GATE 2026 Geology and Geophysics - Geophysics (GG-2) Question Paper with Solutions

General Instructions

(i) This booklet contains 65 questions, each provided with a complete, step-by-step solution.

(ii) It comprises 37 single-correct multiple-choice questions, 9 multiple-correct questions and 19 numerical / integer-type questions.

(iii) Attempt each question on your own before reviewing the given solution.

(iv) For multiple-correct questions, all correct options must be selected to earn full marks.

(v) For numerical questions, report the answer rounded exactly as asked.

1. Suresh said, "I did it yesterday." Which one of the following options is the correct form of this sentence in indirect speech?

(A) Suresh said that I did it yesterday.

(B) Suresh says I did it yesterday.

(C) Suresh says that he did it the day before.

(D) Suresh said that he had done it the day before.

Correct Answer: (D) Suresh said that he had done it the day before.

Solution:

This question tests the rules of converting direct speech into indirect (reported) speech. When the reporting verb is in the past tense, three things change together: the pronoun, the verb tense, and any time word that points to "now" from the speaker's original moment.

The direct sentence is: Suresh said, "I did it yesterday." Since the reporting verb "said" is in the past tense, the reported clause must also move one step back in tense, from simple past to past perfect. The pronoun "I" refers to Suresh himself, so in reported speech it becomes "he". The word "yesterday" is tied to the day the statement was made, so in reported speech it becomes "the day before" (or "the previous day"), because we are no longer speaking on that same day.

1. **Suresh said that I did it yesterday.**: The reporting verb "said" is correctly past tense, but the pronoun "I" and the time word "yesterday" are left unchanged. Reported speech never keeps the original speaker's pronoun as "I" when a third person is reporting the sentence, so this option is wrong.
2. **Suresh says I did it yesterday.**: This keeps the reporting verb in the present tense ("says"), which does not match the original, since Suresh "said" it in the past. The pronoun and time word are also unchanged. Wrong on three counts.
3. **Suresh says that he did it the day before.**: The pronoun "he" and time word "the day before" are correct, but the reporting verb "says" is present tense, so it does not match the original past reporting verb "said". Wrong.
4. **Suresh said that he had done it the day before.**: The reporting verb "said" stays past tense, the pronoun correctly changes from "I" to "he", the verb backshifts from simple past "did" to past perfect "had done", and "yesterday" correctly becomes "the day before". Every rule is applied correctly.

The correct option is (D), because it is the only one that backshifts the tense, changes the pronoun, and converts the time expression together, exactly as the rules of indirect speech require.

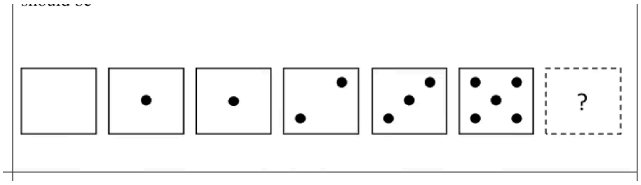
Let's summarize:

- A past tense reporting verb ("said") forces the reported verb one step further back in tense (simple past becomes past perfect).
- Pronouns change to match who is now speaking about whom.
- Time words tied to "today" shift to match the new point of view: yesterday becomes the day before.

So the correct indirect speech version is option (D): "Suresh said that he had done it the day before."

Quick Tip: A past tense reporting verb pushes the reported verb one step further back in tense, and time words like "yesterday" must also shift.

2. To continue the sequence of tiles shown, the tile indicated by the question mark should be



(A)

(B)

(C)

(D)

Correct Answer: (C)

Solution:

Step 1: Count the dots on each tile.

Reading the six given tiles from left to right, the number of dots on each tile is 0, 1, 1, 2, 3, and 5.

Step 2: Identify the pattern.

Each count after the first two is the sum of the two counts right before it: $1 = 0 + 1$, $2 = 1 + 1$, $3 = 1 + 2$, $5 = 2 + 3$. This is the Fibonacci rule, where every term is the sum of the two terms before it.

Step 3: Apply the pattern to find the missing tile.
Following the same rule, the next term after 3 and 5 is $3 + 5 = 8$. So the tile marked with a question mark must carry exactly 8 dots.

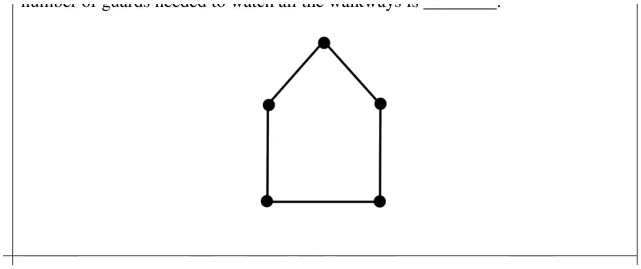
Step 4: Match against the options.
Option (A) shows 4 dots, option (B) shows 6 dots, option (C) shows 8 dots, and option (D) shows 9 dots. Only option (C) has the required count of 8 dots, so it is the tile that continues the sequence.

Final Answer:
The missing tile has 8 dots, matching option (C).

8 dots, option (C)

Quick Tip: Count the dots on each tile and look for a running-sum (Fibonacci-style) pattern: each count is the sum of the two before it.

3. Consider an art gallery whose walkways are shown as lines in the diagram. A black dot represents a junction of two walkways. A guard may be placed at a junction to watch over the walkways that join at that junction. The minimum number of guards needed to watch all the walkways is _____.



(A) 2

(B) 3

(C) 4

(D) 5

Correct Answer: (B) 3

Solution:

Step 1: Understanding the Question.

The diagram has 5 junctions (dots) joined by 5 walkways (lines), forming a closed loop: the top junction connects to the two middle junctions, each middle junction connects down to one bottom junction, and the two bottom junctions connect to each other. A guard placed at a junction watches every walkway that touches that junction. We need the fewest junctions where guards must stand so every walkway has at least one guard at one of its two ends.

Step 2: Key Formula or Approach.

This is a minimum vertex cover problem: pick the smallest set of junctions (vertices) so every walkway (edge) has at least one chosen junction at one of its ends. Since every junction here touches exactly 2 walkways, one guard can cover at most 2 walkways. With 5 walkways in total, at least $\lceil 5/2 \rceil = 3$ guards are needed, because 2 guards can cover at most $2 \times 2 = 4$ walkways, which is not enough for 5.

Step 3: Detailed Explanation.

Label the junctions: T (top), L and R (the two middle junctions), and P and Q (the two bottom junctions). The 5 walkways are T-L, T-R, L-P, R-Q, and P-Q.

Try 2 guards first: no matter which 2 junctions are picked, together

they touch at most 4 walkways (since each junction touches only 2), so one walkway is always left unwatched. So 2 guards, option (A), can never be enough.
Now try 3 guards, placed at T, P, and Q. Junction T covers walkways T-L and T-R. Junction P covers walkways L-P and P-Q. Junction Q covers walkways R-Q and P-Q. Between T, P, and Q, all 5 walkways are watched, with P-Q watched twice. So 3 guards are enough.
Since 2 guards are proved impossible and 3 guards are shown to work, the minimum is exactly 3.

Step 4: Final Answer.
The minimum number of guards needed is 3, which is option (B). Options (C) and (D) use more guards than necessary, so they are not the minimum even though they would also work.

$\boxed{3}$

Quick Tip: Each junction touches 2 walkways, so count how many walkways a guard can cover, then find the fewest junctions that together cover all 5 walkways.

4. The 2nd of June is a Thursday in a certain year. Which day of the week is the 3rd of July in that year?

(A) Thursday

(B) Friday

(C) Saturday

(D) Sunday

Correct Answer: (D) Sunday

Solution:

Step 1: Understanding the Question.

We know 2nd June falls on a Thursday. We need to find the day of the week for 3rd July of the same year.

Step 2: Key Formula or Approach.

The day of the week repeats every 7 days. To move from one date to another, count the total number of days between them, divide by 7, and the remainder tells us how many days of the week to move forward from the starting day.

Step 3: Detailed Explanation.

June has 30 days. From 2nd June to 2nd July is exactly 30 days later, since June has 30 days and adding 30 days to 2nd June lands exactly on 2nd July. From 2nd July to 3rd July is 1 more day. So the total gap from 2nd June to 3rd July is $30 + 1 = 31$ days.

Now find the remainder when 31 is divided by 7: $31 = 4 \times 7 + 3$, so the remainder is 3. This means 3rd July falls exactly 3 days of the week after 2nd June.

Starting from Thursday and counting 3 days forward: Thursday plus 1 day is Friday, plus 2 days is Saturday, plus 3 days is Sunday.

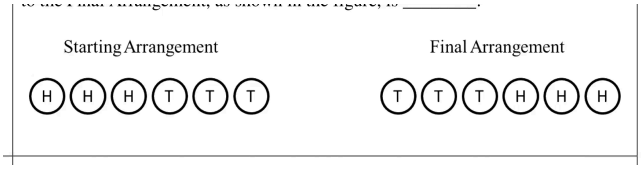
Step 4: Final Answer.

3rd July falls on a Sunday, which is option (D).

$$\boxed{\text{Sunday}}$$

Quick Tip: Count the total number of days between the two dates and find the remainder after dividing by 7 to shift forward from Thursday.

5. A coin with heads facing up is shown as H and a coin with tails facing up is shown as T.
Six coins are placed in the Starting Arrangement, as shown in the figure below. A "step" is defined as interchanging a pair of adjacent coins without flipping them. The minimum number of steps needed to go from the Starting Arrangement to the Final Arrangement, as shown in the figure, is _____.



(A) 3

(B) 6

(C) 9

(D) 12

Correct Answer: (C) 9

Solution:

Step 1: Understanding the Question.
The starting row of six coins is H H H T T T, and the final row is T T T H H H. A step swaps two coins that are sitting next to each other, without flipping either coin. We need the smallest number of such swaps to turn the starting row into the final row.

