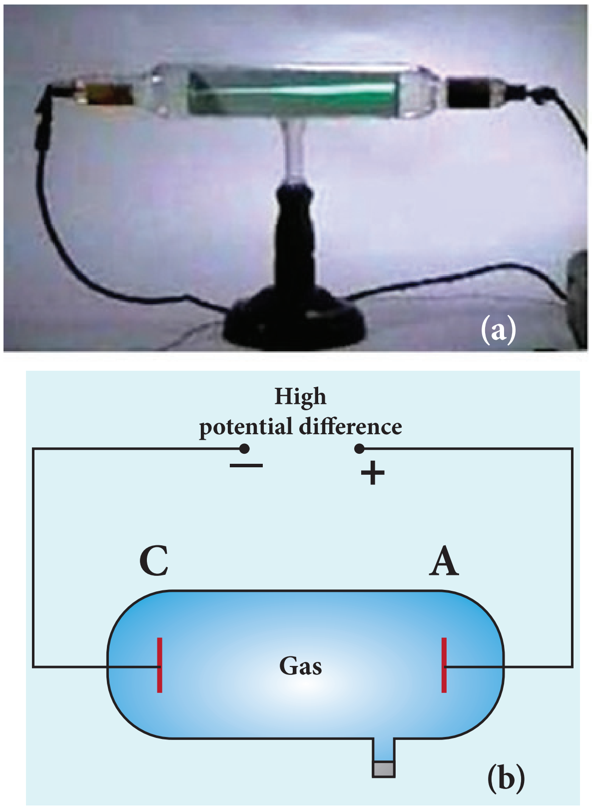


Figure 8.2 Discharge tube (a) real picture (b) schematic diagram

Suppose the pressure of the gas in discharge tube is reduced to around 110 mm of Hg using vacuum pump, it is observed that no discharge takes place. When the pressure is kept near 100 mm of Hg, the discharge of electricity through the tube takes place. Consequently, irregular streaks of light appear and also crackling sound is produced. When the pressure is reduced to the order of 10 mm of Hg, a luminous column known as positive column is formed from anode to cathode.

When the pressure reaches to around 0.01 mm of Hg, positive column disappears. At this time, a dark space is formed between anode and cathode which is often called Crooke's dark space and the walls of the tube appear with green colour. At this stage, some invisible rays emanate from cathode called cathode rays, which are later found be a beam of electrons.

Properties of cathode rays

(1) Cathode rays possess energy and momentum and travel in a straight line with high speed of the order of 10^7m s^{-1}. It can be deflected by application of electric and magnetic fields. The direction of deflection indicates that they are negatively charged particles.

(2) When the cathode rays are allowed to fall on matter, they produce heat. They affect the photographic plates and also produce fluorescence when they fall on certain crystals and minerals.

(3) When the cathode rays fall on a material of high atomic weight, x-rays are produced.

(4) Cathode rays ionize the gas through which they pass.

(5) The speed of cathode rays is up to $\left(\frac{1}{10}\right)^{th}$ of the speed of light.

8.2.1 Determination of specific charge $\left(\frac{e}{m}\right)$ of an electron – Thomson's experiment

Thomson's experiment is considered as one among the landmark experiments for the birth of modern physics. In 1887, J. J. Thomson made remarkable improvement in the scope of study of gases in discharge tubes. In the presence of electric and magnetic fields, the cathode rays are deflected. By the variation of electric and magnetic fields, mass normalized charge or the specific charge (charge per unit mass) of the cathode rays is measured.

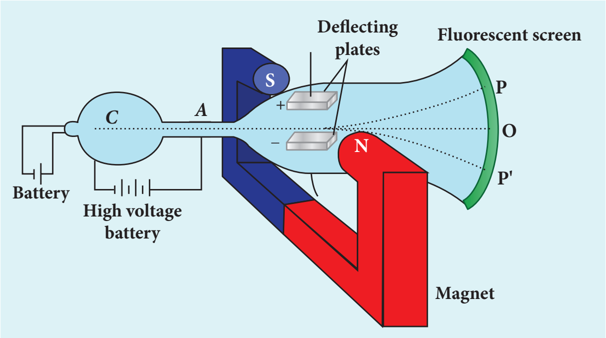


Figure 8.3 Arrangement of J.J. Thomson experiment to determine the specific charge of an electron

The arrangement of J. J. Thomson's experiment is shown in Figure 8.3. A highly evacuated discharge tube is used and cathode rays (electron beam) produced at cathode are attracted towards anode disc A. Anode disc is made with pin hole in order to allow only a narrow beam of cathode rays. These cathode rays are now allowed to pass through the parallel metal plates, maintained at high voltage as shown in Figure 8.3. Further, this gas discharge tube is kept in between pole pieces of magnet such that both electric and magnetic fields are perpendicular to each other. When the cathode rays strike the screen, they produce scintillation and hence bright spot is observed. This is achieved by coating the screen with zinc sulphide.

(i) Determination of velocity of cathode rays

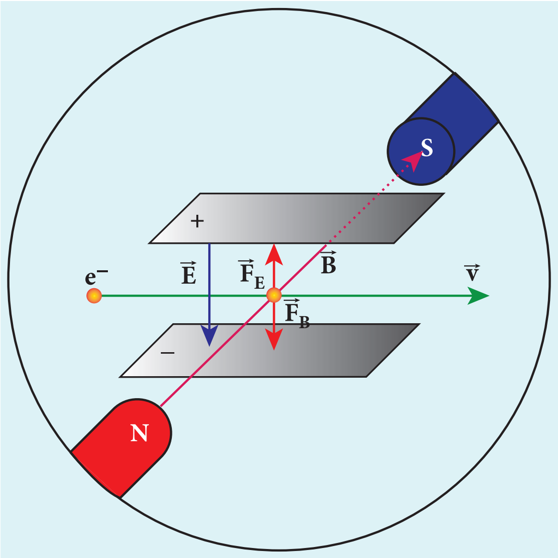


Figure 8.4 Electric force balancing the magnetic force – the path of electron beam is a straight line

For a fixed electric field between the plates, the magnetic field is adjusted such that the cathode rays (electron beam) strike at the original position O (Figure 8.3). This means that the magnitude of electric force is balanced by the magnitude of force due to magnetic field as shown in Figure 8.4. Let e be the charge of the cathode rays, then

$$eE = eB\upsilon$$

$$\Rightarrow \upsilon = \frac{E}{B} \quad (8.1)$$

(ii) Determination of specific charge

Since the cathode rays (electron beam) are accelerated from cathode to anode, the potential energy of the electron beam at the cathode is converted into kinetic energy of the electron beam at the anode. Let V be the potential difference between anode and cathode, then the potential energy is eV. Then from law of conservation of energy,

$$eV = \frac{1}{2}m\upsilon^2 \Rightarrow \frac{e}{m} = \frac{\upsilon^2}{2V}$$

Substituting the value of velocity from equation (8.1), we get

$$\frac{e}{m} = \frac{1}{2V}\frac{E^2}{B^2} \quad (8.2)$$

Substituting the values of E, B and V, the specific charge can be determined as

$$\frac{e}{m} = 1.7 \times 10^{11}\,\mathrm{C\,kg^{-1}}$$

(iii) Deflection of charge only due to uniform electric field

When the magnetic field is turned off, the deflection is only due to electric field. The deflection in vertical direction is due to the electric force.

$$F_e = eE \quad (8.3)$$

Let m be the mass of the electron and by applying Newton's second law of motion, acceleration of the electron is

$$a_e = \frac{1}{m}F_e \quad (8.4)$$

Substituting equation (8.4) in equation (8.3),

$$a_e = \frac{1}{m}eE = \frac{e}{m}E$$

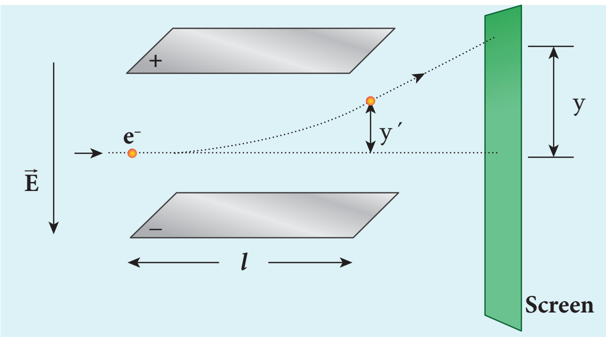


Figure 8.5 Deviation of path by applying uniform electric field

Let y be the deviation produced from original position on the screen as shown in Figure 8.5. Let the initial upward velocity of cathode ray be $u=0$ before entering the parallel electric plates. Let t be the time taken by the cathode rays to travel in electric field. Let l be the length of one of the plates, then the time taken is

$$t=\frac{l}{\upsilon} \quad (8.5)$$

Hence, the deflection y' of cathode rays is (note: $u=0$ and $a_e=\frac{e}{m}E$)

$$y'=ut+\frac{1}{2}at^2 \Rightarrow y'=ut+\frac{1}{2}a_e t^2$$

$$=\frac{1}{2}\left(\frac{e}{m}E\right)\left(\frac{l}{\upsilon}\right)^2$$

$$y'=\frac{1}{2}\frac{e}{m}\frac{l^2B^2}{E} \quad (8.6)$$

Therefore, the deflection y on the screen is

$$y \propto y' \Rightarrow y=Cy'$$

where C is proportionality constant which depends on the geometry of the discharge tube and substituting y' value in equation 8.6, we get

$$y=C\frac{1}{2}\frac{e}{m}\frac{l^2B^2}{E} \quad (8.7)$$

Rearranging equation (8.7) as

$$\frac{e}{m}=\frac{2yE}{Cl^2B^2} \quad (8.8)$$

Substituting the values on RHS, the value of specific charge is calculated as

$$\frac{e}{m}=1.7\times10^{11}\,\mathrm{C\,kg^{-1}}$$

(iv) Deflection of charge only due to uniform magnetic field

Suppose that the electric field is switched off and only the magnetic field is switched on. Now the deflection occurs only due to magnetic field. The force experienced by the electron in uniform magnetic field applied perpendicular to its path is

$$F_m=e\upsilon B \qquad \text{(in magnitude)}$$

Since this force provides the centripetal force, the electron beam undergoes a semi-circular path. Therefore, we can equate F_m to centripetal force $\frac{m\upsilon^2}{R}$.

$$F_m=e\upsilon B=m\frac{\upsilon^2}{R}$$

where v is the velocity of electron beam at the point where it enters the magnetic field and R is the radius of the circular path traversed by the electron beam.

$$eB=m\frac{\upsilon}{R}\Rightarrow\frac{e}{m}=\frac{\upsilon}{BR} \quad (8.9)$$

Further, substituting equation (8.1) in equation (8.10), we get

$$\frac{e}{m}=\frac{E}{B^2R} \quad (8.10)$$

By knowing the values of electric field, magnetic field and the radius of circular path, the value of specific charge $\left(\frac{e}{m}\right)$ can be calculated, which is also consistant with other two methods.

Note

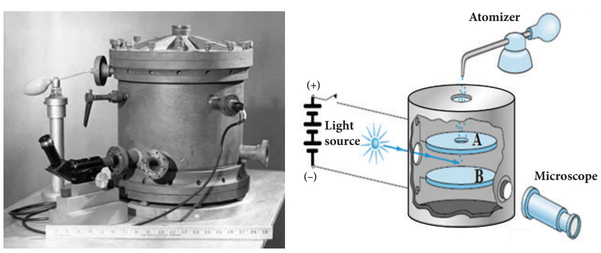
The specific charge is independent of
(a) gas used
(b) nature of the electrodes

8.2.2 Determination of charge of an electron – Millikan's oil drop experiment

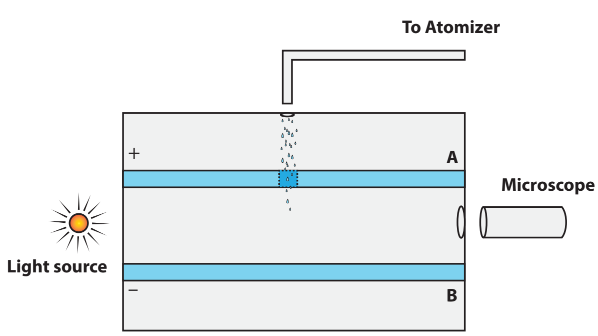
Millikan's oil drop experiment is another important experiment in modern physics which is used to determine one of the fundamental constants of nature known as charge of an electron (Figure 8.6 (a)).

By adjusting electric field suitably, the motion of oil drop inside the chamber can be controlled – that is, it can be made to move up or down or even kept balanced in the field of view for sufficiently long time.

Z7X6Q7



(a)



(b)

Figure 8.6 Millikan's experiment (a) real picture and schematic picture (b) Side view picture

The experimental arrangement is shown in Figure 8.6 (b). The apparatus consists of two horizontal circular metal plates A and B each with diameter around 20 cm and are separated by a small distance 1.5 cm. These two parallel plates are enclosed in a chamber with glass walls. Further, plates A and B are given a high potential difference around 10 kV such that electric field acts vertically downward. A small hole is made at the centre of the upper plate A and atomizer is kept exactly above the hole to spray the liquid. When a fine droplet of highly viscous liquid (like glycerine) is sprayed using atomizer, it falls freely downward through the hole of the top plate only under the influence of gravity.

Few oil drops in the chamber can acquire electric charge (negative charge) because of friction with air or passage of x-rays in between the parallel plates. Further the chamber is illuminated by light which is passed horizontally and oil drops can be seen clearly using microscope placed perpendicular to the light beam. These drops can move either upwards or downward.

Let m be the mass of the oil drop and q be its charge. Then the forces acting on the droplet are

(a) gravitational force $F_g = mg$
(b) electric force $F_e = qE$
(c) buoyant force F_b
(d) viscous force F_v

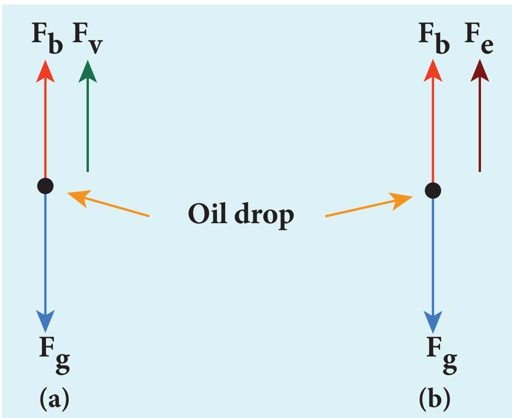


Figure 8.7 Free body diagram of the oil drop – (a) without electric field (b) with electric field

(a) Determination of radius of the droplet

When the electric field is switched off, the oil drop accelerates downwards. Due to the presence of air drag forces, the oil drops easily attain its terminal velocity and moves with constant velocity. This velocity can be carefully measured by noting down the time taken by the oil drop to fall through a predetermined distance. The free body diagram of the oil drop is shown in Figure 8.7 (a), we note that viscous force and buoyant force balance the gravitational force.

Let the gravitational force acting on the oil drop (downward) be $F_g = mg$

Let us assume that oil drop to be spherical in shape. Let ρ be the density of the oil drop, and r be the radius of the oil drop, then the mass of the oil drop can be expressed in terms of its density as

$$\rho = \frac{m}{V}$$

$$\Rightarrow m = \rho\left(\frac{4}{3}\pi r^3\right) \qquad \left(\because \text{volume of the sphere, } V = \frac{4}{3}\pi r^3\right)$$

The gravitational force can be written in terms of density as

$$F_g = mg \Rightarrow F_g = \rho\left(\frac{4}{3}\pi r^3\right)g$$

Let σ be the density of the air, the upthrust force experienced by the oil drop due to displaced air is

$$F_b = \sigma\left(\frac{4}{3}\pi r^3\right)g$$

Once the oil drop attains a terminal velocity v, the net downward force acting on the oil drop is equal to the viscous force acting opposite to the direction of motion of the oil drop. From Stokes law, the viscous force on the oil drop is

$$F_v = 6\pi r\upsilon\eta$$

From the free body diagram as shown in Figure 8.7 (a), the force balancing equation is

$$F_g = F_b + F_\upsilon$$

$$\rho\left(\frac{4}{3}\pi r^3\right)g = \sigma\left(\frac{4}{3}\pi r^3\right)g + 6\pi r\upsilon\eta$$

$$\frac{4}{3}\pi r^3(\rho-\sigma)g = 6\pi r\upsilon\eta$$

$$\frac{2}{3}\pi r^3(\rho-\sigma)g = 3\pi r\upsilon\eta$$

$$r = \left[\frac{9\eta\upsilon}{2(\rho-\sigma)g}\right]^{\frac{1}{2}} \qquad (8.11)$$

Thus, equation (8.11) gives the radius of the oil drop.

(b) Determination of electric charge

When the electric field is switched on, charged oil drops experience an upward electric force (qE). Among many drops, one particular drop can be chosen in the field of view of microscope and strength of the electric field is adjusted to make that particular drop to be stationary. Under these circumstances, there will be no viscous force acting on the oil drop. Then, from the free body diagram shown Figure 8.7 (b), the net force acting on the oil droplet is

$$F_e + F_b = F_g$$

$$\Rightarrow qE + \frac{4}{3}\pi r^3\sigma g = \frac{4}{3}\pi r^3\rho g$$

$$\Rightarrow qE = \frac{4}{3}\pi r^3(\rho-\sigma)g$$

$$\Rightarrow q = \frac{4}{3E}\pi r^3(\rho-\sigma)g \qquad (8.12)$$

Substituting equation (8.11) in equation (8.12), we get

$$q = \frac{18\pi}{E}\left(\frac{\eta^3\upsilon^3}{2(\rho-\sigma)g}\right)^{\frac{1}{2}}$$

Millikan repeated this experiment several times and computed the charges on oil drops. He found that the charge of any oil drop can be written as integral multiple of a basic value, $-1.6\times10^{-19}\,C$, which is nothing but the charge of an electron.

8.3 ATOM MODELS

Introduction

Around 400 B.C, Greek philosophers Leucippus and Democretus proposed the concept of atom, 'Every object on continued subdivision ultimately yields atoms'. Later, many physicists and chemists tried to understand the nature with the idea of atoms. Many theories were proposed to explain the properties (physical and chemical) of bulk materials on the basis of atomic model.

For instance, J. J. Thomson proposed a theoretical atom model which is based on static distribution of electric charges. Since this model fails to explain the stability of atom, one of his students E. Rutherford proposed the first dynamic model of an atom. Rutherford gave atom model which is based on results of an experiment done by his students (Geiger and Marsden). But this model also failed to explain the stability of the atom.

Later, Niels Bohr who is also a student of Rutherford proposed an atomic model for hydrogen atom which is more successful than other two models. Niels Bohr atom model could explain the stability of the atom and also the origin of line spectrum. There are other atom models, such as Sommerfeld's atom model and atom model from wave mechanics (quantum mechanics). But we will restrict ourselves only to very simple (mathematically simple) atom model in this section.

8.3.1 J. J. Thomson's Model (Water melon model)

In this model, the atoms are visualized as homogeneous spheres which contain uniform distribution of positively charged particles (Figure 8.8 (a)). The negatively charged particles known as electrons are embedded in it like seeds in water melon as shown in Figure 8.8 (b).

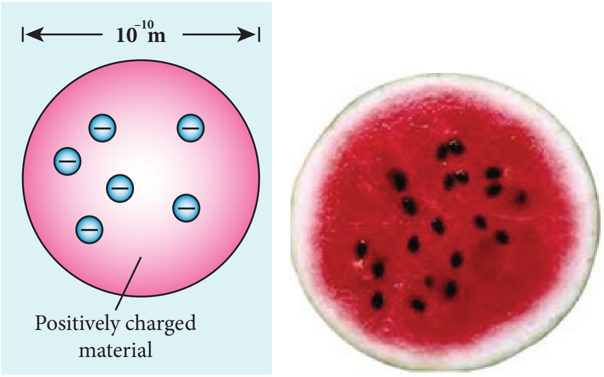


Figure 8.8 (a) Atom (b) Water melon

The atoms are electrically neutral, this implies that the total positive charge in an atom is equal to the total negative charge. According to this model, all the charges are assumed to be at rest. But from classical electrodynamics, no stable equilibrium points exist in electrostatic configuration (this is known as Earnshaw's theorem) and hence such an atom cannot be stable. Further, it fails to explain the origin of spectral lines observed in the spectrum of hydrogen atom and other atoms.

8.3.2 Rutherford's model

In 1911, Geiger and Marsden did a remarkable experiment based on the advice of their teacher Rutherford, which is known as scattering of alpha particles by gold foil.

The experimental arrangement is shown in Figure 8.9. A source of alpha particles (radioactive material, example polonium) is

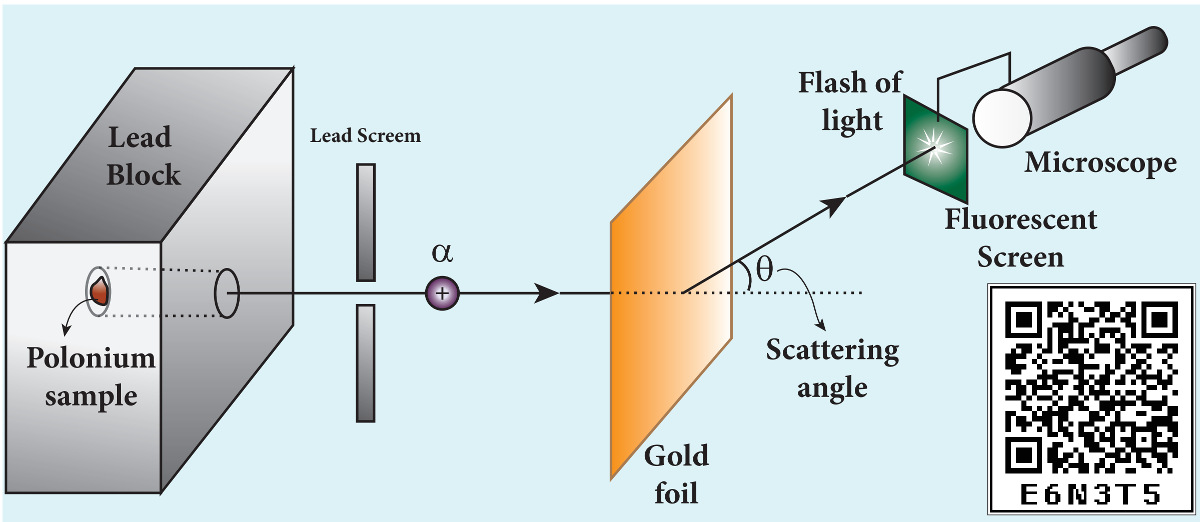


Figure 8.9 Schematic diagram for scattering of alpha particles experiment by Rutherford

kept inside a thick lead box with a fine hole as seen in Figure 8.9. The alpha particles coming through the fine hole of lead box pass through another fine hole made on the lead screen. These particles are now allowed to fall on a thin gold foil and it is observed that the alpha particles passing through gold foil are scattered through different angles. A movable screen (from 0° to 180°) which is made up of zinc sulphide (ZnS) is kept on the other side of the gold foil to collect the alpha particles. Whenever alpha particles strike the screen, a flash of light is observed which can be seen through a microscope.

Rutherford proposed an atom model based on the results of alpha scattering experiment. In this experiment, alpha particles (positively charged particles) are allowed to fall on the atoms of a metallic gold foil. The results of this experiment are given below and are shown in Figure 8.10, Rutherford expected the nuclear model to be as seen in Figure 8.10 (a) but the experiment showed the model as in Figure 8.10 (b).

(a) Most of the alpha particles are un-deflected through the gold foil and went straight.
(b) Some of the alpha particles are deflected through a small angle.
(c) A few alpha particles (one in thousand) are deflected through the angle more than 90°

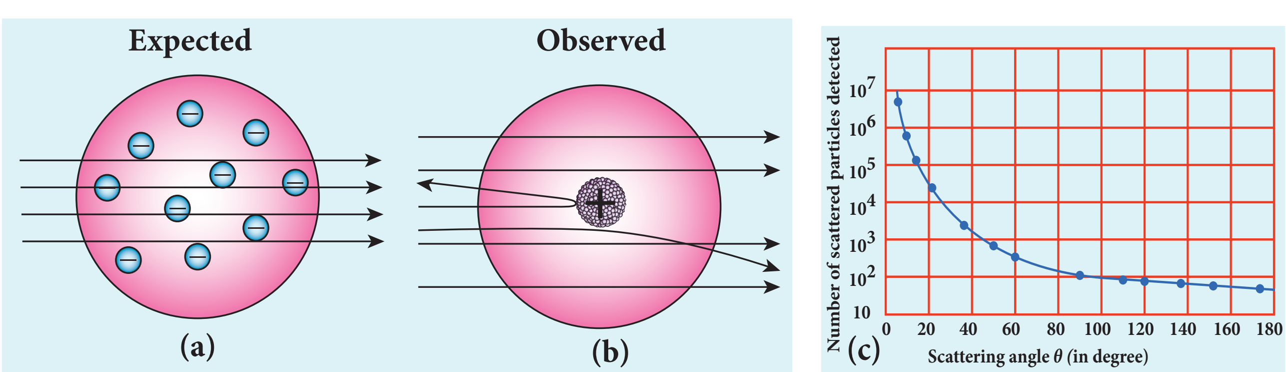


Figure 8.10 In alpha scattering experiment – (a) Rutherford expected (b) experiment result (c) The variation of alpha particles scattered N(θ) with scattering angle θ

(d) Very few alpha particles returned back (back scattered) –that is, deflected back by 180°

In Figure 8.10 (c), the dotted points are the alpha scattering experiment data points obtained by Geiger and Marsden and the solid curve is the prediction from Rutherford's nuclear model. It is observed that the Rutherford's nuclear model is in good agreement with the experimental data.

Conclusion made by Rutherford based on the above observation

From the experimental observations, Rutherford proposed that an atom has a lot of empty space and contains a tiny matter known as nucleus whose size is of the order of 10^{-14}m. The nucleus is positively charged and most of the mass of the atom is concentrated in nucleus. The nucleus is surrounded by negatively charged electrons. Since static charge distribution cannot be in a stable equilibrium, he suggested that the electrons are not at rest and they revolve around the nucleus in circular orbits like planets revolving around the sun.

(a) Distance of closest approach

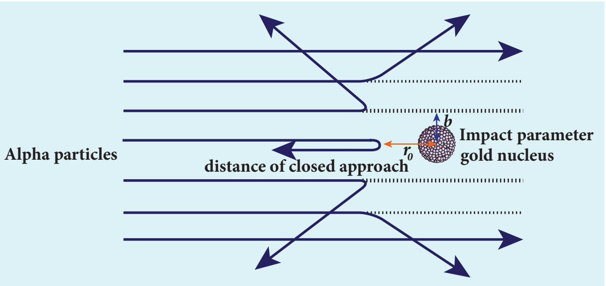


Figure 8.11 Distance of closest approach and impact parameter

When an alpha particle moves straight towards the nucleus, it reaches a point where it comes to rest momentarily and returns back as shown in Figure 8.11. **The minimum distance between the centre of the nucleus and the alpha particle just before it gets reflected back through 180° is defined as the distance of closest approach r_0 (also known as contact distance).** At this distance, all the kinetic energy of the alpha particle will be converted into electrostatic potential energy (Refer unit 1, volume 1 of +2 physics text book).

$$\frac{1}{2}m\upsilon_0^2 = \frac{1}{4\pi\varepsilon_0}\frac{(2e)(Ze)}{r_0}$$

$$\Rightarrow r_0 = \frac{1}{4\pi\varepsilon_0}\frac{2Ze^2}{\left(\frac{1}{2}m\upsilon_0^2\right)} = \frac{1}{4\pi\varepsilon_0}\frac{2Ze^2}{E_k}$$

where E_k is the kinetic energy of the alpha particle.This is used to estimate the size of the nucleus but size of the nucleus is always lesser than the distance of closest approach. Further, Rutherford calculated the radius of the nucleus for different nuclei and found that it ranges from 10^{-14} m to 10^{-15} m.

(b) Impact parameter

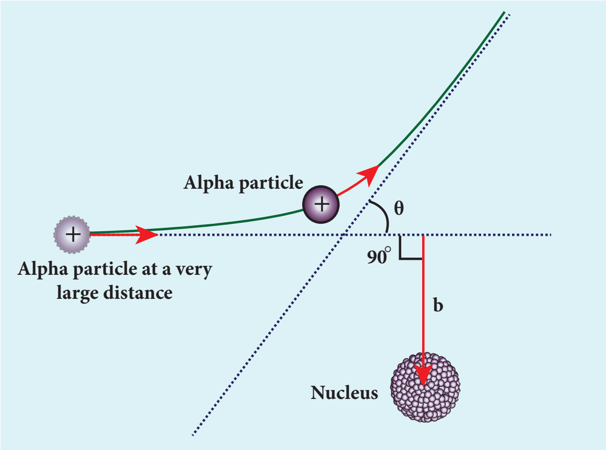


Figure 8.12 Impact parameter

The impact parameter (b) (see Figure 8.12) **is defined as the perpendicular distance between the centre of the gold nucleus and the direction of velocity vector of alpha particle when it is at a large distance.** The relation between impact parameter and scattering angle can be shown as

$$b \propto \cot\left(\frac{\theta}{2}\right) \Rightarrow b = K \cot\left(\frac{\theta}{2}\right) \qquad (8.13)$$

where $K = \frac{1}{4\pi\varepsilon_0}\frac{2Ze^2}{m\upsilon_0^2}$ and θ is called scattering angle. Equation (8.13) implies that when impact parameter increases, the scattering angle decreases. Smaller the impact parameter, larger will be the deflection of alpha particles.

Drawbacks of Rutherford model

Rutherford atom model helps in the calculation of the diameter of the nucleus and also the size of the atom but has the following limitations:

(a) This model fails to explain the distribution of electrons around the nucleus and also the stability of the atom.

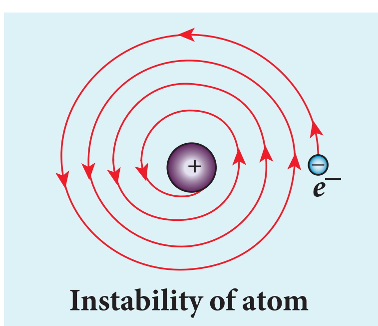


Figure 8.13 Spiral in motion of an electron around the nucleus

According to classical electrodynamics, any accelerated charge emits electromagnetic radiations. Due to emission of radiations, it loses its energy. Hence, it can no longer sustain the circular motion. The radius of the orbit, therefore, becomes smaller and smaller (undergoes spiral motion) as shown in Figure 8.13 and finally the electron should fall into the nucleus and the atoms should disintegrate. But this does not happen. Hence, Rutherford model could not account for the stability of atoms.

(b) According to this model, emission of radiation must be continuous and must give continuous emission spectrum but experimentally we observe only line (discrete) emission spectrum for atoms.

8.3.3 Bohr atom model

In order to overcome the limitations of the Rutherford atom model in explaining the stability and also the line spectrum observed for a hydrogen atom (Figure 8.14), Niels Bohr made modifications of Rutherford atom model. He is the first person to give better theoretical model of the structure of an atom to explain the line spectrum of hydrogen atom. The following are the assumptions (postulates) made by Bohr.

Figure 8.14 The line spectrum of hydrogen

Postulates of Bohr atom model:

(a) The electron in an atom moves around nucleus in circular orbits under the influence of Coulomb electrostatic force of attraction. This Coulomb force gives necessary centripetal force for the electron to undergo circular motion.

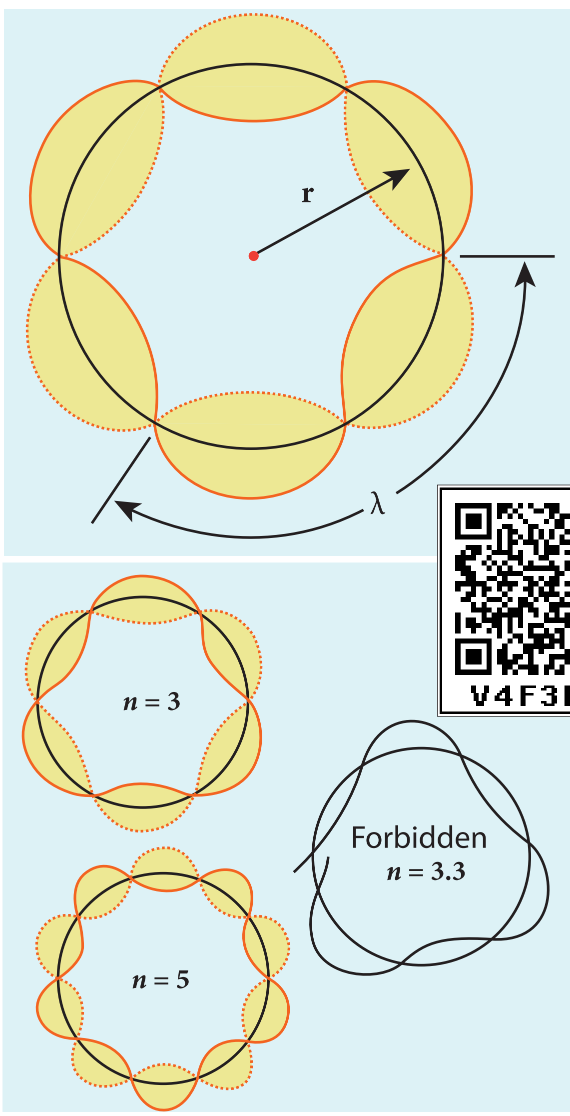
(b) Electrons in an atom revolve around the nucleus only in certain discrete orbits called stationary orbits where it does not radiate electromagnetic energy. Only those discrete orbits allowed are stable orbits.

The angular momentum of the electron in these stationary orbits are quantized – that is, it can be written as integer or integral multiple of $\frac{h}{2\pi}$ called as reduced Planck's constant – that is, $\hbar$ (read it as h-bar) and the integer n is called as principal quantum number of the orbit.

$$l = n\hbar \qquad \text{where } \hbar = \frac{h}{2\pi}$$

This condition is known as angular momentum quantization condition.

According to quantum mechanics, particles like electrons have dual nature (Refer unit 7, volume 2 of +2 physics text book). The standing wave pattern of the de Broglie wave associated with orbiting electron in a stable orbit is shown in Figure 8.15.



V4F3E7

Figure 8.15 Standing wave pattern for electron in a stable orbit

The circumference of an electron's orbit of radius r must be an integral multiple of de Broglie wavelength – that is,

$$2\pi r = n\lambda \qquad (8.14)$$

where n = 1,2,3,......

But the de Broglie wavelength (λ) for an electron of mass m moving with velocity v is $\lambda = \frac{h}{m\upsilon}$ where h is called Planck's constant. Thus from equation (8.14),

$$2\pi r = n\left(\frac{h}{m\upsilon}\right)$$

$$m\upsilon r = n\frac{h}{2\pi}$$

For any particle of mass m undergoing circular motion with radius r and velocity v, the magnitude of angular momentum l is given by

$$l = r(m\upsilon)$$

$$m\upsilon r = l = n\hbar$$

(c) Energy of orbits are not continuous but discrete. This is called the quantization of energy. An electron can jump from one orbit to another orbit by absorbing or emitting a photon whose energy is equal to the difference in energy (ΔE) between the two orbital levels (Figure 8.16)

$$\Delta E = E_{final} - E_{initial} = h\nu = h\frac{c}{\lambda}$$

where c is the speed of light and λ is the wavelength of the radiation used and ν is the frequency of the radiation. Thus, the frequency of the radiation emitted is related only to change in atom's energy and it does not depend on frequency of electron's orbital motion.

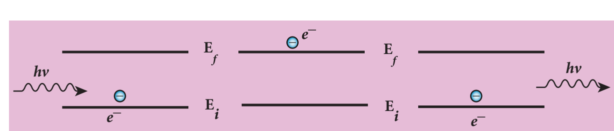


Figure 8.16 Absorption and emission of radiation

EXAMPLE 8.1

The radius of the 5th orbit of hydrogen atom is 13.25 Å. Calculate the wavelength of the electron in the 5th orbit.

Solution:

$2\pi r = n\lambda$

$2 \times 3.14 \times 13.25\text{Å} = 5 \times \lambda$

$\therefore \lambda = 16.64\text{Å}$

EXAMPLE 8.2

Find the (i) angular momentum (ii) velocity of the electron in the 5th orbit of hydrogen atom.

($h = 6.6 \times 10^{-34}$ Js, $m = 9.1 \times 10^{-31}$ kg)

Solution

(i) Angular momentum is given by

$$l = n\hbar = \frac{nh}{2\pi}$$

$$= \frac{5 \times 6.6 \times 10^{-34}}{2 \times 3.14} = 5.25 \times 10^{-34} \text{ kgm}^2\text{s}^{-1}$$

(ii) Velocity is given by

$$\text{Velocity } \upsilon = \frac{l}{mr}$$

$$= \frac{(5.25 \times 10^{-34} \ \cancel{\text{kg}}\text{m}^{\cancel{2}}\text{s}^{-1})}{(9.1 \times 10^{-31} \ \cancel{\text{kg}})(13.25 \times 10^{-10} \ \cancel{\text{m}})}$$

$$\upsilon = 4.4 \times 10^5 \text{ms}^{-1}$$

Radius of the orbit of the electron and velocity of the electron

Consider an atom which contains the nucleus at rest and an electron revolving around the nucleus in a circular orbit of radius r_n as shown in Figure 8.17. Nucleus is made up of protons and neutrons. Since proton is positively charged and neutron is electrically neutral, the charge of a nucleus is purely the total charge of protons.

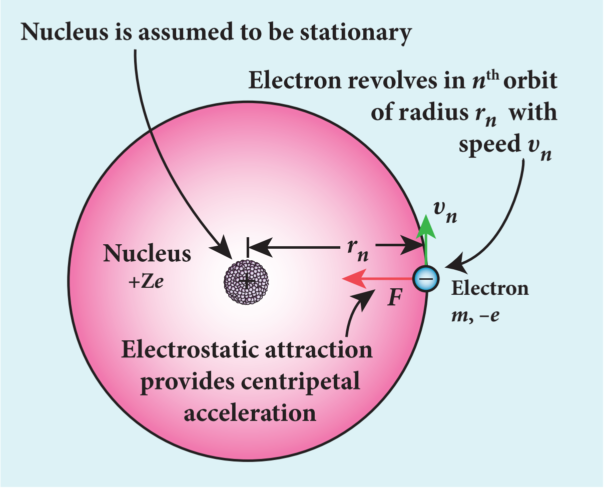


Figure 8.17 Electron revolving around the nucleus

Let Z be the atomic number of the atom, then $+Ze$ is the charge of the nucleus. Let $-e$ be the charge of the electron. From Coulomb's law, the force of attraction between the nucleus and the electron is

$$\vec{F}_{coloumb} = \frac{1}{4\pi\varepsilon_0}\frac{(+Ze)(-e)}{r_n^2}\hat{r}$$

$$= -\frac{1}{4\pi\varepsilon_0}\frac{Ze^2}{r_n^2}\hat{r}$$

This force provides necessary centripetal force

$$\vec{F}_{centripetal} = \frac{m\upsilon_n^2}{r_n}\hat{r}$$

where m be the mass of the electron that moves with a velocity υ_n in a circular orbit. Therefore,

$$\left|\vec{F}_{coloumb}\right| = \left|\vec{F}_{centripetal}\right|$$

$$\frac{1}{4\pi\varepsilon_0}\frac{Ze^2}{r_n^2} = \frac{m\upsilon_n^2}{r_n}$$

$$r_n = \frac{4\pi\varepsilon_0 (m\upsilon_n r_n)^2}{Zme^2} \quad (8.15)$$

From Bohr's assumption, the angular momentum quantization condition, $m\upsilon_n r_n = l_n = n\hbar$,

$$\therefore r_n = \frac{4\pi\varepsilon_0 (m\upsilon_n r_n)^2}{Zme^2}$$

$$r_n = \frac{4\pi\varepsilon_0 (n\hbar)^2}{Zme^2} = \frac{4\pi\varepsilon_0 n^2\hbar^2}{Zme^2}$$

$$r_n = \left(\frac{\varepsilon_0 h^2}{\pi me^2}\right)\frac{n^2}{Z} \qquad \left(\because \hbar = \frac{h}{2\pi}\right) \quad (8.16)$$

where $n \in \mathbb{N}$. Since, ε_0, h, e and π are constants. Therefore, the radius of the orbit becomes

$$r_n = a_0 \frac{n^2}{Z}$$

where $a_0 = \frac{\varepsilon_0 h^2}{\pi me^2} = 0.529\,\text{Å}$. This is known as Bohr radius which is the smallest radius of the orbit in an atom. Bohr radius is also used as unit of length called Bohr. 1 Bohr = 0.53 Å. For hydrogen atom (Z = 1), the radius of n^{th} orbit is

$$r_n = a_0 n^2$$

For n = 1 (first orbit or ground state),

$$r_1 = a_0 = 0.529\text{ Å}$$

For n = 2 (second orbit or first excited state),

$$r_2 = 4a_0 = 2.116\text{ Å}$$

For n = 3 (third orbit or second excited state),

$$r_3 = 9a_0 = 4.761\text{ Å}$$

and so on.

Thus the radius of the orbit from centre increases with n, that is, $r_n \propto n^2$ as shown in Figure 8.18.

Further, Bohr's angular momentum quantization condition leads to

$$m\upsilon_n r_n = m\upsilon_n a_0 n^2 = n\frac{h}{2\pi}$$

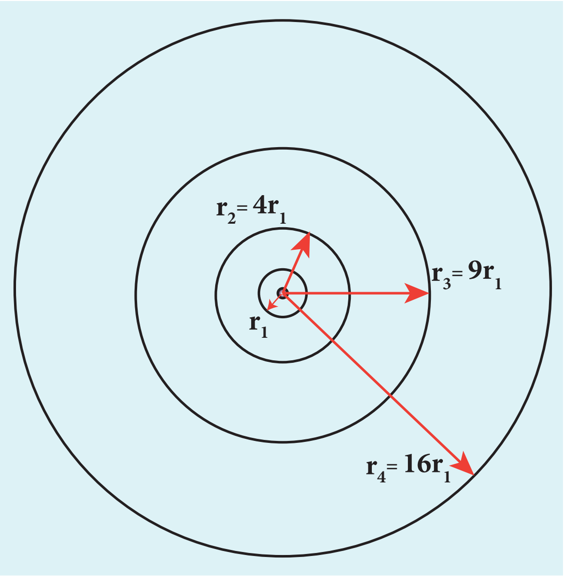


Figure 8.18 Variation of radius of the orbit with principal quantum number

$$\upsilon_n = \frac{h}{2\pi m a_0}\frac{Z}{n}$$

$$\upsilon_n \propto \frac{1}{n}$$

Note that the velocity of electron decreases as the principal quantum number increases as shown in Figure 8.19. This curve is the rectangular hyperbola. This implies that the velocity of electron in ground state is maximum when compared to excited states.