Step 2: Key Formula or Approach.
Since all three H coins are identical to each other, and all three T coins are identical to each other, the minimum number of adjacent swaps needed equals the number of (H, T) pairs that are in the wrong

relative order compared to the target. In the starting row every H sits to the left of every T, but in the final row every T must sit to the left of every H. So every one of the 3 H coins has to cross past every one of the 3 T coins at least once, and each crossing needs exactly one adjacent swap.

Step 3: Detailed Explanation.

Count the number of (H, T) pairs that must swap their relative order. There are 3 H coins (call them H_1, H_2, H_3) and 3 T coins (T_1, T_2, T_3). In the starting row, every H_i is to the left of every T_j. In the final row, every T_j must be to the left of every H_i. So each of the 3 H coins needs to move past each of the 3 T coins, giving $3 \times 3 = 9$ required crossings. Each adjacent swap can fix exactly one such crossing (swapping one H past one T that is immediately next to it), and no single swap can fix two crossings at once. So the minimum number of adjacent swaps needed is exactly 9, no fewer.

This matches the standard result that reversing the relative order of an m-block and an n-block using only adjacent swaps takes exactly $m \times n$ swaps; here $m = n = 3$, giving $3 \times 3 = 9$.

Step 4: Final Answer.

The minimum number of steps needed is 9, which is option (C).

$$\boxed{9}$$

Quick Tip: Each of the 3 heads must cross past each of the 3 tails at least once, and each adjacent swap fixes exactly one such crossing.

6. Exacerbate : Mitigate :: _____

Choose the option with the correct pair of words to fill the blank.

(A) Aggravate : Alleviate

(B) Alleviate : Precipitate

(C) Aggravate : Precipitate

(D) Emancipate : Exonerate

Correct Answer: (A) Aggravate : Alleviate

Solution:

This question tests vocabulary through analogy. We first need to work out the relationship between the given pair of words, "exacerbate" and "mitigate", and then find the option whose two words share the exact same relationship.

"Exacerbate" means to make a bad situation worse. "Mitigate" means to make a bad situation less severe, or to lessen its harm. So the given pair is a pair of opposites (antonyms): one word means to worsen something, the other means to ease or reduce it.

1. **Aggravate : Alleviate**: "Aggravate" means to make something worse, and "alleviate" means to make something easier to bear or less severe. This is exactly the same worsen-versus-ease relationship as exacerbate and mitigate. This matches.
2. **Alleviate : Precipitate**: "Alleviate" means to ease a problem, but "precipitate" means to cause something to happen suddenly. These two words are not opposites of each other, so the relationship does not match.
3. **Aggravate : Precipitate**: "Aggravate" means to worsen, but "precipitate" means to bring on suddenly, not to ease or lessen.

These words are not antonyms of each other, so this pair fails the test.

4. **Emancipate : Exonerate**: "Emancipate" means to set free from control, and "exonerate" means to clear someone of blame. These two words are close in meaning, not opposites, so this pair does not fit.

Only option (A), Aggravate : Alleviate, is a genuine pair of opposites, matching the worsen-versus-ease relationship of exacerbate and mitigate.

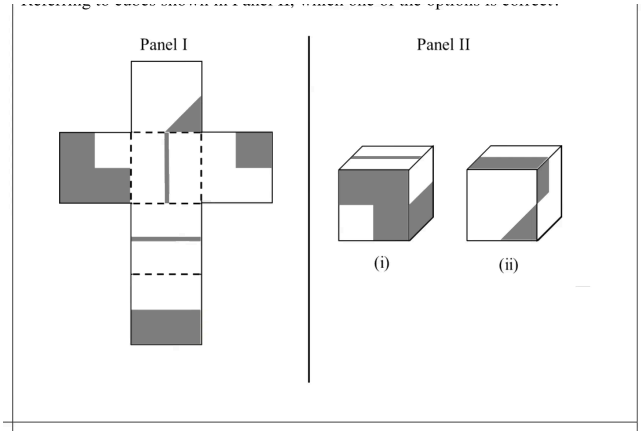
Let's summarize:

- Exacerbate (worsen) and mitigate (ease) are antonyms.
- Among the options, only aggravate (worsen) and alleviate (ease) form the same kind of antonym pair.
- The other three options either pair unrelated words or pair words that are close in meaning rather than opposite.

So the correct option is (A): Aggravate : Alleviate.

Quick Tip: Exacerbate means to worsen and mitigate means to lessen, so look for an option that is also a worsen-versus-lessen pair of opposites.

7. A paper shown in Panel I is folded along the dashed lines (- - -) to construct a cube. The shaded regions shown in Panel I appear on the outer surface of the cube. Referring to cubes shown in Panel II, which one of the options is correct?



(A) Only (i) can correspond to the unfolded cube in Panel I.

(B) Only (ii) can correspond to the unfolded cube in Panel I.

(C) Both (i) and (ii) can correspond to the unfolded cube in Panel I.

(D) Neither (i) nor (ii) can correspond to the unfolded cube in Panel I.

Correct Answer: (B) Only (ii) can correspond to the unfolded cube in Panel I.

Solution:

Step 1: Understanding the Question.

Panel I shows a flat six-square paper cutout (a cube net) with certain regions shaded. When this net is folded along the dashed lines, it forms a cube, and the shaded regions land on specific outer faces. We must check which of the two isometric cubes in Panel II, (i) or (ii), can actually be produced by folding the net exactly as drawn, without flipping the sheet over.

Step 2: Key Formula or Approach.

Two rules make this kind of question solvable without physically

cutting and folding paper. First, in a net, a face and the face directly two squares away in a straight line become opposite faces of the cube once folded, while directly touching squares become adjacent faces. Second, and most useful here, folding a flat sheet is a rigid operation: it never mirrors or flips the sheet over. So the rotational order (clockwise or anticlockwise) in which markings sit around any shared corner or edge in the flat net must stay the same after folding. If a candidate cube shows the same markings in the opposite rotational order, that cube is a mirror image and cannot be produced by folding this particular net.

Step 3: Detailed Explanation.

Pick the dashed central square in Panel I as the reference (front) face; the squares directly above and below it fold to become the top and bottom faces, and the two squares attached on its left and right fold to become the left and right faces, while the square at the far end of the vertical strip becomes the back face, opposite the front.

Two shaded regions are the most distinctive markings on the net: the gray L-shaped patch with a white notch on the left-arm square, and the gray diagonal triangle on the topmost square. Once folded, these two shaded faces sit next to each other around a shared corner of the cube, and the straight edge of the L-shaped patch meets the diagonal cut of the triangular patch at a fixed rotational sense, set entirely by how they are drawn touching each other in the flat net.

Checking cube (i): the diagonal shaded face and the L-shaped shaded face meet at their shared edge in the reversed rotational sense compared to the net, which only happens if the net were flipped over before folding. Since we are only folding, never flipping, cube (i) cannot be the correct result.

Checking cube (ii): the shaded faces meet at their shared edge in the same rotational sense as they appear in the flat net in Panel I, exactly

what a genuine fold, with no flip, produces. So cube (ii) is a valid, achievable folding of the net.

Step 4: Final Answer.
Only cube (ii) can be obtained by folding the paper shown in Panel I, so the correct option is (B).

Only (ii)

Quick Tip: Fold the net without flipping it over: the order in which shaded faces meet at a shared corner in the flat net must stay the same after folding, never mirrored.

8. In a population, patients who have high cholesterol also have high blood pressure (BP). Some patients with high BP also have diabetes. There are no patients who have both high cholesterol and diabetes. Furthermore,
1. the total number of patients with at least one of these conditions is 75,
2. the number of patients with high cholesterol is 10,
3. the number of patients with high BP is 45, and
4. the number of patients with only high BP and no other conditions is 20.
Then the number of patients who have both diabetes and high BP is

_____

(A) 0

(B) 15

(C) 20

(D) 10

Correct Answer: (B) 15

Solution:

Step 1: Set up the three groups.

Let C stand for patients with high cholesterol, B for patients with high BP, and D for patients with diabetes. The question tells us every patient with high cholesterol also has high BP, so C sits fully inside B. It also tells us no patient has both cholesterol and diabetes, so C and D never overlap.

Step 2: Break high BP into its non overlapping parts.

Since C is completely inside B, and C never touches D, the group B splits into three clean, non overlapping parts: patients with only high BP, patients with high BP and cholesterol together, and patients with high BP and diabetes together.

We are told the "only high BP" part has 20 patients.

Because every cholesterol patient already has high BP, the "high BP and cholesterol" part is just the full cholesterol group, which has 10 patients.

Call the "high BP and diabetes" part y, the value we want to find.

Step 3: Write the equation for the total high BP count.

Adding the three parts of B gives the total high BP count of 45:

$$20 + 10 + y = 45$$

$$y = 45 - 30 = 15$$

Step 4: Check with the total of 75.

Since every cholesterol patient is already counted inside high BP, the

group with at least one condition is really just high BP together with diabetes. Using this, the diabetes only group works out to $75 - 45 = 30$, which does not clash with anything given, so $y = 15$ is consistent.

Final Answer:
The number of patients with both diabetes and high BP is 15, so the correct option is (B). Option (A), 0, ignores the given 45 high BP total. Option (C), 20, comes from wrongly leaving out the cholesterol overlap. Option (D), 10, mixes up the diabetes and BP overlap with the cholesterol count.

$\boxed{15}$

Quick Tip: Cholesterol is a subset of BP and never overlaps diabetes, so split the 45 BP patients into only BP, BP with cholesterol and BP with diabetes.

9. Four people P, Q, R, and S, of different ages, make the following observations.
P: I am younger than S.
Q: I am neither the youngest nor the oldest.
R: P is older than me.
Based on these observations, the youngest person is _____.

(A) P

(B) Q

(C) R

(D) S

Correct Answer: (C) R

Solution:

Step 1: Turn each statement into an age comparison.
P says "I am younger than S", so P is below S in age.
R says "P is older than me", so P is above R in age, which means R is younger than P.
Putting these two together in one line: R is younger than P, and P is younger than S.

Step 2: Chain the comparisons.
From the two facts above we get the order R, then P, then S, going from younger to older:

$$R < P < S$$

So among these three people, R is currently the youngest.

Step 3: Place Q using Q's own statement.
Q says Q is neither the youngest nor the oldest of the whole group of four. That rules out Q being the overall youngest.
Since R is already the smallest value among R, P and S, and Q cannot be smaller than everyone (Q is not the youngest), Q must sit somewhere above R, for example between R and P, or between P and S.

Step 4: Identify the overall youngest.
With Q ruled out from the bottom spot, R stays the smallest of all four people.

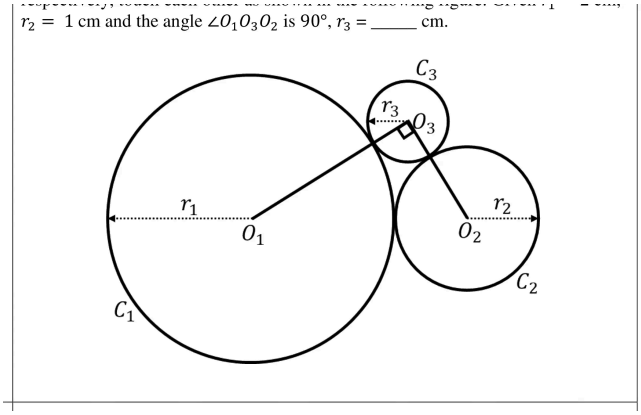
Final Answer:

R is the youngest person, so the correct option is (C). Option (A), P, is wrong since R is younger than P. Option (B), Q, is wrong because Q directly says Q is not the youngest. Option (D), S, is wrong since S is the oldest of the three linked by the first two statements.

$\boxed{R}$

Quick Tip: Combine P's and R's statements to get R younger than P younger than S, then use Q's statement to rule Q out of the youngest spot.

10. Circles C_1, C_2, and C_3, with centers O_1, O_2, and O_3, and radii r_1, r_2, and r_3, respectively, touch each other as shown in the following figure. Given $r_1 = 2$ cm, $r_2 = 1$ cm and the angle $\angle O_1O_3O_2$ is 90 degrees, $r_3 =$ _____ cm.



(A) $\frac{1}{2}\left(-3+\sqrt{17}\right)$

(B) $\frac{1}{2}\left(3+\sqrt{17}\right)$

(C) $\frac{1}{2}\left(-2+\sqrt{17}\right)$

(D) $\frac{1}{2}\left(-3+2\sqrt{17}\right)$

Correct Answer: (A) $\frac{1}{2}\left(-3+\sqrt{17}\right)$

Solution:

Step 1: Write the tangency conditions as distances.

When two circles touch each other from outside, the distance between their centers equals the sum of their radii. Applying this to each touching pair:

$$O_1O_2 = r_1 + r_2 = 2 + 1 = 3$$

$$O_1O_3 = r_1 + r_3 = 2 + r_3$$

$$O_2O_3 = r_2 + r_3 = 1 + r_3$$

Step 2: Use the right angle at O_3.

We are told $\angle O_1O_3O_2 = 90°$, so triangle $O_1O_3O_2$ is right angled at O_3. By the Pythagorean theorem, the square of the side opposite the right angle equals the sum of the squares of the other two sides:

$$O_1O_3^2 + O_2O_3^2 = O_1O_2^2$$

$$(2 + r_3)^2 + (1 + r_3)^2 = 3^2$$

Step 3: Expand and simplify.

$$(4 + 4r_3 + r_3^2) + (1 + 2r_3 + r_3^2) = 9$$

$$2r_3^2 + 6r_3 + 5 = 9$$

$$2r_3^2 + 6r_3 - 4 = 0$$

Dividing every term by 2:

$$r_3^2 + 3r_3 - 2 = 0$$

Step 4: Solve the quadratic.

Using the quadratic formula with $a = 1, b = 3, c = -2$:

$$r_3 = \frac{-3 \pm \sqrt{9 + 8}}{2} = \frac{-3 \pm \sqrt{17}}{2}$$

A radius cannot be negative, and $\sqrt{17}$ is close to 4.12, so only the plus sign gives a positive value near 0.56 cm. So we keep:

$$r_3 = \frac{-3 + \sqrt{17}}{2}$$

Final Answer:

$r_3 = \frac{1}{2}\left(-3 + \sqrt{17}\right)$ cm, which is option (A). Option (B) flips the sign in front of 3. Option (C) uses -2 instead of -3 inside the bracket, a slip in the constant term. Option (D) doubles the surd term by mistake.

$$\boxed{r_3 = \frac{1}{2}\left(-3 + \sqrt{17}\right) \text{ cm}}$$

Quick Tip: Use tangent circle distances (sum of radii) with the Pythagorean theorem on the right angle at O3, then solve the resulting quadratic.

11. Which one of the following increases over time when the soil is fully saturated?

(A) Infiltration

(B) Runoff

(C) Porosity

(D) Permeability

Correct Answer: (B) Runoff

Solution:

Step 1: Understanding the Question:

The question asks which quantity keeps growing with time once a soil has already reached full saturation, meaning every pore space in it is filled with water.

Step 2: Key Concept:

Infiltration is the rate at which water enters the soil surface. It is high when soil is dry and keeps falling as the soil fills up with water, settling near a low, roughly steady rate once the soil is saturated, it does not keep climbing. Porosity is the fraction of open space between soil grains, a fixed geometric property of the soil that does not change

just because water is filling it. Permeability, also called hydraulic conductivity, depends on how well connected the pore spaces are, another fixed material property of the soil that stays essentially constant once the soil is already saturated. Runoff is the part of rainfall or applied water that cannot soak in and instead flows across the surface.

Step 3: Detailed Explanation:
As a storm continues on already saturated ground, the soil has no more spare pore space to take in extra water, so its infiltration rate stays pinned near its lowest value instead of rising. Since less and less of the incoming water can enter the ground, more of it is forced to move over the surface as runoff, so runoff keeps building up with time on saturated ground. Porosity and permeability belong to the soil grains and structure themselves, not to how wet the soil currently is, so neither of them grows just because the soil stays saturated for longer.

Step 4: Final Answer:
Runoff is the quantity that increases with time once the soil is fully saturated, so the correct option is (B). Infiltration actually falls to a steady minimum rather than rising, ruling out (A). Porosity, option (C), and permeability, option (D), are fixed properties of the soil material and are not affected by how long the soil stays saturated.

Quick Tip: Once soil cannot absorb more water, incoming water has to go somewhere. Think about what happens to the excess.

12. In which one of the following layers inside the Earth, the velocity of the P-wave exceeds 13 km/s?

(A) Upper mantle

(B) Outer core

(C) Lower mantle

(D) Inner core

Correct Answer: (C) Lower mantle

Solution:

Step 1: Understanding the Question:
We need to identify in which of the Earth's internal layers the P-wave (primary, compressional seismic wave) speed goes above 13 km/s.

Step 2: Key Concept:
P-wave speed in the Earth depends mainly on the rigidity, incompressibility and density of the material it travels through, and it changes with depth as pressure and composition change. As a general trend, P-wave speed rises steadily through the solid mantle with depth, then drops sharply at the boundary into the liquid outer core because a liquid cannot support shear the way a solid can, and then rises again gradually through the outer core before reaching the solid inner core.

Step 3: Detailed Explanation:
In the upper mantle, P-wave speed is around 8 to 10 km/s, well under 13 km/s.
Moving down through the lower mantle, the increasing pressure keeps pushing the P-wave speed higher, and it climbs to about 13.5 to 13.7 km/s just above the boundary with the core, so the lower mantle

is where the speed actually crosses past 13 km/s.
At the core mantle boundary, the material changes from solid rock to liquid iron alloy, and the P-wave speed drops abruptly to around 8 km/s at the top of the outer core, then increases only gradually to about 10 km/s by the base of the outer core, staying below 13 km/s throughout.
In the solid inner core, the P-wave speed is around 11 km/s, higher than the outer core but still short of 13 km/s.

Step 4: Final Answer:
The lower mantle is the layer where P-wave velocity exceeds 13 km/s, so the correct option is (C). The upper mantle, option (A), stays near 8 to 10 km/s. The outer core, option (B), drops sharply at the core mantle boundary and stays near 8 to 10 km/s. The inner core, option (D), sits around 11 km/s, still below 13 km/s.

Quick Tip: P-wave speed rises through the solid mantle and peaks just before the core mantle boundary, then drops on entering the liquid outer core.

13. Which one of the following is NOT a volcanic igneous rock?

(A) Dacite

(B) Syenite

(C) Trachyte

(D) Komatiite

Correct Answer: (B) Syenite

Solution:

Step 1: Understanding the Question:
We need to pick the rock that is NOT formed as a volcanic, meaning extrusive, igneous rock, from among Dacite, Syenite, Trachyte and Komatiite.

Step 2: Key Concept:
Igneous rocks form from cooling magma or lava and are grouped by where that cooling happens. Volcanic (extrusive) rocks cool quickly at or near the surface, giving them a fine grained texture. Plutonic (intrusive) rocks cool slowly deep underground, giving them a coarse grained texture. Many plutonic rocks have a fine grained, chemically matching extrusive twin, for example granite pairs with rhyolite, and gabbro pairs with basalt.