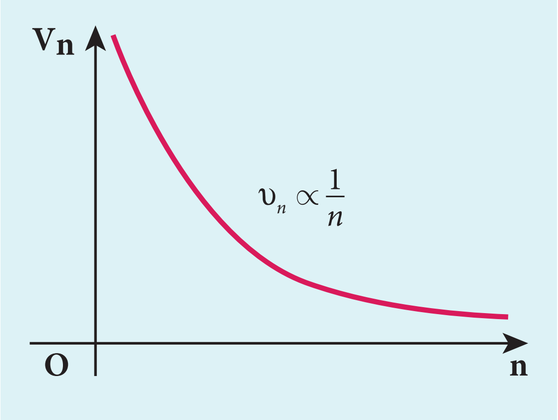


Figure 8.19 Variation of velocity of the electron in the orbit with principal quantum number

The energy of an electron in the n^{th} orbit

Since the electrostatic force is a conservative force, the potential energy for the n^{th} orbit is

$$U_n = \frac{1}{4\pi\varepsilon_0}\frac{(+Ze)(-e)}{r_n} = -\frac{1}{4\pi\varepsilon_0}\frac{Ze^2}{r_n}$$

$$= -\frac{1}{4\varepsilon_0^2}\frac{Z^2me^4}{h^2n^2} \quad \left(\because r_n = \frac{\varepsilon_0 h^2}{\pi m e^2}\frac{n^2}{Z}\right)$$

The kinetic energy for the n^{th} orbit is

$$KE_n = \frac{1}{2}m\upsilon_n^2 = \frac{me^4}{8\varepsilon_0^2h^2}\frac{Z^2}{n^2}$$

This implies that $U_n = -2\ KE_n$. Total energy in the n^{th} orbit is

$$E_n = KE_n + U_n = KE_n - 2KE_n = -KE_n$$

$$E_n = -\frac{me^4}{8\varepsilon_0^2h^2}\frac{Z^2}{n^2}$$

For hydrogen atom (Z = 1),

$$E_n = -\frac{me^4}{8\varepsilon_0^2h^2}\frac{1}{n^2}\ \text{joule} \qquad (8.17)$$

where n stands for principal quantum number. The negative sign in equation (8.17) indicates that the electron is bound to the nucleus.

Substituting the values of mass and charge of an electron (m and e), permittivity of free space ε_0 and Planck's constant h and expressing in terms of eV, we get

$$E_n = -13.6\frac{1}{n^2}\,eV$$

For the first orbit (ground state), the total energy of electron is $E_1 = -13.6\ eV$.

For the second orbit (first excited state), the total energy of electron is $E_2 = -3.4\ eV$.

For the third orbit (second excited state), the total energy of electron is $E_3 = -1.51\ eV$ and so on.

Notice that the energy of the first excited state is greater than the ground state, second excited state is greater than the first excited state and so on. Thus, the orbit which is closest to the nucleus (r_1) has lowest energy (minimum energy compared with other orbits). So, it is often called ground state energy (lowest energy state). The ground state energy of hydrogen ($-13.6\ eV$) is used as a unit of energy called Rydberg (1 Rydberg = $-13.6\ eV$).

The negative value of this energy is because of the way the zero of the potential energy is defined. When the electron is taken away to an infinite distance (very far distance) from nucleus, both the potential energy and kinetic energy terms vanish and hence the total energy also vanishes.

The energy level diagram along with the shape of the orbits for increasing values of n are shown in Figure 8.20. It shows that the energies of the excited states come closer and closer together when the principal quantum number n takes higher values.

EXAMPLE 8.3

(a) Show that the ratio of velocity of an electron in the first Bohr orbit to the speed of light c is a dimensionless number.

(b) Compute the velocity of electrons in ground state, first excited state and second excited state in Bohr atom model.

Solution

(a) The velocity of an electron in n^{th} orbit is

$$\upsilon_n = \frac{h}{2\pi m a_0}\frac{Z}{n}$$

where $a_0 = \frac{\varepsilon_0 h^2}{\pi m e^2}$ = Bohr radius. Substituting for a_0 in υ_n,

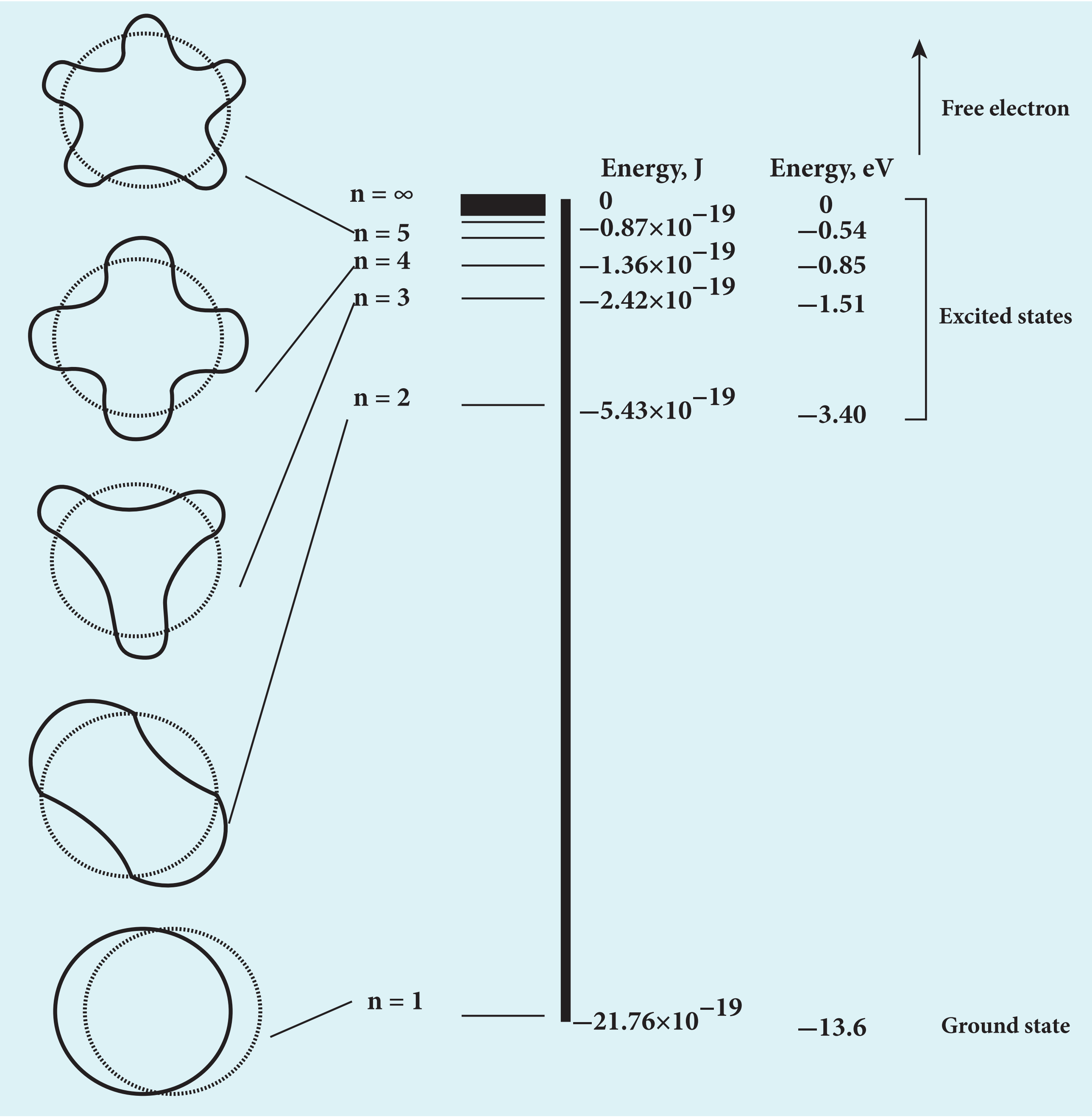


Figure 8.20 **Energy levels of a hydrogen atom**

$$\upsilon_n = \frac{e^2}{2\varepsilon_0 h}\frac{Z}{n} = c\left(\frac{e^2}{2\varepsilon_0 hc}\right)\frac{Z}{n} = \frac{\alpha c Z}{n}$$

where c is the speed of light in free space or vacuum and its value is $c = 3 \times 10^8$ m s^{-1}and α is called fine structure constant.

For a hydrogen atom, $Z = 1$ and for the first orbit, $n = 1$, the ratio of velocity of electron in first orbit to the speed of light in vacuum or free space is

$$\frac{\upsilon_1}{c} = \alpha = \frac{e^2}{2\varepsilon_0 hc}$$

$$\alpha = \frac{(1.6\times10^{-19}\,\mathrm{C})^2}{2\times(8.854\times10^{-12}\,\mathrm{C^2N^{-1}m^{-2}})} \frac{(1.6\times10^{-19}\,\mathrm{C})^2}{(6.6\times10^{-34}\,\mathrm{Nms})\times(3\times10^8\,\mathrm{ms^{-1}})}$$

$$\approx \frac{1}{136.9} = \frac{1}{137}$$ which is a dimensionless number

$$\Rightarrow \alpha = \frac{1}{137}$$

(b) Using fine structure constant, the velocity of electron can be written as

$$\upsilon_n = \frac{\alpha c Z}{n}$$

For hydrogen atom ($Z = 1$) the velocity of electron in n^{th} orbit is

$$\upsilon_n = \frac{c}{137}\frac{1}{n} = (2.19\times10^6)\frac{1}{n}\ \text{ms}^{-1}$$

For the first orbit (ground state), the velocity of electron is

$$\upsilon_1 = 2.19\times10^6\,\text{ms}^{-1}$$

For the second orbit (first excited state), the velocity of electron is

$$\upsilon_2 = 1.095\times10^6\,\text{ms}^{-1}$$

For the third orbit (second excited state), the velocity of electron is

$$\upsilon_3 = 0.73\times10^6\,\text{ms}^{-1}$$

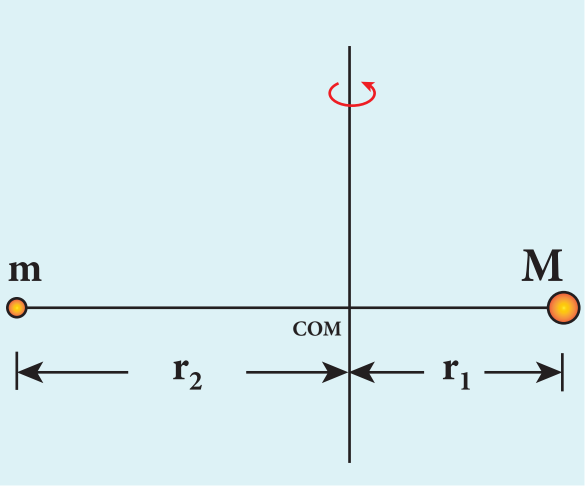
Here, $\upsilon_1 > \upsilon_2 > \upsilon_3$

EXAMPLE 8.4

The Bohr atom model is derived with the assumption that the nucleus of the atom is stationary and only electrons revolve around the nucleus. Suppose the nucleus is also in motion, then calculate the energy of this new system.

Solution

Let the mass of the electron be m and mass of the nucleus be M. Since there is no external force acting on the system, the centre of mass of hydrogen atom remains at rest. Hence, both nucleus and electron move about the centre of mass as shown in figure.



Let V be the velocity of the nuclear motion and υ be the velocity of electron motion. Since the total linear momentum of the system is zero,

$$-m\upsilon + M\upsilon = 0 \text{ or}$$

$$MV = m\upsilon = p$$

$$\vec{p}_e + \vec{p}_n = \vec{0} \text{ or}$$

$$|\vec{p}_e| = |\vec{p}_n| = p$$

Hence, the kinetic energy of the system is

$$KE = \frac{p_n^2}{2M} + \frac{p_e^2}{2m} = \frac{p^2}{2}\left(\frac{1}{M} + \frac{1}{m}\right)$$

Let $\frac{1}{M} + \frac{1}{m} = \frac{1}{\mu_m}$. Here the reduced mass is, $\mu_m = \frac{mM}{M+m}$

Therefore, the kinetic energy of the system now is $KE = \frac{p^2}{2\mu_m}$

Since the potential energy of the system is same, the total energy of the hydrogen can be expressed by replacing mass by reduced mass, which is

$$E_n = -\frac{\mu_m e^4}{8\varepsilon_0^2 h^2}\frac{1}{n^2}$$

Since the nucleus is very heavy compared to the electron, the reduced mass is closer to the mass of the electron.

Note In 1931, H.C. Urey and co-workers noticed that in the shorter wavelength region of the hydrogen spectrum lines, faint companion lines are observed. From the isotope displacement effect (isotope shift), the isotope of the same element will have slightly different spectral lines. The presence of these faint lines confirmed the existence of isotopes of hydrogen atom (which is named as Deuterium).

On calculating wavelength or wave number difference between the faint and bright spectral lines, atomic mass of deuterium is measured to be twice that of atomic mass of hydrogen atom. Bohr atom model could not explain this isotopic shift. Thus by considering nuclear motion (although the movement of the nucleus is much smaller, it is observed) into account in the Bohr atom model, the wave number or wavelength difference between hydrogen atom and deuterium is theoretically calculated which perfectly agreed with the spectroscopic measured values.

The difference between hydrogen atom and deuterium is in the number of neutron. Hydrogen atom contains an electron and a proton, whereas deuterium has an electron, a proton and a neutron.

Excitation energy and excitation potential

The energy required to excite an electron from lower energy state to any higher energy state is known as excitation energy.

The excitation energy for an electron from ground state (n = 1) to first excited state (n = 2) is called first excitation energy, which is

$$E_I = E_2 - E_1 = -3.4\ eV - (-13.6\ eV) = 10.2\ eV$$

Similarly, the excitation energy for an electron from ground state (n = 1) to second excited state (n = 3) is called second excitation energy, which is

$$E_{II} = E_3 - E_1 = -1.51\ eV - (-13.6\ eV) = 12.1\ eV$$

and so on.

Excitation potential is defined as excitation energy per unit charge.

First excitation potential is,

$$E_{\mathrm{I}} = eV_{\mathrm{I}} \Rightarrow V_{\mathrm{I}} = \frac{1}{e}E_{\mathrm{I}} = 10.2 \text{ volt}$$

Second excitation potential is,

$$E_{\mathrm{II}} = eV_{\mathrm{II}} \Rightarrow V_{\mathrm{II}} = \frac{1}{e}E_{\mathrm{II}} = 12.1 \text{ volt}$$

and so on.

Ionization energy and ionization potential

An atom is said to be ionized when an electron is completely removed from the atom – that is, it reaches the state with energy $E_{n\to\infty}$. **The minimum energy required to remove an electron from an atom in the ground state is known as binding energy or ionization energy**.

$$E_{\text{ionizaation}} = E_\infty - E_1 = 0 - (-13.6\,eV)$$
$$= 13.6\,eV$$

When an electron is in n^{th} state of an atom, the energy spent to remove an electron from that state – that is, its ionization energy is

$$E_{\text{ionization}} = E_\infty - E_n = 0 - \left(-\frac{13.6}{n^2}Z^2\ \mathrm{eV}\right)$$
$$= \frac{13.6}{n^2}Z^2\ \mathrm{eV}$$

At normal room temperature, the electron in a hydrogen atom (Z=1) spends most of its time in the ground state. **The amount of energy spent to remove an electron from the ground state of an atom**

Table 8.1

Physical quantity	Ground state	First excited state	Second excited state
Radius ($r_n \propto n^2$)	0.529 Å	2.116 Å	4.761 Å
Velocity ($v_n \propto n^{-1}$)	2.19×10^6 m s^{-1}	1.095×10^6 m s^{-1}	0.73×10^6 m s^{-1}
Total Energy ($E_n \propto n^{-2}$)	-13.6 *eV*	-3.4 *eV*	-1.51 *eV*

***(E = 0* for *n→∞)* is known as first ionization energy (13.6 *eV*).** Then, the hydrogen atom is said to be in ionized state or simply called as hydrogen ion, denoted by H^+. If we supply more energy than the ionization energy, the excess energy will be the kinetic energy of the free electron.

Ionization potential is defined as ionization energy per unit charge.

$$V_{\text{ionization}} = \frac{1}{e} E_{\text{ionization}} = \frac{13.6}{n^2} Z^2 \, V$$

Thus, for a hydrogen atom (Z=1), the ionization potential is

$$V = \frac{13.6}{n^2} \text{ volt}$$

The radius, velocity and total energy in ground state, first excited state and second excited state is listed in Table 8.1.

EXAMPLE 8.5

Suppose the energy of a hydrogen–like atom is given as $E_n = -\frac{54.4}{n^2}$ *eV* where $n \in \mathbb{N}$. Calculate the following:

(a) Sketch the energy levels for this atom and compute its atomic number.

(b) If the atom is in ground state, compute its first excitation potential and also its ionization potential.

(c) When a photon with energy 42 *eV* and another photon with energy 56 *eV* are made to collide with this atom, does this atom absorb these photons?

(d) Determine the radius of its first Bohr orbit.

(e) Calculate the kinetic and potential energies in the ground state.

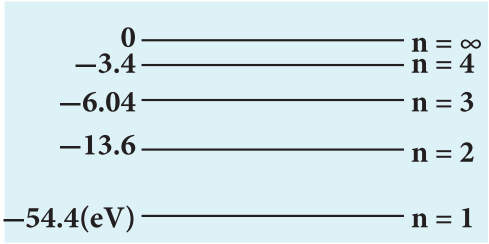
Solutions

(a) Given that

$$E_n = -\frac{54.4}{n^2} \text{ eV}$$

For $n = 1$, the ground state energy $E_1 = -54.4$ eV and for $n = 2$, $E_2 = -13.6$ *eV*. Similarly, $E_3 = -6.04$ *eV*, $E_4 = -3.4$ *eV* and so on.

For large value of principal quantum number – that is, $n = \infty$, we get $E_\infty = 0$ eV.



(b) For a hydrogen-like atom, ground state energy is

$$E_1 = -\frac{13.6}{n^2} Z^2 \text{ eV}$$

where Z is the atomic number. Hence, comparing this energy with given energy, we get, $-13.6\,Z^2 = -54.4 \Rightarrow Z = \pm 2$. Since, atomic number cannot be negative number, $Z = 2$.

(c) The first excitation energy is

$$E_I = E_2 - E_1 = -13.6\,eV - (-54.4\,eV)$$
$$= 40.8\,eV$$

Hence, the first excitation potential is

$$V_I = \frac{1}{e}E_I = \frac{(40.8\,eV)}{e}$$
$$= 40.8\text{ volt}$$

The first ionization energy is

$$E_{\text{ionization}} = E_\infty - E_1 = 0 - (-54.4\,eV)$$
$$= 54.4\,eV$$

Hence, the first ionization potential is

$$V_{\text{ionization}} = \frac{1}{e}E_{\text{ionization}} = \frac{(54.4\,eV)}{e}$$
$$= 54.4\text{ volt}$$

(d) Consider two photons to be A and B.

Given that photon A with energy 42 eV and photon B with energy 51 eV

From Bohr assumption, difference in energy levels is equal to photon energy, then atom will absorb energy, otherwise, not.

$$E_2 - E_1 = -13.6\,eV - (-54.4\,eV)$$
$$= 40.8\,eV \approx 41\,eV$$

Similarly,

$$E_3 - E_1 = -6.04\,eV - (-54.4\,eV)$$
$$= 48.36\,eV$$

$$E_4 - E_1 = -3.4\,eV - (-54.4\,eV)$$
$$= 51\,eV$$

$$E_3 - E_2 = -6.04\,eV - (-13.6\,eV)$$
$$= 7.56\,eV$$

and so on.

But note that $E_2 - E_1 \neq 42$ eV, $E_3 - E_1 \neq 42$ eV, $E_4 - E_1 \neq 42$ eV and $E_3 - E_2 \neq 42$ eV.

For all possibilities, no difference in energy is an integer multiple of photon energy. Hence, photon A is not absorbed by this atom. But for Photon B, $E_4 - E_1 = 51$ eV, which means, Photon B can be absorbed by this atom.

(e) Since total energy is equal to negative of kinetic energy in Bohr atom model, we get

$$KE_n = -E_n = -\left(-\frac{54.4}{n^2}eV\right)$$
$$= \frac{54.4}{n^2}eV$$

Potential energy is negative of twice the kinetic energy, which means,

$$U_n = -2KE_n = -2\left(\frac{54.4}{n^2}eV\right)$$
$$= -\frac{108.8}{n^2}eV$$

For a ground state, put n =1

Kinetic energy is KE_1 = 54.4 eV and Potential energy is $U_1 = -108.8$ eV

8.3.4 Atomic spectra

Materials in the solid, liquid and gaseous states emit electromagnetic radiations when they are heated up and these emitted radiations usually belong to continuous spectrum. For example, when white light is examined through a spectrometer, electromagnetic radiations of all wavelengths are observed which is a continuous spectrum.

In early twentieth century, many scientists spent considerable time in understanding the characteristic radiations emitted by the atoms of individual elements exposed to a flame or electrical discharge. When they are viewed

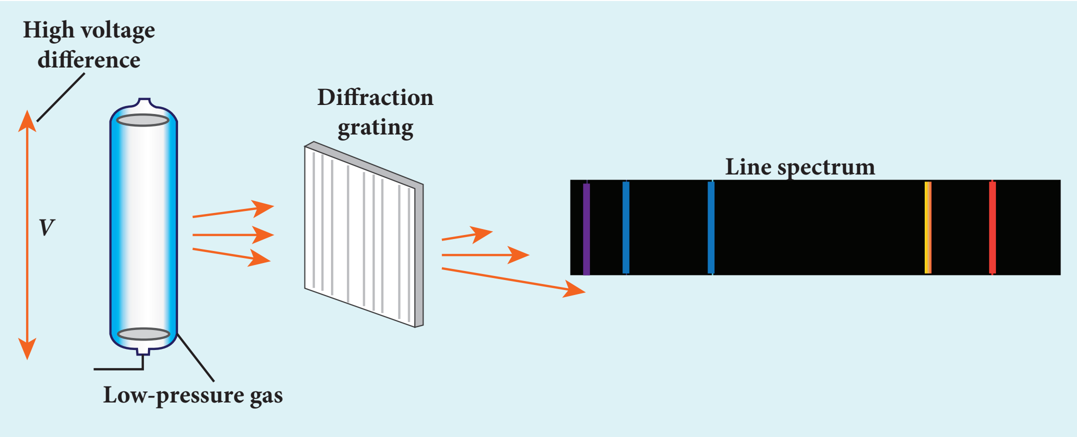


Figure 8.21 Spectrum of an atom

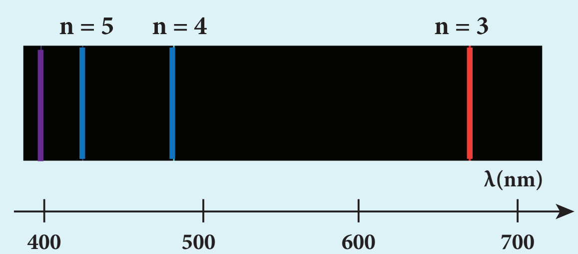
or photographed, instead of a continuous spectrum, the radiation contains of a set of discrete lines, each with characteristic wavelength. In other words, the wavelengths of the light obtained are well defined and the positions and intensities are characteristic of the element as shown in Figure 8.21.

This implies that these spectra are unique to each element and can be used to identify the element of the gas (like finger print used to identify a person) – that is, it varies from one gas to another gas. This uniqueness of line spectra of elements made the scientists to determine the composition of stars, sun and also used to identify the unknown compounds.

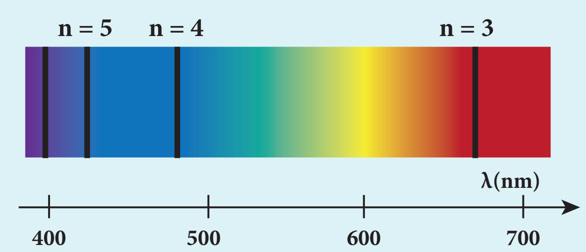
Hydrogen spectrum

When the hydrogen gas enclosed in a tube is heated up, it emits electromagnetic radiations of certain sharply-defined characteristic wavelength (line spectrum), called hydrogen emission spectrum (Refer unit 5, volume 1 of +2 physics text book). The emission spectra of hydrogen are shown in Figure 8.22(a).

When any gas is heated up, the thermal energy is supplied to excite the electrons. Similarly by passing light on the atoms, electrons can be excited by absorbing photons. Once the electrons get sufficient energy as given by Bohr's postulate (c), it absorbs energy with particular wavelength (or frequency) and jumps from its stationary state (original state) to higher energy state. Those wavelengths (or frequencies) for which the colours are not observed are seen as dark lines in the absorption spectrum as shown in Figure 8.22 (b).



(a) Emission spectrum of hydrogen (in wavelength)



(b) Absorption spectrum of hydrogen (in wavelength)

Figure 8.22 Hydrogen spectrum (a) emission (b) absorption

Since electrons in excited states have very small life time, these electrons jump back to ground state through spontaneous emission in a short duration of time (approximately 10^{-8} s) by emitting the radiation with same wavelength (or frequency) corresponding to the colours it absorbed (Figure 8.22 (a)). This is called emission spectroscopy.

The wavelengths of these lines can be calculated with great precision. Further, the emitted radiation contains wavelengths both lesser and greater than the visible spectrum.

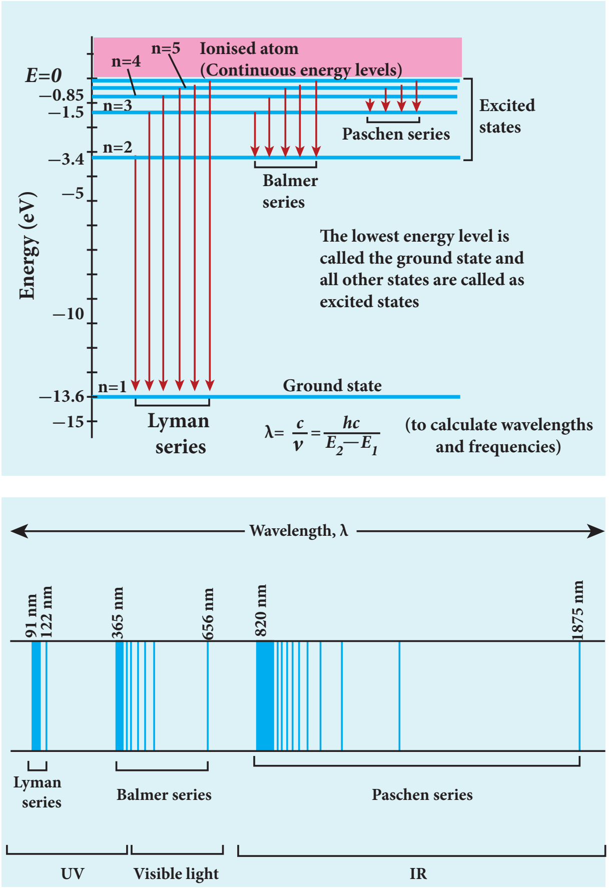


Figure 8.23 Spectral series – Lyman, Balmer, Paschen series

Notice that the spectral lines of hydrogen as shown in Figure 8.23 are grouped in separate series. In each series, the distance of separation between the consecutive wavelengths decreases from higher wavelength to the lower wavelength, and also wavelength in each series approach a limiting value known as the series limit. These series are named as Lyman series, Balmer series, Paschen series, Brackett series, Pfund series, etc. The wavelengths of these spectral lines perfectly agree with the equation derived from Bohr atom model.

$$\frac{1}{\lambda} = R\left(\frac{1}{n^2} - \frac{1}{m^2}\right) = \bar{\nu} \qquad (8.18)$$

where $\bar{\nu}$ **is known as wave number which is inverse of wavelength,** R is known as Rydberg constant whose value is 1.09737 x 10^7 m^{-1} and m and n are positive integers such that $m > n$. The various spectral series are discussed below:

(a) Lyman series

Put n = 1 and m = 2,3,4....... in equation (8.18). The wave number or wavelength of spectral lines of Lyman series which lies in ultra-violet region is

$$\bar{\nu} = \frac{1}{\lambda} = R\left(\frac{1}{1^2} - \frac{1}{m^2}\right)$$

(b) Balmer series

Put n = 2 and m = 3,4,5....... in equation (8.18). The wave number or wavelength of spectral lines of Balmer series which lies in visible region is

$$\bar{\nu} = \frac{1}{\lambda} = R\left(\frac{1}{2^2} - \frac{1}{m^2}\right)$$

(c) Paschen series

Put n = 3 and m = 4,5,6....... in equation (8.18). The wave number or wavelength of spectral lines of Paschen series which lies in infra-red region (near IR) is

$$\bar{\nu} = \frac{1}{\lambda} = R\left(\frac{1}{3^2} - \frac{1}{m^2}\right)$$

(d) Brackett series

Put n = 4 and m = 5,6,7........ in equation (8.18). The wave number or wavelength of spectral lines of Brackett series which lies in infra-red region (middle IR) is

$$\bar{v} = \frac{1}{\lambda} = R\left(\frac{1}{4^2} - \frac{1}{m^2}\right)$$

(e) Pfund series

Put n = 5 and m = 6,7,8........ in equation (8.18). The wave number or wavelength of spectral lines of Pfund series which lies in infra-red region (far IR) is

$$\bar{v} = \frac{1}{\lambda} = R\left(\frac{1}{5^2} - \frac{1}{m^2}\right)$$

Different spectral series are listed in Table 8.2.

Table 8.2

n	*m*	Series Name	Region
1	2,3,4.....	Lyman	Ultraviolet
2	3,4,5.....	Balmer	Visible
3	4,5,6.....	Paschen	Infrared
4	5,6,7......	Brackett	Infrared
5	6,7,8.....	Pfund	Infrared

Limitations of Bohr atom model

The following are the drawbacks of Bohr atom model

(a) Bohr atom model is valid only for hydrogen atom or hydrogen like-atoms but not for complex atoms.

(b) When the spectral lines are closely examined, individual lines of hydrogen spectrum is accompanied by a number of faint lines. These are often called **fine structure**. This is not explained by Bohr atom model.

(c) Bohr atom model fails to explain the intensity variations in the spectral lines.

(d) The distribution of electrons in atoms is not completely explained by Bohr atom model.

8.4 NUCLEI

Introduction

In the previous section, we have discussed various preliminary atom models, Rutherford's alpha particle scattering experiment and Bohr atom model. These played a vital role to understand the structure of the atom and the nucleus. In this section, the structure and the properties of the nucleus, and its classifications are discussed.

8.4.1 Composition of nucleus

Atoms have a nucleus surrounded by electrons. The nucleus contains protons and neutrons. The neutrons are electrically neutral ($q = 0$) and the protons have positive charge ($q = +e$) equal in magnitude of the charge of the electron ($q = -e$). **The number of protons in the nucleus is called the atomic number** and it is denoted by *Z***.** The number of neutrons in the nucleus is called neutron number **(*N*). The total number of neutrons and protons in the nucleus is called the mass number** and it is denoted by *A*. Hence, $A = Z + N$.

The two constituents of nucleus namely neutrons and protons, are collectively called nucleons. The mass of a proton is 1.6726×10^{-27} kg which is roughly 1836 times the mass of the electron. The mass of a neutron is slightly greater than the mass of the proton and it is equal to 1.6749×10^{-27} kg.

To specify the nucleus of any element, we use the following general notation

$$_Z^AX$$

where X is the chemical symbol of the element, A is the mass number and Z is the atomic number. For example, the nitrogen nucleus is represented by ${}^{15}_{7}N$. It implies that nitrogen contains 15 nucleons of which 7 are protons ($Z=7$) and 8 are neutrons ($N=A-Z=8$). Note that once the element is specified, the value of Z is known and subscript Z is sometimes omitted. For example, nitrogen nucleus is simply denoted as ${}^{15}N$ and we call it as 'nitrogen fifteen'.

Since the nucleus is made up of positively charged protons and electrically neutral neutrons, the overall charge of the nucleus is positive and it has the value of $+Ze$. But the atom is electrically neutral which implies that the number of electrons in the atom is equal to the number of protons in the nucleus.

8.4.2 Isotopes, isobars, and isotones

Isotopes:

In nature, there are atoms of a particular element whose nuclei have same number of protons but different number of neutrons. These kinds of atoms are called isotopes. In other words, **isotopes are atoms of the same element having same atomic number *Z*, but different mass number *A*.** For example, hydrogen has three isotopes and they are represented as ${}^{1}_{1}H$ (hydrogen), ${}^{2}_{1}H$ (deuterium),and ${}^{3}_{1}H$ (tritium). Note that all the three nuclei have one proton and, hydrogen has no neutron, deuterium has 1 neutron and tritium has 2 neutrons.

The number of isotopes for the particular element and their relative abundances (percentage) vary with each element. For example, carbon has four main isotopes: ${}^{11}_{6}C$, ${}^{12}_{6}C$, ${}^{13}_{6}C$ and ${}^{14}_{6}C$. But in nature, the percentage of ${}^{12}_{6}C$ is approximately 98.9%, that of ${}^{13}_{6}C$ is 1.1% and that of ${}^{14}_{6}C$ is 0.0001%. The other carbon isotope ${}^{11}_{6}C$, do not occur naturally and it can be produced only in nuclear reactions in the laboratory or by cosmic rays.

The chemical properties of any atom are determined only by electrons, the isotopes of any element have same electronic structure and same chemical properties. So the isotopes of the same element are placed in the same location in the periodic table.

Isobars:

Isobars are the atoms of different elements having the same mass number *A*, but different atomic number *Z*. In other words, isobars are the atoms of different chemical element which has same number of nucleon. For example ${}^{40}_{16}S$, ${}^{40}_{17}Cl$, ${}^{40}_{18}Ar$, ${}^{40}_{19}K$ and ${}^{40}_{20}Ca$ are isobars having same mass number 40 and different atomic number. Unlike isotopes, isobars are chemically different elements. They have different physical and chemical properties.

Isotones:

Isotones are the atoms of different elements having same number of neutrons. ${}^{12}_{5}B$ and ${}^{13}_{6}C$ are examples of isotones which 7 neutrons.

8.4.3 Atomic and nuclear masses

The mass of nuclei is very small when it expressed in SI units (about 10^{-25} kg or less). Therefore, it is more convenient to express it in terms of another unit namely, the *atomic mass unit (u)*. **One atomic mass unit (*u*) is defined as the 1/12$^{\text{th}}$ of the mass of the isotope of carbon ${}^{12}_{6}C$,** the most abundant naturally occurring isotope of carbon.

In other words

$$1\,u = \frac{\text{mass of } {}^{12}_{6}C \text{ atom}}{12} = \frac{1.9926\times10^{-26}}{12}$$
$$= 1.660\times10^{-27}\,\text{kg}$$

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In terms of this atomic mass unit, the mass of the neutron = 1.008665 u, the mass of the proton =1.007276 u, the mass of the hydrogen atom = 1.007825 u and the mass of ${}^{12}_{6}C = 12u$. Note that usually mass specified is the mass of the atoms, not mass of the nucleus. To get the nuclear mass of particular nucleus, the mass of electrons has to be subtracted from the corresponding atomic mass. Experimentally the atomic mass is determined by the instrument called Bainbridge mass spectrometer. If we determine the atomic mass of the element without considering the effect of its isotopes, we get the mass averaged over different isotopes weighted by their abundances.

EXAMPLE 8.6

Calculate the average atomic mass of chlorine if no distinction is made between its different isotopes?

Solution

The element chlorine is a mixture of 75.77% of ${}^{35}_{17}Cl$ and 24.23% of ${}^{37}_{17}Cl$. So the average atomic mass will be

$$\frac{75.77}{100} \times 34.96885u + \frac{24.23}{100} \times 36.96593u$$
$$= 35.453u$$

In fact, the chemist uses the average atomic mass or simply called chemical atomic weight (35.453 u for chlorine) of an element. So it must be remembered that the atomic mass which is mentioned in the periodic table is basically averaged atomic mass.

8.4.4 Size and density of the nucleus

The alpha particle scattering experiment and many other measurements using different methods have been carried out on the nuclei of various atoms. The nuclei are found to be approximately spherical in shape. It is experimentally found that radius of nuclei for $Z > 10$, satisfies the following empirical formula

$$R = R_0 A^{\frac{1}{3}} \qquad (8.19)$$

Here A is the mass number of the nucleus and the constant R_0 = 1.2 F, where 1 F = 1×10^{-15} m. The unit fermi (F) is named after Enrico Fermi.

EXAMPLE 8.7

Calculate the radius of ${}^{197}_{79}Au$ nucleus.

Solution

According to the equation (8.19),

$$R = 1.2 \times 10^{-15} \times (197)^{\frac{1}{3}} = 6.97 \times 10^{-15} m$$

Or $\quad R = 6.97$ F

EXAMPLE 8.8

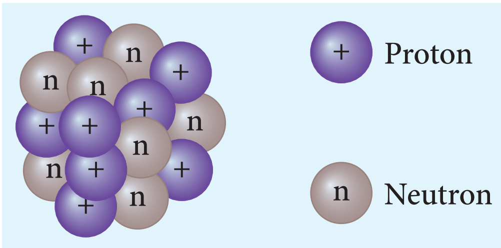
Calculate the density of the nucleus with mass number A.

Solution

From equation (8.19), the radius of the nuclei satisfy the equation $= R_0 A^{\frac{1}{3}}$. Then the volume of the nucleus

$$V = \frac{4}{3}\pi R^3 = \frac{4}{3}\pi R_0^3 A$$

By ignoring the mass difference between the proton and neutron, the total mass of the nucleus having mass number A is equal to $A.m$ where m is mass of the proton and is equal to 1.6726 x 10^{-27} kg.



Nuclear density

$$\rho = \frac{\text{mass of the nuclei}}{\text{Volume of the nuclei}} = \frac{A.m}{\frac{4}{3}\pi R_0^3 A} = \frac{m}{\frac{4}{3}\pi R_0^3}$$

The above expression shows that the nuclear density is independent of the mass number A. In other words, all the nuclei ($Z > 10$) have the same density and it is an important characteristics of the nuclei.

We can calculate the numerical value of this density by substituting the corresponding values.

$$\rho = \frac{1.67\times10^{-27}}{\frac{4}{3}\pi\times(1.2\times10^{-15})^3} = 2.3\times10^{17}\,\text{kg m}^{-3}.$$

It implies that nucleons are extremely tightly packed in the nucleus and compare this density with the density of water which is 10^3 kg m^{-3}.

DO YOU KNOW? A single teaspoon of nuclear matter would weigh about trillion tons.

8.4.5 Mass defect and binding energy

It is experimentally found out that the mass of any nucleus is always less than the sum of the mass of its individual constituents. For example, consider the carbon-12 nucleus which is made up of 6 protons and 6 neutrons.

Mass of 6 neutrons $= 6\times1.00866\ u = 6.05196\,u$

Mass of 6 protons $= 6\times1.00727\,u = 6.04362\,u$

Mass of 6 electrons $= 6\times0.00055\,u = 0.0033\,u$

The expected mass of carbon-12 nucleus $= 6.05196\,u + 6.04362\,u = 12.09558\,u$

But using mass spectroscopy, the atomic mass of carbon-12 atom is found to be 12 u. So if we subtract the mass of 6 electrons (0.0033 u) from 12 u, we get the carbon-12 nuclear mass which is equal to 11.9967 u. Note that the experimental mass of carbon-12 nucleus is less than the total mass of its individual constituents by $\Delta m = 0.09888\,u$. This difference in mass Δm is called mass defect. In general, if M, m_p, and m_n are mass of the nucleus (^A_ZX), the mass of a proton and the mass of a neutron respectively, then the mass defect is given by

$$\Delta m = \left(Zm_p + Nm_n\right) - M \qquad (8.20)$$

Where has this mass disappeared? The answer was provided by Albert Einstein with the help of famous mass-energy relation $(E = mc^2)$. According to this relation, the mass can be converted into energy and energy can be converted into mass. In the case of the carbon-12 nucleus, when 6 protons and 6 neutrons combine to form carbon-12 nucleus, mass equal to mass defect disappears and the corresponding energy is released. This is called the binding energy of the nucleus (BE) and is equal to $(\Delta m)c^2$. In fact, to separate the carbon-12 nucleus into individual constituents, we must supply the energy equal to binding energy of the nucleus.

We can write the equation (8.20) in terms of binding energy

$$BE = \left(Zm_p + Nm_n - M\right)c^2 \qquad (8.21)$$

It is always convenient to work with the mass of the atom than the mass of the nucleus. Hence by adding and subtracting the mass of the Z electrons, we get

$$BE = \left(Zm_p + Zm_e + Nm_n - M - Zm_e\right)c^2 \quad (8.22)$$

$$BE = \left[Z\left(m_p + m_e\right) + Nm_n - M - Zm_e\right]c^2$$

where $m_p + m_e = m_H$ (mass of hydrogen atom)

$$BE = \left[Zm_H + Nm_n - (M + Zm_e)\right]c^2 \quad (8.23)$$

Here $M + Zm_e = M_A$ where M_A is the mass of the atom of an element ${}^{A}_{Z}X$.

Finally, the binding energy in terms of the atomic masses is given by

$$BE = \left[Zm_H + Nm_n - M_A\right]c^2 \quad (8.24)$$

Note

Using Einstein's mass-energy equivalence, the energy equivalent of one atomic mass unit $1u = 1.66\times10^{-27}\times(3\times10^{8})^2 = 14.94\times10^{-11}J \approx 931MeV$

EXAMPLE 8.9

Compute the binding energy of ${}^{4}_{2}He$ nucleus using the following data: Atomic mass of Helium atom, $M_A(He) = 4.00260\,u$ and that of hydrogen atom, $m_H = 1.00785\,u$.

Solution:

Binding energy $BE = \left[Zm_H + Nm_n - M_A\right]c^2$

For helium nucleus, $Z = 2, N = A - Z = 4 - 2 = 2$

Mass defect

$$\Delta m = \left[(2\times1.00785u) + (2\times1.008665u) - 4.00260u\right] \quad \Delta m = 0.03038u$$

$$B.E = 0.03038u\times c^2$$

$$B.E = 0.03038\times931MeV = 28\,MeV$$

$$\left[\because 1uc^2 = 931MeV\right]$$

The binding energy of the ${}^{4}_{2}He$ nucleus is 28 MeV.

8.4.6 Binding energy curve

In the previous section, the origin of the binding energy is discussed. Now we can find the average binding energy per nucleon $\overline{BE}$. It is given by

$$\overline{BE} = \frac{\left[Zm_H + Nm_n - M_A\right]c^2}{A} \quad (8.25)$$

The average binding energy per nucleon is the energy required to separate single nucleon from the particular nucleus. $\overline{BE}$ is plotted against A of all known nuclei. It gives a curve as seen in Figure 8.24.

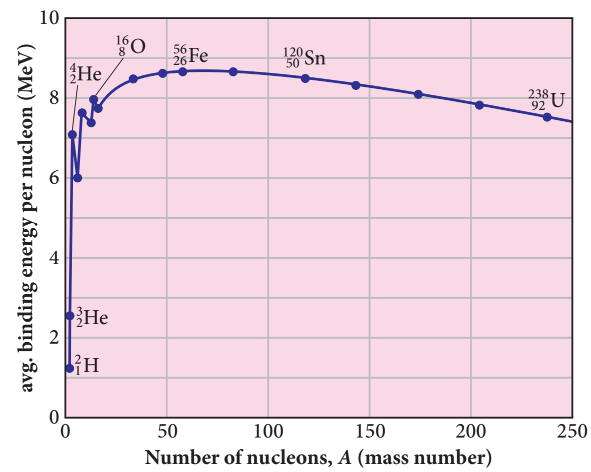


Figure 8.24 Avg. binding energy of the nucleons

Important inferences from of the average binding energy curve:

(1) The value of $\overline{BE}$ rises as the mass number increases until it reaches a maximum value of 8.8 MeV for A = 56 (iron) and then it slowly decreases.

(2) The average binding energy per nucleon is about 8.5 MeV for nuclei having mass number between A = 40 and 120. These elements are comparatively more stable and not radioactive.

(3) For higher mass numbers, the curve reduces slowly and $\overline{BE}$ for uranium is about 7.6 MeV. They are unstable and radioactive.
From Figure 8.24, if two light nuclei with A<28 combine with a nucleus with A<56, the binding energy per nucleon is more for final nucleus than initial nuclei. Thus, if the lighter elements combine to produce a nucleus of medium value A, a large amount of energy will be released. This is the basis of nuclear fusion and is the principle of the hydrogen bomb.