Step 3: Detailed Explanation:
Dacite is a fine grained, extrusive rock of intermediate composition, chemically between andesite and rhyolite, formed from lava, so it is volcanic.
Trachyte is the fine grained, extrusive equivalent of syenite, also formed from lava at the surface, so it is volcanic too.
Komatiite is an ultramafic lava rock, mostly known from ancient Archean eruptions that were hot enough to produce this unusual, very magnesium rich lava, so it is also volcanic.
Syenite, unlike the other three, is a coarse grained rock that crystallizes slowly deep inside the crust. It is the plutonic, intrusive counterpart of trachyte, sharing a similar chemical composition but a different, slow cooling origin.

Step 4: Final Answer:
Syenite is the plutonic rock in the list, not a volcanic one, so the

correct option is (B). Dacite, option (A), Trachyte, option (C), and Komatiite, option (D), are all extrusive lava rocks.

Quick Tip: Three of the four are fine grained lava rocks. One is coarse grained and forms slowly underground instead.

14. If r_1 and r_2, respectively, denote the equatorial and polar radii of a reference ellipsoid, then the radius, R, of its equivalent sphere is given by

(A) $\left(r_1 r_2^2\right)^{2/3}$

(B) $r_1 r_2^2$

(C) $\left(r_1^2 r_2\right)^{1/3}$

(D) $r_1^2 r_2$

Correct Answer: (C) $\left(r_1^2 r_2\right)^{1/3}$

Solution:

Step 1: Understanding the Question:

A reference ellipsoid used in geodesy is an oblate spheroid, meaning it is a sphere squashed slightly at the poles. Its equatorial radius r_1 is the same in every direction around the equator, while its polar radius r_2 is shorter, measured from center to pole. We need the radius R of a sphere that has the same volume as this ellipsoid, called the equivalent sphere.

Step 2: Key Formula or Approach:

An oblate spheroid can be treated as a solid with two equal semi axes of length r_1 and one semi axis of length r_2. Its volume is:

$$V_{ellipsoid} = \frac{4}{3}\pi r_1^2 r_2$$

The volume of a sphere of radius R is:

$$V_{sphere} = \frac{4}{3}\pi R^3$$

Setting the two volumes equal, since the equivalent sphere by definition has the same volume as the ellipsoid, gives the formula for R.

Step 3: Detailed Explanation:
Equating the two volumes:

$$\frac{4}{3}\pi R^3 = \frac{4}{3}\pi r_1^2 r_2$$

The $\frac{4}{3}\pi$ on both sides cancels, leaving:

$$R^3 = r_1^2 r_2$$

Taking the cube root of both sides:

$$R = \left(r_1^2 r_2\right)^{1/3}$$

Step 4: Final Answer:
The radius of the equivalent sphere is $R = \left(r_1^2 r_2\right)^{1/3}$, matching option (C). Option (A) raises the wrong grouping to the power 2/3, which does not even come out to a length. Option (B) and option (D) both

leave the volume expression without taking the needed cube root, so they are not radii at all.

$$\boxed{R = \left(r_1^2 r_2\right)^{1/3}}$$

Quick Tip: Equate the ellipsoid's volume (4/3 pi r1^2 r2) with a sphere's volume (4/3 pi R^3) and solve for R.

15. Which of the following sources is/are used in land seismic surveys?

(A) Vibrosis

(B) Dynamite

(C) Airgun

(D) Thumper

Correct Answer: (A) Vibrosis ; (B) Dynamite ; (D) Thumper

Solution:

Step 1: Understand what a seismic source does.
A seismic source is the device used to send energy, in the form of vibrations, into the ground. The waves that bounce back off rock layers are picked up by receivers called geophones, and this data is used to map the subsurface. Different sources work in different settings. Some only work on land, some only work in water.

Step 2: Check option (A), Vibrosis.
This refers to the Vibroseis method, where a heavy truck fitted with a vibrating plate presses onto the ground and sends a long, controlled sweep of vibrations into the earth. It is one of the most common

sources for land seismic surveys, especially in areas where explosives cannot be used, such as near towns. So option (A) is correct.

Step 3: Check option (B), Dynamite.
Dynamite, or other explosive charges, is placed in a shallow drilled hole and set off to create a sharp, strong burst of seismic energy. This has been used in land seismic surveys for many decades and is still used where the terrain allows shot holes to be drilled. So option (B) is correct.

Step 4: Check option (C), Airgun.
An airgun releases a burst of compressed air into water, which makes a sharp pressure pulse. It only works underwater since it needs water around it to transmit the pulse well, so it is used in marine seismic surveys, not on land. Option (C) does not belong here.

Step 5: Check option (D), Thumper.
A thumper, also called a weight drop source, is a truck mounted heavy weight that is lifted and dropped onto a steel plate on the ground, sending a sharp impact into the earth. This is a classic land seismic source, used where vibroseis trucks or explosives are not practical. So option (D) is correct.

Step 6: Final answer.
The vibrator truck, dynamite and the thumper all generate seismic waves through direct contact with the ground, while the airgun needs a body of water to work and belongs to marine surveys.

(A), (B) and (D)

Quick Tip: Think about which sources need solid ground contact to work, and which one needs a body of water.

16. A pebble has an average diameter of 8 mm. Its ϕ-value on the Udden-Wentworth scale is _____ (answer in integer).

Correct Answer: -3

Solution:

Step 1: Recall the phi scale formula.

The Udden-Wentworth scale describes grain size using the phi (ϕ) value, defined as

$$\phi = -\log_2(d)$$

where d is the grain diameter in millimetres. The minus sign means coarser grains get smaller, often negative, phi values, and finer grains get larger phi values.

Step 2: Put in the given diameter.

The pebble has a diameter of

$$d = 8 \text{ mm}$$

so

$$\phi = -\log_2(8)$$

Step 3: Simplify the logarithm.

Since $8 = 2^3$,

$$\log_2(8) = 3$$

so

$$\phi = -3$$

Step 4: Final answer.

An 8 mm pebble has a phi value of -3, which fits the pebble and gravel range on the Udden-Wentworth scale, where coarse fragments carry negative phi values.

$$\boxed{\phi = -3}$$

Quick Tip: Use $\phi = -\log_2(d)$, with the diameter d in millimetres.

17. A quartz vein in a granitic outcrop has an initial length of 20 cm. If it undergoes uniform stretching resulting in a longitudinal strain of 0.45, the length of the vein post deformation is _____cm (answer in integer).

Correct Answer: 29

Solution:

Step 1: Recall the definition of longitudinal strain.

Longitudinal strain, also called extension, compares the change in length of an object to its original length. It is written as

$$e = \frac{L_f - L_i}{L_i}$$

where L_i is the original, undeformed length and L_f is the final, deformed length.

Step 2: Note down the given values.

$$L_i = 20\ \text{cm}, \quad e = 0.45$$

Step 3: Rearrange for the final length.

$$L_f = L_i(1 + e)$$

Step 4: Substitute and calculate.

$$L_f = 20 \times (1 + 0.45) = 20 \times 1.45$$

$$L_f = 29\ \text{cm}$$

Step 5: Final answer.

After the uniform stretching, the quartz vein grows from 20 cm to 29 cm.

$$\boxed{L_f = 29\ \text{cm}}$$

Quick Tip: Final length equals initial length multiplied by (1 + strain).

18. Which one of the following is the CORRECT pair of radioactive parent isotope and the corresponding radiogenic daughter isotope?

(A) $^{87}Rb \rightarrow {}^{86}Sr$

(B) $^{147}Sm \rightarrow {}^{143}Nd$

(C) $^{235}U \rightarrow {}^{206}Pb$

(D) $^{238}U \rightarrow {}^{207}Pb$

Correct Answer: (B) $^{147}Sm \rightarrow {}^{143}Nd$

Solution:

Step 1: Recall the main radiometric dating systems used in geology.
Radiometric dating relies on a radioactive parent isotope decaying at a fixed rate into a stable daughter isotope. Each dating system has its own correct parent-daughter pair, and geologists must know these pairs to date rocks correctly.

Step 2: Check option (A), the Rb-Sr system.
In the rubidium-strontium system, radioactive rubidium-87 decays by beta decay into strontium-87, not strontium-86. Strontium-86 is a separate, stable, non-radiogenic isotope of strontium used as the reference isotope in Rb-Sr dating, not the daughter product. So option (A) is incorrect.

Step 3: Check option (B), the Sm-Nd system.
In the samarium-neodymium system, radioactive samarium-147 decays by alpha decay into neodymium-143. This is the correct

parent-daughter pair used in Sm-Nd dating of old igneous and metamorphic rocks. So option (B) is correct.

Step 4: Check option (C), the U-Pb system with U-235.
Uranium-235 belongs to the actinium decay series and decays through a chain of steps into lead-207, not lead-206. Lead-206 is instead the end product of the uranium-238 decay series. So option (C) is incorrect.

Step 5: Check option (D), the U-Pb system with U-238.
Uranium-238 belongs to the uranium decay series and decays through a long chain into lead-206, not lead-207. The daughter isotope given here actually belongs to the uranium-235 chain. So option (D) is incorrect.

Step 6: Final answer.
Only the samarium-neodymium pair is written correctly among the four options.

$$\boxed{^{147}Sm \rightarrow {}^{143}Nd}$$

Quick Tip: Recall the standard parent-daughter isotope pairs used in radiometric dating: Rb-87 to Sr-87, Sm-147 to Nd-143, U-238 to Pb-206, U-235 to Pb-207.

19. Which one of the following pairs of geophysical signatures most commonly indicates volcanogenic massive sulphide deposits?

(A) High gravity and low resistivity

(B) Low gravity and high resistivity

(C) Low gravity and low resistivity

(D) High gravity and high resistivity

Correct Answer: (A) High gravity and low resistivity

Solution:

Step 1: Recall what a VMS deposit is made of.
Volcanogenic massive sulphide (VMS) deposits form on or near the ancient sea floor, where hot hydrothermal fluids linked to submarine volcanic activity deposit thick, massive layers of sulphide minerals such as pyrite, chalcopyrite, sphalerite and galena.

Step 2: Think about the density of these minerals.
Sulphide minerals are much denser than the surrounding volcanic and sedimentary host rocks. A dense mass of ore in the subsurface pulls the local gravitational field up slightly, so a gravity survey over a VMS body shows a positive gravity anomaly, that is, high gravity.

Step 3: Think about the electrical behaviour of these minerals.
Massive sulphides, especially pyrite and chalcopyrite, are excellent electrical conductors compared with the normal silicate host rocks. A good conductor lets current pass through it easily, which shows up as low resistivity, or equivalently high conductivity, in resistivity or electromagnetic surveys.

Step 4: Combine both signatures.
So a VMS deposit is expected to show high gravity together with low resistivity, since it is both denser and more conductive than its

surroundings.

Step 5: Rule out the other options.
Low gravity would mean a low density body, which does not fit a massive sulphide accumulation. High resistivity would mean poor conductivity, which does not fit sulphide minerals either. So options (B), (C) and (D) all go against the physical properties of massive sulphides.

Step 6: Final answer.

High gravity and low resistivity

Quick Tip: Massive sulphides are denser and far more conductive than the host rock around them.

20. Which one of the following is a physical weathering process?

(A) Oxidation

(B) Hydrolysis

(C) Exfoliation

(D) Carbonation

Correct Answer: (C) Exfoliation

Solution:

Step 1: Recall the two broad types of weathering.
Weathering breaks rock down at or near the surface, and it happens in two main ways. Physical, or mechanical, weathering breaks rock

into smaller pieces without changing its chemical make-up. Chemical weathering changes the minerals themselves through reactions with water, oxygen and acids.

Step 2: Check option (A), Oxidation.
Oxidation is a chemical reaction in which minerals, especially those containing iron, react with oxygen to form new minerals such as iron oxides. This changes the chemistry of the rock, so it is a chemical weathering process, not physical.

Step 3: Check option (B), Hydrolysis.
Hydrolysis is a chemical reaction between minerals, commonly feldspars, and water, which breaks the mineral structure apart and forms new minerals like clays. This is also a chemical process.

Step 4: Check option (C), Exfoliation.
Exfoliation happens when outer layers or sheets of rock peel away from a larger rock mass, usually because of the release of confining pressure, called unloading, as overlying rock erodes away, or through repeated heating and cooling. No new minerals form here, the rock just breaks apart physically into curved sheets or slabs. This makes exfoliation a physical, mechanical, weathering process.

Step 5: Check option (D), Carbonation.
Carbonation is the reaction of carbon dioxide dissolved in water with minerals such as calcite, forming soluble bicarbonates. This is a chemical process too, common in limestone weathering.

Step 6: Final answer.
Among the four choices, only exfoliation breaks rock apart without

changing its mineral chemistry, so it is the physical weathering process.

Exfoliation

Quick Tip: Look for the option where no new mineral forms, only mechanical breakup of the rock.

21. Which of the following statements is/are CORRECT according to the laws of electromagnetic induction in the Earth?

(A) Current density in a region of finite conductivity is solenoidal

(B) Magnetic vector potential is irrotational

(C) Magnetic field is solenoidal

(D) Curl of electric field is negative time rate of change of curl of magnetic vector potential

Correct Answer: (A) Current density in a region of finite conductivity is solenoidal ; (C) Magnetic field is solenoidal ; (D) Curl of electric field is negative time rate of change of curl of magnetic vector potential

Solution:

Step 1: Set up the governing equations.

Electromagnetic induction inside the Earth is described with Maxwell's equations in the quasi-static, low frequency form used in geophysics, where displacement current is ignored next to conduction current. The key relations are Gauss's law for magnetism, Faraday's law, and the definition of the magnetic vector potential $\mathbf{A}$, where

$$\mathbf{B} = \nabla \times \mathbf{A}$$

Step 2: Check option (A), current density is solenoidal.
In a region of finite, non-zero, conductivity, charge does not keep piling up as the field varies in the quasi-static regime used for EM induction, so the continuity equation gives

$$\nabla \cdot \mathbf{J} = 0$$

A vector field with zero divergence is called solenoidal. So the current density here is indeed solenoidal, and option (A) is correct.

Step 3: Check option (B), magnetic vector potential is irrotational.
A field is irrotational only if its curl is zero. But by definition, the curl of the magnetic vector potential gives the magnetic field itself:

$$\nabla \times \mathbf{A} = \mathbf{B} \neq 0$$

Since the curl of $\mathbf{A}$ is not zero in general, $\mathbf{A}$ is not irrotational. Option (B) is incorrect.

Step 4: Check option (C), magnetic field is solenoidal.
Gauss's law for magnetism states that magnetic field lines never start or end, meaning there are no magnetic monopoles:

$$\nabla \cdot \mathbf{B} = 0$$

This is exactly the definition of a solenoidal field, so option (C) is correct.

Step 5: Check option (D), curl of E versus curl of A.

Faraday's law of induction states

$$\nabla \times \mathbf{E} = -\frac{\partial \mathbf{B}}{\partial t}$$

Substituting $\mathbf{B} = \nabla \times \mathbf{A}$ gives

$$\nabla \times \mathbf{E} = -\frac{\partial}{\partial t}(\nabla \times \mathbf{A})$$

which says the curl of the electric field equals the negative time rate of change of the curl of the magnetic vector potential. This matches option (D) exactly, so it is correct.

Step 6: Final answer.

Statements (A), (C) and (D) all follow directly from Maxwell's equations applied to EM induction in the Earth, while (B) goes against the very definition of the vector potential.

$$\boxed{\text{(A), (C) and (D)}}$$

Quick Tip: Apply Maxwell's equations: $\nabla \cdot \mathbf{B} = 0$, $\mathbf{B} = \nabla \times \mathbf{A}$, and Faraday's law.

22. Which of the following is/are REE-bearing mineral(s)?

(A) Monazite

(B) Natrolite

(C) Spodumene

(D) Gibbsite

Correct Answer: (A) Monazite

Solution:

Step 1: Recall what makes a mineral REE bearing.
A mineral is called REE bearing when rare earth elements such as cerium, lanthanum and neodymium sit inside its own crystal structure as essential, not trace, components. We check each option against this rule.

Step 2: Check Monazite.
Monazite is a phosphate mineral with the general formula $(Ce, La, Th)PO_4$. Cerium and lanthanum are essential parts of its structure, so monazite is a true REE bearing mineral and one of the main ores of rare earth elements.

Step 3: Check Natrolite.
Natrolite is a zeolite with formula $Na_2Al_2Si_3O_{10} \cdot 2H_2O$. It is a hydrated sodium aluminosilicate. No rare earth element sits in its structure, so natrolite is not REE bearing.

Step 4: Check Spodumene.
Spodumene has formula $LiAlSi_2O_6$. It is a lithium pyroxene and the chief ore of lithium in rare element pegmatites. Its structure holds lithium and aluminium, not rare earth elements, so it is not REE bearing.

Step 5: Check Gibbsite.
Gibbsite is aluminium hydroxide, $Al(OH)_3$, and forms during lateritic weathering as an ore of bauxite. It carries aluminium and hydroxide groups only, with no rare earth element in its formula.

Step 6: Final conclusion.
Only monazite carries rare earth elements as essential structural components.

Monazite

Quick Tip: Only phosphate minerals like monazite host rare earth elements as essential structural components.

23. Which of the following pair(s) of well logs does/do NOT exhibit crossover for gas-bearing zones in hydrocarbon reservoirs?

(A) Neutron and density

(B) Caliper and resistivity

(C) Neutron and self-potential

(D) Caliper and gamma-ray

Correct Answer: (B) Caliper and resistivity ; (C) Neutron and self-potential ; (D) Caliper and gamma-ray

Solution:

Step 1: Recall the neutron density crossover effect.
In a gas bearing zone, the neutron log reads an apparent porosity lower than the true porosity because gas has a very low hydrogen index compared to oil or water. At the same depth the density log reads an apparent porosity higher than true porosity because gas is much lighter than the rock matrix. When both curves are plotted on compatible scales, the neutron curve moves toward lower porosity and the density curve moves toward higher porosity, so the two curves swing toward each other and cross. This is called the neutron density crossover, and it is the standard, well known gas indicator in log interpretation.

Step 2: Analyze option (A), neutron and density.
As shown in Step 1, this pair does show a clear crossover in gas zones. So option (A) is excluded from the answer because it DOES show crossover.

Step 3: Analyze option (B), caliper and resistivity.
The caliper log records the physical diameter of the borehole and detects washouts or mud cake buildup. The resistivity log records the electrical resistivity of the formation and fluids. These two logs respond to completely different physical properties and have no defined crossover relationship tied to the presence of gas. So option (B) does NOT exhibit a gas crossover.

Step 4: Analyze option (C), neutron and self-potential.
The self-potential or SP log responds to the electrochemical and electrokinetic potential difference between the borehole mud and the formation water, and is mainly used to pick shale versus permeable sand boundaries. It has no scale-based crossover behavior with the neutron log in gas zones. So option (C) does NOT exhibit a gas

crossover.

Step 5: Analyze option (D), caliper and gamma-ray.
The gamma ray log measures natural radioactivity and is mainly a shale or clay indicator, while the caliper again measures hole diameter. Neither log carries a scale calibrated to show a gas crossover, so no crossover effect is defined between them. So option (D) does NOT exhibit a gas crossover.

Step 6: Final conclusion.
Only the neutron and density pair is designed to show a crossover in gas zones. The pairs that do NOT exhibit crossover are caliper-resistivity, neutron-self potential and caliper-gamma ray.

$$\boxed{\text{(B), (C) and (D)}}$$

Quick Tip: Crossover in gas zones is a hallmark of the neutron-density log pair only.

24. If the diameter of a scaled-down model of the Earth is 45 cm, then the equivalent length on the surface of the Earth for 1 cm on the model is _____ km (rounded off to two decimal places).
[Use: Radius of the Earth = 6371 km]

Correct Answer: 283.16

Solution:

Step 1: Find the actual diameter of the Earth.