(4) If a nucleus of heavy element is split (fission) into two or more nuclei of medium value A, the energy released would again be large. The atom bomb is based on this principle and huge energy of atom bombs comes from this fission when it is uncontrolled. Fission is explained in the section 8.7

EXAMPLE 8.10

Compute the binding energy per nucleon of ${}_{2}^{4}He$ nucleus.

Solution

From example 8.9, we found that the BE of ${}_{2}^{4}He$ =28 Mev

Binding energy per nucleon $= \overline{B.E} =$ 28 MeV/4 = 7 MeV.

8.5 NUCLEAR FORCE

Nucleus contains protons and neutrons. From electrostatics, we learnt that like charges repel each other. In the nucleus, the protons are separated by a distance of about a few Fermi ($10^{-15} m$), they must exert on each other a very strong repulsive force. For example, the electrostatic repulsive force between two protons separated by a distance $10^{-15} m$

$$F = k \times \frac{q^2}{r^2} = 9 \times 10^9 \times \frac{\left(1.6 \times 10^{-19}\right)^2}{\left(10^{-15}\right)^2} \approx 230N$$

The acceleration experienced by a proton due to the force of 230 N is

$$a = \frac{F}{m} = \frac{230N}{1.67 \times 10^{-27} kg} \approx 1.4 \times 10^{29} m s^{-2}.$$

This is nearly 10^{28} times greater than the acceleration due to gravity. So if the protons in the nucleus experience only the electrostatic force, then the nucleus would fly apart in an instant. Then how protons are held together in nucleus?

From this observation, it was concluded that there must be a strong attractive force between protons to overcome the repulsive Coulomb's force. This attractive force which holds the nucleus together is called strong nuclear force. The properties of strong nuclear force were understood through various experiments carried out between 1930s and 1950s. A few properties of strong nuclear force are

(i) The strong nuclear force is of very short range, acting only up to a distance of a few Fermi. But inside the nucleus, the repulsive Coulomb force or attractive gravitational forces between two protons are much weaker than the strong nuclear force between two protons. Similarly, the gravitational force between two neutrons is also much weaker than strong nuclear force between the neutrons. So nuclear force is the strongest force in nature.

(ii) The strong nuclear force is attractive and acts with an equal strength between proton-proton, proton-neutron, and neutron – neutron.

(iii) Strong nuclear force does not act on the electrons. So it does not alter the chemical properties of the atom.

8.6 RADIOACTIVITY

In the binding energy curve, the stability of the nucleus that has $Z > 82$ starts to decrease and these nuclei are called unstable nuclei. Some of the unstable nuclei decay naturally by emitting some kind of particles to form a stable nucleus. The elements of atomic number $Z > 82$ and isotopes of lighter nuclei belong to naturally-occurring radioactive nuclei. Each of these radioactive nuclei decays to another nucleus by the emission of ${}_{2}^{4}He$ nucleus ($\alpha-$decay) or electron or positron ($\beta-$decay) or gamma rays ($\gamma-$decay).

The phenomenon of spontaneous emission of highly penetrating radiations such as α, β and γ rays by an element is called radioactivity and the substances which emit these radiations are called radioactive elements. These radioactive elements can be heavy elements ($Z > 82$), isotopes of lighter and heavy elements and these isotopes are called radioisotopes. For example, carbon isotope ${}_{6}^{14}C$ is radioactive but ${}_{6}^{12}C$ is not.

Radioisotopes have a variety of applications such as carbon dating, cancer treatment, etc. When radioactive nucleus undergoes decay, the mass of the system decreases – that is, the mass of the initial nucleus before decay is always greater than the sum of the mass of the final nucleus and that of the emitted particle. When this difference in mass $\Delta m < 0$, it appears as the energy according to Einstein's relation $E = |\Delta m| c^2$.

The phenomenon of radioactivity was first discovered by Henri Becquerel in 1896. Later, Marie Curie and her husband Pierre Curie did a series of experiments in detail to understand the phenomenon of radioactivity. In India, Saha Institute of Nuclear Physics (SINP), Kolkata is the premier institute pursuing active research in nuclear physics.

Note During early days of nuclear physics research, the term 'radiation' was used to denote the emanations from radioactive nuclei. Now we know that α rays are in fact ${}_{2}^{4}He$ nuclei and β rays are electrons or positrons. Certainly, they are not electromagnetic radiation. The γ ray alone is electromagnetic radiation.

8.6.1 Alpha decay

When unstable nuclei decay by emitting an α-particle (${}_{2}^{4}He$ nucleus), it loses two protons and two neutrons. As a result, its atomic number Z decreases by 2, the mass number decreases by 4. We write the alpha decay process symbolically in the following way

$$ {}_{Z}^{A}X \rightarrow {}_{Z-2}^{A-4}Y + {}_{2}^{4}He \qquad (8.26)$$

Here X is called the parent nucleus and Y is called the daughter nucleus.

Example: Decay of Uranium ${}_{92}^{238}U$ to thorium ${}_{90}^{234}Th$ with the emission of ${}_{2}^{4}He$ nucleus (α-particle)

$$ {}_{92}^{238}U \rightarrow {}_{90}^{234}Th + {}_{2}^{4}He $$

As already mentioned, the total mass of the daughter nucleus and ${}_{2}^{4}He$ nucleus is always less than that of the parent nucleus. The difference in mass $(\Delta m = m_X - m_Y - m_\alpha)$ is released as energy called **disintegration energy** Q and is given by

$$Q = (m_X - m_Y - m_\alpha)c^2 \qquad (8.27)$$

Note that for spontaneous decay (natural radioactivity) $Q > 0$. In alpha decay process, the disintegration energy is certainly positive ($Q > 0$). In fact, the disintegration energy Q is also the net kinetic energy gained in the decay process or if the parent nucleus is at rest, Q is the total kinetic energy of daughter nucleus and the ${}^4_2He$ nucleus. Suppose $Q < 0$, then the decay process cannot occur spontaneously and energy must be supplied to induce the decay.

Note In alpha decay, why does the unstable nucleus emit ${}^4_2He$ nucleus? Why it does not emit four separate nucleons? After all ${}^4_2He$ consists of two protons and two neutrons. For example, if ${}^{238}_{92}U$ nucleus decays into ${}^{234}_{90}Th$ by emitting four separate nucleons (two protons and two neutrons), then the disintegration energy Q for this process turns out to be negative. It implies that the total mass of products is greater than that of parent(${}^{238}_{92}U$) nucleus. This kind of process cannot occur in nature because it would violate conservation of energy. In any decay process, the conservation of energy, conservation of linear momentum and conservation of angular momentum must be obeyed.

EXAMPLE 8.11

(a) Calculate the disintegration energy when stationary ${}^{232}_{92}U$ nucleus decays to thorium ${}^{228}_{90}Th$ with the emission of α particle. The atomic masses are of ${}^{232}_{92}U = 232.037156u$, ${}^{228}_{90}Th = 228.028741u$ and ${}^4_2He = 4.002603u$

(b) Calculate kinetic energies of ${}^{228}_{90}Th$ and α-particle and their ratio.

Solution

The difference in masses

$\Delta m = (m_U - m_{Th} - m_\alpha)$

= (232.037156–228.028741 – 4.002603)u

The mass lost in this decay = 0.005812 u

Since 1u = 931MeV, the energy Q released is

$$Q = (0.005812u) \times (931 MeV / u)$$
$$= 5.41 MeV$$

This disintegration energy Q appears as the kinetic energy of α particle and the daughter nucleus.

In any decay, the total linear momentum must be conserved.

Total linear momentum of the parent nucleus = total linear momentum of the daughter nucleus +α particle

Since before decay, the uranium nucleus is at rest, its momentum is zero.

By applying conservation of momentum, we get

$$0 = m_{Th}\vec{\upsilon}_{Th} + m_\alpha \vec{\upsilon}_\alpha$$

$$m_\alpha \vec{\upsilon}_\alpha = -m_{Th}\vec{\upsilon}_{Th}$$

It implies that the alpha particle and daughter nucleus move in opposite directions.

In magnitude $m_\alpha \upsilon_\alpha = m_{Th}\upsilon_{Th}$

The velocity of α particle $\upsilon_\alpha = \frac{m_{Th}}{m_\alpha}\upsilon_{Th}$

Note that $\frac{m_{Th}}{m_\alpha} > 1$, so $\upsilon_\alpha > \upsilon_{Th}$. The ratio of the kinetic energy of α particle to the daughter nucleus

$$\frac{K.E_\alpha}{K.E_{Th}} = \frac{\frac{1}{2} m_\alpha \upsilon^2_\alpha}{\frac{1}{2} m_{Th}\upsilon^2_{Th}}$$

By substituting, the value of υ_α into the above equation, we get

$$\frac{K.E_\alpha}{K.E_{Th}} = \frac{m_{Th}}{m_\alpha} = \frac{228.02871}{4.002603} = 57$$

The kinetic energy of α particle is 57 times greater than the kinetic energy of the daughter nucleus ($^{228}_{90}Th$).

The disintegration energy Q = total kinetic energy of products

$$K.E_\alpha + K.E_{Th} = 5.41\, MeV$$

$$57K.E_{Th} + K.E_{Th} = 5.41 MeV$$

$$K.E_{Th} = \frac{5.41}{58} MeV = 0.093 MeV$$

$$K.E_\alpha = 57K.E_{Th} = 57 \times 0.093 = 5.301\, MeV$$

In fact, 98% of total kinetic energy is taken by the α particle.

8.6.2 Beta decay

In beta decay, a radioactive nucleus emits either electron or positron. If electron (e^-) is emitted, it is called β^- decay and if positron (e^+) is emitted, it is called β^+ decay. The positron is an anti-particle of an electron whose mass is same as that of electron and charge is opposite to that of electron – that is, $+e$. Both positron and electron are referred to as beta particles.

β^- decay:

In β^- decay, the atomic number of the nucleus increases by one but mass number remains the same. This decay is represented by

$$^A_ZX \rightarrow {}^{A}_{Z+1}Y + e^- + \bar{\nu} \qquad (8.28)$$

It implies that the element X becomes Y by giving out an electron and antineutrino ($\bar{\nu}$). In otherwords, in each β^- decay, one neutron in the nucleus of X is converted into a proton by emitting an electron (e^-) and antineutrino. It is given by

$$n \rightarrow p + e^- + \bar{\nu}$$

Where p -proton, $\bar{\nu}$ -antineutrino.

Example: Carbon ($^{14}_6C$) is converted into nitrogen ($^{14}_7N$) through β^- decay.

$$^{14}_6C \rightarrow {}^{14}_7N + e^- + \bar{\nu}$$

β^+ decay:

In β^+ decay, the atomic number is decreased by one and the mass number remains the same. This decay is represented by

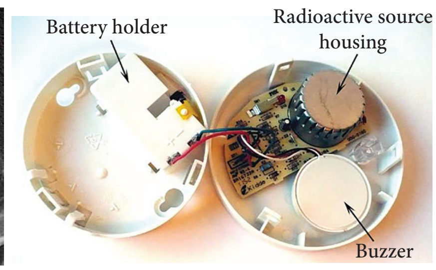
$$^A_ZX \rightarrow {}^{A}_{Z-1}Y + e^+ + \nu \qquad (8.29)$$

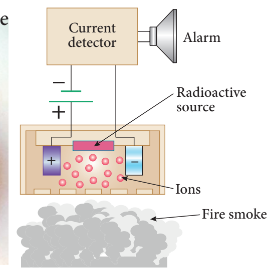
It implies that the element X becomes Y by giving out an positron and neutrino (ν). In otherwords, for each β^+ decay, a proton in the nucleus of X is converted into a neutron by emitting a positron (e^+) and a neutrino. It is given by

$$p \rightarrow n + e^+ + \nu$$

However a single proton (not inside any nucleus) cannot have β^+ decay due to energy conservation, because neutron mass is larger than proton mass. But a single neutron (not inside any nucleus) can have β^- decay.

A very interesting application of alpha decay is in smoke detectors which prevent us from any hazardous fire.





The smoke detector uses around 0.2 *mg* of man-made weak radioactive isotope called americium ($^{241}_{95}Am$). This radioactive source is placed between two oppositely charged metal plates and α radiations from $^{241}_{95}Am$ continuously ionize the nitrogen, oxygen molecules in the air space between the plates. As a result, there will be a continuous flow of small steady current in the circuit. If smoke enters, the radiation is being absorbed by the smoke particles rather than air molecules. As a result, the ionization and along with it the current is reduced. This drop in current is detected by the circuit and alarm starts.

The radiation dosage emitted by americium is very much less than safe level, so it can be considered harmless.

Example: Sodium ($^{22}_{11}Na$) is converted into neon ($^{22}_{10}Ne$) through β^+ decay.

$$^{22}_{11}Na \rightarrow {}^{22}_{10}Ne + e^+ + \nu$$

It is important to note that the electron or positron which comes out from nuclei during beta decay never present inside the nuclei rather they are produced during the conversion of neutron into proton or proton into neutron inside the nucleus.

Neutrino:

Initially, it was thought that during beta decay, a neutron in the parent nucleus is converted to the daughter nuclei by emitting only electron as given by

$$^{A}_{Z}X \rightarrow {}_{Z+1}^{A}Y + e^- \qquad (8.30)$$

But the kinetic energy of electron coming out of the nucleus did not match with the experimental results. In alpha decay, the alpha particle takes only certain allowed discrete energies whereas in beta decay, it was found that the beta particle (i.e, electron) have a continuous range of energies. But the conservation of energy and momentum gives specific single values for electron energy and the recoiling nucleus Y. It seems that the conservation of energy, momentum are violated and could not be explained why energy of beta particle have continuous range of values. So beta decay remained as a puzzle for several years.

After a detailed theoretical and experimental study, in 1931 W.Pauli proposed a third particle which must be present in beta decay to carry away missing energy and momentum. Fermi later named this particle the *neutrino* (little neutral one) since it has no charge, have very little mass. For many years, the neutrino (symbol ν , Greek *nu*) was hypothetical and could not be verified experimentally. Finally, the neutrino was detected experimentally in 1956 by Fredrick Reines and Clyde Cowan. Later Reines received Nobel prize in physics in the year 1995 for his discovery.

The neutrino has the following properties

- It has zero charge
- It has an antiparticle called anti-neutrino.
- Recent experiments showed that the neutrino has very tiny mass.
- It interacts very weakly with the matter. Therefore, it is very difficult to detect. In fact, in every second, trillions of neutrinos coming from the sun are passing through our body without any interaction.

8.6.3 Gamma decay

In α and β decay, the daughter nucleus is in the excited state most of the time. The typical life time of excited state is approximately 10^{-11}s. So this excited state nucleus immediately returns to the ground state or lower energy state by emitting highly energetic photons called γ rays. In fact, when the atom is in the excited state, it returns to the ground state by emitting photons of energy in the order of few eV. But when the excited state nucleus returns to its ground state, it emits a highly energetic photon (γ rays) of energy in the order of MeV. The gamma decay is given by

$$ {}_{Z}^{A}X^{*} \rightarrow {}_{Z}^{A}X + \text{gamma}(\gamma)\text{rays} \qquad (8.31)$$

Here the asterisk(*) means excited state nucleus. In gamma decay, there is no change in the mass number or atomic number of the nucleus.

Boron (${}_{5}^{12}B$) has two beta decay modes as shown in Figure 8.25:

(1) it undergoes beta decay directly into ground state carbon (${}_{6}^{12}C$) by emitting an electron of maximum of energy 13.4 MeV.

(2) it undergoes beta decay to an excited state of carbon (${}_{6}^{12}C^{*}$) by emitting an electron of maximum energy 9.0 MeV followed by gamma decay to ground state by emitting a photon of energy 4.4 MeV.

It is represented by

$$ {}_{5}^{12}B \rightarrow {}_{6}^{12}C + e^{-} + \bar{\nu}$$

$$ {}_{6}^{12}C^{*} \rightarrow {}_{6}^{12}C + \gamma$$

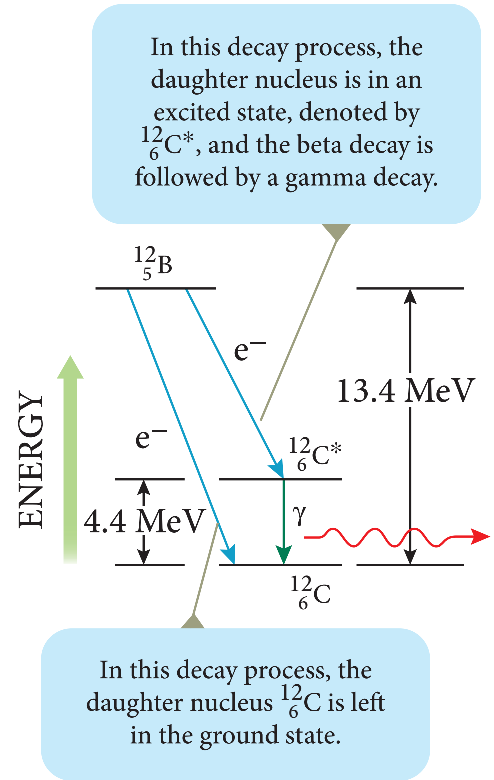


Figure 8.25 Gamma decay

8.6.4 Law of radioactive decay

In the previous section, the decay process of a single radioactive nucleus was discussed. In practice, we have bulk material of radioactive sample which contains a vast number of the radioactive nuclei and not all the radioactive nucleus in a sample decay at the same time. It decays over a period of time and this decay is basically a random process. It implies that we cannot predict which nucleus is going to decay or rather we can determine like probabilistic basis (like tossing a coin). We can calculate approximately how many nuclei in a sample are decayed over a period of time.

At any instant t, the number of decays per unit time, called rate of decay $\left(\frac{dN}{dt}\right)$ is proportional to the number of nuclei (N) at the same instant.

$$\frac{dN}{dt} \propto N$$

By introducing a proportionality constant, the relation can be written as

$$\frac{dN}{dt} = -\lambda N \qquad (8.32)$$

Here proportionality constant λ is called decay constant which is different for different radioactive sample and the negative sign in the equation implies that the N is decreasing with time.

By rewriting the equation (8.32), we get

$$dN = -\lambda N dt \qquad (8.33)$$

Here dN represents the number of nuclei decaying in the time interval dt.

Let us assume that at time $t=0$ s, the number of nuclei present in the radioactive sample is N_0. By integrating the equation (8.33), we can calculate the number of undecayed nuclei N at any time t.

From equation (8.33), we get

$$\frac{dN}{N} = -\lambda\, dt \qquad (8.34)$$

$$\int_{N_0}^{N} \frac{dN}{N} = -\int_0^t \lambda\, dt$$

$$\left[\ln N\right]_{N_0}^{N} = -\lambda t$$

$$\ln\left[\frac{N}{N_0}\right] = -\lambda t$$

Taking exponentials on both sides, we get

$$N = N_0\, e^{-\lambda t} \qquad (8.35)$$

[Note: $e^{\ln x} = e^y \Rightarrow x = e^y$]

Equation (8.35) is called the law of radioactive decay. Here N denotes the number of undecayed nuclei present at any time t and N_0 denotes the number of nuclei at initial time $t=0$. Note that the number of atoms is decreasing exponentially over the time. This implies that the time taken for all the radioactive nuclei to decay will be infinite. Equation (8.35) is plotted in Figure 8.26.

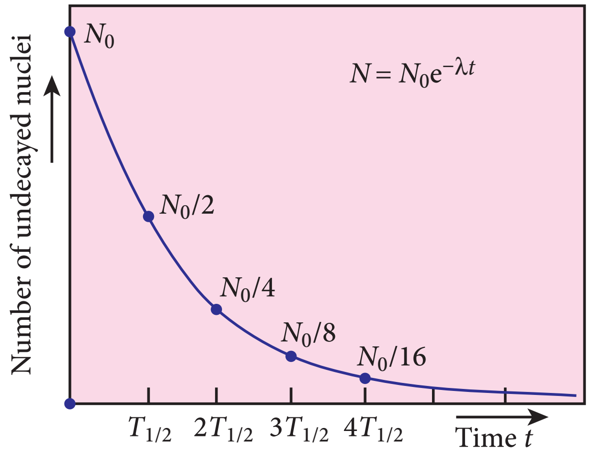


Figure 8.26 Law of radioactive decay

We can also define another useful quantity called activity (R) or decay rate which is the number of nuclei decayed per second and it is denoted as $R = \left|\frac{dN}{dt}\right|$. Note that activity R is a positive quantity.

From equation (8.35), we get

$$R = \left|\frac{dN}{dt}\right| = \lambda N_0\, e^{-\lambda t} \qquad (8.36)$$

$$R = R_0\, e^{-\lambda t} \qquad (8.37)$$

where $R_0 = \lambda N_0$

The equation (8.37) is also equivalent to radioactive law of decay. Here R_0 is the activity of the sample at *t=0* and R is the activity of the sample at any time t. From equation (8.37), activity also shows exponential decay behavior. The activity R also can be expressed in terms of number of undecayed atoms present at any time t.

From equation (8.37), since $N = N_0 e^{-\lambda t}$, we write

$$R = \lambda N \qquad (8.38)$$

Equation (8.35) implies that the activity at any time t is equal to the product of decay constant and number of undecayed nuclei at the same time t. Since N decreases over time, R also decreases.

The SI unit of activity R is Becquerel and one Becquerel (Bq) is equal to one decay per second. There is also another standard unit for the activity called Curie(Ci).

1 Curie =1 Ci $= 3.7 \times 10^{10}$ decays per second

1 Ci $= 3.7 \times 10^{10}$ Bq

Note Initially one curie was defined as number of decays per second in 1 g of radium and it is equal to 3.7×10^{10} decays/s.

8.6.5 Half-life

It is difficult to calculate the time taken by a given a sample of N atoms to decay. However, we can calculate the time taken by the given sample of atoms to reduce some fraction of the initial amount.

We can define the half-life $T_{1/2}$ as the time required for the number of atoms initially present to reduce to one half of the initial amount.

The half-life is the important characteristic of every radioactive sample. Some radioactive nuclei are known to have half-life as long as 10^{14} years and some nucleus have very shorter life time (10^{-14}s).

We can express half-life in terms of the decay constant. At $t = T_{1/2}$, the number of undecayed nuclei $N = \frac{N_0}{2}$.

By substituting this value in to the equation (8.35), we get

$$\frac{N_0}{2} = N_0 e^{-\lambda T_{1/2}}$$

$$\frac{1}{2} = e^{-\lambda T_{1/2}} \text{ or } e^{\lambda T_{1/2}} = 2$$

Taking logarithm on both sides and rearranging the terms,

$$T_{1/2} = \frac{\ln 2}{\lambda} = \frac{0.6931}{\lambda} \qquad (8.39)$$

Note One should not think that shorter half-life material is safer than longer half-life material because it will not last long. The shorter half-life sample will have higher activity and it is more 'radioactive' which is more harmful.

If the number of atoms present at t=0 is N_0, then $\frac{N_0}{2}$ atoms remain undecayed in first half-life and $\frac{N_0}{4}$ atoms remain undecayed after second half life and so on. In general, after n half-lives, the number of nuclei remaining undecayed is given by

$$N = \left(\frac{1}{2}\right)^n N_0 \qquad (8.40)$$

where n can be integer or non-integer. Since the activity of radioactive sample also obeys the exponential decay law, we can also write an equation for an activity similar to equation (8.36).

After n half-lives, the activity or decay rate of any radioactive sample is

$$R = \left(\frac{1}{2}\right)^n R_0 \qquad (8.41)$$

Mean life (τ):

When the radioactive nucleus undergo the decay, the nucleus which disintegrates first has zero life time and the nucleus which decay last has an infinite lifetime. The actual life time for each nucleus varies from zero to infinity. Therefore, it is meaningful to define average life or mean life time τ, that the nucleus survives before it decays.

The mean life time of the nucleus is the ratio of sum or integration of life times of all nuclei to the total number nuclei present initially.

The total number of nuclei decaying in the time interval from t to $t+\Delta t$ is equal to $R\Delta t = \lambda N_0 e^{-\lambda t}\Delta t$. It implies that until the time t, this $R\Delta t$ number of nuclei lived. So the life time of these $R\Delta t$ nuclei is equal to be $tR\Delta t$. In the limit $\Delta t \to 0$, the total life time of all the nuclei would be the integration of $tRdt$ from the limit $t=0$ to $t=\infty$.

Mean life

$$\tau = \frac{\int_0^\infty t[Rdt]}{N_0} = \frac{\int_0^\infty t\left[\lambda N_0 e^{-\lambda t}\, dt\right]}{N_0} \qquad (8.42)$$

After a few integration (refer box item), the expression for mean life time,

$$\tau = \frac{1}{\lambda} \qquad (8.43)$$

Note that mean life and decay constant is inversely proportional to each other.

Using mean life, the half-life can be rewritten as

$$T_{1/2} = \tau \ln 2 = 0.6931\tau \qquad (8.44)$$

Mean life : Not for examination

The integration in the equation (8.42) can be performed using integration by parts.

$$\tau = \frac{\int_0^\infty \lambda N_0 t\, e^{-\lambda t} dt}{N_0} = \frac{\lambda N_0 \int_0^\infty t\, e^{-\lambda t} dt}{N_0}$$

$$\tau = \lambda \int_0^\infty t\, e^{-\lambda t} dt$$

$$u = t \qquad dv = e^{-\lambda t}\, dt$$

$$\tau = \lambda \int_0^\infty t\, e^{-\lambda t} dt = \lambda \left[\frac{t\, e^{-\lambda t}}{-\lambda}\right]_0^\infty - \lambda \int_0^\infty \left[\frac{e^{-\lambda t}}{-\lambda}\right] dt$$

By substituting the limits, the first term in the above equation becomes zero.

$$\tau = \int_0^\infty e^{-\lambda t}\, dt = -\frac{1}{\lambda}\left[e^{-\lambda t}\right]_0^\infty = \frac{1}{\lambda}$$

EXAMPLE 8.12

Calculate the number of nuclei of carbon-14 undecayed after 22,920 years if the initial number of carbon-14 atoms is 10,000. The half-life of carbon-14 is 5730 years.

Solution

To get the time interval in terms of half-life,

$$n = \frac{t}{T_{1/2}} = \frac{22{,}920\, yr}{5730\, yr} = 4$$

The number of nuclei remaining undecayed after 22,920 years,

$$N=\left(\frac{1}{2}\right)^{n}N_{0}=\left(\frac{1}{2}\right)^{4}\times 10,000$$

$$N=625$$

EXAMPLE 8.13

A radioactive sample has $2.6\mu g$ of pure ${}^{13}_{7}N$ which has a half-life of 10 minutes. (a) How many nuclei are present initially? (b) What is the activity initially? (c) What is the activity after 2 hours? (d) Calculate mean life of this sample.

Solution

(a) To find N_0, we have to find the number of ${}^{13}_{7}N$ atoms in $2.6\mu g$. The atomic mass of nitrogen is 13. Therefore, 13 g of ${}^{13}_{7}N$ contains Avogadro number (6.02×10^{23}) of atoms.

In 1 g, the number of ${}^{13}_{7}N$ is equal to be $\frac{6.02\times10^{23}}{13}$ atoms. So the number of ${}^{13}_{7}N$ atoms in $2.6\mu g$ is

$$N_0=\frac{6.02\times10^{23}}{13}\times 2.6\times10^{-6}=12.04\times10^{16}\text{ atoms}$$

(b) To find the initial activity R_0, we have to evaluate decay constant λ

$$\lambda=\frac{0.6931}{T_{1/2}}=\frac{0.6931}{10\times60}=1.155\times10^{-3}\text{s}^{-1}$$

Therefore

$$R_0=\lambda N_0=1.155\times10^{-3}\times12.04\times10^{16}$$

$$=13.90\times10^{13}\text{ decays/s}$$

$$=13.90\times10^{13}\text{ Bq}$$

In terms of a curie,

$$R_0=\frac{13.90\times10^{13}}{3.7\times10^{10}}=3.75\times10^{3}\text{ Ci}$$

since $1\text{Ci}=3.7\times10^{10}\text{Bq}$

(c) Activity after 2 hours can be calculated in two different ways:

Method 1: $R=R_0e^{-\lambda t}$

At $t=2\text{ hr}=7200$ s

$$R=3.75\times10^{3}\times e^{-7200\times1.155\times10^{-3}}$$

$$R=3.75\times10^{3}\times2.4\times10^{-4}=0.9\text{ Ci}$$

Method 2: $R=\left(\frac{1}{2}\right)^{n}R_0$

Here $n=\frac{120\text{ min}}{10\text{ min}}=12$

$$R=\left(\frac{1}{2}\right)^{12}\times3.75\times10^{3}\approx0.9\text{Ci}$$

(d) mean life $\tau=\frac{T_{1/2}}{0.6931}=\frac{10\times60}{0.6931}$

$=865.67\,s$

8.6.6 Carbon dating

The interesting application of beta decay is radioactive dating or carbon dating. Using this technique, the age of an ancient object can be calculated. All living organisms absorb carbon dioxide (CO_2) from air to synthesize organic molecules. In this absorbed CO_2, the major part is ${}^{12}_{6}C$ and very small fraction (1.3×10^{-12}) is radioactive ${}^{14}_{6}C$ whose half-life is 5730 years.

Carbon-14 in the atmosphere is always decaying but at the same time, cosmic rays from outer space are continuously bombarding the atoms in the atmosphere which produces ${}^{14}_{6}C$. So the continuous production and decay of ${}^{14}_{6}C$ in the atmosphere keep the ratio of ${}^{14}_{6}C$ to ${}^{12}_{6}C$ always constant. Since our human body, tree or any living organism continuously absorb CO_2 from the atmosphere, the ratio of ${}^{14}_{6}C$ to ${}^{12}_{6}C$ in the living organism is also nearly constant. But when the organism dies, it stops absorbing CO_2. Since ${}^{14}_{6}C$ starts to decay, the ratio of ${}^{14}_{6}C$ to ${}^{12}_{6}C$ in a dead organism or specimen decreases over the years. Suppose the ratio of ${}^{14}_{6}C$ to ${}^{12}_{6}C$ in the ancient tree pieces excavated is known, then the age of the tree pieces can be calculated.

EXAMPLE 8.14

Keezhadi (கீழடி), a small hamlet, has become one of the very important archeological places of Tamilandu. It is located in Sivagangai district. A lot of artefacts (gold coins, pottery, beads, iron tools, jewellery and charcoal, etc.) have been unearthed in Keezhadi which have given substantial evidence that an ancient urban civilization had thrived on the banks of river Vaigai. To determine the age of those materials, the charcoal of 200 g sent for carbon dating is given in the following figure (b). The activity of ${}^{14}_{6}C$ is found to be 38 decays/s. Calculate the age of charcoal.

Figure (a) Keezhadi – excavation site

Figure (b) – Characol which was sent for carbon dating

Solution

To calculate the age, we need to know the initial activity (R_0) of the characol (when the sample was alive).

The activity R of the sample

$$R = R_0 e^{-\lambda t} \qquad (1)$$

To find the time t, rewriting the above equation (1), $e^{\lambda t} = \frac{R_0}{R}$

By taking the logarithm on both sides, we get $t = \frac{1}{\lambda}\ln\left(\frac{R_0}{R}\right)$ (2)

Here R = 38 decays/s=38 Bq.

To find decay constant, we use the equation

$$\lambda = \frac{0.6931}{T_{1/2}} = \frac{0.6931}{5730\,\text{yr} \times 3.156 \times 10^7\,\text{s/yr}}$$

$[\because 1\text{yr} = 365.25 \times 24 \times 60 \times 60\,\text{s} = 3.156 \times 10^7\,\text{s}]$

$$\lambda = 3.83 \times 10^{-12}\,s^{-1}$$

To find the initial activity R_0, we use the equation $R_0 = \lambda N_0$. Here N_0 is the number of carbon-14 atoms present in the sample when it was alive. The mass of the characol is 200 g. In 12 g of carbon, there are 6.02×10^{23} carbon atoms. So 200 g contains,

$$\frac{6.02 \times 10^{23}\,\text{atoms}/mol}{12\,g/mol} \times 200 \approx 1 \times 10^{25}\,\text{atoms}$$

When the tree(sample) was alive, the ratio of ${}^{14}_{6}C$ to ${}^{12}_{6}C$ is 1.3×10^{-12}. So the total number of carbon-14 atoms is given by

$$N_0 = 1 \times 10^{25} \times 1.3 \times 10^{-12} = 1.3 \times 10^{13}\,\text{atoms}$$

The initial activity

$$R_0 = 3.83 \times 10^{-12} \times 1.3 \times 10^{13} \approx 50\,\text{decays}/s$$
$$= 50\,Bq$$

By substituting the value of R_0 and λ in the equation (2), we get

$$t = \frac{1}{3.83\times10^{-12}}\times\ln\left[\frac{50}{38}\right]$$

$$t = \frac{0.27}{3.83}\times10^{12} \approx 7\times10^{10}\text{ sec}$$

In years

$$t = \frac{7\times10^{10}s}{3.156\times10^{7}s/yr} \approx 2200\text{ years}$$

In fact, the excavated materials were sent for carbon dating to USA by Archeological Department of Tamilnadu and the report confirmed that the age of Keezhadi artefacts lies between 2200 years to 2500 years (Sangam era- 400 BC to 200 BC). The Keezhadi excavations experimentally proved that urban civilization existed in Tamil Nadu even 2000 years ago!

8.6.7 Discovery of Neutrons

In 1930, two German physicists Bothe and Becker found that when beryllium was bombarded with α particles, highly penetrating radiation was emitted. This radiation was capable of penetrating the thick layer of lead and was unaffected by the electric and magnetic fields. Initially, it was thought as γ radiation. But in the year 1932, James Chadwick discovered that those radiations are not EM waves but they are particles of mass little greater than the mass of the proton and had no charge. He called them as neutrons. The above reaction can be written as

$$^{9}_{4}Be + ^{4}_{2}He \rightarrow ^{12}_{6}C + ^{1}_{0}n$$

where $^{1}_{0}n$ denotes neutron.

Neutrons are stable inside the nucleus. But outside the nucleus they are unstable. If the neutron comes out of the nucleus (free neutron), it decays with emission of proton, electron, and antineutrino with the half life of 13 minutes.

Neutrons are classified according to their kinetic energy as (i) slow neutrons (0 to 1000 eV) (ii) fast neutrons (0.5 MeV to 10 MeV). The neutrons with average energy of about 0.025 eV in thermal equilibrium are called thermal neutron, because at 298K, the thermal energy kT $0.025eV$. Slow and fast neutrons play a vital role in nuclear reactors.

8.7 NUCLEAR FISSION

In 1939, German scientists Otto Hahn and F. Strassman discovered that when uranium nucleus is bombarded with a neutron, it breaks up into two smaller nuclei of comparable masses with the release of energy. **The process of breaking up of the nucleus of a heavier atom into two smaller nuclei with the release of a large amount of energy is called nuclear fission**. The fission is accompanied by the release of neutrons. The energy that is released in the nuclear fission is of many orders of magnitude greater than the energy released in chemical reactions.

Uranium undergoes fission reaction in 90 different ways. The most common fission reactions of $^{235}_{92}U$ nuclei are shown here.

$$^{235}_{92}U + ^{1}_{0}n \rightarrow ^{236}_{92}U^{*} \rightarrow ^{141}_{56}Ba + ^{92}_{36}Kr + 3^{1}_{0}n + Q \qquad (8.45)$$

$$^{235}_{92}U + ^{1}_{0}n \rightarrow ^{236}_{92}U^{*} \rightarrow ^{140}_{54}Xe + ^{94}_{38}Sr + 2^{1}_{0}n + Q \qquad (8.46)$$

Here Q is energy released during the decay of each uranium nuclei. When the slow neutron is absorbed by the uranium nuclei, the mass number increases by one and goes to an excited state $^{236}_{92}U^{*}$. But this excited state does not last longer than 10^{-12}s and decay into two daughter nuclei along with 2 or 3 neutrons. From each reaction, on an average, 2.5 neutrons are emitted. It is shown in Figure 8.27

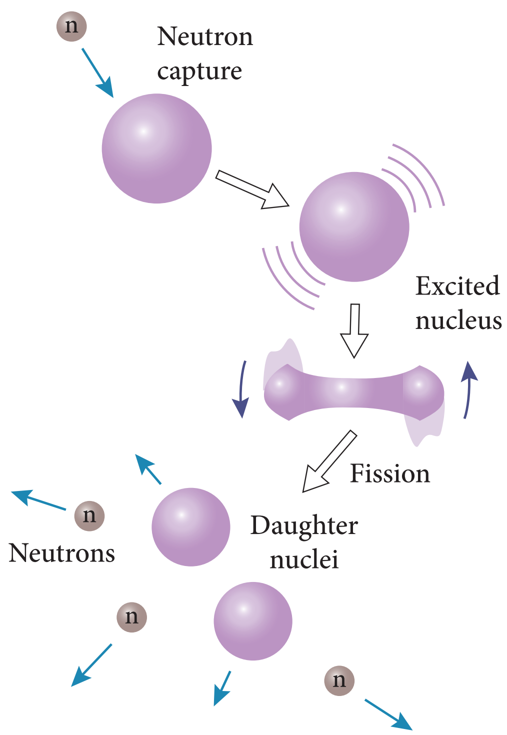


Figure 8.27 Nuclear fission

Energy released in fission:

We can calculate the energy (Q) released in each uranium fission reaction. We choose the most favorable fission which is given in the equation (8.45).

$$^{235}_{92}U + ^{1}_{0}n \rightarrow ^{236}_{92}U^{*} \rightarrow ^{141}_{56}Ba + ^{92}_{36}Kr + 3^{1}_{0}n + Q$$

Mass of $^{235}_{92}U$ = 235.045733 u

Mass of $^{1}_{0}n$ = 1.008665 u

Total mass of reactant = 236.054398 u

Mass of $^{141}_{56}Ba$ = 140.9177 u

Mass of $^{92}_{36}Kr$ = 91.8854 u

Mass of 3 neutrons = 3.025995 u

The total mass of products = 235.829095 u

Mass defect $\Delta m = 236.054398\ u - 235.829095\ u$ $= 0.225303\ u$

So the energy released in each fission = $0.225303 \times 931 MeV \approx 200\ MeV$

This energy first appears as kinetic energy of daughter nuclei and neutrons. But later, this kinetic energy is transferred to the surrounding matter as heat.

Chain reaction:

When one $^{235}_{92}U$ nucleus undergoes fission, the energy released might be small. But from each fission reaction, three neutrons are released. These three neutrons cause further fission in another three $^{235}_{92}U$ nuclei which in turn produce nine neutrons. These nine neutrons initiate fission in another 27 $^{235}_{92}U$ nuclei and so on. This is called a chain reaction and the number of neutrons goes on increasing almost in geometric progression. It is shown in Figure 8.28.

There are two kinds of chain reactions: (i) uncontrolled chain reaction (ii) controlled chain reaction. In an uncontrolled chain reaction, the number of neutrons multiply indefinitely and the entire amount of energy released in a fraction of second.

The atom bomb is an example of nuclear fission in which uncontrolled chain reaction occurs. Atom bombs produce massive destruction for mankind. During World War II, in the year 1946 August 6 and 9, USA dropped two atom bombs in two places of Japan, Hiroshima and Nagasaki. As a result, lakhs of people were killed and the two cities were completely destroyed. Even now the people who are living in those places have side effects caused by the explosion of atom bombs.

It is possible to calculate the typical energy released in a chain reaction. In the

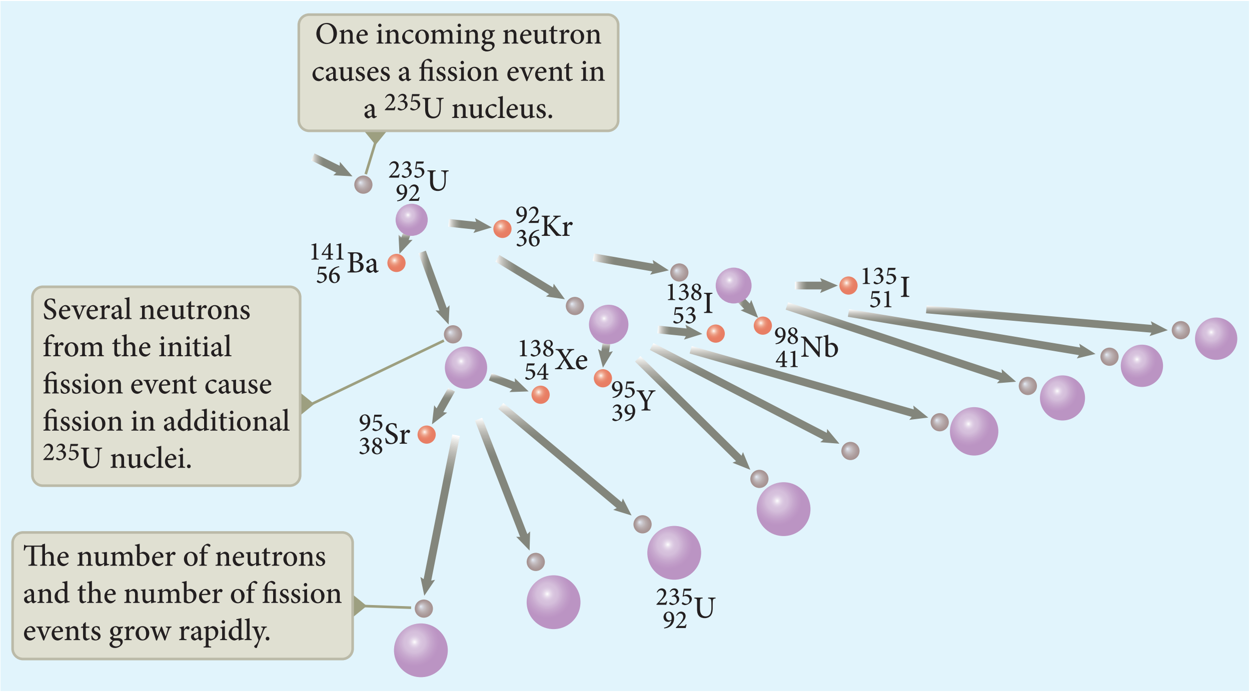


Figure 8.28 Nuclear chain reaction

first step, one neutron initiates the fission of one nucleus by producing three neutrons and energy of about 200 MeV. In the second step, three nuclei undergo fission, in third step nine nuclei undergo fission, in fourth step 27 nucleus undergo fission and so on. In the 100th step, the number of nuclei which undergoes fission is around 2.5×10^{40}. The total energy released after 100th step is $2.5\times10^{40}\times200\,MeV = 8\times10^{29}\,J$. It is really an enormous amount of energy which is equivalent to electrical energy required in Tamilnadu for several years.

If the chain reaction is controllable, then we can harvest an enormous amount of energy for our needs. It is achieved in a controlled chain reaction. In the controlled chain reaction, the average number of neutron released in each stage is kept as one such that it is possible to store the released energy. In nuclear reactors, the controlled chain reaction is achieved and the produced energy is used for power generation or for research purpose.

EXAMPLE 8.15

Calculate the amount of energy released when 1 kg of $^{235}_{92}U$ undergoes fission reaction.

Solution

235 g of $^{235}_{92}U$ has 6.02×10^{23} atoms. In one gram of $^{235}_{92}U$, the number of atoms is equal to $\frac{6.02\times10^{23}}{235} = 2.56\times10^{21}$.

So the number of atoms in 1 kg of $^{235}_{92}U$ = $2.56\times10^{21}\times1000 = 2.56\times10^{24}$

Each $^{235}_{92}U$ nucleus releases 200 MeV of energy during the fission. The total energy released by 1kg of $^{235}_{92}U$ is

$$Q = 2.56\times10^{24}\times200\,MeV = 5.12\times10^{26}\,MeV$$

By converting in terms of joules,

$$Q = 5.12\times10^{26}\times1.6\times10^{-13}\,J = 8.192\times10^{13}\,J.$$

In terms of Kilowatt hour,

$$Q = \frac{8.192\times10^{13}}{3.6\times10^{6}} = 2.27\times10^{7}\,kWh$$

This is enormously large energy which is enough to keep 100 W light bulb operating for 30,000 years. To produce this much energy through chemical reaction, around 20,000 tons of TNT(tri nitro toluene) has to be exploded.