We are given the radius of the Earth as 6371 km. The diameter is twice the radius, so

$$D_{earth} = 2 \times 6371 = 12742 \text{ km}$$

Step 2: Set up the scale ratio.

The scaled-down model has a diameter of 45 cm, and this 45 cm represents the full 12742 km of the real Earth. The scale of the model is therefore

$$\text{Scale} = \frac{12742 \text{ km}}{45 \text{ cm}}$$

Step 3: Find the length represented by 1 cm.

Dividing both sides by 45, the length on the real Earth equivalent to 1 cm on the model is

$$\frac{12742}{45} = 283.1555\ldots \text{ km}$$

Step 4: Round to two decimal places.

Rounding 283.1555 to two decimal places gives

$$283.16 \text{ km}$$

Step 5: Final conclusion.

So 1 cm on the scaled-down model equals about 283.16 km on the

actual surface of the Earth. This value lies within the accepted range of 283.00 to 283.25 km.

$$\boxed{283.16 \text{ km}}$$

Quick Tip: Divide the real Earth's diameter by the model's diameter to get km per cm.

25. The wavelength of a certain portion of the electromagnetic spectrum ranges from 2000 nm to 3000 nm. The highest frequency associated with the above portion of the spectrum is_____$\times 10^8$ MHz (rounded off to one decimal place).

Correct Answer: 1.5

Solution:

Step 1: Identify which wavelength gives the highest frequency.
Frequency and wavelength are inversely related through the wave equation

$$f = \frac{c}{\lambda}$$

where c is the speed of light and λ is the wavelength. Since frequency goes up as wavelength goes down, the highest frequency in the given range corresponds to the shortest wavelength, which is 2000 nm.

Step 2: Convert the wavelength to metres.

$$\lambda = 2000 \text{ nm} = 2000 \times 10^{-9} \text{ m} = 2 \times 10^{-6} \text{ m}$$

Step 3: Apply the wave equation with $c = 3 \times 10^8$ m/s.

$$f = \frac{c}{\lambda} = \frac{3 \times 10^8}{2 \times 10^{-6}} = 1.5 \times 10^{14} \text{ Hz}$$

Step 4: Convert the frequency from Hz to MHz.

Since 1 MHz $= 10^6$ Hz,

$$f = \frac{1.5 \times 10^{14}}{10^6} \text{ MHz} = 1.5 \times 10^8 \text{ MHz}$$

Step 5: Final conclusion.

So the highest frequency in the given range of the electromagnetic spectrum is

$$\boxed{1.5 \times 10^8 \text{ MHz}}$$

The required coefficient, rounded to one decimal place, is **1.5**.

Quick Tip: Highest frequency comes from the shortest wavelength; use f = c/lambda.

26. Consider that the upper continental crust is 10 km thick and made up of granitic rock having density of 2800 kg/m^3. The surface heat flow due to radiogenic heat from the granitic rock having heat production value of 10^{-9} W/kg is_____ mW/m^2 (answer in integer).

Correct Answer: 28

Solution:

Step 1: Recall how radiogenic heat production relates to heat flow.
Radiogenic heat production H is usually given per unit mass, in W/kg. To get the heat generated per unit volume, we multiply it by the density ρ of the rock:

$$A = \rho \times H$$

where A is the volumetric heat production in W/m^3.

Step 2: Compute the volumetric heat production.
Given $\rho = 2800$ kg/m^3 and $H = 10^{-9}$ W/kg,

$$A = 2800 \times 10^{-9} = 2.8 \times 10^{-6} \text{ W/m}^3$$

Step 3: Relate heat flow to heat production and layer thickness.
For a layer of thickness z that produces heat uniformly through its own volume, the surface heat flow contributed by that layer, assuming the base of the layer contributes no heat flow of its own, is

$$q = A \times z$$

This comes from integrating the volumetric heat production over the

thickness of the layer.

Step 4: Convert the thickness to metres and substitute.

$$z = 10\ \text{km} = 10000\ \text{m}$$

$$q = 2.8 \times 10^{-6} \times 10000 = 2.8 \times 10^{-2}\ \text{W/m}^2$$

Step 5: Convert to mW/m^2.

$$q = 2.8 \times 10^{-2}\ \text{W/m}^2 = 28 \times 10^{-3}\ \text{W/m}^2 = 28\ \text{mW/m}^2$$

Step 6: Final conclusion.

The surface heat flow contributed by radiogenic heat from the granitic upper crust is

$$\boxed{28\ \text{mW/m}^2}$$

Quick Tip: Multiply density by heat production per kg to get heat production per volume, then multiply by thickness.

27. What is the value of magnetotelluric impedance phase over a homogeneous half-space?

(A) 0°

(B) 90°

(C) 45°

(D) 180°

Correct Answer: (C) 45°

Solution:

For a uniform (1-D) conducting half-space, the electric and magnetic fields of a plane EM wave diffusing downward obey the quasi-static Maxwell equations (displacement currents neglected in the Earth):

$$\frac{\partial E_x}{\partial z} = -i\omega\mu_0 H_y, \quad \frac{\partial H_y}{\partial z} = -\sigma E_x$$

Combining these gives the diffusion equation $\frac{\partial^2 E_x}{\partial z^2} = i\omega\mu_0\sigma E_x = k^2 E_x$, so the field decaying into the half-space is $E_x(z) = E_x(0)e^{-kz}$ with

$$k = \sqrt{i\omega\mu_0\sigma} = \sqrt{\omega\mu_0\sigma}\, e^{i\pi/4}$$

since $i = e^{i\pi/2}$ and $\sqrt{i} = e^{i\pi/4}$. From Faraday's law, $H_y(z) = \frac{k}{i\omega\mu_0} E_x(z)$, so the magnetotelluric impedance is

$$Z = \frac{E_x}{H_y} = \frac{i\omega\mu_0}{k} = \sqrt{\frac{i\omega\mu_0}{\sigma}}$$

The phase of the numerator $i\omega\mu_0$ is $90°$ and the phase of k is $45°$, so the phase of Z is $90° - 45° = 45°$. This $45°$ phase is independent of frequency, resistivity or permeability — it is a universal signature of a homogeneous (1-D) half-space, and MT interpreters use departures from $45°$ as a diagnostic of lateral resistivity heterogeneity.

Impedance phase = 45°

Quick Tip: For a homogeneous half-space, impedance $Z = \sqrt{i\omega\mu_0\rho}$; taking the square root of $i = e^{i\pi/2}$ halves its phase angle.

28. If A_b and A_s denote the amplitudes of the body and surface waves, respectively, at a distance, r, from the source, then the relation between them is given by

(A) $\frac{A_b}{A_s} \propto \sqrt{r}$

(B) $\frac{A_b}{A_s} \propto r$

(C) $\frac{A_b}{A_s} \propto \frac{1}{r}$

(D) $\frac{A_b}{A_s} \propto \frac{1}{\sqrt{r}}$

Correct Answer: (D) $\frac{A_b}{A_s} \propto \frac{1}{\sqrt{r}}$

Solution:

Body waves radiate outward from a point source through the full 3-D volume of the Earth, so at distance r their energy is spread over the surface of an expanding sphere, whose area grows as $4\pi r^2 \propto r^2$. By conservation of energy, the energy flux (energy per unit area) falls off as $1/r^2$:

$$E_b \propto \frac{1}{r^2}$$

Since seismic energy density is proportional to the square of the displacement amplitude, $E_b \propto A_b^2$, so

$$A_b \propto \frac{1}{r}$$

Surface waves, in contrast, are trapped within roughly one wavelength of the free surface and spread out only along an expanding circle (cylindrical/2-D spreading); the "area" over which their energy is shared grows only as the circumference, $2\pi r \propto r$:

$$E_s \propto \frac{1}{r} \Rightarrow A_s \propto \frac{1}{\sqrt{r}}$$

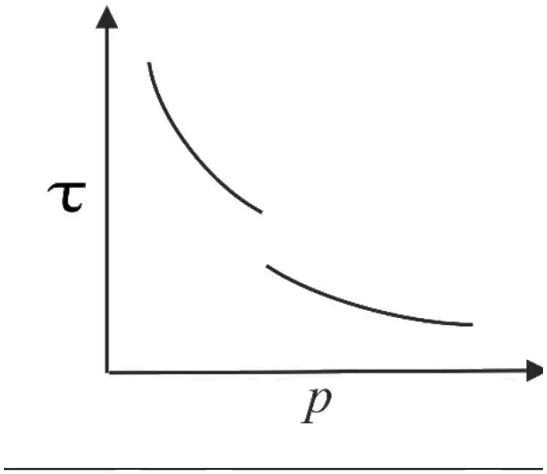
Taking the ratio of the two amplitude decay laws:

$$\frac{A_b}{A_s} \propto \frac{1/r}{1/\sqrt{r}} = \frac{\sqrt{r}}{r} = \frac{1}{\sqrt{r}}$$

$\boxed{\frac{A_b}{A_s} \propto \frac{1}{\sqrt{r}}}$ — this is exactly why, far from an earthquake source, surface waves (which decay more slowly) dominate the seismogram even though body waves arrive first.

Quick Tip: Body waves spread over an expanding sphere (amplitude ~1/r); surface waves spread over an expanding circle near the surface (amplitude ~1/√r).

29. If τ and p denote intercept time and slowness, respectively, in the $\tau - p$ diagram, then which one of the following is CORRECT for a P-wave propagating inside the Earth?