Nuclear reactor:

Nuclear reactor is a system in which the nuclear fission takes place in a self-sustained controlled manner and the energy produced is used either for research purpose or for power generation. The first nuclear reactor was built in the year 1942 at Chicago, USA by physicist Enrico Fermi. The main parts of a nuclear reactor are fuel, moderator and control rods. In addition to this, there is a cooling system which is connected with power generation set up.

Fuel: The fuel is fissionable material, usually uranium or plutonium. Naturally occurring uranium contains only 0.7% of ${}^{235}_{92}U$ and 99.3% are only ${}^{238}_{92}U$. So the ${}^{238}_{92}U$ must be enriched such that it contains at least 2 to 4% of ${}^{235}_{92}U$. In addition to this, a neutron source is required to initiate the chain reaction for the first time. A mixture of beryllium with plutonium or polonium is used as the neutron source. During fission of ${}^{235}_{92}U$, only fast neutrons are emitted but the probability of initiating fission by it in another nucleus is very low. Therefore, slow neutrons are preferred for sustained nuclear reactions.

Moderators: The moderator is a material used to convert fast neutrons into slow neutrons. Usually the moderators are chosen in such a way that it must be very light nucleus having mass comparable to that of neutrons. Hence, these light nuclei undergo collision with fast neutrons and the speed of the neutron is reduced (Note that a billiard ball striking a stationary billiard ball of equal mass would itself be stopped but the same billiard ball bounces off almost with same speed when it strikes a heavier mass. This is the reason for using lighter nuclei as moderators). Most of the reactors use water, heavy water (D_2O) and graphite as moderators. The blocks of uranium stacked together with blocks of graphite (the moderator) to form a large pile is shown in the Figure 8.29 (a) & (b).

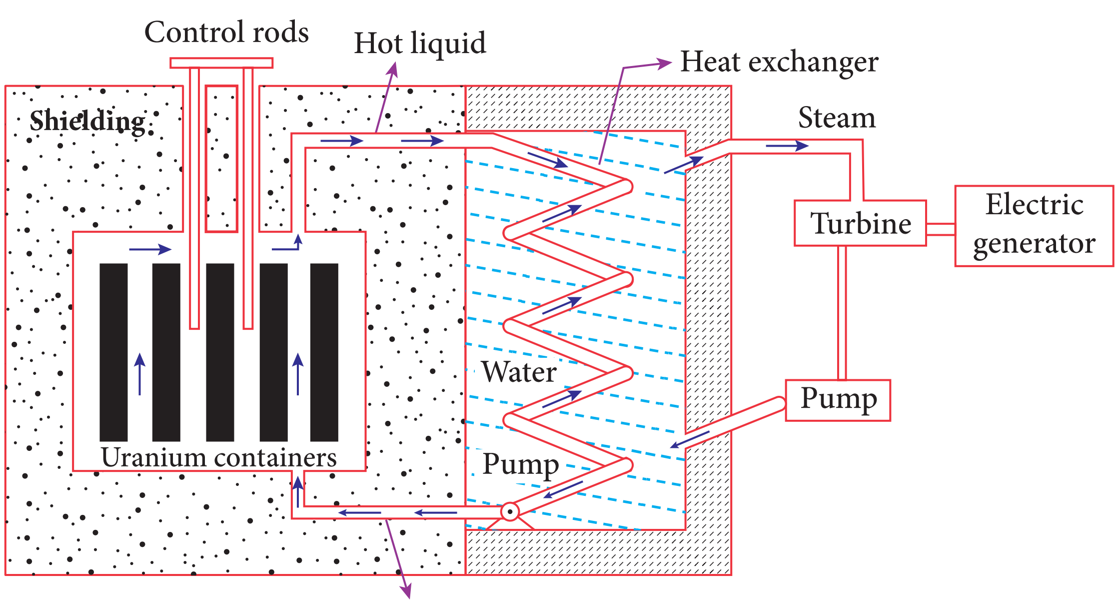


Figure 8.29 (a) Block diagram of Nuclear reactor

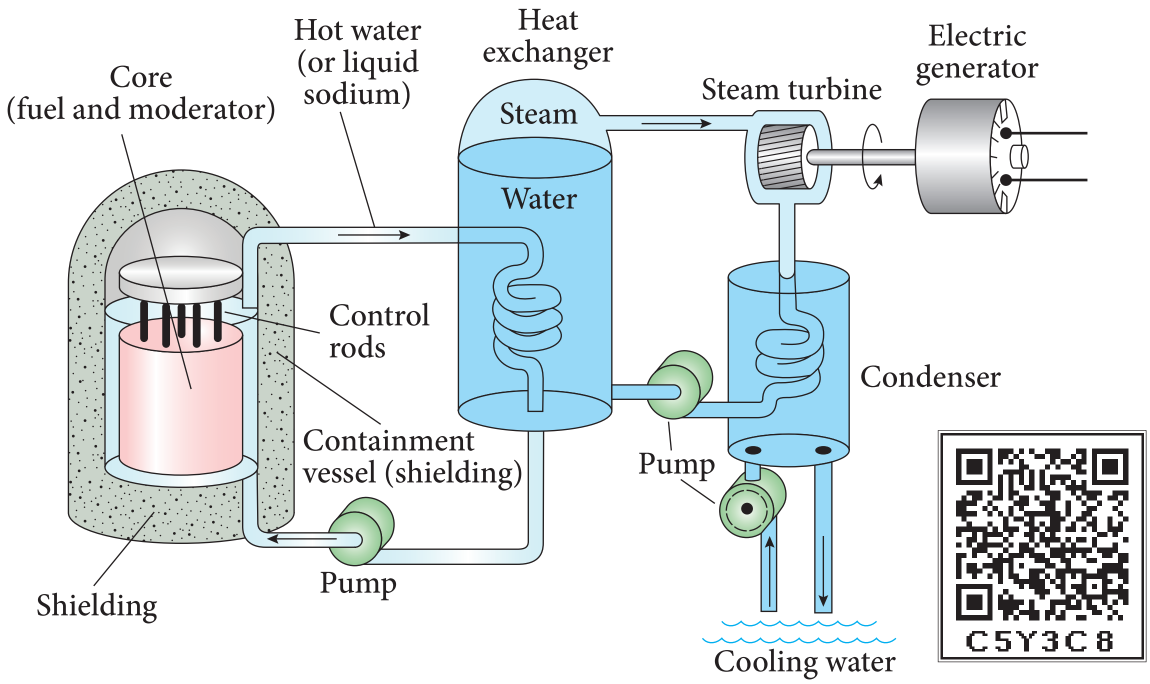


Figure 8.29 (b) Schematic diagram of nuclear reactor

Control rods: The control rods are used to adjust the reaction rate. During each fission, on an average 2.5 neutrons are emitted and in order to have the controlled chain reactions, only one neutron is allowed to cause another fission and the remaining neutrons are absorbed by the control rods.

Usually cadmium or boron acts as control rod material and these rods are inserted into the uranium blocks as shown in the Figure 8.29 (a) and (b). Depending on the insertion depth of control rod into the uranium, the average number of neutrons produced per fission is set to be equal to one or greater than one. If the average number of neutrons produced per fission is equal to one, then reactor is said to be in critical state. In fact, all the nuclear reactors are maintained in critical state by suitable adjustment of control rods. If it is greater than one, then reactor is said to be in super-critical and it may explode sooner or may cause massive destruction.

Shielding: For a protection against harmful radiations, the nuclear reactor is surrounded by a concrete wall of thickness of about 2 to 2.5 m.

Cooling system: The cooling system removes the heat generated in the reactor core. Ordinary water, heavy water and liquid sodium are used as coolant since they have very high specific heat capacity and have large boiling point under high pressure. This coolant passes through the fuel block and carries away the heat to the steam generator through heat exchanger as shown in Figure 8.29(a) and (b). The steam runs the turbines which produces electricity in power reactors.

Note India has 22 nuclear reactors in operation. Nuclear reactors are constructed in two places in Tamilnadu, Kalpakkam and Kudankulam. Even though nuclear reactors are aimed to cater to our energy need, in practice nuclear reactors now are able to provide only 2% of energy requirement of India.

8.8 NUCLEAR FUSION

When two or more light nuclei (A<20) combine to form a heavier nucleus, then it is called nuclear fusion. In the nuclear fusion, the mass of the resultant nucleus is less than the sum of the masses of original light nuclei. The mass difference appears as energy. The nuclear fusion never occurs at room temperature unlike nuclear fission. It is because when two light nuclei come closer to combine, it is strongly repelled by the coulomb repulsive force.

To overcome this repulsion, the two light nuclei must have enough kinetic energy to move closer to each other such that the nuclear force becomes effective. This can be achieved if the temperature is very much greater than the value 10^7 K. When the surrounding temperature reaches around 10^7K, lighter nuclei start fusing to form heavier nuclei and this resulting reaction is called thermonuclear fusion reaction.

Energy generation in stars:

The natural place where nuclear fusion occurs is the core of the stars, since its temperature is of the order of 10^7K. In fact, the energy generation in every star is only through thermonuclear fusion. Most of the stars including our Sun fuse hydrogen into helium and some stars even fuse helium into heavier elements.

The early stage of a star is in the form of cloud and dust. Due to their own gravitational pull, these clouds fall inward. As a result, its gravitational potential energy is converted to kinetic energy and finally into heat. When the temperature is high enough to initiate the thermonuclear fusion, they start to release enormous energy which tends to stabilize the star and prevents it from further collapse.

The sun's interior temperature is around $1.5\times10^7\ K$. The sun is converting $6\times10^{11}\ kg$ hydrogen into helium every second and it has enough hydrogen such that these fusion lasts for another 5 billion years. When the hydrogen is burnt out, the sun will enter into new phase called red giant where helium will fuse to become carbon. During this stage, sun will expand greatly in size and all its planets will be engulfed in it.

According to Hans Bethe, the sun is powered by **proton-proton cycle** of fusion reaction. This cycle consists of three steps and the first two steps are as follows:

$$ {}_1^1H + {}_1^1H \rightarrow {}_1^2H + e^+ + \nu \qquad (8.44)$$

$$ {}_1^1H + {}_1^2H \rightarrow {}_2^3He + \gamma \qquad (8.45)$$

A number of reactions are possible in the third step. But the dominant one is

$$ {}_2^3He + {}_2^3He \rightarrow {}_2^4He + {}_1^1H + {}_1^1H \qquad (8.46)$$

The overall energy production in the above reactions is about 27 MeV. The radiation energy we received from the sun is due to these fusion reactions.

Elementary particles:

An atom has a nucleus surrounded by electrons and nuclei is made up of protons and neutrons. Till 1960s, it was thought that protons, neutrons and electrons are fundamental building blocks of matter. In 1964, physicist Murray Gellman and George

Zweig theoretically proposed that protons and neutrons are not fundamental particles; in fact they are made up of quarks. These quarks are now considered elementary particles of nature. Electrons are fundamental or elementary particles because they are not made up of anything. In the year 1968, the quarks were discovered experimentally by Stanford Linear Accelerator Center (SLAC), USA. There are six quarks namely, up, down, charm, strange, top and bottom and their antiparticles. All these quarks have fractional charges. For example, charge of up quark is $+\frac{2}{3}e$ and that of down quark is $-\frac{1}{3}e$.

According to quark model, proton is made up of two up quarks and one down quark and neutron is made up of one up quark and two down quarks as shown in the Figure 8.30.

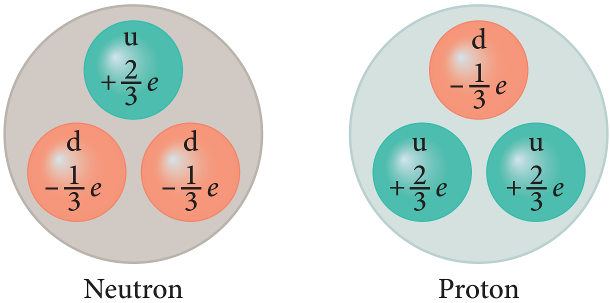


Figure 8.30 Constituents of nucleons

The study of elementary particles is called particle physics and it is an active area of research even now. Till date, more than 20 Nobel prizes have been awarded in the field of particle physics.

Fundamental forces of nature:

It is known that there exists gravitational force between two masses and it is universal in nature. Our planets are bound to the sun through gravitational force of the sun. In +2 volume 1, we have learnt that between two charges there exists electromagnetic force and it plays major role in most of our day-to-day events. In this unit, we have learnt that between two nucleons, there exists a strong nuclear force and this force is responsible for stability of the nucleus. In addition to these three forces, there exists another fundamental force of nature called the weak force. This weak force is even shorter in range than nuclear force. This force plays an important role in beta decay and energy production of stars. During the fusion of hydrogen into helium in sun, neutrinos and enormous radiations are produced through weak force. The detailed mechanism of weak force is beyond the scope of this book and for further reading, appropriate books can be referred.

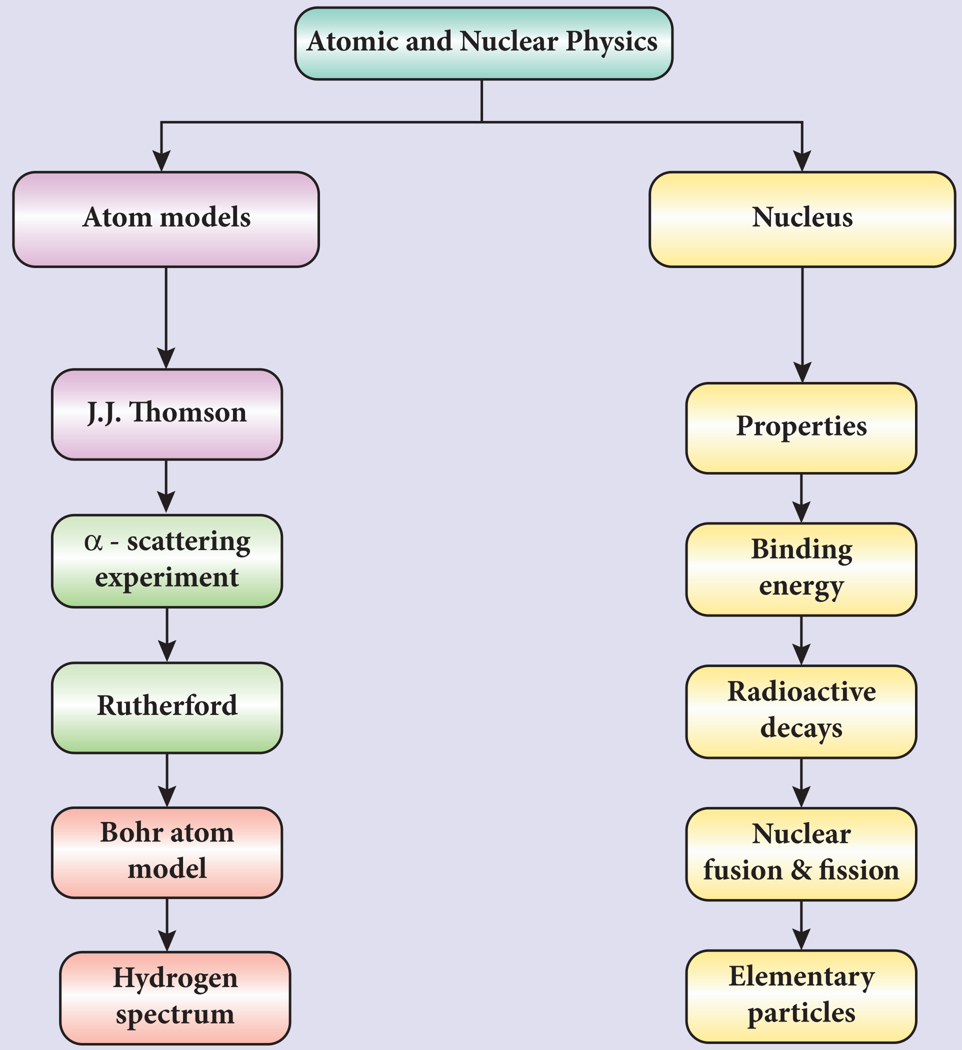
Gravitational, electromagnetic, strong and weak forces are called fundamental forces of nature. It is very interesting to realize that, even for our day-to-day life, we require these four fundamental forces. To put it in simple words: We are in the Earth because of Earth's gravitational attraction on our body. We are standing on the surface of the earth because of the electromagnetic force between atoms of the surface of the earth with atoms in our foot. The atoms in our body are stable because of strong nuclear force. Finally, the lives of species in the earth depend on the solar energy from the sun and it is due to weak force which plays vital role during nuclear fusion reactions going on in the core of the sun.

SUMMARY

- A device used to study the conduction of electricity through gases is known as gas discharge tube
- Charge per unit mass is known as specific charge or normalized charge, and it is independent of gas used and also nature of electrodes used
- The minimum distance between alpha particle and centre of the nucleus just before it gets reflected back by 180° is defined as distance of closest approach r0
- The impact parameter (b) (see Figure 8.12) is defined as the perpendicular distance between the centre of the gold nucleus and the direction of velocity vector of alpha particle when it is at a large distance.
- According to Bohr atom model, angular momentum is quantized.
- The radius of the orbit in Bohr atom model is $r_n = a_0 \frac{n^2}{Z}$
- The radius of first orbit is $a_0 = \frac{\varepsilon_0 h^2}{\pi m e^2} = 0.529\,\text{Å}$ also known as Bohr radius
- The velocity of electron in nth orbit is $\upsilon_n = \frac{h}{2\pi m a_0}\frac{Z}{n}$
- The fine structure constant is $\alpha = \frac{1}{137}$ which is a dimensionless constant
- The total energy of electron in the nth orbit is $E_n = -\frac{me^4}{8\varepsilon_0^2 h^2}\frac{Z^2}{n^2} = -13.6\frac{1}{n^2}eV$
- The energy required to excite an electron from the lower energy state to any higher energy state is known as excitation energy and corresponding potential supplied is known as excitation potential.
- The minimum energy required to remove an electron from an atom which is in ground state is known as ionization energy.
- The potential difference through which an electron should be accelerated to get ionization energy is known as ionization potential.
- The wavelength of spectral lines of Lyman series lies in ultra-violet region
- The wavelength of spectral lines of Balmer series lies in visible region while those of Paschen and Brackett series lie in infra-red region
- The nucleus of element X having atomic number Z and mass number A is represented by ${}_Z^A X$
- The radius of nucleus (Z > 10) of mass number A is given by $R = R_0 A^{1/3}$ where $R_0 = 1.2\,F$
- The density of nucleus $\rho = 2.3 \times 10^{17}\, kg\, m^{-3}$
- If M, m_p and m_n are masses of a nucleus (${}_Z^A X$), proton and neutron respectively , then the mass defect is $\Delta m = \left(Zm_p + Nm_n\right) - M$
- The binding energy of nucleus $B.E = \left(Zm_p + Nm_n - M\right)c^2$
- The binding energy per nucleon is maximum for iron which is 8.8 MeV.

- Alpha decay: ${}_{Z}^{A}X \rightarrow {}_{Z-2}^{A-4}Y + {}_{2}^{4}He$
- β^- decay: ${}_{Z}^{A}X \rightarrow {}_{Z+1}^{A}Y + e^- + \overline{\nu}$
- β^+ decay: ${}_{Z}^{A}X \rightarrow {}_{Z-1}^{A}Y + e^+ + \nu$
- Gamma decay: ${}_{Z}^{A}X^* \rightarrow {}_{Z}^{A}X + \gamma$
- Law of radioactive decay: $N = N_0 e^{-\lambda t}$
- In general, after n half lives, the number of nuclei undecayed is $N = \left(\frac{1}{2}\right)^n N_0$
- The relation between half-life and decay constant $T_{1/2} = \frac{\ln 2}{\lambda}$
- If a heavier nucleus decays into lighter nuclei, it is called nuclear fission
- If two lighter nuclei fuse to heavier nuclei, it is called nuclear fusion
- In nuclear reactors, the nuclear chain reaction is controlled. In stars, the energy generation is through nuclear fusion.

CONCEPT MAP


Atomic and Nuclear Physics
Atom models
J.J. Thomson
α - scattering experiment
Rutherford
Bohr atom model
Hydrogen spectrum
Nucleus
Properties
Binding energy
Radioactive decays
Nuclear fusion & fission
Elementary particles

EVALUATION

I Multiple Choice Questions

1. Suppose an alpha particle accelerated by a potential of V volt is allowed to collide with a nucleus whose atomic number is Z, then the distance of closest approach of alpha particle to the nucleus is

(a) $14.4\frac{Z}{V}$ Å (b) $14.4\frac{V}{Z}$ Å

(c) $1.44\frac{Z}{V}$ Å (d) $1.44\frac{V}{Z}$ Å

2. In a hydrogen atom, the electron revolving in the fourth orbit, has angular momentum equal to

(a) h (b) $\frac{h}{\pi}$

(c) $\frac{4h}{\pi}$ (d) $\frac{2h}{\pi}$

U7J6Z1

3. Atomic number of H-like atom with ionization potential 122.4 V for $n = 1$ is

(a) 1 (b) 2 (c) 3 (d) 4

4. The ratio between the first three orbits of hydrogen atom is

(a) 1:2:3 (b) 2:4:6

(c) 1:4:9 (d) 1:3:5

5. The charge of cathode rays is

(a) positive (b) negative

(c) neutral (d) not defined

6. In J.J. Thomson e/m experiment, a beam of electron is replaced by that of muons (particle with same charge as that of electrons but mass 208 times that of electrons). No deflection condition is achieved only if

(a) B is increased by 208 times

(b) B is decreased by 208 times

(c) B is increased by 14.4 times

(d) B is decreased by 14.4 times

7. The ratio of the wavelengths for the transition from $n = 2$ to $n = 1$ in Li^{++}, He^{+} and H is

(a)1: 2: 3 (b) 1: 4: 9

(c) 3:2:1 (d) 4: 9: 36

8. The electric potential between a proton and an electron is given by $V = V_0 \ln\left(\frac{r}{r_0}\right)$, where r_0 is a constant. Assume that Bohr atom model is applicable to potential, then variation of radius of n^{th} orbit r_n with the principal quantum number n is

(a) $r_n \propto \frac{1}{n}$ (b) $r_n \propto n$

(c) $r_n \propto \frac{1}{n^2}$ (d) $r_n \propto n^2$

9. If the nuclear radius of ^{27}Al is 3.6 fermi, the approximate nuclear radius of ^{64}Cu is

(a) 2.4 (b) 1.2

(c) 4.8 (d)3.6

10. The nucleus is approximately spherical in shape. Then the surface area of nucleus having mass number A varies as

(a) $A^{2/3}$ (b) $A^{4/3}$

(c) $A^{1/3}$ (d) $A^{5/3}$

11. The mass of a $_3^7Li$ nucleus is 0.042 u less than the sum of the masses of all its nucleons. The binding energy per nucleon of $_3^7Li$ nucleus is nearly

(a) 46 MeV (b) 5.6 MeV

(c) 3.9MeV (d)23 MeV

12. M_p denotes the mass of the proton and M_n denotes mass of a neutron. A given nucleus of binding energy B, contains Z protons and N neutrons. The mass M(N,Z) of the nucleus is given by(where c is the speed of light)
   (a) $M(N,Z) = NM_n + ZM_p - Bc^2$
   (b) $M(N,Z) = NM_n + ZM_p + Bc^2$
   (c) $M(N,Z) = NM_n + ZM_p - B/c^2$
   (d) $M(N,Z) = NM_n + ZM_p + B/c^2$

13. A radioactive nucleus (initial mass number A and atomic number Z) emits 2α and 2 positrons. The ratio of number of neutrons to that of proton in the final nucleus will be
   (a) $\frac{A-Z-4}{Z-2}$ (b) $\frac{A-Z-2}{Z-6}$
   (c) $\frac{A-Z-4}{Z-6}$ (d) $\frac{A-Z-12}{Z-4}$

14. The half-life period of a radioactive element A is same as the mean life time of another radioactive element B. Initially both have the same number of atoms. Then
   (a) A and B have the same decay rate initially
   (b) A and B decay at the same rate always
   (c) B will decay at faster rate than A
   (d) A will decay at faster rate than B.

15. A system consists of N_0 nucleus at t=0. The number of nuclei remaining after half of a half-life (that is, at time $t = \frac{1}{2}T_{\frac{1}{2}}$)
   (a) $\frac{N_0}{2}$ (b) $\frac{N_0}{\sqrt{2}}$
   (c) $\frac{N_0}{4}$ (d) $\frac{N_0}{8}$

Answers

1) c	2) d	3) c	4) c	5) b
6) c	7) d	8) b	9) c	10) A
11) b	12) c	13) b	14) c	15) b

II Short answer questions

1. What are cathode rays?.
2. Write the properties of cathode rays.
3. Give the results of Rutherford alpha scattering experiment.
4. Write down the postulates of Bohr atom model.
5. What is meant by excitation energy.
6. Define the ionization energy and ionization potential.
7. Write down the draw backs of Bohr atom model.
8. What is distance of closest approach?
9. Define impact parameter.
10. Write a general notation of nucleus of element X. What each term denotes?
11. What is isotope? Give an example.
12. What is isotone? Give an example.
13. What is isobar? Give an example.
14. Define atomic mass unit *u*.
15. Show that nuclear density is almost constant for nuclei with $Z > 10$.
16. What is mass defect?
17. What is binding energy of a nucleus? Give its expression.
18. Calculate the energy equivalent of 1 atomic mass unit.
19. Give the physical meaning of binding energy per nucleon.
20. What is meant by radioactivity?

21. Give the symbolic representation of alpha decay, beta decay and gamma decay.
22. In alpha decay, why the unstable nucleus emits $^{4}_{2}He$ nucleus? Why it does not emit four separate nucleons?
23. What is mean life of nucleus? Give the expression.
24. What is half-life of nucleus? Give the expression.
25. What is meant by activity or decay rate? Give its unit.
26. Define curie.
27. What are the constituent particles of neutron and proton?

III Long answer questions

1. Explain the J.J. Thomson experiment to determine the specific charge of electron.
2. Discuss the Millikan's oil drop experiment to determine the charge of an electron.
3. Derive the energy expression for hydrogen atom using Bohr atom model.
4. Discuss the spectral series of hydrogen atom.
5. Explain the variation of average binding energy with the mass number by graph and discuss its features.
6. Explain in detail the nuclear force.
7. Discuss the alpha decay process with example.
8. Discuss the beta decay process with examples.
9. Discuss the gamma decay process with example.
10. Obtain the law of radioactivity.
11. Discuss the properties of neutrino and its role in beta decay.
12. Explain the idea of carbon dating.
13. Discuss the process of nuclear fission and its properties.
14. Discuss the process of nuclear fusion and how energy is generated in stars?
15. Describe the working of nuclear reactor with a block diagram.
16. Explain in detail the four fundamental forces.
17. Briefly explain the elementary particles of nature.

Exercises

1. Consider two hydrogen atoms H_A and H_B in ground state. Assume that hydrogen atom H_A is at rest and hydrogen atom H_B is moving with a speed and make head-on collide on the stationary hydrogen atom H_A. After the strike, both of them move together. What is minimum value of the kinetic energy of the moving hydrogen atom H_B, such that any one of the hydrogen atoms reaches one of the excitation state.

   [Ans: 20.4 eV]

2. In the Bohr atom model, the frequency of transitions is given by the following expression

   $v = Rc\left(\frac{1}{n^2} - \frac{1}{m^2}\right)$, where $n < m$, Consider the following transitions:

Transitions	$m \to n$
1	$3 \to 2$
2	$2 \to 1$
3	$3 \to 1$

Show that the frequency of these transitions obey sum rule (which is known as Ritz combination principle)

[Ans: $v_{3\to2} + v_{2\to1} = v_{3\to1}$]

3. (a) A hydrogen atom is excited by radiation of wavelength 97.5 nm. Find the principal quantum number of the excited state.

(b) Show that the total number of lines in emission spectrum is $\frac{n(n-1)}{2}$ and compute the total number of possible lines in emission spectrum.

[Ans: (a) n =4 (b) 6 possible transitions]

4. Calculate the radius of the earth if the density of the earth is equal to the density of the nucleus.[mass of earth $5.97\times10^{24}\,kg$].

[Ans: $180\,\text{m}$]

5. Calculate the mass defect and the binding energy per nucleon of the $^{108}_{47}Ag$ nucleus. [atomic mass of Ag = 107.905949]

Ans: $\begin{bmatrix} \Delta m = 0.990391u \text{ and} \\ \overline{B.E} = 8.5 MeV / A \end{bmatrix}$

6. Half lives of two radioactive elements A and B are 20 minutes and 40 minutes respectively. Initially, the samples have equal number of nuclei. Calculate the ratio of decayed numbers of A and B nuclei after 80 minutes.

[Ans: 5:4]

7. On your birthday, you measure the activity of the sample ^{210}Bi which has a half-life of 5.01 days. The initial activity that you measure is $1\mu Ci$. (a) What is the approximate activity of the sample on your next birthday? Calculate (b) the decay constant (c) the mean life (d) initial number of atoms.

[Ans: (a) $10^{-22}\mu Ci$ (b) $1.6\times10^{-6}\,s^{-1}$

(c) 7.24days (d) 2.31×10^{10}]

8. Calculate the time required for 60% of a sample of radon undergo decay. Given $T_{1/2}$ of radon =3.8 days

[Ans: 5.022 days]

9. Assuming that energy released by the fission of a single $^{235}_{92}U$ nucleus is 200MeV, calculate the number of fissions per second required to produce 1 watt power.

[Ans: 3.125×10^{10}]

10. Show that the mass of radium ($^{226}_{88}Ra$) with an activity of 1 curie is almost a gram. Given $T_{1/2}$ =1600 years.

11. Characol pieces of tree is found from an archeological site. The carbon-14 content of this characol is only 17.5% that of equivalent sample of carbon from a living tree. What is the age of tree?

[Ans: 1.44×10^4 yr]

BOOKS FOR REFERENCE

1. Introduction to Modern Physics, H.S. Mani and G.K. Mehta, East-West Press, New Delhi
2. Concepts of Modern Physics, Arthur Beiser, McGraw Hill, 6^{th} edition
3. Concepts of Physics – H. C. Verma, Volume 2, Bharati Bhawan Publisher
4. Fundamentals of Physics, Halliday, Resnick and Walker, Wiley Publishers, 10th edition
5. Physics for scientist and engineers with modern physics, Serway and Jewett, Brook/Coole publishers, 8^{th} edition
6. Physics for scientist and engineers with modern physics, Paul Tipler and Gene Mosca, Sixth edition, W.H.Freeman and Company

ICT CORNER

Atomic and Nuclear physics

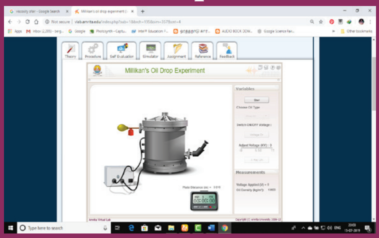
In this activity you will be able to(i) experimentally demonstrate the concept of Millikan's oil drop experiment (ii) find the terminal velocity of the drop and (iii) find the charge on a drop.

Topic: Millikan's oil drop experiment

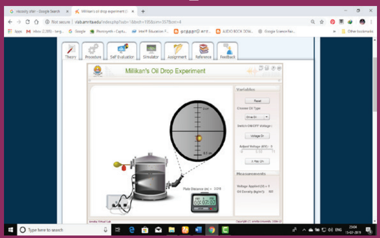
STEPS:

- Open the browser and type"vlab.amrita.edu" in the address bar.
- Click 'Physical Sciences'tab. Thenclick 'Modern Physics Virtual Lab' and then click 'Millikan's oil drop experiment'. Go to "simulator" tab to do the experiment.
- Click on 'START' button. Click on Combo box to choose the oil.
- Click 'START' button of stop watch and notice the time taken t_1 by a drop, to travel distance l_1 between any two points. Calculate the terminal velocity $\upsilon_1 = \dfrac{l_1}{t_1}$
- Click 'Voltage On' to suspend the same oil drop in air, which is the balancing voltage V.
- Click the 'X Ray ON' button and notice the time taken t_2 by same drop to travel distance l_2 between any two points. Calculate the terminal velocity $\upsilon_2 = \dfrac{l_2}{t_2}$
- Charge of drop is calculated using the equation q = $\dfrac{6\pi\eta r(\upsilon_1 + \upsilon_2)d}{V}$. r-radius of oil drop (can be measure using telescope), η-viscosity of air (1.81×10^{-5} $kgm^{-1}s^{-1}$), d is the distance between the plates.

Step1



Step2



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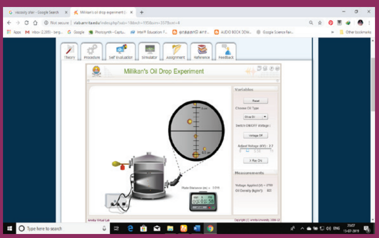
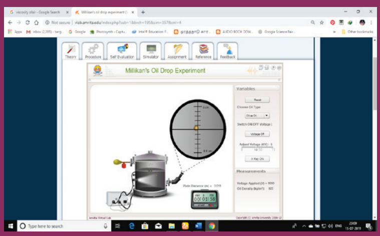


Step4



Note:

1. One time sign up is needed to do simulation. Then login using that username and password.
2. Install flash player in your browserif it is not there.

URL:

http://vlab.amrita.edu/index.php?sub=1&brch=195&sim=357&cnt=4

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* Pictures are indicative only.

* If browser requires, allow **Flash Player** or **Java Script** to load the page.

UNIT 9

SEMICONDUCTOR ELECTRONICS

Electronics is clearly the winner of the day
– John Ford.

Learning Objectives

In this unit, the students are exposed to,

- Energy band diagram in semiconductors
- Types of semiconductors
- Formation of p-n junction diode and its V-I characteristics
- Rectification process
- Special purpose diodes
- Transistors and their immediate applications
- Digital and analog signals
- Logic gates
- Boolean algebra
- De Morgan's theorem

U9M9D7

9.1 INTRODUCTION

Electronics has become a part of our daily life. All gadgets like mobile phones, computers, televisions, music systems etc work on the electronic principles. Electronic circuits are used to perform various operations in devices like air conditioners, microwave oven, dish washers and washing machines. Besides this, its applications are widespread in all fields like communication systems, medical diagnosis and treatments and even handling money through ATMs.

Evolution of Electronics:

The history of electronics began with the invention of vacuum diode by J.A. Fleming in 1897. This was followed by a vacuum triode implemented by Lee De Forest to control electrical signals. This led to the introduction of tetrode and pentode tubes.

Subsequently, the transistor era began with the invention of bipolar junction transistor by Bardeen, Brattain and Shockley in 1948 for which Nobel prize was awarded in 1956. The emergence of Germanium and Silicon semiconductor materials made this transistor gain popularity, in turn its application in different electronic circuits.

The following years witnessed the invention of the integrated circuits (ICs) that helped to integrate the entire electronic circuit on a single chip which is small in size and cost-effective. Since 1958 ICs capable

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of holding several thousand electronic components on a single chip such as small-scale, medium-scale, large-scale, and very-large scale integration started coming into existence. Digital integrated circuits became another robust IC development that enhanced the architecture of computers.All these radical changes led to the introduction of microprocessor in 1969 by Intel.

The electronics revolution, in due course of time, accelerated the computer revolution. Now the world is on its way towards small particles of nano-size, far too small to see. This helps in the miniaturization to an unimaginable size. A room-size computer during its invention has now emerged as a laptop, palmtop, iPad, etc. In the recent past, IBM revealed the smallest computer whose size is comparable with the tip of the rice grain, measuring just 0.33 mm on each side.

Electronics is the branch of physics incorporated with technology towards the design of circuits using transistors and microchips. It depicts the behaviour and movement of electrons and holes in a semiconductor, electrons and ions in vacuum, or gas. Electronics deals with electrical circuits that involve active components such as transistors, diodes, integrated circuits, and sensors, associated with the passive components like resistors, inductors, capacitors, and transformers.

This chapter deals with semiconductor devices like p-n junction diodes, bipolar junction transistors and logic circuits.

Note

Passive Components: components that cannot generate power in a circuit.

Active components: components that can generate power in a circuit.

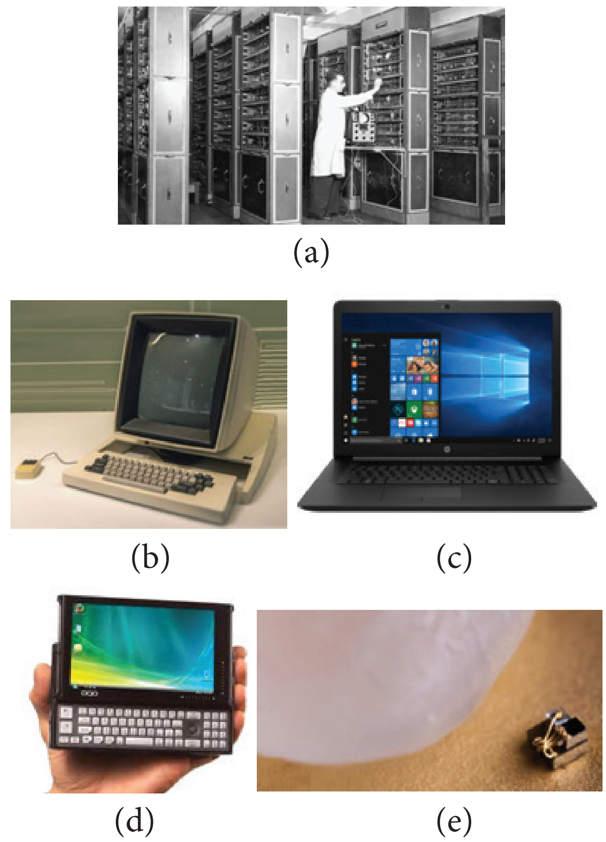


Figure: 9.1 Evolution of computers
(a) one of the world's first computers
(b) desktop computer (c) laptop computer
(d) palmtop computer (e) Thinnest computer revealed by IBM

DO YOU KNOW?

The world's first computer 'ENIAC' was invented by J. Presper Eckert and John Mauchly at the University of Pennsylvania. The construction work started in 1943 and got over in 1946. It occupied an area of around 1800 square feet. It had 18,000 vacuum tubes and it weighed around 50 tons.

9.1.1 Energy band diagram of solids

In an isolated atom, the electronic energy levels are widely separated and are far apart and the energy of the electron is decided by the orbit in which it revolves around the nucleus. However, in the case of a solid, the atoms are closely spaced and hence the electrons in the outermost energy levels of nearby atoms influence each other. This changes the nature of the electron motion in a solid from that of an isolated atom to a large extent.

The valence electrons in an atom are responsible for the bonding nature. Let us consider an atom with one electron in the outermost orbit. It means that the number of valence electrons is one. When two such atoms are brought close to each other, the valence orbitals are split into two. Similarly the unoccupied orbitals of each atom will also split into two. The electrons have the choice of choosing any one of the orbitals as the energy of both the orbitals is the same. When the third atom of the same element is brought to this system, the valence orbitals of all the three atoms are split into three. The unoccupied orbitals also will split into three.

In reality, a solid is made up of millions of atoms. When millions of atoms are brought close to each other, the valence orbitals and the unoccupied orbitals are split according to the number of atoms. In this case, the energy levels will be closely spaced and will be difficult to differentiate the orbitals of one atom from the other and they look like a band as shown in Figure 9.2. **This band of very large number of closely spaced energy levels in a very small energy range is known as energy band.**

The energy band formed due to the valence orbitals is called valence band and that formed due to the unoccupied orbitals to which electrons can jump when energised is called the conduction band. The energy gap between the valence band and the conduction band is called forbidden energy gap. Electrons cannot exist in the forbidden energy gap.

The representation of the valence band and conduction band is shown in Figure 9.2(a). E_V represents the maximum energy of the valence band and E_C represents minimum energy of the conduction band. The forbidden energy gap, $E_g = E_C - E_V$. The kinetic energy of the electron increases from bottom to top (near the nucleus to the farthest) and the potential energy decreases indicating that the electrons in the orbitals closer to the nucleus are bound with large potential energy. Hence, the electrons closer to nucleus require a lot of energy to be excited. The electrons in the valence band are less bound to the nucleus and can be easily excited.

Note The energy levels of the orbiting electrons are measured in electron volts, (eV).

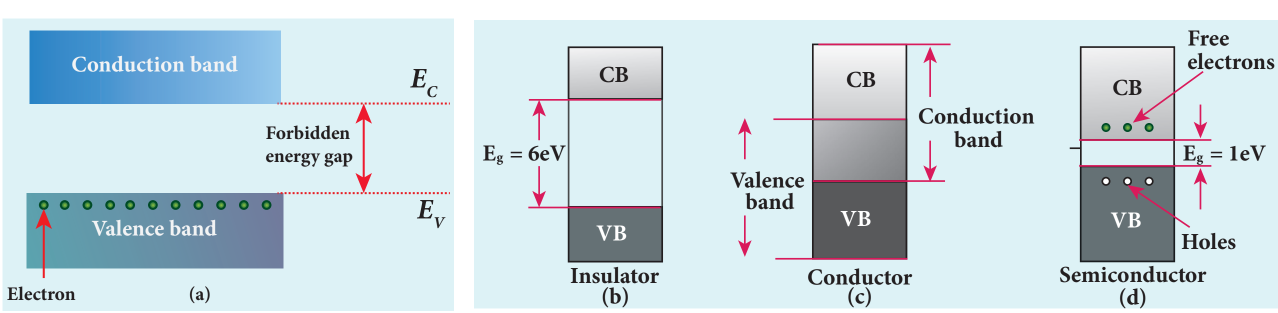


Figure: 9.2 (a) Schematic representation of valence band, conduction band, and forbidden energy gap. Energy band structure of (b) Insulator (c) Conductor (d) Semiconductor

9.1.2 Classification of materials

The classification of solids into insulators, metals, and semiconductors can be explained with the help of the energy band diagram.

9.1.2.1 Insulators

The energy band structure of insulators is shown in Figure 9.2(b). The valence band and the conduction band are separated by a large energy gap. The forbidden energy gap is approximately 6 eV in insulators. The gap is very large that electrons from valence band cannot move into conduction band even on the application of strong external electric field or the increase in temperature. Therefore, the electrical conduction is not possible as the free electrons are almost nil and hence these materials are called insulators. Its resistivity is in the range of 10^{11}–$10^{19}\,\Omega$m.

9.1.2.2 Conductors

In condutors, the valence band and conduction band overlap as shown in Figure 9.2(c). Hence, electrons can move freely into the conduction band which results in a large number of free electrons in the conduction band. Therefore, conduction becomes possible even at low temperatures. The application of electric field provides sufficient energy to the electrons to drift in a particular direction to constitute a current. For condutors, the resistivity value lies between 10^{-2} and $10^{-8}\,\Omega$m.

9.1.2.3 Semiconductors

In semiconductors, there exists a narrow forbidden energy gap $\left(E_g < 3\,eV\right)$ between the valence band and the conduction band. At a finite temperature, thermal agitations in the solid can break the covalent bond between the atoms (covalent bond is formed due to the sharing of electrons to attain stable electronic configuration). This releases some electrons from valence band to conduction band. Since free electrons are small in number, the conductivity of the semiconductors is not as high as that of the conductors. The resistivity value of semiconductors is from 10^{-5} to $10^{6}\,\Omega$m.

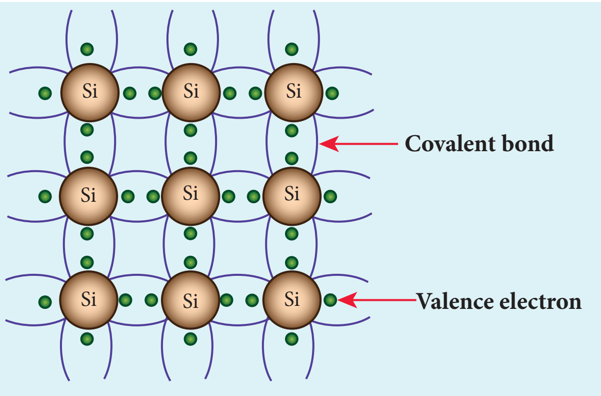
Note In semiconductors, electrons in the valence band are bound electrons which cannot move. Hence, they cannot contribute for conduction.

When the temperature is increased further, more number of electrons is promoted to the conduction band and increases the conduction. Thus, we can say that the electrical conduction increases with the increase in temperature. In other words, resistance decreases with increase in temperature. Hence, semiconductors are said to have negative temperature coefficient of resistance. The most important elemental semiconductor materials are Silicon (Si) and Germanium (Ge). The forbidden energy gaps for Si and Ge at at room temperature are 1.1 eV and 0.7 eV respectively.

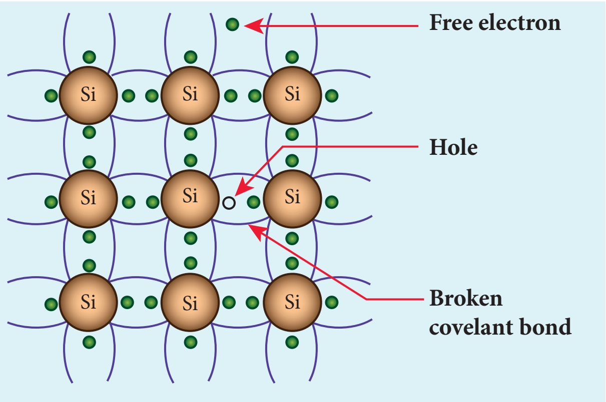
9.2 TYPES OF SEMICONDUCTRORS

9.2.1 Intrinsic semiconductors

A semiconductor in its pure form without impurity is called an intrinsic semiconductor. Here, impurity means any



(a)



(b)

Figure: 9.3 (a) Two dimensional crystal lattice of Silicon (b) The presence of free electron, hole, and broken covalent bond in the intrinsic Silicon crystal.

other atom in the crystal lattice. The Silicon lattice is shown in Figure 9.3(a). Each Silicon atom has four electrons in the outermost orbit and is covalently bonded with the neighbouring atoms to form the lattice. The band diagram for this case is show in Figure 9.4(a).

A small increase in temperature is sufficient enough to break some of the covalent bonds and release the electrons free from the lattice as shown in Figure 9.3(b). As a result, some states in the valence band become empty and the same number of states in the conduction band will be occupied as shown in Figure 9.4(b). The vacancies produced in the valence band are called holes. As the holes are deficiency of electrons, they are treated to possess positive charges. Hence, electrons and holes are the two charge carriers in semiconductors.

In intrinsic semiconductors, the number of electrons in the conduction band is equal to the number of holes in the valence band. The conduction is due to the electrons in the

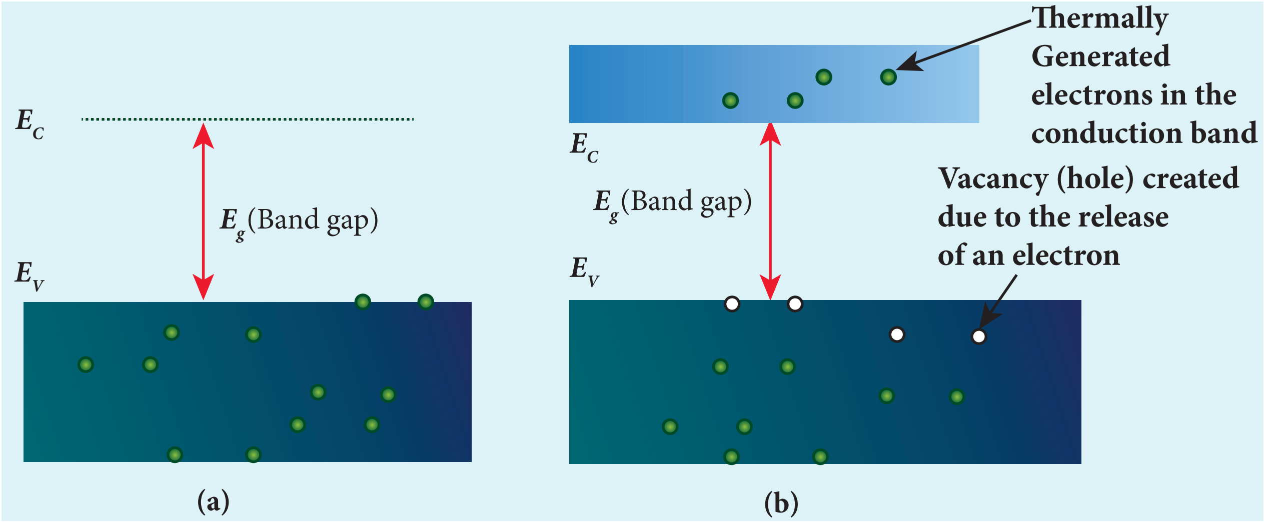


Figure: 9.4 (a) Valence band and conduction band of intrinsic semiconductor. (b) Presence of thermally generated electrons in the conduction band and vacancy created due to the shift of electron from valence band to conduction band at room temperature.

conduction band and holes in the valence band. These currents are represented as I_e and I_h respectively.

Note Definition of a hole: When an electron is excited, covalent bond is broken. Now octet rule will not be satisfied. Thus each excited electron leaves a vacancy to complete bonding. This 'deficiency' of electron is termed as a 'hole'

The total current (I) is always the sum of the electron current (I_e) and the hole current (I_h). $I = I_e + I_h$

An intrinsic semiconductor behaves like an insulator at 0 K. The increase in temperature increases the number of charge carriers (electrons and holes). The schematic diagram of the intrinsic semiconductor in band diagram is shown in Figure 9.4(b). The intrinsic carrier concentration is the number electron in the conduction band or the number of holes in the valence band in an intrinsic semiconductor.

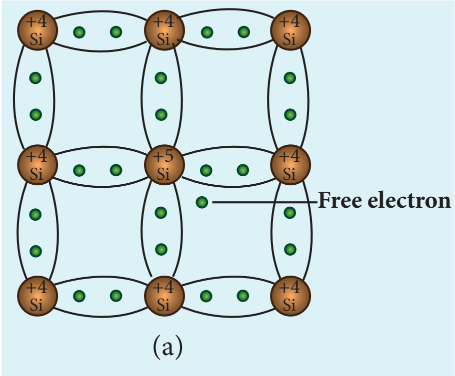
9.2.2 Extrinsic semiconductors

The carrier concentration in an intrinsic semiconductor is not sufficient enough to develop efficient electronic devices. Another way of increasing the carrier concentration in an intrinsic semiconductor is by adding impurity atoms. **The process of adding impurities to the intrinsic semiconductor is called doping.** It increases the concentration of charge carriers (electrons and holes) in the semiconductor and in turn, its electrical conductivity. The impurity atoms are called dopants and its order is approximately 100 ppm (parts per million).

9.2.2.1 n-type semiconductor

A n-type semiconductor is obtained by doping a pure Germanium (or Silicon) crystal with a dopant from group V pentavalent elements like Phosphorus, Arsenic, and Antimony as shown in Figure 9.5(a). The dopant has five valence electrons while the Germanium atom has four valence electrons. During the process of doping, a few of the Germanium atoms are replaced by the group V dopants. Four of the five valence electrons of the impurity atom are bound with the 4 valence electrons of the neighbouring replaced Germanium atom. The fifth valence electron of the impurity atom will be loosely attached with the nucleus as it has not formed the covalent bond.

The energy level of the loosely attached fifth electron from the dopant is found just below the conduction band edge and



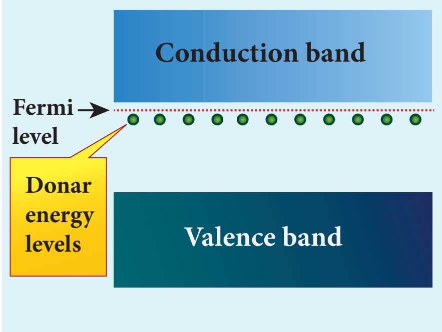


Figure 9.5 n-type extrinsic semiconductor: (a) Free electron which is loosely attached to the lattice (b) Representation of donar energy level.

is called the donor energy level as shown in Figure9.5(b). At room temperature, these electrons can easily move to the conduction band with the absorption of thermal energy. It is shown in the Figure 9.6. Besides, an external electric field also can set free the loosely bound electrons and lead to conduction.

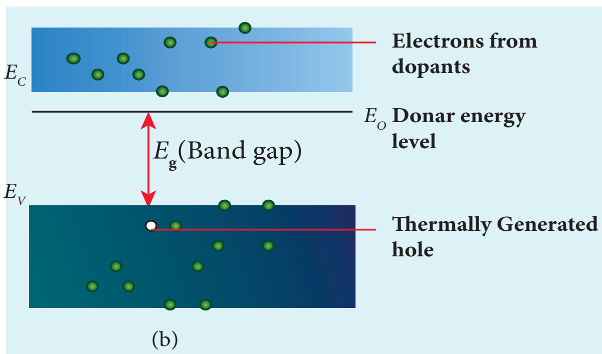


Figure 9.6 Thermally generated hole in the valence band and the free electrons generated by the dopants in the conduction band (n-type semiconductor).

It is important to note that the energy required for an electron to jump from the valence band to the conduction band (E_g) in an intrinsic semiconductor is 0.7 eV for Ge and 1.1 eV for Si, while the energy required to set free a donor electron is only 0.01 eV for Ge and 0.05 eV for Si.

The group V pentavalent impurity atoms donate electrons to the conduction band and are called donor impurities. Therefore, each impurity atom provides one extra electron to the conduction band in addition to the thermally generated electrons. These thermally generated electrons leave holes in valence band. Hence, the majority carriers of current in an n-type semiconductor are electrons and the minority carriers are holes. Such a semiconductor doped with a pentavalent impurity is called an n-type semiconductor.

9.2.2.2 p-type semiconductor

Here, a trivalent atom from group III elements such as Boron, Aluminium, Gallium and Indium is added to the Germanium or Silicon substrate. The dopant with three valence electrons are bound with the neighbouring Germanium atom as shown in Figure 9.7(a). As Germanium atom has four valence electrons, one electron position of the dopant in the Germanium crystal lattice will remain vacant. The missing electron position in the covalent bond is denoted as a hole.

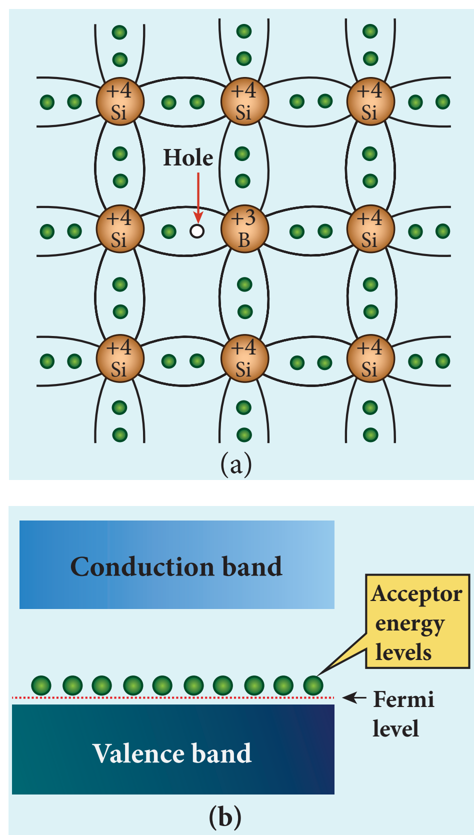


Figure 9.7 P-type extrinsic semiconductor
(a) Hole generated by the dopant
(b) Representation of acceptor energy level.

To make complete covalent bonding with all four neighbouring atoms, the dopant is in need of one more electron. These dopants can accept electrons from the neighbouring atoms. Therefore, this impurity is called an acceptor impurity. The energy level of the hole created by each

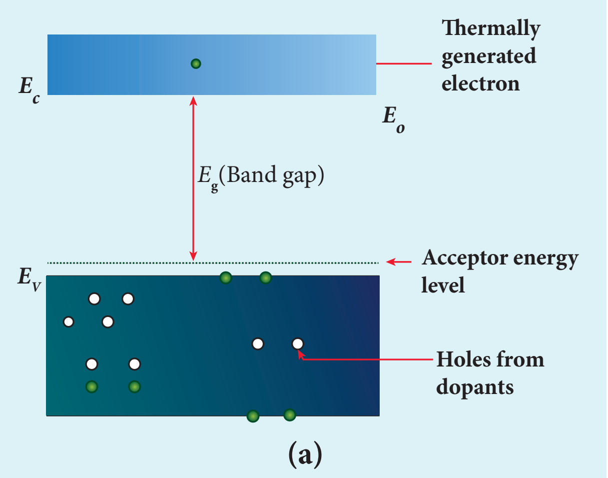


Figure 9.8 Thermally generated electron in the conduction band and the holes generated by the dopants in the valence band (p-type semiconductor).

impurity atom is just above the valence band and is called the acceptor energy level, as shown in Figure 9.7(b).

For each acceptor atom, there will be a hole in the valence band in addition to the thermally generated holes. In such an extrinsic semiconductor, holes are the majority carriers and thermally generated electrons are minority carriers as shown in Figure 9.8. The semiconductor thus formed is called a p-type semiconductor.

> **DO YOU KNOW?** The n-type and p-type semiconductor are neutral as we are topping neutral atoms to the intrinsic semiconductors.

9.3 DIODES

9.3.1 P-N Junction formation

9.3.1.1 Formation of depletion layer

A *p-n* junction is formed by joining *n*-type and *p*-type semiconductor materials as shown in Figure 9.9(a). Since the *n*-region has a high electron concentration and the *p*-region a high hole concentration, electrons diffuse from the *n*-side to the *p*-side. This causes *diffusion current* which exists due to the concentration difference of electrons. The electrons diffusing into the p-region may occupy holes in that region and make it negative. The holes left behind by these electrons in the *n*-side are equivalent to the diffusion of holes from the *p*-side to the *n*-side. If the electrons and holes were not charged, this diffusion process would continue until the concentration of electrons

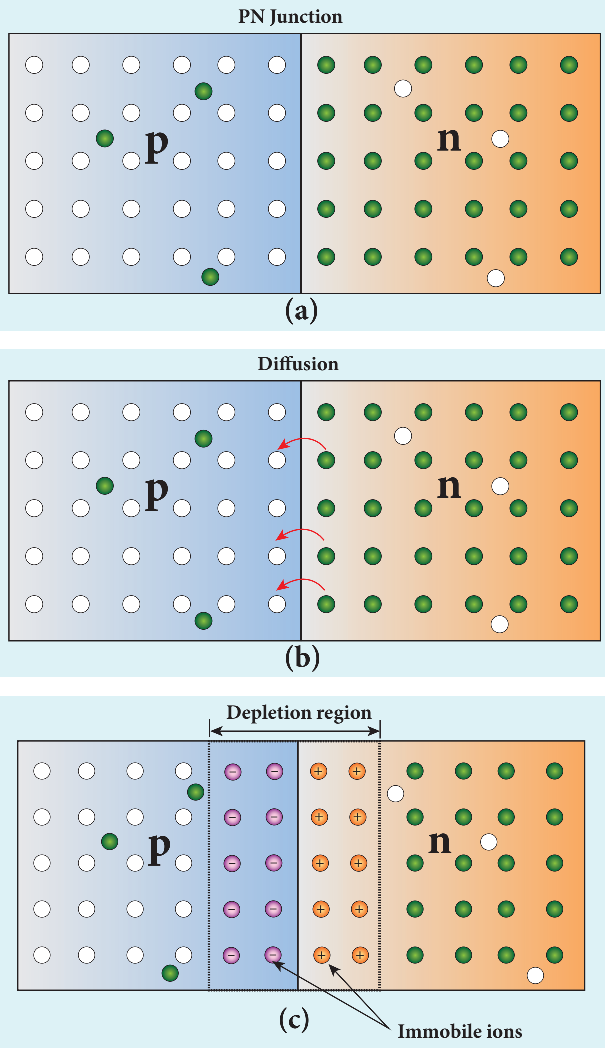


Figure: 9.9 (a) P-N junction (b) Diffusion of electrons across the junction (c) Presence of immobile ions in the depletion region and holes on the two sides were the same, as happens if two gasses come into contact with each other.

But, in a p-n junction, when the electrons and holes move to the other side of the junction, they leave behind exposed charges on dopant atom sites, which are fixed in the crystal lattice and are unable to move. On the *n*-side, positive ion cores are exposed and on the *p*- side, negative ion cores are exposed as shown in Figure 9.9(b). An electric field E forms between the positive ion cores in the *n*-type material and negative ion cores in the *p*-type material. The electric field sweeps free carriers out of this region and hence it is called depletion region as it is depleted of free carriers. A barrier potential V_b due to the electric field E is formed at the junction as shown in Figure 9.9(c).

As this diffusion of charge carriers from both sides continues, the negative ions form a layer of negative space charge region along the p-side. Similarly, a positive space charge region is formed by positive ions on the n-side. The positive space charge region attracts electrons from p-side to n-side and the negative space charge region attracts holes from n-side to p-side. This moment of carriers happen in this region due to the formed electric field and it constitutes a current called drift current. The diffusion current and drift current flow in the opposite direction and at one instant they both become equal. Thus, a p-n junction is formed.

9.3.1.2 Junction potential or barrier potential

The recombination of charge carriers takes place only to a certain point beyond which the depletion layer acts like a barrier to further diffusion of free charges across the junction. This is due to the fact that the immobile ions on both sides establish an electric potential difference across the junction. Therefore, an electron trying to diffuse into the interior of the depletion region encounters a negative wall of ions repelling it backwards. If the free electron has enough energy, it can break through the wall and enter into the p-region, where it can recombine with a hole and create another negative ion.

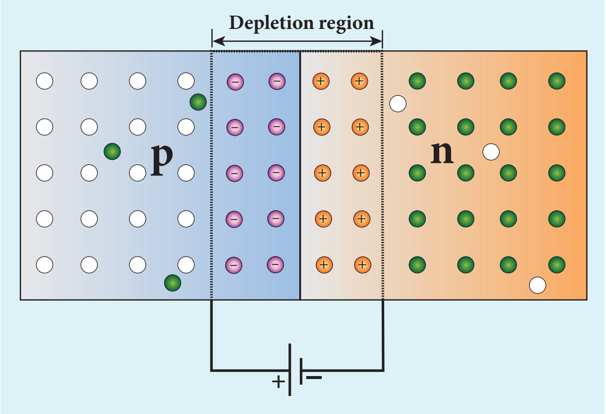


Figure 9.10 Barrier potential formed across the junction

The strength of the electric potential difference across the depletion region keeps increasing with the crossing of each electron until equilibrium is reached; at this point, the internal repulsion of the depletion layer stops further diffusion of free electrons across the junction. This difference in potential across the depletion layer is called the **barrier potential** as shown in Figure 9.10. At 25 ^{0}C, this barrier potential approximately equals 0.7 V for Silicon and 0.3 V for Germanium.

9.3.2 P-N Junction diode

A p-n junction diode is formed when a p-type semiconductor is fused with an-type semiconductor. It is a device with single p-n junction as shown in Figure 9.11(a). The circuit symbol is shown in Figure 9.11(b).

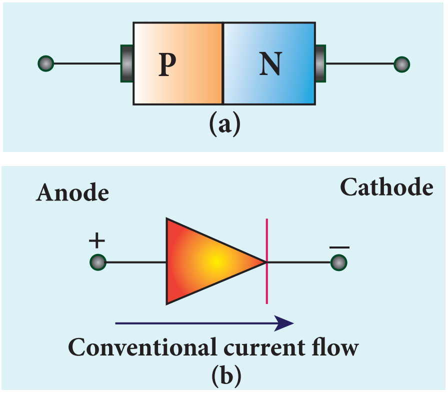


Figure 9.11 p-n junction diode
(a) Schematic representation
(b) Circuit symbol

9.3.2.1 Biasing a diode

Biasing means providing external energy to charge carriers to overcome the barrier potential and make them move in a particular direction. The charge carriers can either move towards the junction or away from the junction. **The external voltage applied to the p-n junction is called bias voltage**. Depending on the polarity of the external source to the p-n junction we have two types of biasing

1. Forward bias
2. Reverse bias

Forward Bias

If the positive terminal of the external voltage source is connected to the p-side and the negative terminal to the n-side, it is called forward biased as shown in Figure 9.12. The application of a forward bias potential makes the electrons move into the n-side and the holes into the p-side. This initiates the recombination with the ions near the junction which in turn reduces the width of the depletion region and hence the barrier potential.

The electron from the n-side is now accelerated towards the p-side as it experiences a reduced barrier potential at the junction. In addition, the accelerated electrons experience a strong attraction by the positive potential applied to the p-side. This results in the movement of electrons towards the p-side and in turn, holes towards the n-side. When the applied voltage is increased, the width of the depletion region and hence the barrier potential are further reduced. This results in a large number of electrons passing through the junction resulting in an exponential rise in current through the junction.

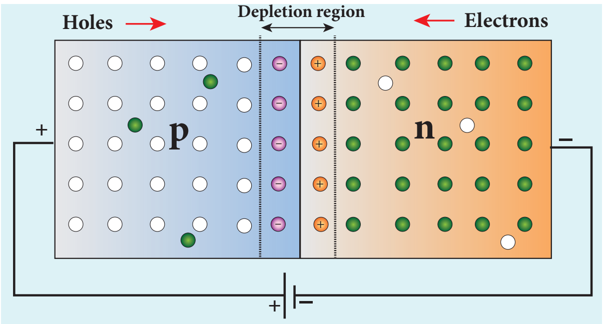


Figure 9.12 Schematic representation of a p-n junction diode under forward bias

Reverse Bias

If the positive terminal of the battery is connected to the n-side and the negative potential to the p-side, the junction is said to be reverse biased as shown in Figure 9.13.

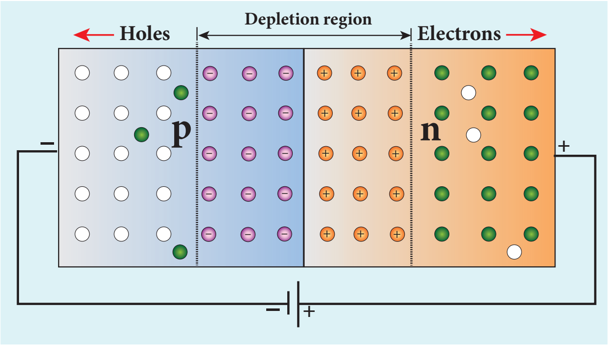


Figure 9.13 Schematic representation of a p-n junction diode under reverse bias

As the positive potential is connected to the n-type material, the electrons in the n-type material are attracted towards the positive terminal in turn, the holes in the p-type material move towards the negative terminal (both away from the junction). It increases the immobile ions at the junction. The net effect is the widening of the depletion region. This leads to an increase in the barrier potential. Consequently, the majority charge carriers from both sides experience a great barrier to cross the junction. This reduces the diffusion current across the junction effectively.

Yet, a small current flows across the junction due to the minority charge carriers in both regions. The reverse bias for majority charge carriers serves as the forward bias for minority charge carriers. The current that flows under a reverse bias is called the reverse saturation current**.** It is represented as I_s.

The reverse saturation current is independent of the applied voltage and it depends only on the thermally generated minority charge carriers. Even a small voltage is sufficient enough to drive the minority charge carriers across the junction.

Note The reverse saturation current of a silicon diode doubles for every 10°C rise in temperature.

9.3.3 Characteristics of a junction diode

9.3.3.1 Forward characteristics

It is the study of the variation in current through the diode with respect to the applied voltage across the diode when it is forward biased.

The p-n junction diode is forward biased as shown in Figure 9.14(a). An external resistance (R) is used to limit the flow of current through the diode. The voltage across the diode is varied by varying the biasing voltage across the dc power supply. The forward bias voltage and the corresponding forward bias current are noted. A graph is plotted by taking the forward bias voltage (V) along the x-axis and the current (I) through the diode along the y-axis. This graph is called the **forward V-I characteristics** of the p-n junction diode and is shown in Figure 9.14(b). Three inferences can be brought out from the graph:

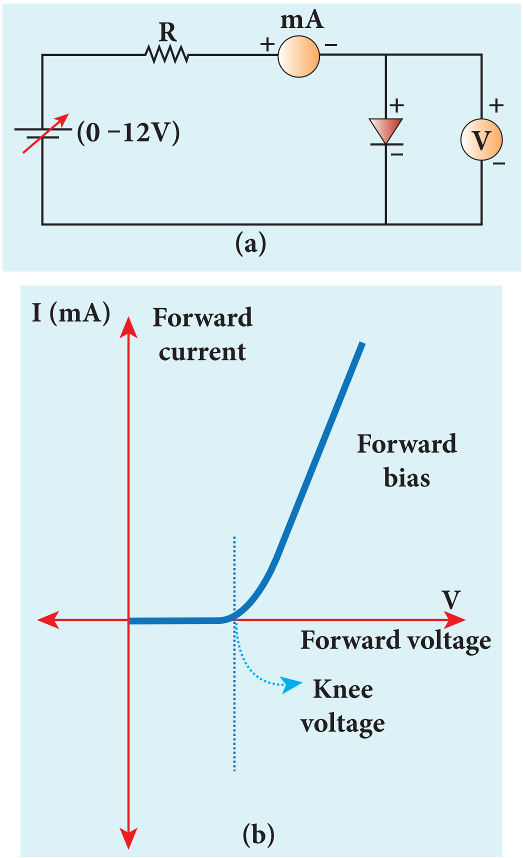


Figure 9.14 p-n junction diode (a) diode under forward bias (b) forward characteristics

(i) At room temperature, a potential difference equal to the barrier potential is required before a reasonable forward current starts flowing across the diode. This voltage is known as **threshold voltage or cut-in voltage or knee voltage** (V_{th}). It is approximately

0.3 V for Germanium and 0.7 V for Silicon. The current flow is negligible when the applied voltage is less than the threshold voltage. Beyond the threshold voltage, increase in current is significant even for a small increase in voltage.

(ii) The graph clearly infers that the current flow is not linear and is exponential. Hence it does not obey Ohm's law.

(iii) The forward resistance (r_f) of the diode is the ratio of the small change in voltage (ΔV) to the small change in current (ΔI), $r_f = \frac{\Delta V}{\Delta I}$.

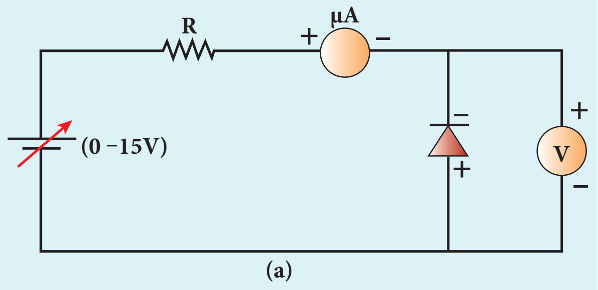
(iv) Thus the diode behaves as a conductor when it is forward biased.

However, if the applied voltage is increased beyond a rated value, it will produce an extremely large current which may destroy the junction due to overheating. This is called as the breakdown of the diode and the voltage at which the diode breaks down is called the breakdown voltage. Thus, it is safe to operate a diode well within the threshold voltage and the breakdown voltage.

9.3.3.2 Reverse characteristics

The circuit to study the reverse characteristics is shown in Figure 9.15(a). In the reverse bias, the p-region of the diode is connected to the negative terminal and n-region to the positive terminal of the dc power supply.

A graph is drawn between the reverse bias voltage and the current across the junction, which is called the reverse characteristics of a p-n junction diode. It is shown in Figure 9.15(b). Under this bias, a very small current in μA, flows across the junction. This is due to the flow of the minority charge carriers called the leakage current or reverse



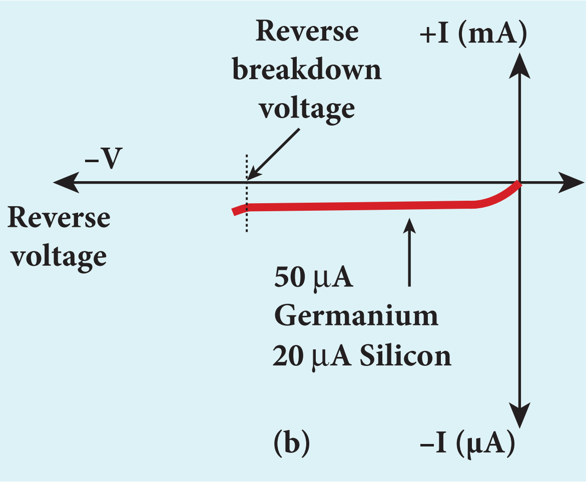


Figure 9.15 p-n junction diode (a) diode under reverse bias (b) reverse characteristics

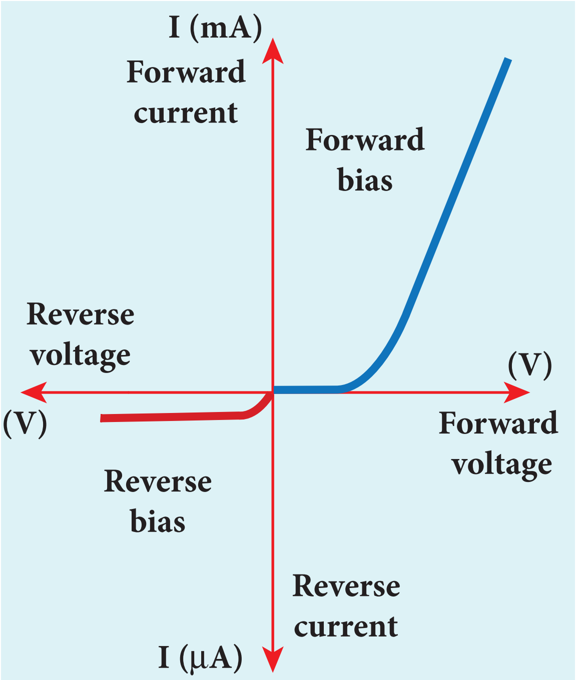


Figure 9.16 Forward and reverse charactristics of a diode

saturation current. Besides, the current is almost independent of the voltage. The reverse bias voltage can be increased only up to the rated value otherwise the diode will enter into the breakdown region.

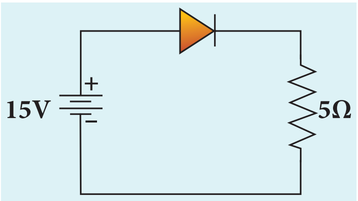
Note Ideal diode: It acts like a conductor when it is forward biased. When it is reverse biased, it acts like an insulator. The barrier potential is assumed to be zero and hence it behaves like a resistor.

The forward and reverse charactristics are given in one graph as shown in Figure 9.16.

L5N1M2

EXAMPLE 9.1

An ideal diode and a 5 Ω resistor are connected in series with a 15 V power supply as shown in figure below. Calculate the current that flows through the diode.



Solution

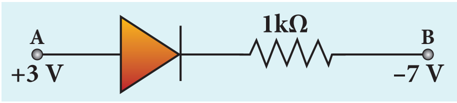
The diode is forward biased and it is an ideal one. Hence, it acts like a closed switch with no barrier voltage. Therefore, current that flows through the diode can be calculated using Ohm's law.

$V = IR$

$$I = \frac{V}{R} = \frac{15}{5} = 3\,A$$

EXAMPLE 9.2

Consider an ideal junction diode. Find the value of current flowing through AB is



Solution

The barrier potential of the diode is neglected as it is an ideal diode.

The value of current flowing through AB can be obtained by using Ohm's law

$$I = \frac{V}{R} = \frac{3-(-7)}{1\times10^{3}} = \frac{10}{10^{3}} = 10^{-2}\,A = 10\,mA$$

9.3.4 Rectification

The process of converting alternating current into direct current is called rectification. In this section, we will discuss two types of rectifiers namely, half wave rectifier and full wave rectifier.

9.3.4.1 Half wave rectifier circuit

The half wave rectifier circuit is shown in Figure 9.17(a). The circuit consists of a transformer, a p-n junction diode and a resistor. In a half wave rectifier circuit, either a positive half or the negative half of the AC input is passed through while the other half is blocked. Only one half of the input wave reaches the output. Therefore, it is called half wave rectifier. Here, a p-n junction diode acts as a rectifier diode.

During the positive half cycle

When the positive half cycle of the ac input signal passes through the circuit, terminal A becomes positive with respect to terminal B. The diode is forward biased and hence it conducts. The current flows through the load resistor R_L and the AC voltage developed across R_L constitutes the output voltage V_0 and the waveform of the diode current is shown in Figure 9.17(b).

During the negative half cycle

When the negative half cycle of the ac input signal passes through the circuit, terminal A is negative with respect to terminal B. Now the diode is reverse biased and does not conduct and hence no current passes through R_L. The reverse saturation current in a diode is negligible. Since there is no voltage drop across R_L, the negative half cycle of ac supply is suppressed at the output. The output waveform is shown in Figure 9.17b.

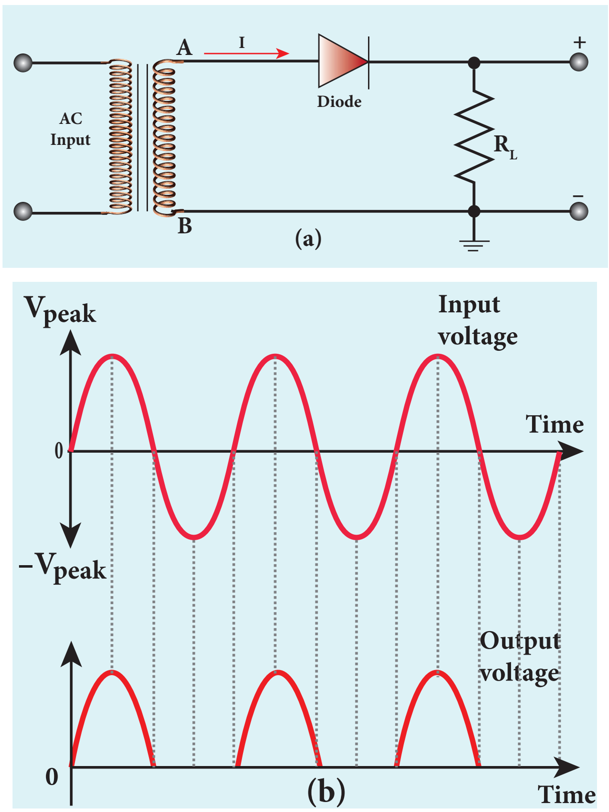


Figure 9.17 (a) Input ac signal (b) half wave rectifier circuit (c) input and output waveforms

The output of the half wave rectifier is not a steady dc voltage but a pulsating wave. This pulsating voltage can not be used for electronic equipments. A constant or a steady voltage is required which can be obtained with the help of filter circuits and voltage regulator circuits.

Efficiency (η) **is the ratio of the output dc power to the ac input power supplied to the circuit**. Its value for half wave rectifier is 40.6 %

Note If the direction of the diode is reversed, the negative half of the ac signal is passed through and the positive half is blocked.

9.3.4.1 Full wave rectifier

The positive and negative half cycles of the AC input signal pass through the full wave rectifier circuit and hence it is called the full wave rectifier. The circuit is shown in Figure 9.18(a). It consists of two p-n junction diodes, a center tapped

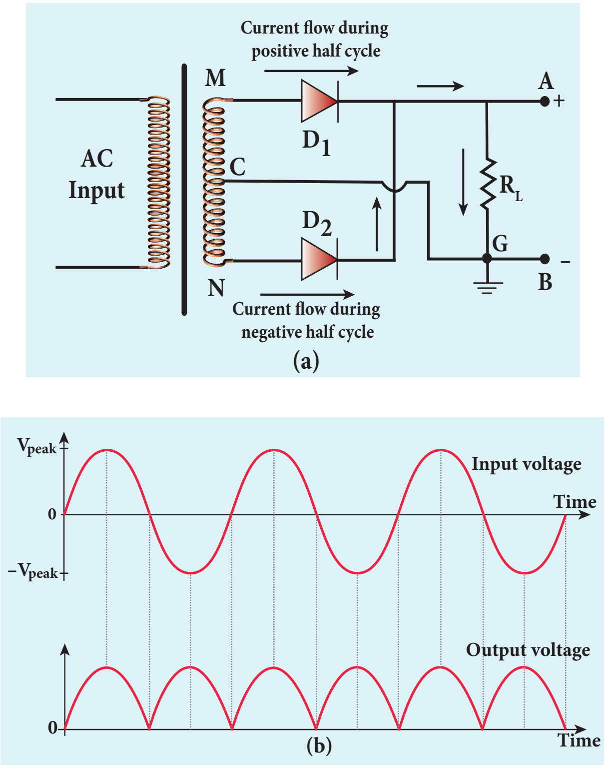


Figure 9.18 (a) Full wave rectifier circuit (b) Input and output waveforms

transformer, and a load resistor (R_L). The centre is usually taken as the ground or zero voltage reference point. Due to the centre tap transformer, the output voltage rectified by each diode is only one half of the total secondary voltage.

During positive half cycle

When the positive half cycle of the ac input signal passes through the circuit, terminal M is positive, G is at zero potential and N is at negative potential. This forward biases diode D_1 and reverse biases diode D_2. Hence, being forward biased, diode D_1 conducts and current flows along the path MD_1AGC. As a result, positive half cycle of the voltage appears across R_L in the direction G to C

During negative half cycle

When the negative half cycle of the ac input signal passes through the circuit, terminal N is positive, G is at zero potential and M is at negative potential. This forward biases diode D_2 and reverse biases diode D_1. Hence, being forward biased, diode D_2 conducts and current flows along the path ND_2BGC. As a result, negative half cycle of the voltage appears across R_L in the same direction from G to C

Hence in a full wave rectifier both postive and negative half cycles of the input signal pass through the load in the same direction as shown in Figure 9.18(b). Though both positive and negative half cycles of ac input are rectified, the output is still pulsating in nature.

The efficiency (η) of full wave rectifier is twice that of a half wave rectifier and is found to be 81.2 %. It is because both the positive and negative half cycles of the ac input source are rectified.

Note Centre tap transformer: There is a facility to tap at halfway point in the secondary windings. This helps to measure the induced voltage from one end of the secondary to the centre point. If the centre tap point is grounded then the voltage applied across the secondary will be divided by half. For example, if the voltage applied across the secondary is 240 V, then the voltage across one end and the centre tap point is +120 V and at the other end it is –120 V.

9.3.5 Breakdown mechanism

The reverse current or the reverse saturation current due to the minority charge carriers is small. If the reverse bias applied to a p-n junction is increased beyond a point, the junction breaks down and the reverse current rises sharply. The voltage at which this happens is called the breakdown voltage and it depends on the width of the depletion region, which in turn depends on the doping level.

A normal p-n junction diode gets damaged at this point. Specially designed diodes like Zener diode can be operated at this region and can be used for the purpose of voltage regulation in circuits. There are two mechanisms that are responsible for breakdown under increasing reverse voltage.

9.3.5.1 Zener breakdown

Heavily doped p-n junctions have narrow depletion layers of the order of $<10^{-6}$ m. When a reverse voltage across this junction is increased to the breakdown limit, a very strong electric field of strength 3×10^7 V m^{-1} is set up across the narrow layer. This electric field is strong enough to break or rupture the covalent bonds in the lattice and thereby generating electron-hole pairs. This effect is called **Zener effect**.

Even a small further increase in reverse voltage produces a large number of charge carriers. Hence the junction has very low resistance in the breakdown region. This process of emission of electrons due to the rupture of bands in from the lattice due to strong electric field is known as **internal field emission or field ionization**. The electric field required for this is of the order of $10^6\ V\ m^{-1}$.

9.3.5.2 Avalanche breakdown

Avalanche breakdown occurs in lightly doped junctions which have wide depletion layers. Here, in this case, the electric field is not strong enough to produce breakdown. Alternatively, the thermally generated minority charge carriers accelerated by the electric field gains sufficient kinetic energy, collide with the semiconductor atoms while passing through the depletion region. This leads to the breaking of covalent bonds and in turn generates electron-hole pairs.

The newly generated charge carriers are also accelerated by the electric field resulting in more collisions and further production of charge carriers. This cumulative process leads to an avalanche of charge carriers across the junction and consequently reduces the reverse resistance. The diode current increases sharply.

Note

For a reverse voltage of,

(i) less than 4V → Zener effect predominates

(ii) greater than 6V → Avalanche effect predominates

(iii) between 4 and 6V → both effects are present.

9.3.6 Zener diode

Zener diode is a heavily doped silicon diode used in reverse biased condition and is named after its inventor C. Zener. It is specially designed to be operated in the breakdown region. The doping level of the Silicon diode can be varied to have a wide range of breakdown voltages from 2 V to over 1000 V.

As explained in the previous section, Zener breakdown occurs due to the breaking of covalent bonds by the strong electric field set up in the depletion region by the reverse voltage. It produces an extremely large number of electrons and holes which constitute the reverse saturation current. The current is limited by both external resistance and power dissipation of the diode. A Zener diodes is shown in Figure 9.19(a) and its circuit symbol of Zener diode is shown in Figure 9.19(b).

It looks like an ordinary p-n junction diode except the cathode lead approximating the shape of a 'z' letter. The arrow head points the direction of conventional current. In Figure 9.19(a), black ring indicates the cathode lead.

Figure 9.19 Zener diode (a) commercial picture(b) circuit symbol

9.3.6.1 V-I Characteristics of Zener diode

The circuit to study the forward and reverse characteristics of a Zener diode is shown in Figure 9.20(a) and Figure 9.20 (b). The V-I characteristics of a Zener diode is shown in Figure 9.20(c). The forward characteristic of a Zener diode is similar to that of an ordinary p-n junction diode. It starts conducting approximately around 0.7 V. However, the reverse characteristics is highly significant in Zener diode. The increase in reverse voltage normally generates very small reverse current. While in Zener diode, when the reverse voltage is increased to the breakdown voltage (V_Z), the increase in current is very sharp. The voltage remains almost constant throughout the breakdown region. In Figure 9.20(c), $I_{Z(max)}$ represents the maximum reverse current. If the reverse current is increased further, the diode will be damaged. The important parameters on the reverse characteristics are

$V_Z \rightarrow$ Zener breakdown voltage

$I_{Z(\min)} \rightarrow$ minimum current to sustain breakdown

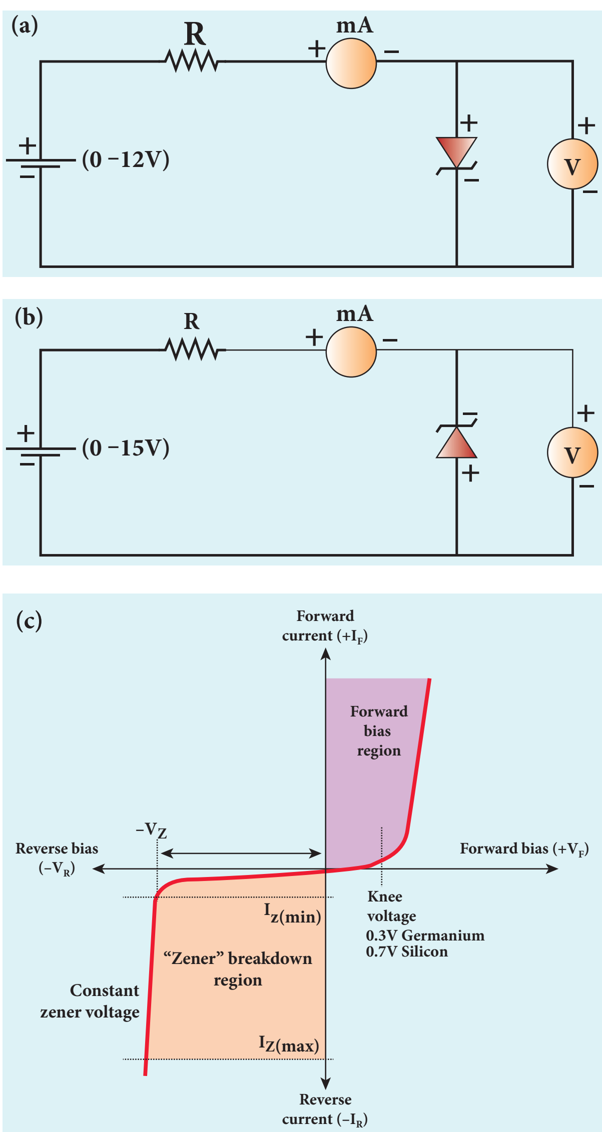


Figure 9.20 Zener diode (a) forward bias (b) reverse bias (c) V-I characteristics

$I_{Z(\max)} \rightarrow$ maximum current limited by maximum power dissipation.