(A) Velocity continuously increases with depth

(B) Velocity continuously decreases with depth

(C) Velocity initially increases then decreases and again increases with depth

(D) Velocity initially decreases then increases and again decreases with depth

Correct Answer: (C) Velocity initially increases then decreases and again increases with depth

Solution:

In $\tau - p$ (intercept time vs ray parameter) analysis, each ray parameter $p = \sin i/v$ corresponds to one turning depth, and $\tau(p)$ is built from $\tau(p) = \int(\eta)\, dz$ type integrals along the ray. For a velocity structure that increases smoothly and monotonically with depth, $\tau(p)$ is a single, smooth, continuously decreasing curve with no breaks — every value of p in the observed range maps to exactly one turning point.

The plotted curve here instead shows two separate, offset branches: a steeper branch at smaller p (shallower turning depth) is followed, after a visible break/step, by a second, flatter branch at larger p sitting at a lower τ. A clean break of this kind in the $\tau(p)$ curve is the classic signature of a low-velocity zone (LVZ) at depth: rays that would turn inside the LVZ cannot do so (velocity has locally decreased, so they instead refract more efficient below it), which removes a contiguous band of ray parameters from the observed branch and creates a shadow zone in the travel-time curve and a corresponding jump/offset in $\tau(p)$.

Physically this requires the P-wave velocity to (i) increase with depth down to the top of the low-velocity zone, (ii) decrease through the LVZ itself, and (iii) increase again with depth below the LVZ once normal compaction/pressure effects resume. This three-stage behaviour — increase, then decrease, then increase again — is exactly option (C).

Velocity initially increases, then decreases, then increases again with depth

Quick Tip: A smooth single τ-p branch means velocity rises monotonically; a break/offset between two branches signals a low-velocity zone sandwiched between two increasing-velocity regions.

30. If $\hat{f}(t)$ denotes the Hilbert transform of $f(t)$, then the Hilbert transform of $\hat{f}(t)$ is equal to

(A) $-\hat{f}(t)$

(B) $f(t)$

(C) $-f(t)$

(D) $-\frac{d}{dt}f(t)$

Correct Answer: (C) $-f(t)$

Solution:

The Hilbert transform is most easily analysed in the frequency domain. If $F(\omega)$ is the Fourier transform of $f(t)$, the Hilbert transform corresponds to multiplying the spectrum by $-i\,\mathrm{sgn}(\omega)$:

$$\hat{f}(t) \;\longleftrightarrow\; -i\,\mathrm{sgn}(\omega)\,F(\omega)$$

This operator leaves the amplitude spectrum unchanged and only shifts every frequency component's phase by $-90°$ for $\omega > 0$ and $+90°$ for $\omega < 0$ (a quadrature filter).

Applying the Hilbert transform a second time multiplies the spectrum by the same factor again:

$$\mathcal{H}\{\hat{f}(t)\} \;\longleftrightarrow\; \big(-i\,\mathrm{sgn}(\omega)\big)^2 F(\omega) = -i^2\,\mathrm{sgn}^2(\omega)\,F(\omega) = -(-1)(1)\,F(\omega)$$

Wait — carefully: $i^2 = -1$ and $\mathrm{sgn}^2(\omega) = 1$ for $\omega \neq 0$, so $\big(-i\,\mathrm{sgn}(\omega)\big)^2 = i^2\,\mathrm{sgn}^2(\omega) = -1$. Hence

$$\mathcal{H}\{\hat{f}(t)\} \;\longleftrightarrow\; -F(\omega)$$

Taking the inverse Fourier transform of both sides gives $\mathcal{H}\{\hat{f}(t)\} = -f(t)$, i.e. applying the $90°$ phase-shift operator twice produces a net $180°$ phase shift, which is simply a sign flip in the time domain.

$$\boxed{\mathcal{H}\{\hat{f}(t)\} = -f(t)}$$

Quick Tip: In the frequency domain the Hilbert transform multiplies by -i·sgn(ω); squaring this factor gives -1, so applying it twice just negates the signal.

31. What happens to the magnetic susceptibility (k) and remanent magnetization (I_r), when the molten rock undergoes rapid cooling?

(A) k increases and I_r decreases

(B) k decreases and I_r increases

(C) Both k and I_r decrease

(D) Both k and I_r increase

Correct Answer: (B) k decreases and I_r increases

Solution:

When molten rock (magma/lava) cools, its magnetic minerals (mainly magnetite/titanomagnetite) crystallise and grow. The rate of cooling directly controls the grain size of these magnetic minerals:

Rapid cooling (e.g. lava chilled at the surface or against cold rock) leaves very little time for crystal growth, so the magnetic grains that form are very small — close to the single-domain (SD) or pseudo-single-domain (PSD) size range, with few or no internal domain walls.

Susceptibility (k): bulk magnetic susceptibility in an applied field is largely controlled by how easily domain walls can move within multidomain (MD) grains — MD grains respond strongly to a weak applied field because their walls shift readily, giving a high induced susceptibility. Fine, near single-domain grains lack mobile domain walls (a SD grain is a single, uniformly magnetized domain), so the

induced/bulk susceptibility of a rock dominated by fine grains is comparatively low. Hence rapid cooling → smaller grains → lower k.

Remanent magnetization (I_r): single-domain and pseudo-single-domain grains are magnetically far more efficient at recording and, crucially, at *retaining* a stable thermoremanent magnetization (TRM) than multidomain grains, because there are no internal domain walls to later relax or realign and demagnetize the grain. As the fine grains cool quickly through their Curie/blocking temperature in the ambient (Earth's) field, they lock in a strong, stable remanence. Hence rapid cooling → fine SD/PSD grains → higher, more stable I_r.

So rapid cooling produces a rock with *lower* susceptibility but *higher* (more stable) remanent magnetization — exactly option (B).

k decreases, I_r increases

Quick Tip: Rapid cooling produces fine, single-domain-like grains: these are magnetically 'hard' — low induced susceptibility but strong, stable remanent magnetization.

32. Between L_1 and L_2 norms, which one of the following is CORRECT in the treatment of outliers in the data?

(A) L_1 norm gives higher weightage to outliers than L_2 norm

(B) L_2 norm gives higher weightage to outliers than L_1 norm

(C) Both L_1 and L_2 norms give equal weightage to outliers

(D) L_1 norm occasionally gives higher weightage to outliers than L_2 norm

Correct Answer: (B) L_2 norm gives higher weightage to outliers than L_1 norm

Solution:

The L_1 and L_2 norms of a residual vector $e_i = d_i^{obs} - d_i^{pred}$ are defined as:

$$\|e\|_1 = \sum_i |e_i|, \qquad \|e\|_2 = \left(\sum_i e_i^2\right)^{1/2}$$

Consider how the contribution of a single residual to the misfit changes as that residual grows large (i.e., becomes an outlier). For the L_1 norm, the contribution of a data point to the total misfit is $|e_i|$, which grows only *linearly* with the size of the residual. For the L_2 (least-squares) norm, the contribution is e_i^2, which grows *quadratically*.

Concretely, if one residual is 10 times larger than the rest, its contribution to the L_1 misfit is only 10× a typical term, but its contribution to the L_2 misfit is 100× a typical term. When an inversion or regression minimizes an L_2-based objective function, a single large outlier can therefore dominate the total misfit and pull the fitted model strongly toward itself (the model "tries hard" to reduce that one huge squared term). An L_1-based (least-absolute-deviations) fit is comparatively insensitive to the same outlier, since its influence only scales linearly — this is exactly why L_1-norm (robust) inversion is preferred in geophysical data processing when the data contain occasional bad picks or spikes.

Hence the L_2 norm gives higher weightage to outliers than the L_1 norm.

Option (B)

Quick Tip: L1 cost grows linearly with residual size, L2 cost grows quadratically — so a single large outlier dominates an L2 (least-squares)

misfit far more than an L1 misfit.

33. Which pairs of the following combinations of the H-, Q-, K- and A-type of resistivity sounding curves is/are NOT possible to generate four-layer models?

(A) HQ and AQ

(B) KQ and QH

(C) KH and AA

(D) HH and KA

Correct Answer: (A) HQ and AQ ; (D) HH and KA

Solution:

The four basic 3-layer (VES) curve types are classified by the trend of resistivity with depth for ρ_1, ρ_2, ρ_3:

Type A: $\rho_1 < \rho_2 < \rho_3$ (rising) | Type K: $\rho_1 < \rho_2 > \rho_3$ (rise then fall) | Type H: $\rho_1 > \rho_2 < \rho_3$ (fall then rise) | Type Q: $\rho_1 > \rho_2 > \rho_3$ (falling)

A 4-layer curve type is written as two letters: the first letter is the 3-layer type of (ρ_1, ρ_2, ρ_3), and the second letter is the 3-layer type of (ρ_2, ρ_3, ρ_4). The key constraint is **continuity at the ρ_2-ρ_3 interface**: the trend from ρ_2 to ρ_3 is a single physical fact and must agree between the two 3-layer sub-curves used to build the first and second letters.

Types A and H both start with $\rho_2 < \rho_3$ (A: $\rho_1 < \rho_2 < \rho_3$; H: $\rho_1 > \rho_2 < \rho_3$) — so if the first letter is A or H, the second letter's own ρ_2 vs ρ_3 relation must also start with $\rho_2 < \rho_3$, which restricts the second letter to A or K only (both begin with $\rho_2 < \rho_3$). It cannot be H or Q, since those begin with $\rho_2 > \rho_3$ at the (now relabelled) first interface.

Similarly, types K and Q both start with $\rho_2 > \rho_3$ — so if the first letter is K or Q, the second letter is restricted to H or Q only.

This gives exactly 8 physically possible 4-layer type-combinations: AA, AK, HA, HK, KH, KQ, QH, QQ. The other 8 combinations are impossible: AH, AQ, HH, HQ, KA, KK, QA, QK.

Checking each option against this rule:

(A) HQ — H requires second letter ∈{A,K}; Q is not allowed → **impossible**. AQ — A requires second letter ∈{A,K}; Q is not allowed → **impossible**. Both members of option (A) are impossible, so (A) is a correct choice.