The Zener diode is operated in the reverse bais having the voltage greater than V_Z and current less than $I_{Z(max)}$. The reverse characteristic is not exactly vertical which means that the diode possesses some small resistance called Zener dynamic impedance. Zener resistance is the inverse of the slope in the breakdown region. It means an increase in the Zener current produces only a very small increase in the reverse voltage. However this can be neglected. The voltage of an ideal Zener diode does not change once it goes into breakdown. It means that V_Z remains almost constant even when I_Z increases considerably.

Note The maximum reverse bias that can be applied before entering into the Zener region is called the Peak inverse voltage. Commercially referred as PIV rating.

Applications

The zener diode can be used as

1. Voltage regulators
2. Calibrating voltages
3. Provide fixed reference voltage in a network for biasing
4. Protection of any gadget against damage from accidental application of excessive voltage.

9.3.6.2 Zener diode as a voltage regulator

A Zener diode working in the breakdown region can serve as a voltage regulator. It maintains a constant output voltage even when input voltage V_i or load current I_L varies. The circuit used for the same is shown in Figure 9.21. Here in this circuit, the input voltage V_i is regulated at a constant voltage,

V_z (Zener voltage) at the output represented as V_0 using a Zener diode. The output voltage is maintained constant as long as the input voltage does not fall below V_z.

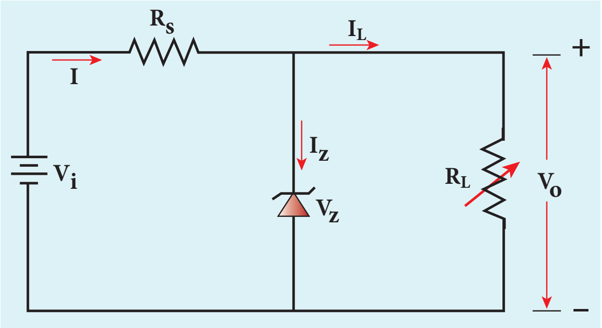


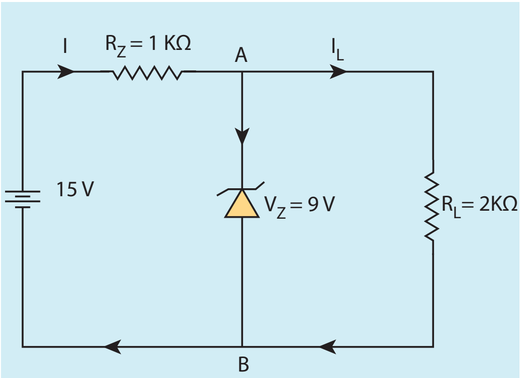
Figure 9.21 Circuit to study voltage regulation by Zener diode

When the potential developed across the diode is greater than V_Z, the diode moves into the Zener breakdown region. It conducts and draws relatively large current through the series resistance R_s. The total current I passing through R_s equals the sum of diode current I_Z and load current I_L $\left(I = I_Z + I_L\right)$. It is to be noted that the total current is always less than the maximum Zener diode current.

Under all conditions $V_o = V_Z$. Thus, output voltage is regulated.

EXAMPLE: 9.3

Find the current through the Zener diode when the load resistance is 1 KΩ. Use diode approximation.



Solution

Voltage across AB is V_Z = 9V

Voltage drop across R = 15 - 9 = 6V

Therefore current through the resistor R,

$$I = \frac{6}{1\times10^3} = 6\ mA$$

Voltage across the load resistor = $V_{AB} = 9V$

Current through load resistor,

$$I_L = \frac{V_{AB}}{R_L} = \frac{9}{2\times10^3} = 4.5\ mA$$

The current through the Zener diode,

$$I_Z = I - I_L = 6mA - 4.5mA = 1.5\ mA$$

9.3.7 Optoelectronic devices

Optoelectronics deals with devices which convert electrical energy into light and light into electrical energy through semiconductors. Optoelectronic device is an electronic device which utilizes light for useful applications. We will discuss some important optoelectronic devices namely, light emitting diodes, photo diodes and solar cells.

9.3.7.1 Light Emitting Diode (LED)

LED is a p-n junction diode which emits visible or invisible light when it is forward biased. Since, electrical energy is converted into light energy, this process is also called electroluminescence. The circuit symbol of LED is shown in Figure 9.22(a).

The cross-sectional view of a commercial LED is shown in Figure 9.22(b). It consists of a p-layer, n-layer and a substrate. A transparent window is used to allow light to travel in the desired direction. An external resistance in series with the biasing source is required to limit the forward current through the LED. In addition, it has two leads; anode and cathode.

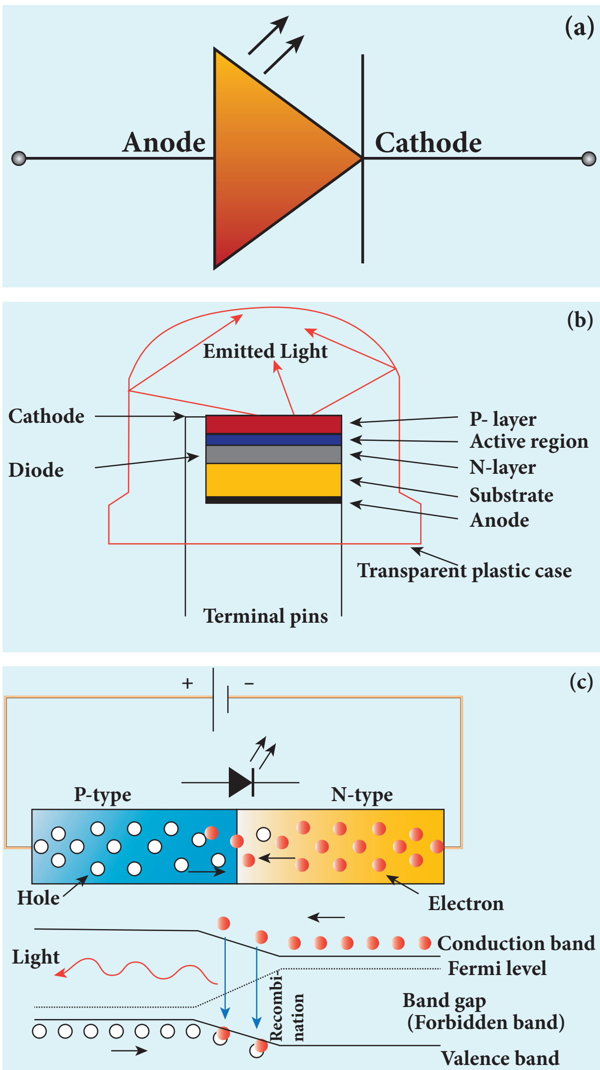


Figure 9.22 (a) Circuit symbol of LED (b) Inside view of LED (c) Schematic diagram to explain recombination process

When the p-n junction is forward biased, the conduction band electrons on n-side and valence band holes on p-side diffuse across the junction. When they cross the junction, they become excess minority carriers (electrons in p-side and holes in n-side). These excess minority carriers recombine with oppositely charged majority carriers in the respective regions, i.e. the electrons in the conduction band recombine with holes in the valence band as shown in the Figure 9.22(c).

During recombination process, energy is released in the form of light (radiative) or heat (non-radiative). For radiative recombination, a photon of energy $h\nu$ is emitted. For non-radiative recombination, energy is liberated in the form of heat.

The colour of the light is determined by the energy band gap of the material. Therefore, LEDs are available in a wide range of colours such as blue (SiC), green (AlGaP) and red (GaAsP). Now a days, LED which emits white light (GaInN) is also available.

Applications

- Indicator lamps on the front panel of the scientific and laboratory equipments.
- Seven-segment displays.
- Traffic signals, emergency vehicle lighting etc.
- Remote control of television, airconditioner etc.

EXAMPLE 9.4

Determine the wavelength of light emitted from LED which is made up of GaAsP semiconductor whose forbidden energy gap is 1.875 eV. Mention the colour of the light emitted (Take $h = 6.6 \times 10^{-34}$ Js).

Solution

$$E_g = \frac{hc}{\lambda}$$

Therefore,

$$\lambda = \frac{hc}{E_g} = \frac{6.6\times10^{-34}\times3\times10^{8}}{1.875\times1.6\times10^{-19}}$$

$$= 660\ nm$$

The wavelength 660 nm corresponds to red colour light.

9.3.7.2 Photodiodes

A p-n junction diode which converts an optical signal into electric signal is known as photodiode. Thus, the operation of photodiode is exactly inverse to that of an LED. Photo diode works in reverse bias. Its circuit symbol is shown in Figure 9.23(a). The direction of arrows indicates that the light is incident on the photo diode.

The device consists of a p-n junction semiconductor made of photosensitive material kept safely inside a plastic case as shown in Figure 9.23(b). It has a small transparent window that allows light to be incident on the p-n junction. Photodiodes can generate current when the p-n junction is exposed to light and hence are called as light sensors.

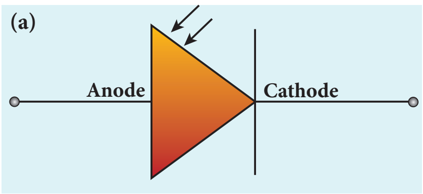


Figure 9.23 (a) Circuit symbol (b) Schematic view of photodiode

When a photon of sufficient energy $(h\upsilon)$ strikes the depletion region of the diode, some of the valence band electrons are elevated into conduction band, in turn holes are developed in the valence band. This creates electron-hole pairs. The amount of electron-hole pairs generated depends on the intensity of light incident on the p-n junction.

These electrons and holes are swept across the p-n junction by the electric field created by reverse voltage before recombination takes place. Thus, holes move towards the n-side and electrons towards the p-side. When the external circuit is made, the electrons flow through the external circuit and constitute the photocurrent.

When the incident light is zero, there exists a reverse current which is negligible. This reverse current in the absence of any incident light is called dark current and is due to the thermally generated minority carriers.

Applications

- Alarm system
- Count items on a conveyer belt
- Photoconductors
- Compact disc players, smoke detectors
- Medical applications such as detectors for computed tomography etc.

9.3.7.2 Solar cell

A solar cell, also known as photovoltaic cell, converts light energy directly into electricity or electric potential difference by **photovoltaic effect**. It is basically a p-n junction which generates emf when solar radiation falls on the p-n junction. A solar cell is of two types: p-type and n-type.

Both types use a combination of p-type and n-type Silicon which together forms the p-n junction of the solar cell. The difference is that p-type solar cells use p-type Silicon as the base with an ultra-thin layer of n-type Silicon as shown in Figure 9.24, while n-type solar cell uses the opposite combination. The other side of the p-Silicon is coated with metal which forms the back electrical contact. On top of the n-type Silicon, metal grid is deposited which acts as the front electrical contact. The top of the solar cell is coated with anti-reflection coating and toughened glass.

In a solar cell, electron–hole pairs are generated due to the absorption of light

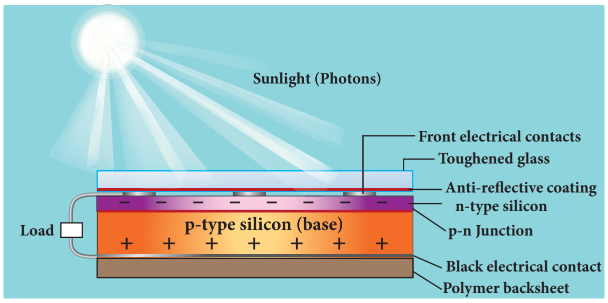


Figure 9.24 Cross-sectional view of a solar cell

near the junction. Then the charge carriers are separated due to the electric field of the depletion region. Electrons move towards n–type Silicon and holes move towards p-type Silicon layer. The electrons reaching the n-side are collected by the front contact and holes reaching p-side are collected by the back electrical contact. Thus a potential difference is developed across solar cell. When an external load is connected to the solar cell, photocurrent flows through the load.

Many solar cells are connected together either in series or in parallel combination to form solar panel or module. Many solar panels are connected with each other to form solar arrays. For high power applications, solar panels and solar arrays are used.

Applications:

- Solar cells are widely used in calculators, watches, toys, portable power supplies, etc.
- Solar cells are used in satellites and space applications
- Solar panels are used to generate electricity.

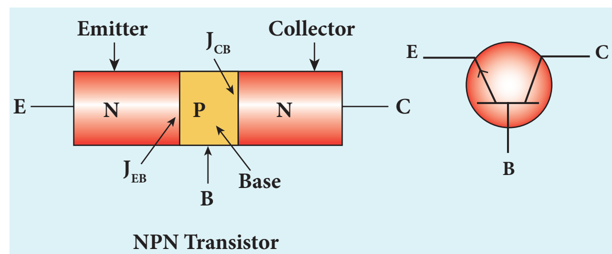
9.4 THE BIPOLAR JUNCTION TRANSISTOR [BJT]

Introduction

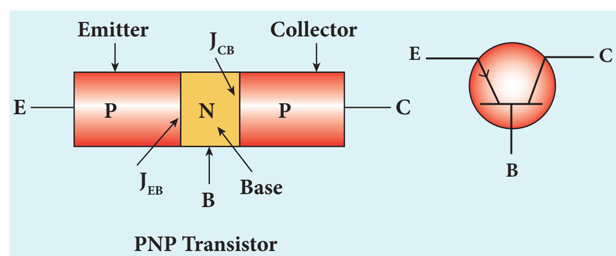
In 1951, William Schockley invented the modern version of transistor. It is a semiconductor device that led to a technological revolution in the twentieth century. The heat loss in transistor is very less. This has laid the foundation of integrated chips which contain thousands of miniaturized transistors. The emergence of the integrated chips led to increasing applications in the fast developing electronics industry.

Bipolar Junction Transistor

The BJT consists of a semiconductor (Silicon or Germanium) crystal in which an n-type material is sandwiched between two p-type materials (PNP transistor) or a p-type material sandwiched between two n-type materials (NPN transistor). To protect it against moisture, it is sealed inside a metal or a plastic case. The two types of transistors with their circuit symbols are shown in Figure 9.25.



(a)



(b)

Figure 9.25 Schematic Diagram of
(a) NPN transistor and circuit symbol
(b) PNP transistor and circuit symbol

The three regions formed are called as emitter, base and collector which are provided with terminals or ohmic contacts labeled as E, B, and C. As a BJT has two p-n junctions, two depletion layers are formed across the emitter-base junction (J_{EB}) and collector-base junction (J_{CB}) respectively. The circuit symbol carries an arrowhead at the emitter lead pointing from p to n indicating the direction of conventional current.

***Emitter*:**

The main function of the emitter is to supply majority charge carriers to the collector region through the base region. Hence, emitter is more heavily doped than the other two regions.

Base:

Base is very thin (10^{-6} m) and very lightly doped compared to the other two regions.

Collector:

The main function of collector is to collect the majority charge carriers supplied by the emitter through the base. Hence, collector is made physically larger than the other two as it has to dissipate more power. Its is modarately dopped.

DO YOU KNOW?

Because of the differing size and the amount of dopping, the emitter and collector cannot be interchanged.

Transistor Biasing

The application of suitable dc voltages across the transistor terminals is called biasing.

Different modes of transistor biasing

Forward Active:

In this bias the emitter-base junction is forward biased and the collector-base junction is reverse biased. The transistor is in the active mode of operation. In this mode, the transistor functions as an amplifier.

Saturation:

Here, the emitter-base junction and collector-base junction are forward biased. The transistor has a very large flow of currents across the junctions. In this mode, transistor is used as a closed switch.

Cut-off:

In this bias, the emitter-base junction and collector-base junction are reverse biased. Transistor in this mode is an open switch.

Note

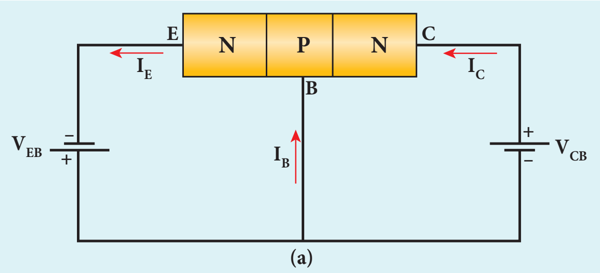
In a PNP transistor, base and collector will be negative with respect to emitter indicated by the middle letter N whereas base and collector will be positive in an NPN transistor [indicated by the middle letter P]

9.4.1 Transistor circuit configurations

There are three types of circuit connections for operating a transistor based on the terminal that is used in common to both input and output circuits.

9.4.1.1 Common-Base (CB) configuration

The base is common to both the input and output circuits. The schematic and circuit symbol are shown in Figure 9.26(a) and 9.26(b). The input current is the emitter current I_E and the output current is the collector current I_C. The input signal is applied between emitter and base, the output is measured between collector and base.



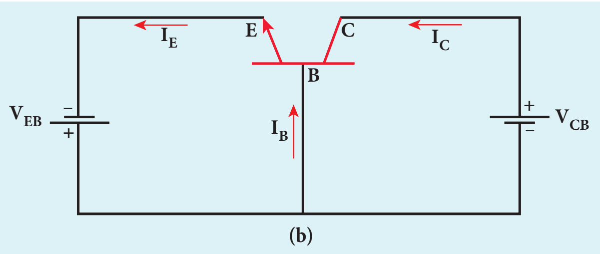


Figure 9.26 NPN transistor in common base configuration (a) schematic circuit diagram (b) circuit symbol

9.4.1.2 Common-Emitter (CE) configuration

In this configuration, the emitter is common to both the input and output loops as shown in Figure 9.27. Base current, I_B is the input current and the collector current, I_C is the output current. The input signal is applied between the emitter and base and the output is measured between the collector and the emitter.

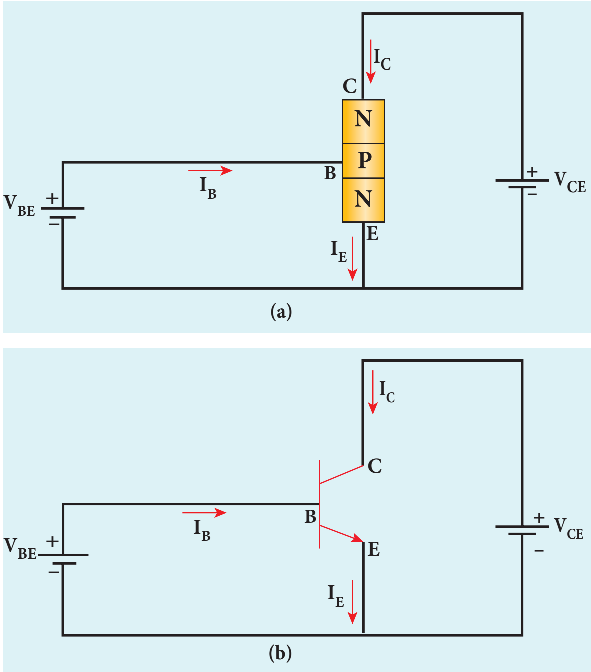


Figure 9.27 NPN transistor in common emitter configuration (a) schematic circuit diagram (b) circuit symbol

9.4.1.3 Common-Collector (CC) configuration

Here, the collector is common to both the input and output circuits as shown in Figure 9.28. The base current I_B is the input current, the emitter current I_E is the output current. The input signal is applied between the base and the collector, the output is measured between the emitter and collector.

Note

As the output is taken from the emitter in common collector configuration, it is called an emitter follower.

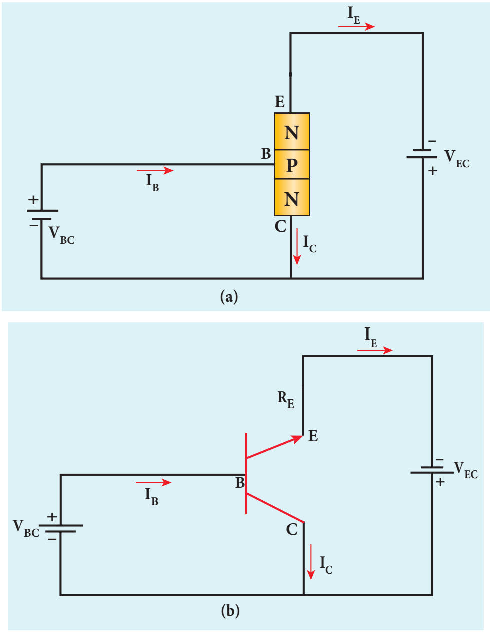


Figure 9.28 NPN transistor in common collector configuration (a) schematic circuit diagram (b) circuit symbol

9.4.2 Transistor action in the common base mode

The operation of an NPN transistor in the common base mode is explained below. The current flow in a common base NPN transistor in the forward active mode is shown in Figure 9.29.

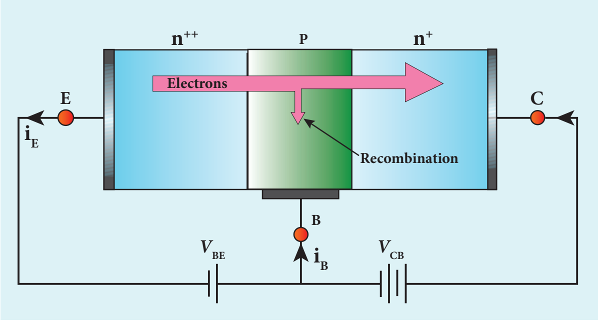


Figure 9.29 Flow of current in a NPN transistor

Basically, a BJT can be considered as two p-n junction diodes connected back-to-back. In the forward active bias of the transistor, the emitter-base junction is forward biased by a dc power supply V_{EB} and the collector-base junction is reverse biased by the bias power supply V_{CB}. The forward bias decreases the depletion region across the emitter-base junction and the reverse bias increases the depletion region across the collector-base junction. Hence, the barrier potential across the emitter-base junction is decreased and the collector-base junction is increased. The voltage across the emitter-base junction is represented as V_{EB} and the collector-base junction as $V_{CB.}$

In an NPN transistor, the majority charge carriers in the emitter are electrons. As it is heavily doped, it has a large number of electrons. The forward bias across the emitter-base junction causes the electrons in the emitter region to flow towards the base region and constitutes the emitter current (I_E). The electrons after reaching the base region recombine with the holes in the base region. Since the base region is very narrow and lightly doped, all the electrons will not have sufficient holes to recombine and hence most of the electrons reach the collector region.

Eventually, the electrons that reach the collector region will be attracted by the collector terminal as it has positive potential and flows through the external circuit. This constitutes the collector current (I_C). The holes that are lost due to recombination in the base region are replaced by the positive potential of the bias voltage V_{EE} and constitute the base current (I_B). The magnitude of the base current will be in micoamperes as against milliamperes for emitter and collector currents.

It is to be noted that if the emitter current is zero, then the collector current is almost zero. It is therefore imperative that a BJT is called a current controlled device. Applying Kirchoff's law, we can write the emitter current as the sum of the collector current and the base current.

$$I_E = I_B + I_C$$

Since the base current is very small, we can write, $I_E \approx I_C$. There is another component of collector current due to the thermally generated electrons called reverse saturation current, denoted as (I_{CO}). This factor is temperature sensitive. Therefore, care must be taken towards the stability of the system at high temperatures.

The ratio of the collector current to the emitter current is called the forward current gain (α_{dc}) of a transistor.

$$\alpha_{dc} = \frac{I_C}{I_E}$$

The α of a transistor is a measure of the quality of a transistor. Higher the value of α better is the transistor. It means that the collector current is closer to the emitter current. The value of α is less than unity and ranges from 0.95 to 0.99. This indicates that the collector current is 95% to 99% of the emitter current.

Note

1. The conventional flow of current is based on the direction of the motion of holes
2. In NPN transistor, current enters from the base into the emitter.
3. In a PNP transistor, current enters from the emitter into the base.
4. The emitter-base junction has low resistance and the collector-base junction has high resistance.

Working of a PNP transistor

The working of a PNP transistor is similar to the NPN transistor except for the fact that the emitter current I_E is due to holes and the base current I_B is due to electrons. However, the current through the external circuit is due to the flow of electrons.

EXAMPLE 9.5

In a transistor connected in the common base configuration, $\alpha = 0.95$, $I_E = 1\,mA$. Calculate the values of I_C and I_B.

Solution

$$\alpha = \frac{I_C}{I_E}$$

$$I_C = \alpha I_E = 0.95 \times 1 = 0.95\,mA$$

$$I_E = I_B + I_C$$

$$\therefore I_B = I_C - I_E = 1 - 0.95 = 0.05\,mA$$

9.4.3 Static Characteristics of Transistor in Common Emitter Mode

The know-how of certain parameters like the input resistance, output resistance, and current gain of a transistor are very important for the effective use of transistors in circuits. The circuit to study the static characteristics of an NPN transistor in the common emitter mode is given in Figure 9.30. The bias supply voltages V_{BB} and V_{CC} bias the base-emitter junction and collector-emitter junction respectively. The junction potential at the base-emitter is represented as V_{BE} and the collector-emitter as V_{CE}. The rheostats R_1 and R_2 are used to vary the base and collector currents respectively.

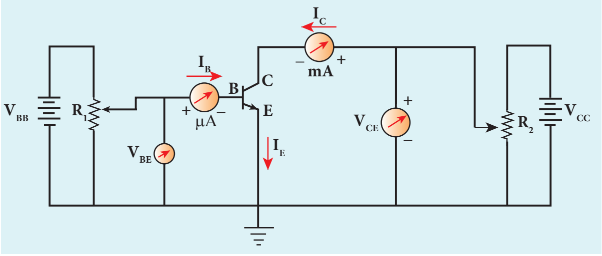


Figure 9.30 Static characteristics of a NPN transistor in common emitter configuration

The static characteristics of the BJT are

1. Input characteristics
2. Output characteristics
3. Transfer characteristics

9.4.3.1 Input Characteristics

Input Characteristics curves give the relationship between the base current (I_B) and base to emitter voltage (V_{BE}) at constant collector to emitter voltage (V_{CE}) and are shown in Figure 9.31.

Initially, the collector to emitter voltage (V_{CE}) is set to a particular voltage (above 0.7 V to reverse bias the junction). Then the base-emitter voltage (V_{BE}) is increased in suitable steps and the corresponding base-current (I_B) is recorded. A graph is plotted with V_{BE} along the x-axis and I_B along the y-axis. The procedure is repeated for different values of V_{CE}.

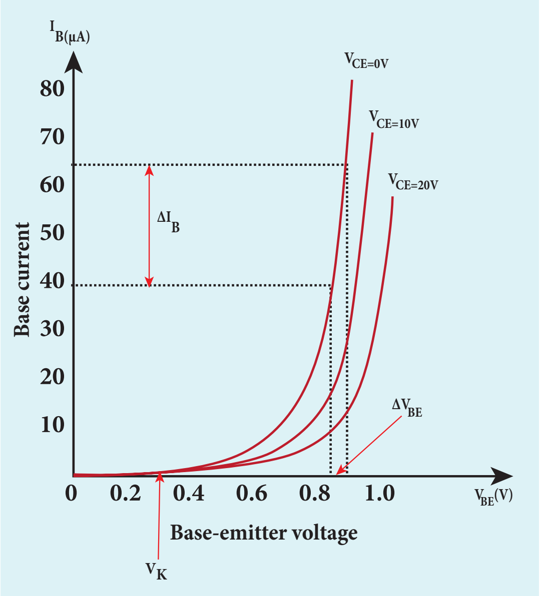


Figure 9.31 Input characteristics of a NPN transistor in common emitter configuration

The following observations are made from the graph.

- The curve looks like the forward characteristics of an ordinary p-n junction diode.

- There exists a threshold voltage or knee voltage (V_k) below which the base current is very small. The value is 0.7 V for Silicon and 0.3 V for Germanium transistors. Beyond the knee voltage, the base current increases with the increase in base-emitter voltage.
- It is also noted that the increase in the collector-emitter voltage decreases the base current. This shifts the curve outward. This is because the increase in collector-emitter voltage increases the width of the depletion region in turn, reduces the effective base width and thereby the base current.

Input impedance

The ratio of the change in base-emitter voltage (ΔV_{BE}) to the change in base current (ΔI_B) at a constant collector-emitter voltage (V_{CE}) is called the input impedance (r_i). The input impedance is not linear in the lower region of the curve.

$$r_i = \left(\frac{\Delta V_{BE}}{\Delta I_B}\right)_{V_{CB}}$$

The input impedance is high for a transistor in common emitter configuration.

9.4.3.2 Output Characteristics

The output characteristics give the relationship between the variation in the collector current (ΔI_C) with respect to the variation in collector-emitter voltage (ΔV_{CE}) at constant input current (I_B) as shown in Figure 9.32.

Initially, the base current (I_B) is set to a particular value. Then collector-emitter voltage (V_{CE}) is increased in suitable steps and the corresponding collector current (I_C) is recorded. A graph is plotted with the V_{CE} along the x-axis and I_C along the y-axis. This procedure is repeated for different values of I_B. The four important regions in the output characteristics are:

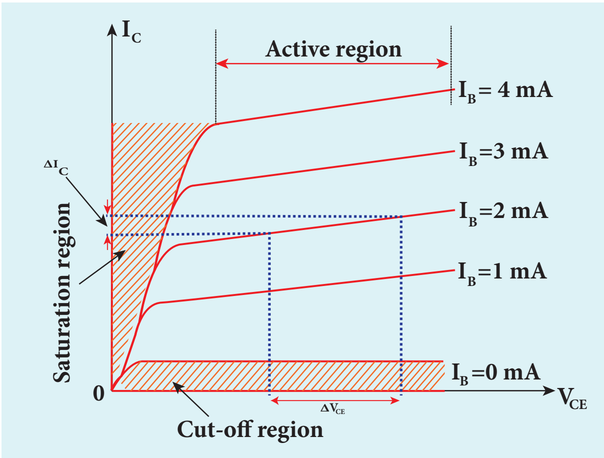


Figure 9.32 Output characteristics of a NPN transistor in common emitter configuration

(i). *Saturation region*

When V_{CE} is increased above 0 V, the I_C increases rapidly to a saturation value almost independent of I_B (Ohmic region, OA) called knee voltage. Transistors are always operated above this knee voltage.

(ii) *Cut-off region*

A small collector current (I_C) exists even after the base current (I_B) is reduced to zero. This current is due to the presence of minority carriers across the collector-base junction and the surface leakage current (I_{CEO}). This region is called as the cut-off region, because the main collector current is cut-off.

(iii) *Active region*

In this region, the emitter-base junction is forward biased and the collector-base junction is reverse biased. The transistor in this region can be used for voltage, current and power amplification.

(iv) *Breakdown region*

If the collector-emitter voltage (V_{CE}) is increased beyond the rated value given by the manufacturer, the collector current (I_C) increases enormously leading to the junction breakdown of the transistor. This avalanche breakdown can damage the transistor.

Output impedance

The ratio of the change in the collector-emitter voltage (ΔV_{CE}) to the corresponding change in the collector current (ΔI_C) at constant base current (I_B) is called output impedance (r_O).

$$r_o = \left(\frac{\Delta V_{CE}}{\Delta I_C}\right)_{I_B}$$

The output impedance for transistor in common emitter configuration is very low.

9.4.3.3 Current transfer characteristics

This gives the variation of collector current (I_C) with changes in base current (I_B) at constant collector-emitter voltage (V_{CE}) as shown in Figure 9.33. It is seen that a small I_C flows even when I_B is zero. This current is called the common emitter leakage current (I_{CEO}), which is due to the flow of minority charge carriers.

Forward current gain

The ratio of the change in collector current (ΔI_C) to the change in base current (ΔI_B) at constant collector-emitter voltage (V_{CE}) is called forward current gain (β).

$$\beta = \left(\frac{\Delta I_C}{\Delta I_B}\right)_{V_{CE}}$$

Its value is very high and it generally ranges from 50 to 200. It depends on the construction of the transistor and will be provided by the manufacturer. There are transistors with β as high as 1000 as well.

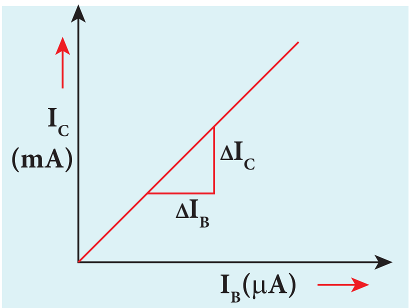


Figure 9.33 Current transfer characteristics of a NPN transistor common emitter configuration

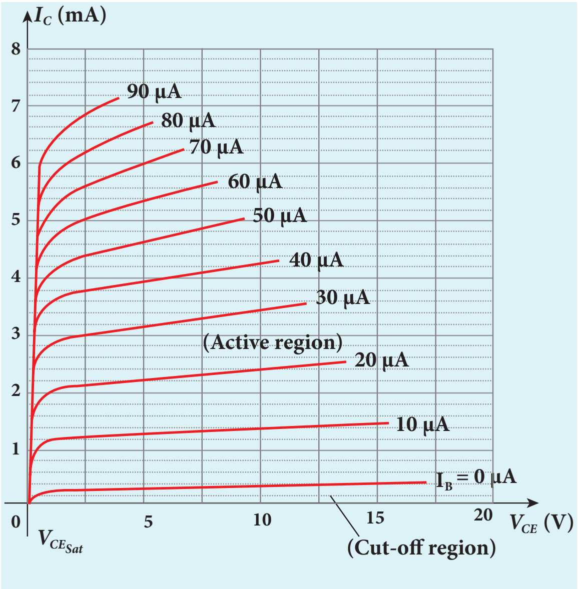
9.4.3.4 Relation between α and β

There is a relation between current gain in the common base configuration α and current gain in the common emitter configuration β which is given below.

$$\alpha = \frac{\beta}{1+\beta} \text{ (or) } \beta = \frac{\alpha}{1-\alpha}$$

EXAMPLE 9.6

The output characteristics of a transistor connected in common emitter mode is shown in the figure. Determine the value of I_C when V_{CE} = 15 V. Also determine the value of I_C when V_{CE} is changed to 10 V



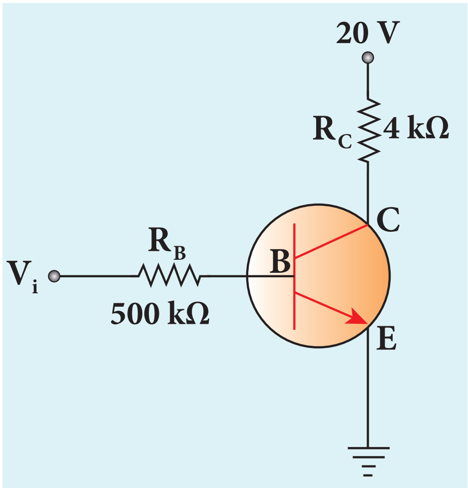
When $V_{CE} = 15\ V, I_C = 1.5\ \mu A$

When V_{CE} is changed to $10\ V, I_C = 1.4\ \mu A$

> **Note** The collector current is independent of the collector-emitter voltage in the active region.

EXAMPLE 9.7

In the circuit shown in the figure, the input voltage V_i is 20 V, $V_{BE} = 0$ V and $V_{CE} = 0$ V. What are the values of I_B, I_C, β?



$$I_B = \frac{V_i}{R_B} = \frac{20\,V}{500\,k\Omega} = 40\,\mu A \left[\because V_{BE} = 0V\right]$$

$$I_C = \frac{V_{CC}}{R_C} = \frac{20\,V}{4\,k\Omega} = 5\,mA \left[\because V_{CE} = 0V\right]$$

$$\beta = \frac{I_C}{I_B} = \frac{5\,mA}{40\,\mu A} = 125$$

9.4.4 Transistor as a switch

The transistor in saturation and cut-off regions functions like an electronic switch that helps to turn ON or OFF a given circuit by a small control signal. The circuit is shown in Figure 9.34.

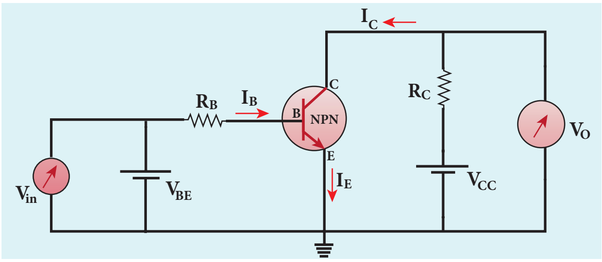


Figure 9.34 Transistor as a switch

- **Presence of dc source at the input (saturation region):**

  When a high input voltage ($V_{in} = +5V$) is applied, the base current (I_B) increases and in turn increases the collector current. The transistor will move into the saturation region (turned ON). The increase in collector current (I_C) increases the voltage drop across R_C, thereby lowering the output voltage, close to zero. The transistor acts like a closed switch and is equivalent to ON condition.

- **Absence of dc source at the input (cut-off region):**

  A low input voltage ($V_{in} = 0V$), decreases the base current (I_B) and in turn decreases the collector current (I_C). The transistor will move into the cut-off region (turned OFF). The decrease in collector current (I_C) decreases the drop across R_C, thereby increasing the output voltage, close to +5 V. The transistor acts as an open switch which is considered as the OFF condition.

It is manifested that, a high input gives a low output and a low input gives a high output. In addition, we can say that the output voltage is opposite to the applied input voltage. Therefore, a transistor can be used as an inverter (NOT gate) in computer logic circuitry.

9.4.5 Operating Point

The operating point is a point where the transistor can be operated efficiently. A line that is drawn with the values V_{CC} (when $I_C = 0$) and I_C (when $V_{CE} = 0$) is called the dc load line. The dc load line superimposed on the output characteristics of a transistor is used to learn the operating point of the transistor as shown in Figure 9.35.

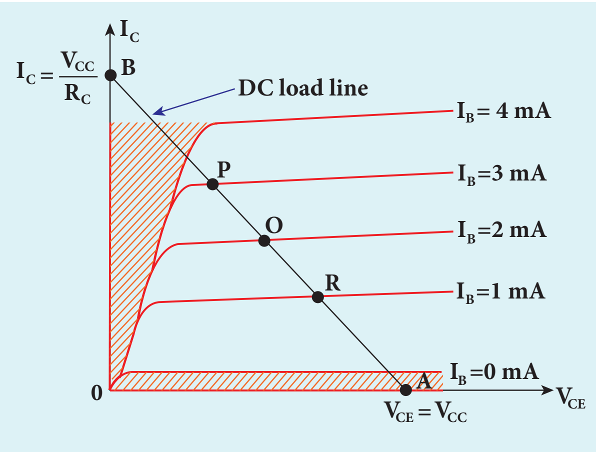


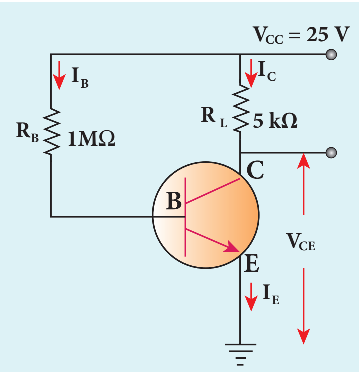
Figure 9.35 Output characteristics of a transistor in common emitter mode with the dc load line

Points P, Q, R in Figure 9.35 are called Q points or quiescent points which determine the operating point or the working point of a transistor. If the operating point is chosen at the middle of the dc load line (point Q), the transistor can effectively work as an amplifier. The operating point determines the maximum signal that can be obtained without being distorted.

For a transistor to work as a open switch, the Q point can be chosen at the cut-off region and to work as a closed switch, the Q point can be chosen in the saturation region.

EXAMPLE:9.8

The current gain of a common emitter transistor circuit shown in figure is 120. Draw the dc load line and mark the Q point on it. (V_{BE} to be ignored).



Solution

$\beta = 120$

Base current $I_B = \dfrac{25V}{1M\Omega} = \dfrac{25}{1\times10^6} = 25\ \mu A$

$\beta = \dfrac{I_C}{I_B}$

$I_C = \beta I_B$

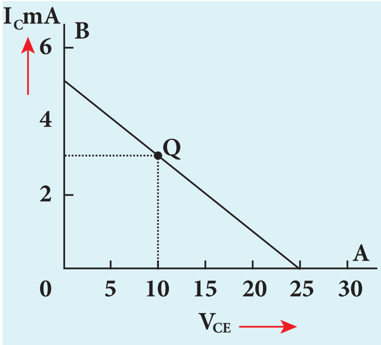
$I_C = 120 \times 25\ \mu A$

$I_C = 3mA$

$V_{CE} = V_{CC} - I_C R_C$

$V_{CE} = 25 - 3\ mA \times 5k$

$V_{CE} = 10V$



9.4.6 Transistor as an amplifier

A transistor operating in the active region has the capability to amplify weak signals. **Amplification is the process of increasing the signal strength (increase in the amplitude).** If a large amplification is required, the transistors are cascaded with coupling elements like resistors, capacitors, and transformers which is called as multistage amplifiers.

Here, the amplification of an electrical signal is explained with a single stage transistor amplifier as shown in Figure 9.36(a). Single stage indicates that the circuit consists of one transistor with the allied components. An NPN transistor is connected in the common emitter configuration.

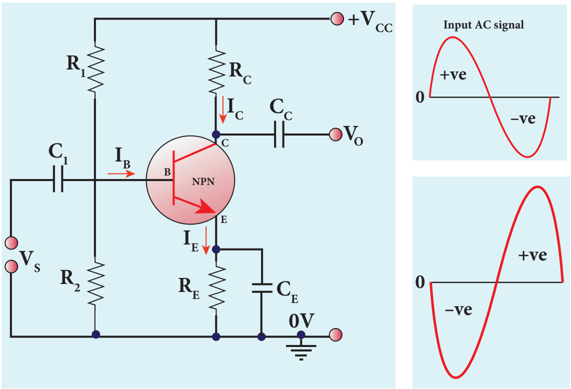


Figure 9.36 (a) Transistor as an amplifier (b) Input and output waveform showing 180^0 phase reversal.

To start with, the Q point or the operating point of the transistor is fixed so as to get the maximum signal swing at the output (neither towards saturation point nor towards cut-off). A load resistance, R_C is connected in series with the collector circuit to measure the output voltage. The capacitor C_1 allows only the ac signal to pass through. The emitter bypass capacitor C_E provides a low reactance path to the amplified ac signal. The coupling capacitor C_C is used to couple one stage of the amplifier with the next stage while constructing multistage amplifiers. V_S is the sinusoidal input signal source applied across the base-emitter. The output is taken across the collector-emitter.

$$\text{Collector current, } I_C = \beta I_B \left[\because \beta = \frac{I_C}{I_B}\right]$$

Applying Kirchhoff's voltage law in the output loop, the collector-emitter voltage is given by

$$V_{CE} = V_{CC} - I_C R_C$$

Working of the amplifier

- *During the positive half cycle*

Input signal (V_s) increases the forward voltage across the emitter-base. As a result, the base current (I_B) increases. Consequently, the collector current (I_C) increases β times. This increases the voltage drop across R_C $(I_c R_c)$ which in turn decreases the collector-emitter voltage (V_{CE}). Therefore, the input signal in the positive direction produces an amplified signal in the negative direction at the output. Hence, the output signal is reversed by 180° as shown in Figure 9.36(b).

- *During the negative half cycle*

Input signal (V_s) decreases the forward voltage across the emitter-base. As a result, base current (I_B) decreases and in turn increases the collector current (I_C). The increase in collector current (I_C) decreases the potential drop across R_C and increases the collector-emitter voltage (V_{CE}). Thus, the input signal in the negative direction produces an amplified signal in the positive direction at the output. Therefore, 180^0 phase reversal is observed during the negative half cycle of the input signal as well as shown in Figure 9.36(b).

9.4.7 Transistor as an oscillator

An electronic oscillator basically converts dc energy into ac energy of high frequency ranging from a few Hz to several MHz. Hence, it is a source of alternating current or voltage. Unlike an amplifier, oscillator does not require any external signal source.

Basically, there are two types of oscillators: **Sinusoidal and non-sinusoidal**. Sinusoidal oscillators generate oscillations in the form of sine waves at constant amplitude and frequency as shown in Figure 9.37(a). Whereas non-sinusoidal oscillators generate complex non-sinusoidal waveforms like

Square-wave, Triangular-wave or Sawtooth-wave as shown in Figure 9.36(b).

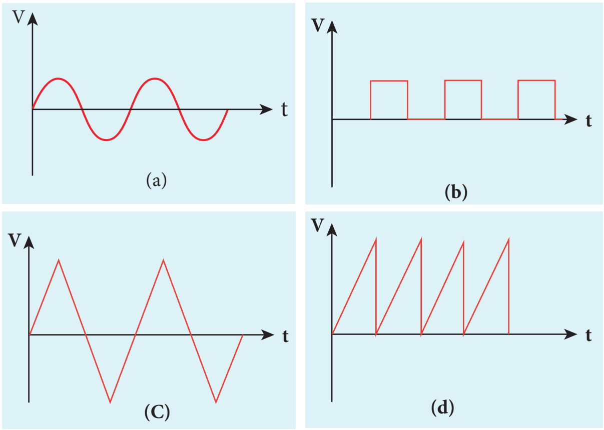


Figure: 9.37 (a) sinusoidal waveform (b) square waveform (c) ramp waveform (d) triangular waveform

Sinusoidal oscillations can be of two types: **Damped and undamped.** If the amplitude of the electrical oscillations decreases with time due to energy loss, it is called damped oscillations as shown in Figure 9.38(a). On the other hand, the amplitude of the electrical oscillations remains constant with time in undamped oscillations as shown in Figure 9.38(b).

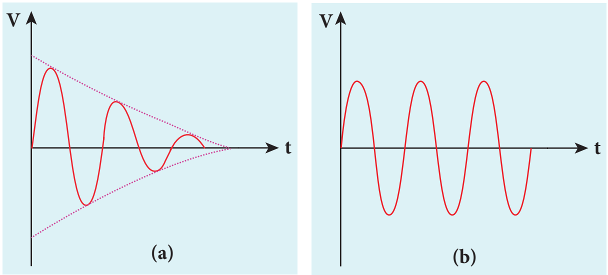


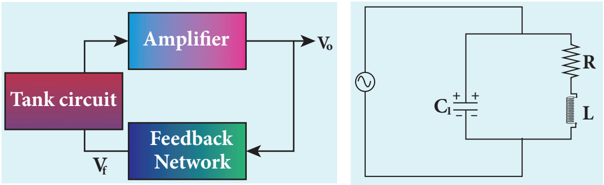
Figure 9.38 (a) Damped oscillations (b) Undamped oscillations

Transistor Oscillator

An oscillator circuit consists of a tank circuit, an amplifier and a feedback circuit as shown in Figure 9.39. The tank circuit generates electrical oscillations and acts as the ac input source to the transistor amplifier. Amplifier amplifies the input ac signal. The feedback circuit provides a portion of the output to the tank circuit to sustain the oscillations without energy loss. Hence, an oscillator does not require an external input signal. The output is said to be self-sustained.

Amplifier

The transistor amplifier circuit is already explained in section {9.4.5}.



(a) (b)

Figure 9.39 (a) Block diagram of an oscillator (b) tank circuit

Feedback network

The circuit used to feedback a portion of the output to the input is called the feedback network. If the portion of the output fed to the input is in phase with the input, then the magnitude of the input signal increases. It is necessary for sustained oscillations.

Tank circuit

The LC tank circuit consists of an inductance and a capacitor connected in parallel as shown in Figure 9.39. Whenever energy is supplied to the tank circuit from a DC source, the energy is stored in inductor and capacitor alternatively. This produces electrical oscillations of definite frequency. (Refer section 4.9.1, Volume 1 of XII std. Physics text book)

But in practical oscillator circuits there will be loss of energy across resistors, inductor coils and capacitors. A small amount of energy is used up in overcoming these losses during every cycle of charging and discharging of the capacitor. Due to this, the amplitude of the oscillations decreases gradually. Hence, the tank circuit produces damped electrical oscillations. Therefore, in order to produce undamped oscillations, a positive feedback is provided from the output circuit to the input circuit.

The frequency of oscillations is determined by the values of L and C using the equation.

$$f = \frac{1}{2\pi\sqrt{LC}}$$

Barkhausen conditions for sustained oscillations

The following condition called Barkhausen conditions should be satisfied for sustained oscillations in the oscillator.

- The loop phase shift must be 0^0 or integral multiples of 2π.
- The loop gain must be unity. $|A\beta| = 1$

Here, A→Voltage gain of the amplifier,
β →feedback ratio; (fraction of the output that is fed back to the input)

There are different types of oscillator circuits based on the different types of tank circuits. Examples: Hartley oscillator, Colpitt's oscillator, Phase shift oscillator, and Crystal oscillator.

Applications of oscillators

- to generate a periodic sinusoidal or non sinusoidal wave forms
- to generate RF carriers
- to generate audio tones
- to generate clock signal in digital circuits
- as sweep circuits in TV sets and CRO

EXAMPLE 9.9

Calculate the range of the variable capacitor that is to be used in a tuned-collector oscillator which has a fixed inductance of 150 μH. The frequency band is from 500 kHz to 1500 kHz.

Resonant frequency, $f_o = \frac{1}{2\pi\sqrt{LC}}$

On simplifying we get $C = \frac{1}{4\pi^2 f_o^2 L}$

When frequency is equal to 500 kHz

$$C = \frac{1}{4\times 3.14^2 \times (500\times 10^3)^2 \times 150\times 10^{-6}}$$

=676 pF

When frequency is equal to 1500 kHz

$$C = \frac{1}{4\times 3.14^2 \times (1500\times 10^3)^2 \times 150\times 10^{-6}}$$

= 75 pF

Therefore, the capacitor range is 75 – 676 pF

9.5 DIGITAL ELECTRONICS

Digital Electronics is the sub-branch of electronics which deals with digital signals. It is increasingly used in numerous applications ranging from high end processor circuits to miniature circuits for signal processing, communication etc. Digital signals are preferred than analog signals due to their better performance, accuracy, speed, flexibility, and immunity to noise.

9.5.1 Analog and Digital Signals

There are 2 different types of signals used in Electronics. They are (i) Analog signals and (ii) Digital signals. An analog signal is a continuously varying voltage or current with respect to time. Such signals have been employed in rectifying circuits and transistor amplifier circuits.

Digital signals are signals which contain only discreate values of voltages. Digital signals need two states: switch ON and OFF.

ON is considered as one state and OFF is considered as the other state. It can also be defined as high (ON) or low (OFF) state, closed (ON) or open (OFF). These high and low states are defined using binary numbers 1or 0 in Boolean Algebra. The state 1 represents the terms: circuit on, high voltage, a closed switch. Similarly a 0 state represents circuit off, low voltage or an open switch.

Positive and Negative Logic

In digital systems, there exists two voltage levels: 5V (high) and 0V (low). In a positive logic system; a binary 1 stands for 5V and 0 stands for 0V while in negative logic system, 1 state for 0V and 0 state for 5V as shown in Figure 9.40.

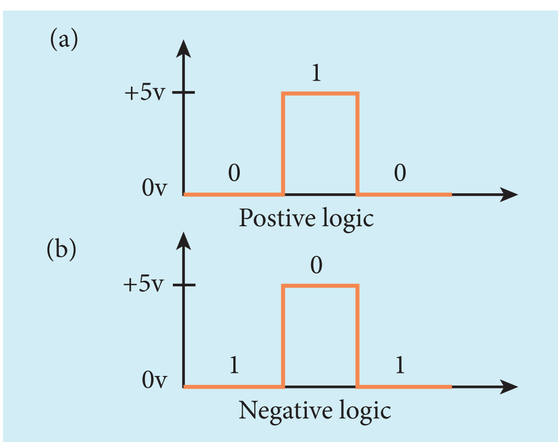


Figure 9.40 (a) Positive (b) Negative logics

9.5.2 Logic gates

A logic gate is an electronic circuit which functions based on digital signals. The logic gates are considered as the basic building blocks of most of the digital systems. It has one output with one or more inputs. There are three types of basic logic gates: AND, OR, and NOT. The other logic gates are Ex-OR, NAND, and NOR. They can be constructed from the basic logic gates.

Digital electronics deals with logical operations. The variables are called logical variables. The operators like logical addition (+) and logical multiplication (·) are called logical operators. When the logical operators (+, .) operate on logical variables (A, B), it gives logical constant (Y). The equation that represents this operation is called logical statement.

For example,

Logical operator: +

Logical variable: A, B

Logical constant: Y

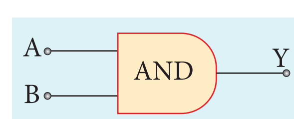
Logical Statement: Y = A + B

The possible combinations of inputs and the corresponding output is given in table called truth table. The circuits which perform the basic logical operations such as logical addition, multiplication and inversion are discussed below.

AND gate

Circuit symbol

The circuit symbol of a two input AND gate is shown in Figure 9.41(a). A and B are inputs and Y is the output. It is a logic gate and hence A, B, and Y can have the value of either 1 or 0.



(a)

Inputs		Output
A	B	$Y = A + B$
0	0	0
0	1	0
1	0	0
1	1	1

(b)

Figure 9.41 (a) Two input AND gate (b) Truth Table

Boolean equation:

$Y = A.B$

It performs logical multiplication and is different from arithmetic multiplication.

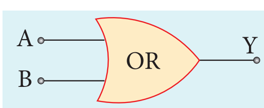
Logic operation

The output of AND gate is high (1) only when all the inputs are high (1). The rest of the cases the output is low. Hence the output of AND gate is high (1) only when all the inputs are high. It is represented in the truth table (Figure 9.41(b)).

OR gate

Circuit Symbol

The circuit symbol of a two input OR gate is shown in Figure 9.42(a). A and B are inputs and Y is the output.



(a)

Inputs		Output
A	B	$Y = A + B$
0	0	0
0	1	1
1	0	1
1	1	1

(b)

Figure 9.42 (a) Two input OR gate (b) Truth Table

Boolean equation:

$Y = A + B$

It performs logical addition and is different from arithmetic addition.

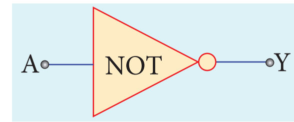
Logic operation

The output of OR gate is high (logic 1 state) when either of the inputs or both are high. The truth table of OR gate is shown in Figure 9.42(b).

NOT gate

Circuit symbol

The circuit symbol of NOT gate is shown in Figure 9.43(a). A is the input and Y is the output.



(a)

Inputs	Output
A	$Y = \overline{A}$
0	1
1	0

(b)

Figure 9.43 (a) NOT gate (b) Truth Table

Boolean equation

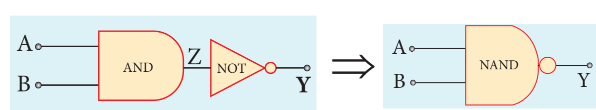
$Y = \overline{A}$

Logic operation

The output is the complement of the input. It is represented with an overbar. It is also called as inverter. The truth table infers that the output Y is 1 when input A is 0 and vice versa. The truth table of NOT is shown in Figure 9.43(b).

NAND gate

The circuit symbol of NAND gate is shown in Figure 9.44(a). A and B are inputs and Y is the output.



(a)

Input		Output (AND)	Output (NAND)
A	B	Z= A.B	$Y = \overline{A.B}$
0	0	0	1
0	1	0	1
1	0	0	1
1	1	1	0

(b)

Figure 9.44 (a)Two input NAND gate (b) Truth Table

Boolean equation

$$Y = \overline{A.B}$$

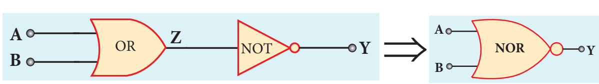
Logic operation

The output Y equals the complement of AND operation. The circuit is an AND gate followed by a NOT gate. Therefore, it is summarized as NAND. The output is at logic zero only when all the inputs are high. The rest of the cases, the output is high (Logic 1 state). The truth table of NAND gate is shown in Figure 9.44(b).

NOR gate

Circuit symbol

The circuit symbol of NOR gate is shown in Figure 9.45(a). A and B are inputs and Y is the output.



(a)

Inputs		Output (OR)	Output (NOR)
A	B	Z=A+B	$Y = \overline{A+B}$
0	0	0	1
0	1	1	0
1	0	1	0
1	1	1	0

(b)

Figure: 9.45 (a) NOR gate (b) Truth Table

Boolean equation

$$Y = \overline{A+B}$$

Logic operation

Y equals the complement of OR operation (A OR B). The circuit is an OR gate followed by a NOT gate and is summarized as NOR. The output is high when all the inputs are low. The output is low for all other combinations of inputs. The truth table of NOR gate is shown in Figure 9.45(b).

Ex-OR gate

Circuit symbol

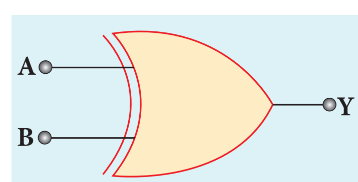
The circuit symbol of Ex-OR gate is shown in Figure 9.46(a). A and B are inputs and Y is the output. The Ex-OR operation is denoted as ⊕.

Boolean equation

$$Y = A.\overline{B} + \overline{A}.B$$
$$Y = A \oplus B$$

Logic operation

The output is high only when either of the two inputs is high. In the case of an Ex-OR gate with more than two inputs, the output will be high when odd number of inputs are high. The truth table of Ex-OR gate is shown in Figure 9.46(b).



(a)

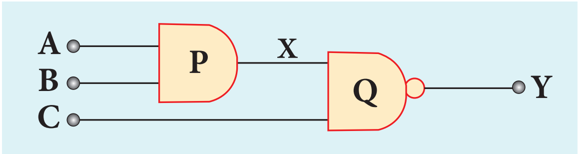
Inputs		Output (Ex-OR)
A	B	$Y = A \oplus B$
0	0	0
0	1	1
1	0	1
1	1	0

(b)

Figure 9.46 (a) Ex-OR gate (b) Truth Table

EXAMPLE 9.10

What is the output Y in the following circuit, when all the three inputs A, B, and C are first 0 and then 1?

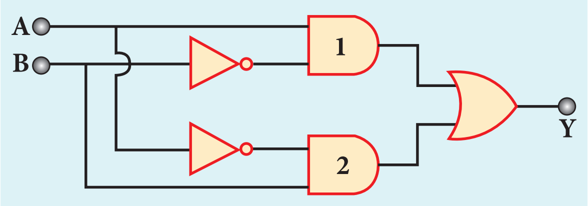


Solution

A	B	C	X = A·B	$Y=\overline{X\cdot C}$
0	0	0	0	1
1	1	1	1	0

EXAMPLE 9.11

In the combination of the following gates, write the Boolean equation for output Y in terms of inputs A and B.



Solution

The output at the 1st AND gate: $A\overline{B}$

The output at the 2nd AND gate: $\overline{A}\,B$

The output at the OR gate: $Y=A.\overline{B}+\overline{A}.B$

9.6 BOOLEAN ALGEBRA

Boolean Algebra is basically a choice between two options (i) yes or no (ii) high or low. These two options in Boolean algebra are represented by binary numbers 0 or 1. It is a concept that relates logic and mathematics which is a century old, made up by George Boole in 1854. Later the importance of Boolean algebra was realized in the design of computer circuits. Today we are in a digital world and most of the comforts that we experience is due to digitization with the foundation based on Boolean algebra.

DO YOU KNOW? The concept of high (1) and low (0) is not a new one. In fact, it was applied in telephone switching circuits by Shannon in 1938.

Laws of Boolean Algebra

The NOT, OR and AND operations discussed in 9.5.2 are the Boolean operations. The results of these operations can be summarised as:

Complement law

A	$Y=\overline{A}$
0	$Y=\overline{0}=1$
1	$Y=\overline{1}=0$

The complement law can be realised as $\overline{\overline{A}}=A$

OR laws

A	B	Y = A+B
0	0	Y = 0+0 = 0
0	1	Y = 0+1 = 1
1	0	Y = 1+0 = 1
1	1	Y = 1+1 = 1

The OR laws can be realised as

1st law	A + 0 = A
2nd law	A + 1 = 1
3rd law	A + A = A
4th law	$A+\overline{A}=1$

AND laws

A	B	Y = A.B
0	0	Y = 0.0 = 0
0	1	Y = 0.1 = 0
1	0	Y = 1.0 = 0
1	1	Y = 1.1 = 1

The AND laws can be realised as

1st law	A . 0 = 0
2nd law	A . 1 = A
3rd law	A . A = A
4th law	$A \cdot \overline{A} = 0$

The Boolean operations obey the following laws.

Commutative laws

A + B = B + A

A . B = B . A

Associative laws

A + (B + C) = (A + B) + C

A . (B . C) = (A .B) . C

Distributive laws

A(B + C) = AB + AC

A + BC = (A + B) (A + C)

The above laws are used to simplify complicated expressions and to simplify the logic circuitry.

9.7 DE MORGAN'S THEOREM

9.7.1 De Morgan's First Theorem

The first theorem states that the complement of the sum of two logical inputs is equal to the product of its complements.

Proof

The Boolean equation for NOR gate is $Y = \overline{A+B}$

The Boolean equation for a bubbled AND gate is $Y = \overline{A}.\overline{B}$

Both cases generate same outputs for same inputs. It can be verified using the following truth table.

A	B	A+B	$\overline{A+B}$	$\overline{A}$	$\overline{B}$	$\overline{A}.\overline{B}$
0	0	0	1	1	1	1
0	1	1	0	1	0	0
1	0	1	0	0	1	0
1	1	1	0	0	0	0

From the above truth table, we can conclude $\overline{A+B} = \overline{A} \cdot \overline{B}$.

Thus De Morgan's First Theorem is proved. It also says that a NOR gate is equal to a bubbled AND gate.

The corresponding logic circuit diagram is shown in Figure 9.47.

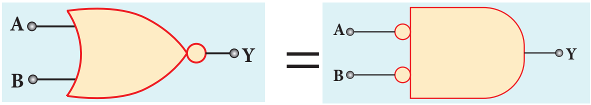


Figure 9.47 NOR gate equals bubbled AND gate

9.7.2 De Morgan's Second Theorem

The second theorem states that the complement of the product of two inputs is equal to the sum of its complements.

Proof

The Boolean equation for NAND gate is $Y = \overline{AB}$

The Boolean equation for bubbled OR gate is $Y = \overline{A} + \overline{B}$

A and B are the inputs and Y is the output. The above two equations produces the same output for the same inputs. It can be verified by using the truth table

A	B	A.B	$\overline{A.B}$	$\overline{A}$	$\overline{B}$	$\overline{A}+\overline{B}$
0	0	0	1	1	1	1
0	1	0	1	1	0	1
1	0	0	1	0	1	1
1	1	1	0	0	0	0

From the above truth table we can conclude $\overline{A.B}=\overline{A}+\overline{B}$

Thus De Morgan's First Theorem is proved. It also says, a NAND gate is equal to a bubbled OR gate.

The corresponding logic circuit diagram is shown in Figure 9.48

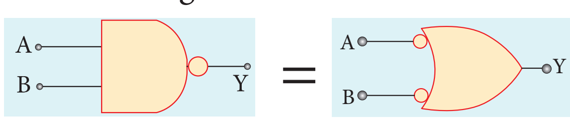


Figure 9.48 NAND gate equals bubbled OR gate

EXAMPLE:9.12

Simplify the Boolean identity

$$AC + ABC = AC$$

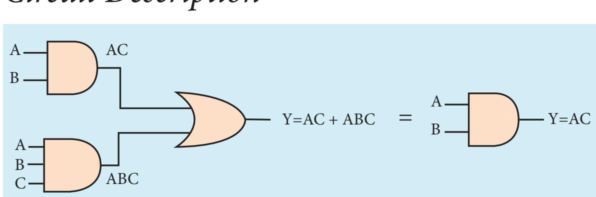
Solution

Step 1: AC (1 + B) = AC.1 [OR law-2]

Step 2: AC . 1 = AC [AND law – 2]

Therefore, AC + ABC = AC

Circuit Description



Thus the given statement is proved.

9.7.3 Integrated Chips

An integrated circuit is also referred as an IC or a chip or a microchip (Figure 9.49). It consists of thousands to millions of transistors, resistors, capacitors, etc. integrated on a small flat piece of semiconductor material that is normally Silicon.

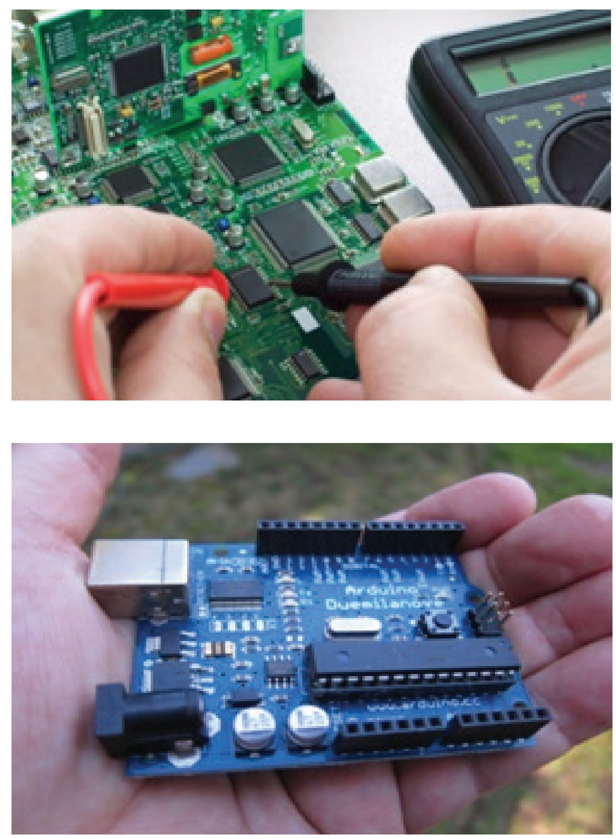


Figure 9.49 Circuits with integrated chips

Integrated circuits (ICs) are the keystone of modern electronics. With the advancement in technology and the emergence of Very Large Scale Integration (VLSI) era it is possible to fit more and more transistors on chips of same piece.

ICs have two main advantages over ordinary circuits: cost and performance. The size, speed, and capacity of chips have progressed enormously with the advancement in technology. Computers, mobile phones, and other digital home appliances are now made possible by the small size and low cost of ICs. ICs can function as an amplifier, oscillator, timer, microprocessor and computer memory.

These extremely small ICs can perform calculations and store data using either digital or analog technology. Digital ICs use logic gates, which work only with values of ones and zeros. A low signal sent to a component on a digital IC will result in a value of 0, while a high signal creates a value of 1.

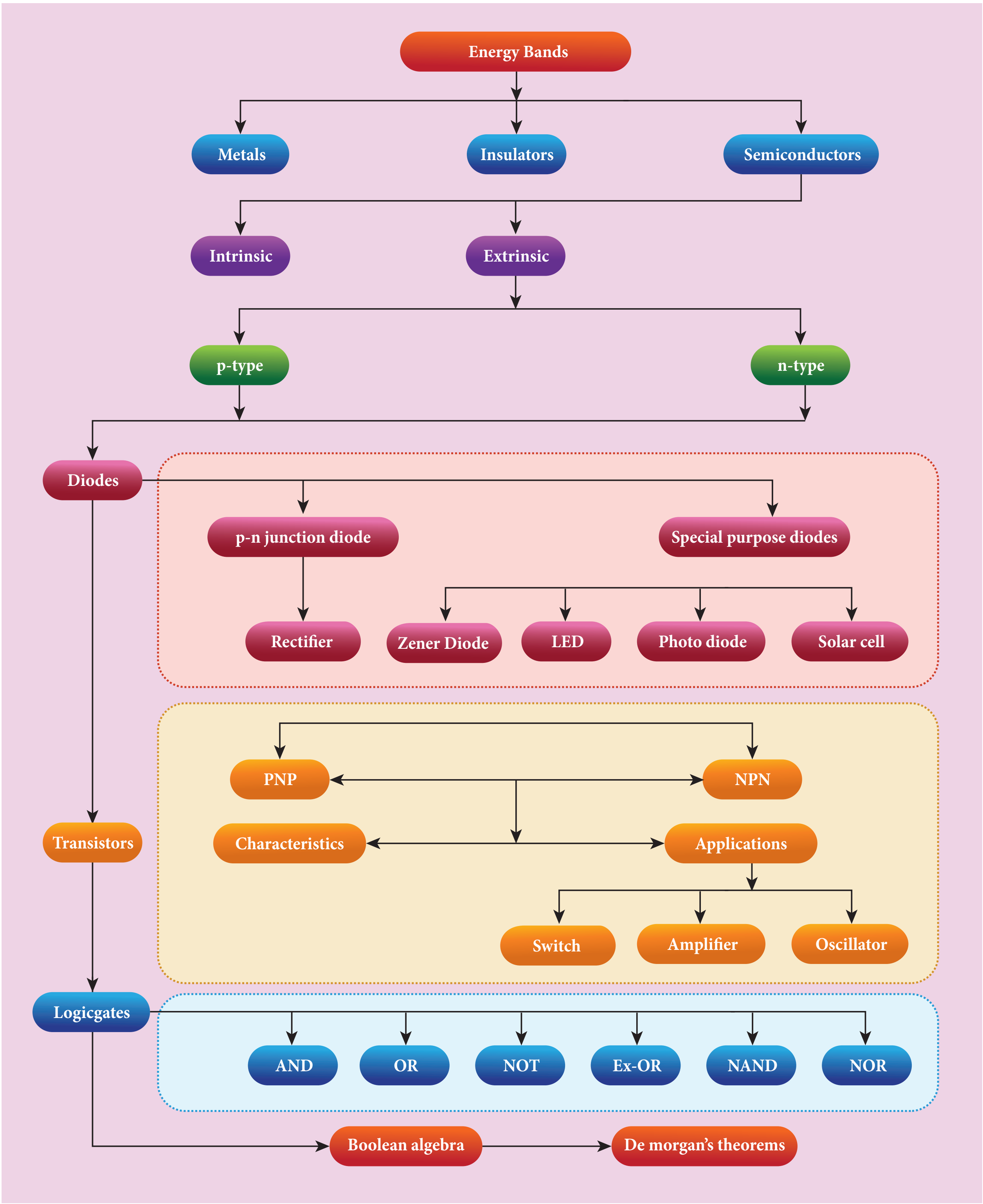
Digital ICs usually find their applications in computers, networking equipment, and most consumer electronics. **Analog ICs** or linear ICs work with continuous values. This means a component on a linear IC can take any value and output another value. Linear ICs are typically used in audio and radio frequency amplification.

SUMMARY

- Energy bands in solids are used to classify them into metals, insulators, and semiconductors
- In a N-type semiconductor, electrons are the majority charge carriers and holes are the minority carriers
- In P-type semiconductor, holes are the majority charge carriers and electrons are the minority charge carriers
- A depletion region is formed in an unbiased PN junction. It is devoid of mobile charge carriers. Instead, it has immobile ions
- When a PN junction diode is forward biased, the depletion region decreases and the diode conducts once after the barrier potential is crossed. It acts like a closed switch.
- A PN junction diode in reverse biased condition functions as a open switch as it does not conduct. The depletion region increases.
- A forward biased PN junction diode functions as a rectifier. Rectification is the process of converting an AC current into DC current
- The half wave rectifier rectifies one half of the input signal and produces a pulsating output.
- Full wave rectifier rectifies both halves of the input signal.
- The efficiency of the full wave rectifier is two times the efficiency of the half wave rectifier
- The two mechanisms that is responsible for breakdown under increasing reverse voltage: Zener and Avalanche breakdown
- Zener breakdown happens in a heavily doped PN junction diode when a strong electric field is applied.
- Avalanche breakdown occurs in lightly doped junctions which have wide depletion layers. It is due to the breaking of covalent bonds by the thermally generated minority charge carriers.

- Zener diode is a heavily doped PN junction diode works in the reverse biased direction
- Light emitting diode is a forward biased semiconductor device that emits emits visible or invisible light when energized. The recombination of minority charge carriers with the majority charge carriers in the respective regions release energy in the form of Photons.
- A PN junction diode made of photosensitive material converts an optical signal into electric signal is called a photodiode.
- When a photon of sufficient energy strikes the diode, it creates an electron-hole pair. These electrons and holes are swept across the p-n junction by the electric field created by reverse voltage before recombination takes place and in turn generates photo current.
- A solar cell is an electrical device that converts the energy of light directly into electricity by the photovoltaic effect.
- A bipolar junction transistor is a semiconductor device is of two types: NPN and PNP.
- BJT has three regions: emitter, base, and collector
- To operate the transistor in the active region, emitter base must be forward biased and collector base must be reverse biased.
- A BJT can be operated in three different configurations: Common base, common emitter, common collector.
- The forward current gain in common base configuration α gives the ratio of the collector current to emitter current.
- The forward current gain in common emitter configuration β gives the ratio of the collector current to the base current
- The BJT connected in common emitter configuration functions as a switch
- The BJT connected in common emitter configuration can be used as an amplifier. There exists a phase reversal of 180^0 between the input signal and the amplified output signal.
- A transistor amplifier combined with a tank circuit and positive feedback acts as an oscillator
- The logic gates are logical circuits provides output only for a combination of inputs.
- The basic logic gates are AND, OR, and NOT gates.
- Boolean algebra is used to simplify complicated expressions and hence to simplify the logic circuit.
- De Morgan's First theorem states that the complement of the sum of two inputs is equal to the product of its complements.
- The second theorem states that the complement of the product of two inputs is equal to the sum of its complements.

CONCEPT MAP



Energy Bands
Metals
Insulators
Semiconductors
Intrinsic
Extrinsic
p-type
n-type
Diodes
p-n junction diode
Special purpose diodes
Rectifier
Zener Diode
LED
Photo diode
Solar cell
PNP
NPN
Transistors
Characteristics
Applications
Switch
Amplifier
Oscillator
Logicgates
AND
OR
NOT
Ex-OR
NAND
NOR
Boolean algebra
De morgan's theorems

EVALUATION

I. Multiple choice questions

1. The barrier potential of a silicon diode is approximately,

   a. 0.7 V b. 0.3V

   c. 2.0 V d. 2.2V

2. Doping a semiconductor results in

   a. The decrease in mobile charge carriers

   b. The change in chemical properties

   c. The change in the crystal structure

   d. The breaking of the covalent bond

3. A forward biased diode is treated as

   a. An open switch with infinite resistance

   b. A closed switch with a voltage drop of 0V

   c. A closed switch in series with a battery voltage of 0.7V

   d. A closed switch in series with a small resistance and a battery.

4. If a half –wave rectified voltage is fed to a load resistor, which part of a cycle the load current will flow?

   a. 0^0–90^0 b. 90^0–180^0

   c. 0^0–180^0 d. 0^0–360^0

5. The primary use of a zener diode is

   a. Rectifier b. Amplifier

   c. Oscillator d. Voltage regulator

6. The principle in which a solar cell operates

   a. Diffusion

   b. Recombination

   c. Photovoltaic action

   d. Carrier flow

7. The light emitted in an LED is due to

   a. Recombination of charge carriers

   b. Reflection of light due to lens action

   c. Amplification of light falling at the junction

   d. Large current capacity.

8. When a transistor is fully switched on, it is said to be

   a. Shorted

   b. Saturated

   c. Cut-off

   d. Open

9. The specific characteristic of a common emitter amplifier is

   a. High input resistance

   b. Low power gain

   c. Signal phase reversal

   d. Low current gain

10. To obtain sustained oscillation in an oscillator,

    a. Feedback should be positive

    b. Feedback factor must be unity

    c. Phase shift must be 0 or 2π

    d. All the above

11. If the input to the NOT gate is A = 1011, its output is

    a. 0100 b. 1000

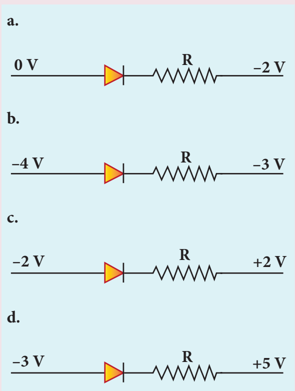
    c. 1100 d. 0011

12. The electrical series circuit in digital form is

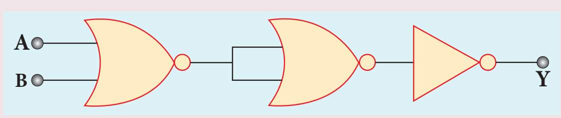
    a. AND b. OR

    c. NOR d. NAND

13. Which one of the following represents forward bias diode? (NEET)



14. The given electrical network is equivalent to (NEET)



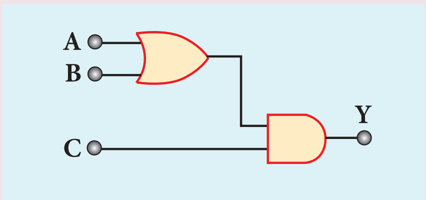
a. AND gate

b. OR gate

c. NOR gate

d. NOT gate

15. The output of the following circuit is 1 when the input ABC is (NEET 2016)



a. 101

b. 100

c. 110

d. 010

Answers

1. a	2. c	3. d	4. c	5. d
6. c	7. a	8. b	9. c	10. d
11. a	12. a	13. a	14. c	15. a

II. Short Answer Questions

1. Define electron motion in a semiconductor.
2. Distinguish between intrinsic and extrinsic semiconductors.
3. What do you mean by doping?
4. How electron-hole pairs are created in a semiconductor material?
5. A diode is called as a unidirectional device. Explain
6. What do you mean by leakage current in a diode?
7. Draw the output waveform of a full wave rectifier.
8. Distinguish between avalanche and zener breakdown.
9. Discuss the biasing polarities in an NPN and PNP transistors.
10. Explain the current flow in a NPN transistor
11. What is the phase relationship between the AC input and output voltages in a common emitter amplifier? What is the reason for the phase reversal?
12. Explain the need for a feedback circuit in a transistor oscillator.
13. Give circuit symbol, logical operation, truth table, and Boolean expression of AND, OR, NOT, NAND, NOR, and EX-OR gates
14. State De Morgan's first and second theorems.

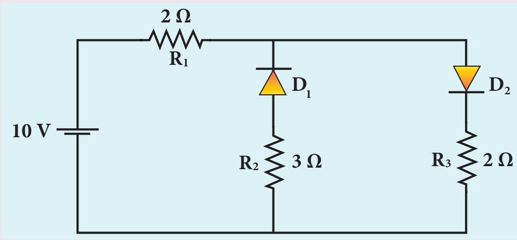
III. Long Answer Questions

1. Elucidate the formation of a N-type and P-type semiconductors.
2. Explain the formation of PN junction diode. Discuss its V–I characteristics.

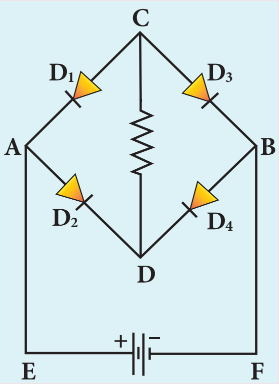
3. Draw the circuit diagram of a half wave rectifier and explain its working
4. Explain the construction and working of a full wave rectifier.
5. What is an LED? Give the principle of operation with a diagram.
6. Write notes on Photodiode.
7. Explain the working principle of a solar cell. Mention its applications.
8. Sketch the static characteristics of a common emitter transistor and bring out the essence of input and output characteristics.
9. Describe the function of a transistor as an amplifier with the neat circuit diagram. Sketch the input and output wave form.
10. Transistor functions as a switch. Explain.
11. State Boolean laws. Elucidate how they are used to simplify Boolean expressions with suitable example.
12. State and prove De Morgan's First and Second theorems.

IV. Numerical Problems

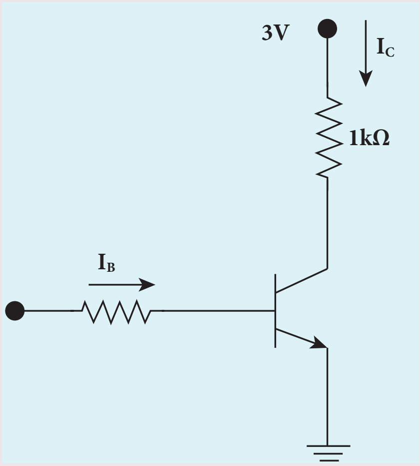
1. The given circuit has two ideal diodes connected as shown in figure below. Calculate the current flowing through the resistance R_1 [Ans: 2.5 A]



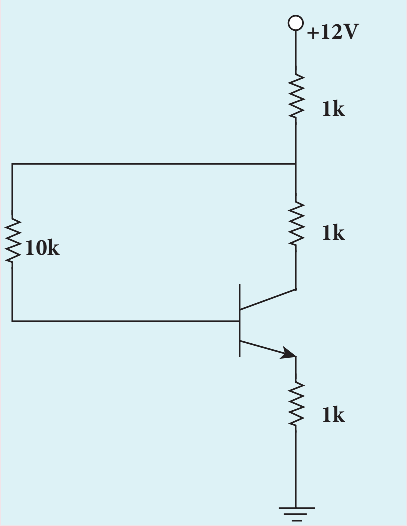
2. Four silicon diodes and a 10 Ω resistor are connected as shown in figure below. Each diode has a resistance of 1Ω. Find the current flows through the 18Ω resistor. [Ans: 0.13 A]



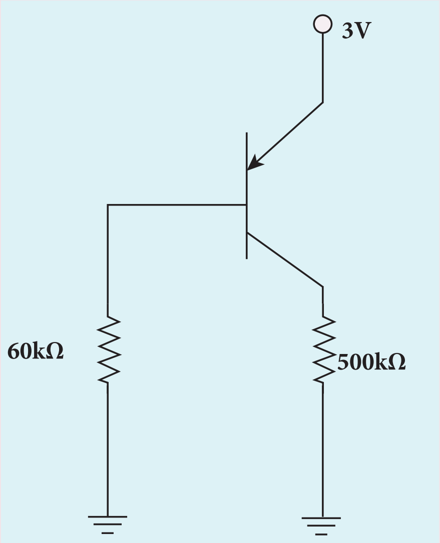
3. Assuming V_{CEsat} = 0.2 V and β = 50, find the minimum base current (I_B) required to drive the transistor given in the figure to saturation. [Ans: 56 μA]



4. A transistor having α =0.99 and V_{BE} = 0.7V, is given in the circuit. Find the value of the collector current.
[Ans: 5.33 mA]



5. In the circuit shown in the figure, the BJT has a current gain (β) of 50. For an emitter – base voltage V_{EB} = 600 mV, calculate the emitter – collector voltage V_{EC} (in volts). [Ans: 2 V]



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ICT CORNER

Semiconductor electronics

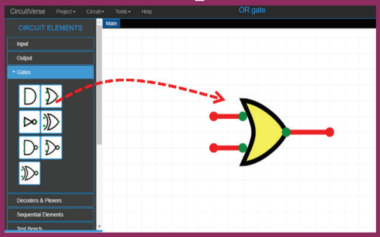
In this activity you will be able to
(i) Construct , manipulate and simulate the logic circuits. (ii) verify the truth tables of AND, OR, NOT, EX-OR, NAND and NOR gates

Topic: Logic gates

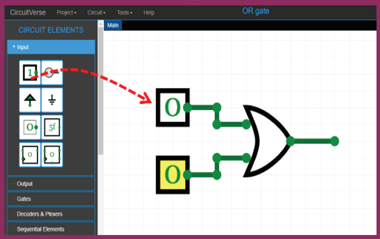
STEPS:

- Open the browser and type "circuitverse.org/simulator" in the address bar.
- Click 'Gates' tab from the circuit elements. Select the gate you want to verify and drag it in to the stage.
- Nodes in the logic gates are connected through wires. Wires can be drawn by dragging from the nodes with the help of mouse.
- Select 'input tool' from input tab. Drag and keep it as two inputs.
- Select 'output tool' or 'digital LED' from output tab. Drag and keep it as output.
- Verify the truth tables of AND, OR, NOT, EX-OR, NAND and NOR gates. You can verify De Morgan's first and second theorems.

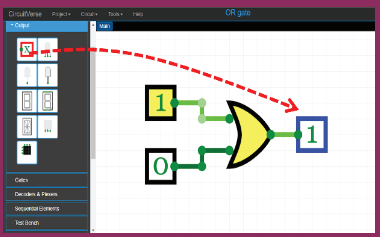
Step1



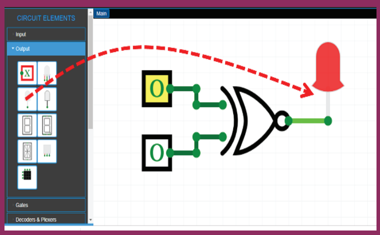
Step2



Step3



Step4



Note:

Login with the help of your mail id if you want to save your project in online.

URL:

https://circuitverse.org/simulator

* Pictures are indicative only.

* If browser requires, allow **Flash Player** or **Java Script** to load the page.

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UNIT 10

COMMUNICATION SYSTEMS

Good communication is the bridge between confusion and clarity

– Nat Turner

In this unit, the students are exposed to

- Basic elements of communication system
- Need for modulation and its types
- Propagation of electromagnetic waves through space
- Satellite communication
- Fiber optic communication
- RADAR
- Internet
- Global positioning system
- Applications of communication technology in fishing, mining, and agriculture sectors.

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10.1 INTRODUCTION

Communication exists since the dawn of life in this world. Growth in science and technology removed the locational disadvantage effectively. Information can be exchanged from one person to another anywhere on this Earth. Right from the developments made in communication by great scientists like J.C. Bose, G. Marconi, and Alexander Graham Bell, communication has witnessed leaps and bounds. The communication industry is one of the largest in size and is the oldest since communication through telegraph (1844), telephone (1876), and Radio (1887) started centuries back. The intensive research in the mid- and late nineteenth century leads to the development of long-distance transmission in the shortest possible time. However, the 20th century witnessed a leap over the development of communication, meeting the demands of speed and secured transfer of data. Every sector in the world is experiencing a significant profit with the advent of Global Positioning System (GPS), satellite, mobile, and optical communications. This unit provides a glimpse of the basic concepts of electronic communication and its applications.

10.2 MODULATION

The transmission of information through short distances does not require complicated techniques. The energy of the information signal is sufficient enough to be sent directly. However if the information, for example, audio frequency (20 to 20,000 Hz) needs to

UNIT-10(XII-Physics_Vol-2).indd 238 8/3/2019 4:54:29 PM

be transmitted to long distances across the world, certain techniques are required to transmit the information without any loss.

For long distance transmission, the low frequency baseband signal (input signal) is superimposed onto a high frequency radio signal by a process called modulation. In the modulation process, a very high frequency signal called carrier signal (radio signal) is used to carry the baseband signal.

As the frequency of the carrier signal is very high, it can be transmitted to long distances with less attenuation. The carrier signal is usually a sine wave signal. Also, the carrier signal will be more compatible with the communication medium like free space and can propagate with greater efficiency.

Note Carrier signal does not have information.

A sinusoidal carrier wave can be represented as $e_c = E_c \sin(2\pi\nu_c t + \phi)$, where E_c is the amplitude, ν_c is the frequency and ϕ is the initial phase of the carrier wave at any instant of time t.

Three characteristics in the carrier signal can be modified by the baseband signal during the process of modulation: amplitude, frequency and phase of the carrier signal.

There are 3 types of modulation based on which parameter is modified. They are (i) amplitude modulation, (ii) frequency modulation and (iii) phase modulation.

10.2.1 AMPLITUDE MODULATION (AM)

If the amplitude of the carrier signal is modified in proportion to the instantaneous amplitude of the baseband signal, then it is called amplitude modulation. Here the frequency and the phase of the carrier signal remain constant. Amplitude modulation is used in radio and TV broadcasting.

The signal shown in Figure 10.1(a) is the baseband signal that carries information. Figure 10.1(b) shows the high-frequency carrier signal and Figure 10.1(c) gives amplitude modulated signal. We can see that amplitude of the carrier is modified in proportion to the amplitude of the baseband signal.

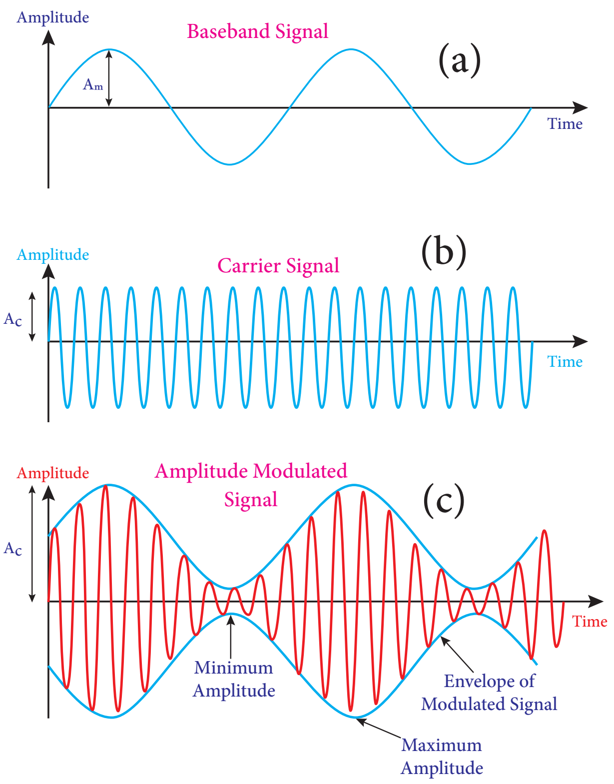

Figure 10.1 Amplitude Modulation (a) baseband signal (b) carrier signal (c) modulated signal

Advantages of AM

i) Easy transmission and reception
ii) Lesser bandwidth requirements
iii) Low cost

Limitations of AM

i) Noise level is high
ii) Low efficiency
iii) Small operating range

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10.2.2 FREQUENCY MODULATION (FM)

The frequency of the carrier signal is modified in proportion to the instantaneous amplitude of the baseband signal in frequency modulation. Here the amplitude and the phase of the carrier signal remain constant. Increase in the amplitude of the baseband signal increases the frequency of the carrier signal and vice versa. This leads to compressions and rarefactions in the frequency spectrum of the modulated wave as shown in Figure 10.2. Louder signal leads to compressions and relatively weaker signals to rarefactions.

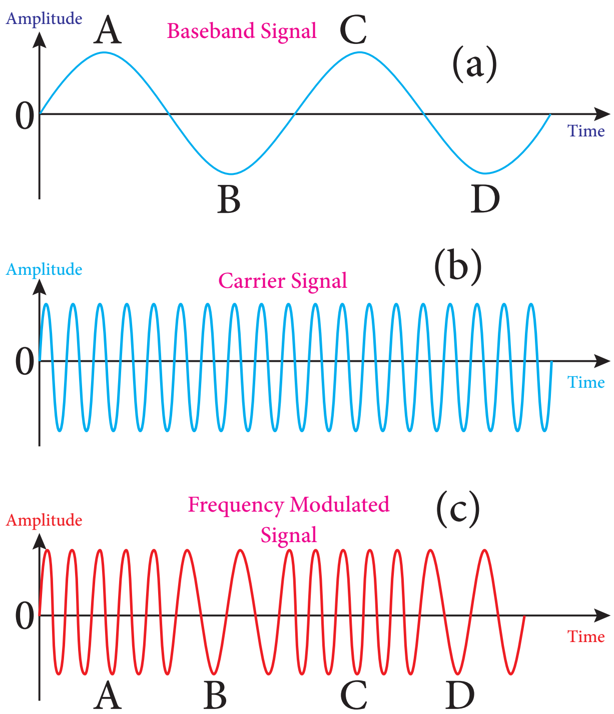


Figure 10.2 Frequency Modulation (a) baseband signal (b) carrier signal (c) frequency modulated signal

When the amplitude of the baseband signal is zero in Figure 10.2(a), the frequency of the modulated signal is the same as the carrier signal. The frequency of the modulated wave increases when the amplitude of the baseband signal increases in the positive direction (A, C). The increase in amplitude in the negative half cycle (B, D) reduces the frequency of the modulated wave (Figure 10.2(c)).

When the frequency of the baseband signal is zero (no input signal), there is no change in the frequency of the carrier wave. It is at its normal frequency and is called as **centre frequency or resting frequency.** Practically this is the allotted frequency of the FM transmitter.

Advantages of FM

i) Large decrease in noise. This leads to an increase in signal-noise ratio.
ii) The operating range is quite large.
iii) The transmission efficiency is very high as all the transmitted power is useful.
iv) FM bandwidth covers the entire frequency range which humans can hear. Due to this, FM radio has better quality compared to AM radio.

Limitations of FM

i) FM requires a much wider channel.
ii) FM transmitters and receivers are more complex and costly.
iii) In FM reception, less area is covered compared to AM.

10.2.3 PHASE MODULATION (PM)

In phase modulation, the instantaneous amplitude of the baseband signal modifies the phase of the carrier signal keeping the amplitude and frequency constant (Figure 10.3). This modulation is used to generate frequency modulated signals. It is similar to frequency modulation except that the phase of the carrier is varied instead of varying frequency.

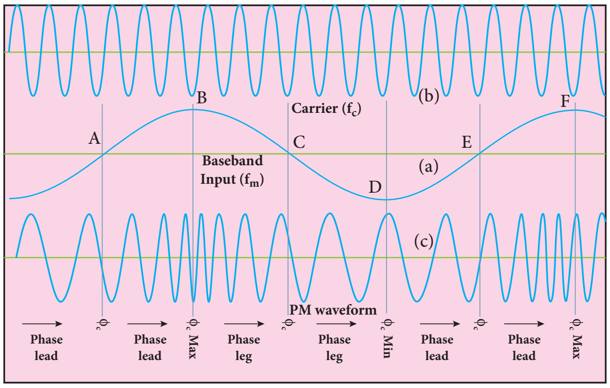


Figure 10.3 Phase Modulation (a) carrier signal (b) baseband signal (c) phase modulated signal

The carrier phase changes according to increase or decrease in the amplitude of the baseband signal. When the modulating signal goes positive, the amount of phase lead increases with the amplitude of the modulating signal. Due to this, the carrier signal is compressed or its frequency is increased.

On the other hand, the negative half cycle of the baseband signal produces a phase lag in the carrier signal. This appears to have stretched the frequency of the carrier wave. Hence similar to frequency modulated wave, phase modulated wave also comprises of compressions and rarefactions. When the signal voltage is zero (A, C and E) the carrier frequency is unchanged.

The frequency shift in carrier wave frequency exists in phase modulation as well. The frequency shift depends on (i) amplitude of the modulating signal and (ii) the frequency of the signal.

Note

- If a square wave is used as the baseband signal, then phase reversal takes place in the modulated signal.
- FM and PM waves are completely different for square wave modulating signal.

Advantages of PM

i) FM signal produced from PM signal is very stable.

ii) The centre frequency called resting frequency is extremely stable.

Note

Comparison between FM and PM

PM wave is similar to FM wave. PM generally uses a smaller bandwidth than FM. In other words, in PM, more information can be sent in a given bandwidth. Hence, phase modulation provides high transmission speed on a given bandwidth.

10.3 THE ELEMENTS OF AN ELECTRONIC COMMUNICATION SYSTEM

Electronics plays a major role in communication. Electronic communication is nothing but the transmission of sound, text, pictures, or data through a medium. Long distance transmission uses free space as a medium. This section provides sufficient information on how voice signal is transmitted by a transmitter through space and received by the receiver at the receiving end.

Elements of an electronic communication system

The elements of the basic communication system are explained with the block diagram shown in Figure 10.4.

1. Information (Baseband or input signal)

Information can be in the form of speech, music, pictures, or computer data.

This information is given as input to the input transducer.

2. Input transducer

A transducer is a device that converts variations in a physical quantity (pressure, temperature, sound) into an equivalent electrical signal or vice versa. In communication system, the transducer converts the information which is in the form of sound, music, pictures or computer data into corresponding electrical signals. **The electrical equivalent of the original information is called the baseband signal.** The best example for the transducer is the microphone that converts sound energy into electrical energy.

3. Transmitter

It feeds the electrical signal from the transducer to the communication channel. It consists of circuits such as amplifier, oscillator, modulator and power amplifier. The transmitter is located at the broadcasting station.

Amplifier: The transducer output is very weak and is amplified by the amplifier.

Oscillator: It generates high-frequency carrier wave (a sinusoidal wave) for long distance transmission into space. As the energy of a wave is proportional to its frequency, the carrier wave has very high energy.

Modulator: It superimposes the baseband signal onto the carrier signal and generates the modulated signal.

Power amplifier: It increases the power level of the electrical signal in order to cover a large distance.

4. Transmitting antenna

It radiates the radio signal into space in all directions. It travels in the form of electromagnetic waves with the velocity of light (3×10^8 m s^{-1}).

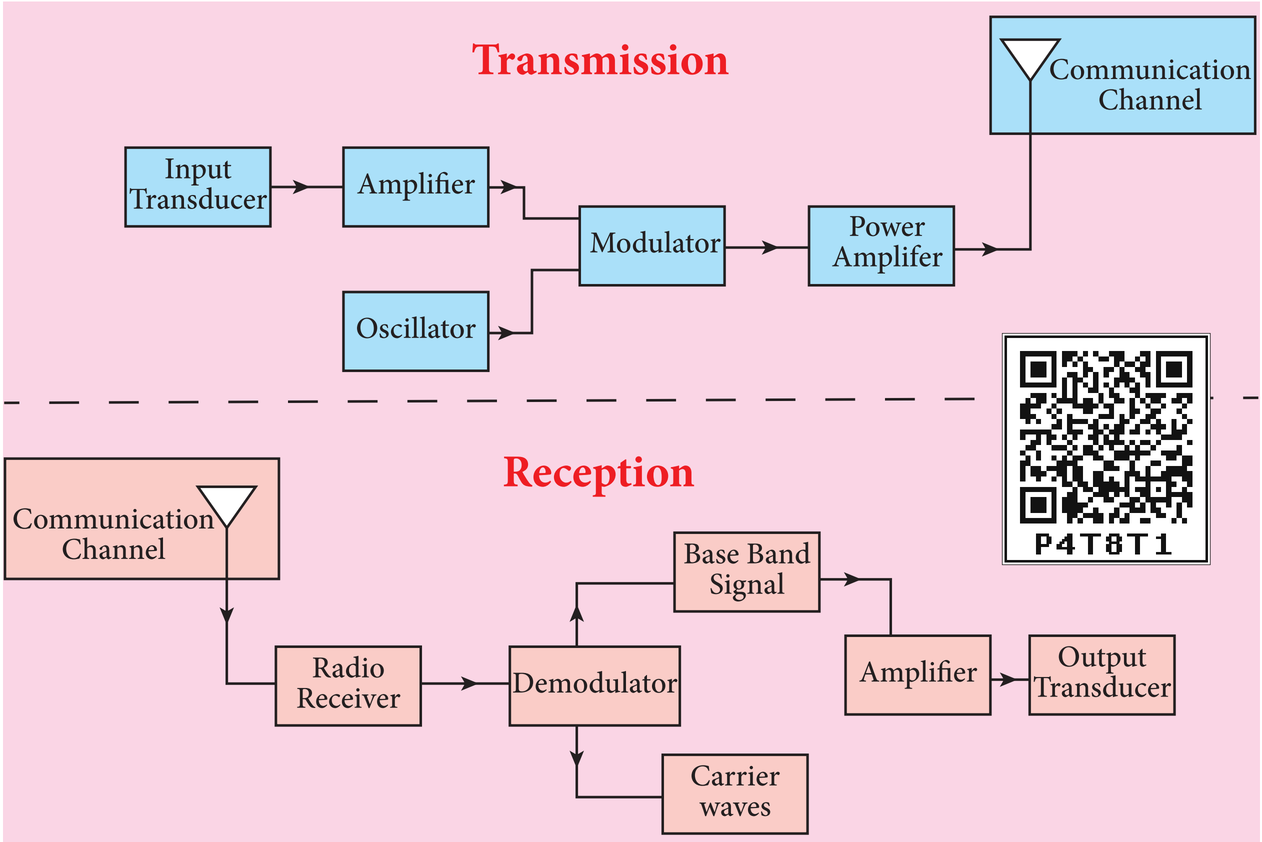


Figure 10.4 Block diagram of transmission and reception of voice signals

5. **Communication channel**

Communication channel is used to carry the electrical signal from transmitter to receiver with less noise or distortion. The communication medium is basically of two types: wireline communication and wireless communication.

Wireline communication (point to point communication) uses mediums like wires, cables and optical fibers. These systems cannot be used for long distance transmission as they are connected physically. Examples are telephone, intercom and cable TV.

Wireless communication uses free space as a communication medium. The signals are transmitted in the form of electromagnetic waves with the help of a transmitting antenna. Hence wireless communication is used for long distance transmission. Examples are mobile, radio or TV broadcasting and satellite communication.

6. **Noise**

It is the undesirable electrical signal that interferes with the transmitted signal. Noise attenuates or reduces the quality of the transmitted signal. It may be man-made (automobiles, welding machines, electric motors etc.) or natural (lightning, radiation from sun and stars and environmental effects). Noise cannot be completely eliminated. However, it can be reduced using various techniques.

7. **Receiver**

The signals that are transmitted through the communication medium are received by a receiving antenna which converts em waves into RF signals and are fed into the receiver. The receiver consists of electronic circuits like demodulator, amplifier, detector etc. The demodulator extracts the baseband signal from the modulated signal. Then the baseband signal is detected and amplified using amplifiers. Finally, it is fed to the output transducer.

8. **Repeaters**

Repeaters are used to increase the range or distance through which the signals are sent. It is a combination of transmitter and receiver. The signals are received, amplified and retransmitted with a carrier signal of different frequency to the destination. The best example is the communication satellite in space.

9. **Output transducer**

It converts the electrical signal back to its original form such as sound, music, pictures or data. Examples of output transducers are loudspeakers, picture tubes, computer monitor, etc.

10. **Attenuation**

The loss of strength of a signal while propagating through a medium is known as attenuation.

11. **Range**

It is the maximum distance between the source and the destination up to which the signal is received with sufficient strength.

10.3.1 BANDWIDTH

The frequency range over which the baseband signals or the information signals such as voice, music, picture etc is transmitted is known as bandwidth. Each of these signals has different frequencies.

The type of communication system depends on the nature of the frequency band for a given signal. Bandwidth gives the difference between the upper and lower frequency limits of the signal. It can also be defined as the portion of the electromagnetic spectrum occupied by the signal. If v_1 and v_2 are the lower and upper-frequency limits of a signal, then the bandwidth, BW = $v_2 - v_1$.

10.3.2 BANDWIDTH OF TRANSMISSION SYSTEM

The range of frequencies required to transmit a piece of specified information in a particular channel is called channel bandwidth or the bandwidth of the transmission system. This corresponds to the spectrum that is assigned to be used by the system. For example, amplitude modulation system requires a channel bandwidth of 10 kHz to transmit a 5 kHz signal, whereas a single side-band system requires only a 5 kHz channel bandwidth for the same 5 kHz signal. This is because in amplitude modulation, the channel bandwidth is twice the signal frequency. Therefore, it is required to reduce the channel bandwidth to accommodate more number of channels in the available electromagnetic spectrum. In some applications, modulation is selected based on this.

10.4 ANTENNA SIZE

Antenna is used at both transmitter and receiver end. Antenna height is an important parameter to be discussed. The height of the antenna must be a multiple of $\frac{\lambda}{4}$.

$$h = \frac{\lambda}{4} \tag{10.1}$$

where λ is wavelength ($\lambda = \frac{c}{v}$), c is the velocity of light and v is the frequency of the signal to be transmitted.

An example

Let us consider two baseband signals. One signal is modulated and the other is not modulated.

The frequency of the original baseband signal is taken as $v = 10\ kHz$ while the modulated signal is $v = 1\ MHz$.

The height of the antenna required to transmit the original baseband signal of frequency $v = 10\ kHz$ is

$$h_1 = \frac{\lambda}{4} = \frac{c}{4v} = \frac{3\times10^8}{4\times10\times10^3} = 7.5km \tag{10.2}$$

The height of the antenna required to transmit the modulated signal of frequency $v = 1\ MHz$ is

$$h_2 = \frac{\lambda}{4} = \frac{c}{4v} = \frac{3\times10^8}{4\times1\times10^6} = 75m \tag{10.3}$$

Comparing equations (10.2) and (10.3), we can infer that it is practically feasible to construct an antenna of height 75 m while the one with 7.5 km is not possible. It clearly manifests that modulated signals reduce the antenna height and are required for long distance transmission.

10.5 PROPAGATION OF ELECTROMAGNETIC WAVES

The information signal modulated with the carrier wave (radio wave) is transmitted by an antenna. This travels through space and is received by the receiving antenna at the other end. The frequencies from 2 kHz to 400 GHz are transmitted through wireless communication. The strength of the electromagnetic wave keeps decreasing while traveling from transmitter to the receiver. The electromagnetic wave transmitted by the transmitter travels in three different modes to reach the receiver according to its frequency range:

- Ground wave propagation (or) surface wave propagation (nearly 2 kHz to 2 MHz)
- Sky wave propagation (or) ionospheric propagation (nearly 3 MHz to 30 MHz)
- Space wave propagation (nearly 30 MHz to 400 GHz)

10.5.1 GROUND WAVE PROPAGATION

If the electromagnetic waves transmitted by the transmitter glide over the surface of the earth to reach the receiver, then the propagation is called ground wave propagation. The corresponding waves are called ground waves or surface waves. The pictorial representation is shown in Figure 10.5(a).

Both transmitting and receiving antennas must be close to the earth. The size of the antenna plays a major role in deciding the efficiency of the radiation of signals.

During transmission, the electrical signals are attenuated over a distance. Some reasons for attenuation are as follows:

- **Increasing distance**: The attenuation of the signal with distance depends on (i) power of the transmitter (ii) frequency of the transmitter and (iii) condition of the Earth surface.
- **Absorption of energy by the Earth**: When the transmitted signal in the form of EM wave is in contact with the Earth, it induces charges in the Earth and constitutes a current. Due to this, the Earth behaves like a leaky capacitor which leads to the attenuation of the wave.
- **Tilting of the wave:** As the wave progresses, the wavefront starts gradually tilting according to the curvature of the Earth. This increase in the tilt decreases the electric field strength of the wave. Finally at some distance, the surface wave dies out due to energy loss.

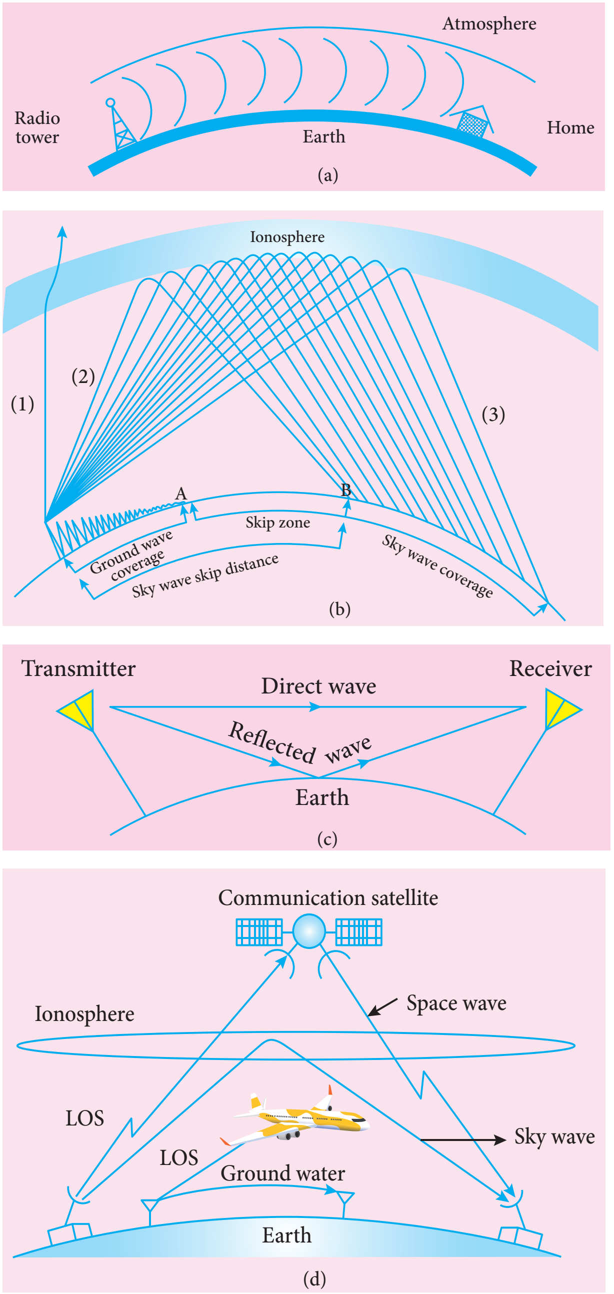


Figure 10.5 Propagation of EM waves (a) Ground wave (b) Sky wave (c) Space wave (d) Summary of all modes of propagation

The frequency of the ground waves is mostly less than 2 MHz as high frequency waves undergo more absorption of energy at the earth's atmosphere. The medium wave signals received during the day time use surface wave propagation.

It is mainly used in local broadcasting, radio navigation, for ship-to-ship, ship-to-shore communication and mobile communication.

10.5.2 SKY WAVE PROPAGATION

The mode of propagation in which the electromagnetic waves radiated from an antenna, directed upwards at large angles, gets reflected by the ionosphere back to earth is called sky wave propagation or ionospheric propagation. The corresponding waves are called sky waves (Figure 10.5(b)).

The frequency range of EM waves in this mode of propagation is 3 to 30 MHz. EM waves of frequency more than 30 MHz can easily penetrate through the ionosphere and does not undergo reflection. It is used for short wave broadcast services. Medium and high frequencies are for long-distance radio communication. Extremely long-distance communication is also possible as the radio waves can undergo multiple reflections between the earth and the ionosphere. A single reflection helps the radio waves to travel a distance of approximately 4000 km.

Ionosphere acts as a reflecting surface. It is at a distance of approximately 50 km and spreads up to 400 km above the Earth surface. Due to the absorption of ultraviolet rays, cosmic ray, and other high energy radiations like α, β rays from sun, the air molecules in the ionosphere get ionized. This produces charged ions and these ions provide a reflecting medium for the reflection of radio waves or communication waves back to Earth within the permitted frequency range. The phenomenon of bending the radio waves back to earth is nothing but the total internal reflection.

This is the reason why the EM waves are transmitted at a critical angle to ensure that the waves undergo total reflection and reaches the ground without escaping into space.

The shortest distance between the transmitter and the point of reception of the sky wave along the surface is called as the skip distance shown in Figure 10.5(b).

The electromagnetic waves are transmitted from the ground at particular angles. When the angle of emission increases, the reception of ground waves decreases. At one point, there will be no reception due to ground waves and marked as A in the Figure 10.5(b).

If the angle of emission is increased further, the reception of sky waves starts at point B in the Figure 10.5(b). **There is a zone (in between A and B) where there is no reception of electromagnetic waves neither ground nor sky, called as skip zone or skip area.**

Note

The higher the frequency, higher is the skip distance and for a frequency less than the critical frequency, skip distance is zero.

10.5.3 SPACE WAVE PROPAGATION

The process of sending and receiving information signal through space is called space wave communication (Figure 10.5(c)). **The electromagnetic waves of very high frequencies above 30 MHz are called as space waves.** These waves travel in a straight line from the transmitter to the receiver. Hence, it is used for a line of sight communication (LOS).

For high frequencies, the transmission towers must be high enough so that the transmitted and received signals (direct waves) will not encounter the curvature of the Earth and hence travel with less attenuation and loss of signal strength. Certain waves reach the receiver after getting reflected from the ground.

The communication systems like television telecast, satellite communication and RADAR are based on space wave propagation. Microwaves having high frequencies (super high frequency band) are used against radio waves due to certain advantages: larger bandwidth, high data rates, better directivity, small antenna size, low power consumption, etc.

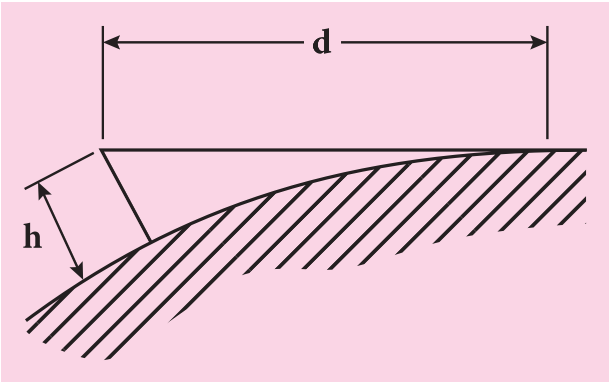


Figure 10.6 Distance of coverage

The range or distance (d) of coverage of the propagation depends on the height (h) of the antenna given by the equation,

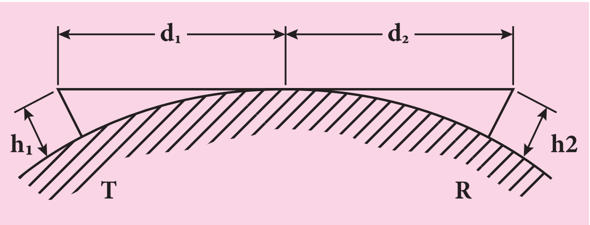
$$d = \sqrt{2Rh} \qquad (10.4)$$

where R is the radius of the Earth and it is 6400 km.

The distance of coverage is shown pictorially in Figure 10.6.

EXAMPLE 10.1

A transmitting antenna has a height of 40 m and the height of the receiving antenna is 30 m. What is the maximum distance between them for line-of-sight communication? The radius of the earth is 6.4×10^{6} m.



Solution:

The total distance d between the transmitting and receiving antennas will be the sum of the individual distances of coverage.

$$d = d_1 + d_2$$

$$= \sqrt{2Rh_1} + \sqrt{2Rh_2}$$

$$= \sqrt{2R}\left(\sqrt{h_1} + \sqrt{h_2}\right)$$

$$= \sqrt{2\times6.4\times10^{6}} \times\left(\sqrt{40} + \sqrt{30}\right)$$

$$= 16\times10^{2}\sqrt{5}\times(6.32+5.48)$$

$$= 42217\,m = 42.217\,km$$

10.6 SATELLITE COMMUNICATION

The satellite communication is a mode of communication of signal between transmitter and receiver via satellite. The message signal from the Earth station is transmitted to the satellite on board via an uplink (frequency band 6 GHz), amplified by a transponder and then retransmitted to another Earth station via a downlink (frequency band 4 GHz) (Figure 10.7).

The high-frequency radio wave signals travel in a straight line (line of sight) may come across tall buildings or mountains or even encounter the curvature of the earth. A communication satellite relays and amplifies such radio signals via transponder to reach distant and far off places using uplinks and downlinks. It is also called as a radio repeater in sky. The applications are found to be in all fields and are discussed below.

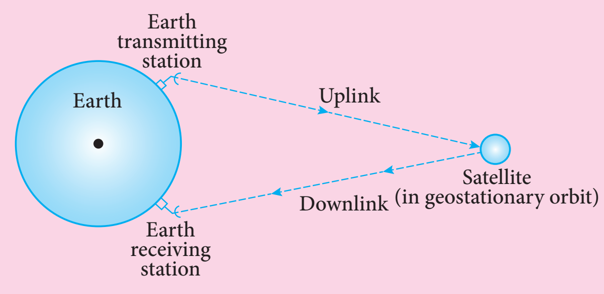


Figure 10.7 Satellite communication system

Applications

Satellites are classified into different types based on their applications. Some satellites are discussed below.

i) **Weather Satellites:** They are used to monitor the weather and climate of Earth. By measuring cloud mass, these satellites enable us to predict rain and dangerous storms like hurricanes, cyclones etc.

ii) **Communication satellites:** They are used to transmit television, radio, internet signals etc. Multiple satellites are used for long distances.

iii) **Navigation satellites:** These are employed to determine the geographic location of ships, aircrafts or any other object.

10.7 FIBRE OPTIC COMMUNICATION

The method of transmitting information from one place to another in terms of light pulses through an optical fiber is called fiber optic communication. It works under the principle of total internal reflection.

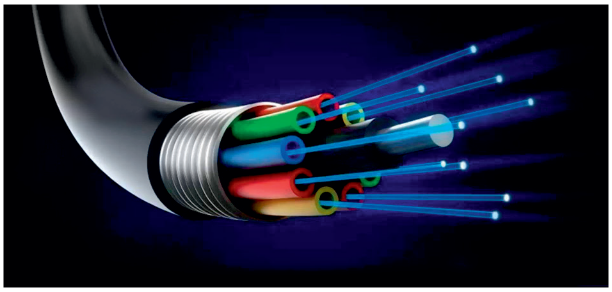


Figure 10.8 Optical fibers

Light has very high frequency (400 THz – 790 THz) than microwave radio systems. The fibers are made up of silica glass or silicon dioxide which is highly abundant on Earth.

Now it has been replaced with materials such as chalcogenide glasses, fluoroaluminate crystalline materials because they provide larger infrared wavelength and better transmission capability.

As fibers are not electrically conductive, it is preferred in places where multiple channels are to be laid and isolation is required from electrical and electromagnetic interference.

Applications

Optical fiber system has a number of applications namely, international communication, inter-city communication, data links, plant and traffic control and defense applications.

Merits

i) Fiber cables are very thin and weigh lesser than copper cables.
ii) This system has much larger band width. This means that its information carrying capacity is larger.
iii) Fiber optic system is immune to electrical interferences.
iv) Fiber optic cables are cheaper than copper cables.

Demerits

i) Fiber optic cables are more fragile when compared to copper wires.
ii) It is an expensive technology.

> **DO YOU KNOW?**
> Fiber optic cables provide the fastest transmission rate compared to any other form of transmission. It can provide data speed of 1 Gbps for homes and business. Multimode fibers operate at the speed of 10 Mbps. Recent developments in optical communication provide the data speed at the rate of 25 Gbps

> **DO YOU KNOW?**
> Most transatlantic telecommunication cables between the United States of America and Europe are fiber optic.

10.8 RADAR AND APPLICATIONS

Radar basically stands for Radio Detection and Ranging System. It is one of the important applications of communication systems and is mainly used to sense, detect, and locate distant objects like aircraft, ships, spacecraft, etc. The angle, range or velocity of the objects that are invisible to the human eye can be determined.

Radar uses electromagnetic waves for communication. The electromagnetic signal is initially radiated into space by an antenna in all directions. When this signal strikes the targeted object, it gets reflected or reradiated in many directions. This reflected (echo) signal is received by the radar antenna which in turn is delivered to the receiver. Then, it is processed and amplified to determine the geographical statistics of the object. The range is determined by calculating the time taken by the signal to travel from RADAR to the target and back.

Applications

Radars find extensive applications in almost all fields. A few are mentioned below.

i) In military, it is used for locating and detecting the targets.
ii) It is used in navigation systems such as ship borne surface search, air search and missile guidance systems.

iii) To measure precipitation rate and wind speed in meteorological observations, Radars are used.

iv) It is employed to locate and rescue people in emergency situations.

10.9 MOBILE COMMUNICATION

Mobile communication is used to communicate with others in different locations without the use of any physical connection like wires or cables. It allows the transmission over a wide range of area without the use of the physical link. It enables the people to communicate with each other regardless of a particular location like office, house etc. It also provides communication access to remote areas.

Figure 10.9 Mobile communication

It provides the facility of roaming – that is, the user may move from one place to another without the need of compromising on the communication. The maintenance and cost of installation of this communication network are also cheap.

Applications

i) It is used for personal communication and cellular phones offer voice and data connectivity with high speed.

ii) Transmission of news across the globe is done within a few seconds.

iii) Using Internet of Things (IoT), it is made possible to control various devices from a single device. Example: home automation using a mobile phone.

iv) It enables smart classrooms, online availability of notes, monitoring student activities etc. in the field of education.

DO YOU KNOW?

Recently, the mobile communication technology has evolved through various stages like 2G, 3G, 4G, 5G, WiMAX, Wibro, EDGE, GPRS and many others. This helps to increase the speed of communication and the range of coverage. The connectivity issues have decreased with reliable and secure connections. The GPS (Global Positioning System) and GSM (Global System for Mobile communication) technology play an important role in mobile communication. This increases the utilization of bandwidth of the network, sharing of the networks, error detections, etc. Many methods like digital switching, TDMA, CDMA have been used to ease the communication process.

10.10 INTERNET

Internet is a fast growing technology in the field of communication system with multifaceted tools. It provides new ways and means to interact and connect with people.

Internet is the largest computer network recognized globally that connects millions of people through computers. It finds extensive applications in all walks of life.

> **DO YOU KNOW?** To store all the information available on the internet, you would need over 1 billion DVDs or 200 million Blu-ray discs.

Applications:

i) **Search engine:** The search engine is basically a web-based service tool used to search for information on World Wide Web.

ii) **Communication:** It helps millions of people to connect with the use of social networking: emails, instant messaging services and social networking tools.

iii) **E-Commerce:** Buying and selling of goods and services, transfer of funds are done over an electronic network.

10.11 GLOBAL POSITIONING SYSTEM

GPS stands for Global Positioning System. It is a *global* navigation satellite system that offers geolocation and time information to a GPS receiver anywhere on or near the Earth.

GPS system works with the assistance of a satellite network. Each of these satellites broadcasts a precise signal like an ordinary radio signal. These signals that convey the location data are received by a low-cost aerial which is then translated by the GPS software. The software is able to recognize the satellite, its location, and the time taken by the signals to travel from each satellite.

Figure 10.10 Constellation of GPS satellites around Earth

The software then processes the data it accepts from each satellite to estimate the location of the receiver.

Applications

Global positioning system is highly useful many fields such as fleet vehicle management (for tracking cars, trucks and buses), wildlife management (for counting of wild animals) and engineering (for making tunnels, bridges etc).

10.12 APPLICATION OF INFORMATION AND COMMUNICATION TECHNOLOGY IN AGRICULTURE, FISHERIES AND MINING

(i) Agriculture

The implementation of information and communication technology (ICT) in agriculture sector enhances the productivity, improves the living standards of farmers and overcomes the challenges and risk factors.

a) ICT is widely used in increasing food productivity and farm management.
b) It helps to optimize the use of water, seeds and fertilizers etc.
c) Sophisticated technologies that include robots, temperature and moisture sensors, aerial images, and GPS technology can be used.
d) Geographic information systems are extensively used in farming to decide the suitable place for the species to be planted.

(ii) Fisheries

a) Satellite vessel monitoring system helps to identify fishing zones.
b) Use of barcodes helps to identify time and date of catch, species name, quality of fish.

(iii) Mining

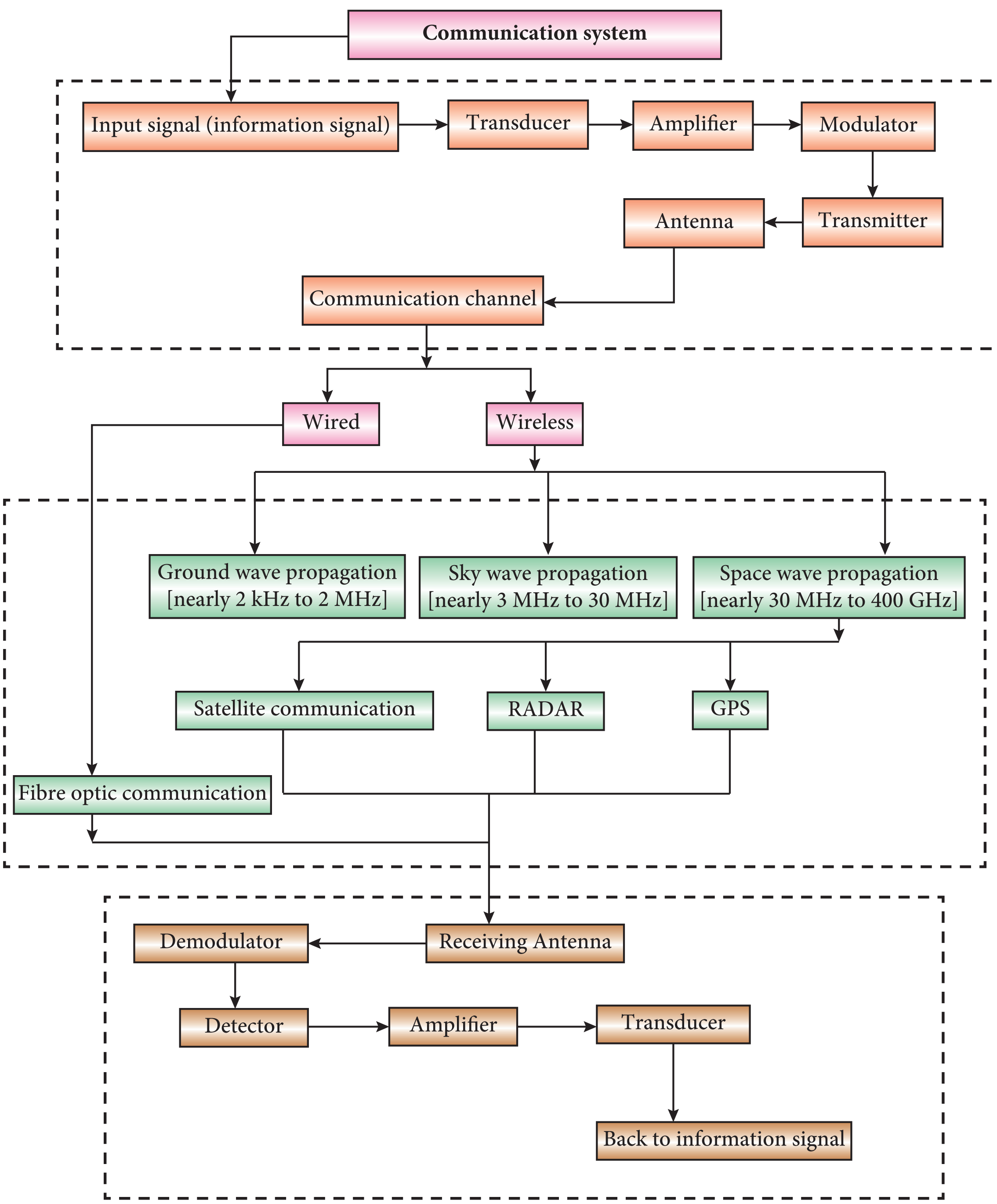
a) ICT in mining improves operational efficiency, remote monitoring and disaster locating system.
b) Information and communication technology provides audio-visual warning to the trapped underground miners.
c) It helps to connect remote sites.

SUMMARY

- The basic elements required for the transmission and reception of a signal through long distance using electromagnetic waves are transducer, amplifier, carrier signal, modulator, power amplifier, medium of transmission, transmitting and receiving antenna, demodulator, detector.
- For long-distance transmission, the baseband signal is modulated with the carrier wave.
- If the amplitude of the carrier signal is modified with the instantaneous amplitude of the baseband signal then it is called amplitude modulation.
- The frequency of the carrier signal is modified with the instantaneous amplitude of the baseband signal in frequency modulation.
- The instantaneous amplitude of the baseband signal modifies the phase of the carrier signal keeping the amplitude and frequency constant is called phase modulation
- The height of the transmitting and receiving antenna must be a multiple of $\frac{\lambda}{4}$.
- If the EM waves transmitted by the transmitter glide over the surface of the earth to reach the receiver, then the propagation of EM waves is called ground wave propagation.
- Electromagnetic waves radiated from an antenna, directed upwards at large angles gets reflected by the ionosphere back to earth is called sky wave propagation.
- The process of sending and receiving information signal through space is called space wave communication.
- The satellite communication is a mode of communication of signal between transmitter and receiver via satellite.
- Fiber-optic communication is a method of transmitting information by sending pulses of light through an optical fiber.

- Radar basically stands for Radio Detection and Ranging System. It is one of the important applications of communication systems for remote sensing.
- Mobile Communication is used to communicate with others in different locations without the use of any physical connection like wires or cables.
- GPS stands for Global Positioning System that offers geolocation and time.
- Communication technology is used extensively in sectors like fisheries, mining, and agriculture.

CONCEPT MAP



EVALUATION

I Multiple Choice Questions

1. The output transducer of the communication system converts the radio signal into ----------------
   (a) Sound
   (b) Mechanical energy
   (c) Kinetic energy
   (d) None of the above

2. The signal is affected by noise in a communication system
   (a) At the transmitter
   (b) At the modulator
   (c) In the channel
   (d) At the receiver

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3. The variation of frequency of carrier wave with respect to the amplitude of the modulating signal is called ------------------
   (a) Amplitude modulation
   (b) Frequency modulation
   (c) Phase modulation
   (d) Pulse width modulation

4. The internationally accepted frequency deviation for the purpose of FM broadcasts.
   (a) 75 kHz (b) 68 kHz
   (c) 80 kHz (d) 70 kHz

5. The frequency range of 3 MHz to 30 MHz is used for
   (a) Ground wave propagation
   (b) Space wave propagation
   (c) Sky wave propagation
   (d) Satellite communication

Answers

1. a 2. c 3. b 4. a 5. c

II Short answers

1. Give the factors that are responsible for transmission impairments.
2. Distinguish between wireline and wireless communication? Specify the range of electromagnetic waves in which it is used.
3. Explain centre frequency or resting frequency in frequency modulation.
4. What does RADAR stand for?
5. What do you mean by Internet of Things?

III Long Answers

1. What is modulation? Explain the types of modulation with necessary diagrams.
2. Elaborate on the basic elements of communication system with the necessary block diagram.
3. Explain the three modes of propagation of electromagnetic waves through space.
4. What do you know about GPS? Write a few applications of GPS.
5. Give the applications of ICT in mining and agriculture sectors.
6. Modulation helps to reduce the antenna size in wireless communication – Explain.
7. Fiber optic communication is gaining popularity among the various transmission media -justify.