(B) KQ — K requires second letter ∈{H,Q}; Q is allowed → possible. QH — Q requires second letter ∈{H,Q}; H is allowed → possible. Both are valid curve types, so (B) is not a correct choice.

(C) KH — K requires second letter ∈{H,Q}; H is allowed → possible. AA — A requires second letter ∈{A,K}; A is allowed → possible. Both valid, so (C) is not a correct choice.

(D) HH — H requires second letter ∈{A,K}; H is not allowed → **impossible**. KA — K requires second letter ∈{H,Q}; A is not allowed → **impossible**. Both members impossible, so (D) is a correct choice.

Options (A) HQ, AQ and (D) HH, KA are NOT possible four-layer curve types

Quick Tip: A 4-layer type 'XY' is valid only if the ρ2-ρ3 trend implied by the end of X matches the ρ2-ρ3 trend implied by the start of Y — A/H (rising ρ2→ρ3) can only be followed by A/K, and K/Q (falling ρ2→ρ3) can only be followed by H/Q.

34. Which of the following is/are the correct explanation(s) for an increase in the adiabatic temperature gradient from the lower mantle (LM) to the outer core (OC)?

(A) OC has lower specific heat capacity and higher temperature than LM

(B) OC has higher specific heat capacity and lower temperature than LM

(C) OC has lower specific heat capacity and lower temperature than LM

(D) OC has higher specific heat capacity and higher temperature than LM

Correct Answer: (A) OC has lower specific heat capacity and higher temperature than LM

Solution:

The steepness of the adiabatic temperature gradient inside the Earth is fixed by the thermodynamic relation

$$\left(\frac{dT}{dr}\right)_{ad} = -\frac{g\,\alpha\,T}{C_p}$$

where g is the local gravity, α is the volumetric thermal expansion coefficient, T is the absolute temperature, and C_p is the specific heat capacity at constant pressure. This tells us that the magnitude of the adiabatic gradient scales directly with T and inversely with C_p: a hotter material, or a material that stores heat less efficiently per unit temperature rise (lower C_p), has a steeper adiabat.

Two independent changes happen simultaneously as we cross the core-mantle boundary (CMB) from the base of the lower mantle into the liquid outer core:

- **Temperature:** A strong thermal boundary layer (the D" region) sits atop the CMB, across which temperature jumps by several hundred to about a thousand kelvin. So the outer-core side of the boundary is hotter than the lower-mantle side, $T_{OC} > T_{LM}$.

- **Specific heat capacity:** The outer core is a dense Fe-Ni-rich liquid alloy, while the lower mantle is silicate (bridgmanite/post-perovskite + ferropericlase). Metallic liquids have a distinctly lower specific heat capacity per unit mass than silicate minerals (roughly 700-800 J/kg/K for the outer core versus about 1200 J/kg/K for lower-mantle silicates), so $C_p^{OC} < C_p^{LM}$.

Because the adiabatic gradient is proportional to T/C_p, an increase in T together with a decrease in C_p both act in the same direction and reinforce each other, producing a markedly steeper (larger) adiabatic gradient on entering the outer core. This is exactly the combination stated in option (A).

OC has lower specific heat capacity and higher temperature than LM

Quick Tip: The adiabatic gradient dT/dr is proportional to T/Cp; check how temperature and specific heat capacity each change going from silicate lower mantle into the metallic liquid outer core.

35. Given the Rayleigh wave velocity V_r, shear wave velocity V_s and the P-wave velocity V_p, which of the following relationships is/are CORRECT?

(A) $V_r < V_s < V_p$

(B) $V_s < V_r < V_p$

(C) $V_s < V_p < V_r$

(D) $V_s = V_r < V_p$

Correct Answer: (A) $V_r < V_s < V_p$

Solution:

Rayleigh waves arise from the constructive interference of P and SV waves trapped near a free surface. Their phase velocity $c = V_r$ is found from the classical Rayleigh secular (free-surface) equation for a homogeneous elastic half-space:

$$\left(2 - \frac{c^2}{\beta^2}\right)^2 = 4\sqrt{1 - \frac{c^2}{\alpha^2}}\sqrt{1 - \frac{c^2}{\beta^2}}$$

where $\alpha = V_p$ and $\beta = V_s$ are the P- and S-wave velocities. For a Poisson solid (Poisson's ratio $\nu = 0.25$, i.e. $\alpha = \sqrt{3}\,\beta$), solving this equation numerically for the root $0 < c < \beta$ gives

$$c = V_r \approx 0.9194\, V_s$$

So the Rayleigh wave is always slightly slower than the shear wave, $V_r < V_s$. Also, for any elastic solid with a positive Poisson's ratio, the shear velocity is always less than the compressional velocity, $V_s = \sqrt{\mu/\rho} < V_p = \sqrt{(K + 4\mu/3)/\rho}$, because the P-wave modulus includes the extra bulk-modulus term. Combining both inequalities:

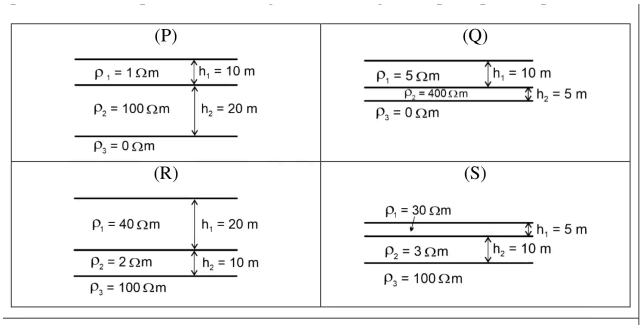
$$V_r < V_s < V_p$$

This matches option (A) exactly, and it also explains the standard seismogram ordering of arrivals: P first, then S, then the (slowest) Rayleigh surface wave.

$\boxed{V_r < V_s < V_p}$

Quick Tip: Rayleigh waves must decay evanescently with depth relative to both P and S potentials, forcing Vr below the smaller of the two body-wave speeds; for a Poisson solid, Vr is about 0.92 Vs.

36. Among the given layered models, labelled as P, Q, R and S, which of the following pairs is/are NOT possible to distinguish according to the principle of equivalence?



Model parameters read off the figure -- (P): $\rho_1 = 1\ \Omega\text{m}, h_1 = 10\ \text{m}; \rho_2 = 100\ \Omega\text{m}, h_2 = 20\ \text{m}; \rho_3 = 0\ \Omega\text{m}.$

(Q): $\rho_1 = 5\ \Omega\text{m}, h_1 = 10\ \text{m}; \rho_2 = 400\ \Omega\text{m}, h_2 = 5\ \text{m}; \rho_3 = 0\ \Omega\text{m}.$

(R): $\rho_1 = 40\ \Omega\text{m}, h_1 = 20\ \text{m}; \rho_2 = 2\ \Omega\text{m}, h_2 = 10\ \text{m}; \rho_3 = 100\ \Omega\text{m}.$

(S): $\rho_1 = 30\ \Omega\text{m}; \rho_2 = 3\ \Omega\text{m}, h_2 = 10\ \text{m}$, with $h_1 = 5\ \text{m}; \rho_3 = 100\ \Omega\text{m}.$

(A) P, R

(B) P, S

(C) P, Q

(D) R, S

Correct Answer: (C) P, Q

Solution:

The principle of equivalence in DC resistivity sounding says that when a middle layer is thin compared to its depth, a range of (thickness, resistivity) combinations for that middle layer produce essentially indistinguishable apparent-resistivity sounding curves, PROVIDED a particular combined parameter is held constant:

- If the middle layer is **resistive** relative to its neighbours (a K-type sequence, $\rho_1 < \rho_2 > \rho_3$), the curves are equivalent when the **transverse resistance** $T = \rho_2 h_2$ is the same (current is forced to flow mostly across/through the resistive layer, so only the total resistance it presents matters).
- If the middle layer is **conductive** relative to its neighbours (an H-type sequence, $\rho_1 > \rho_2 < \rho_3$), the curves are equivalent when the **longitudinal conductance** $S = h_2/\rho_2$ is the same (current channels along the conductive layer, so only its total conductance matters).

Classify and compute for each model:

P: $\rho_1 = 1 < \rho_2 = 100 > \rho_3 = 0$ -- resistive middle layer (K-type). $T_P = \rho_2 h_2 = 100 \times 20 = 2000\ \Omega\text{m}^2$.

Q: $\rho_1 = 5 < \rho_2 = 400 > \rho_3 = 0$ -- resistive middle layer (K-type). $T_Q = \rho_2 h_2 = 400 \times 5 = 2000\ \Omega\text{m}^2$.

R: $\rho_1 = 40 > \rho_2 = 2 < \rho_3 = 100$ -- conductive middle layer (H-type). $S_R = h_2/\rho_2 = 10/2 = 5$ S.

S: $\rho_1 = 30 > \rho_2 = 3 < \rho_3 = 100$ -- conductive middle layer (H-type). $S_S = h_2/\rho_2 = 10/3 \approx 3.33$ S.

P and Q are both K-type with IDENTICAL transverse resistance ($T = 2000\ \Omega\text{m}^2$ for both), so their sounding curves are equivalent and CANNOT be distinguished from one another. R and S are both H-type but their conductances differ (5 S versus 3.33 S), so R and S ARE distinguishable from each other -- ruling out option (D). P and R (option A) and P and S (option B) are of different curve types (K versus H) entirely and are trivially distinguishable, not a matched equivalence pair.

P and Q cannot be distinguished -- option (C)

Quick Tip: For a resistive middle layer (K-type), curves are equivalent when rho2*h2 (transverse resistance) matches; for a conductive middle layer (H-type), they are equivalent when h2/rho2 (conductance) matches. Compute both for all four models.

37. Given data $d = \begin{bmatrix} d_{11} \\ d_{21} \end{bmatrix}$ and the kernel $G = \begin{bmatrix} G_{11} & G_{12} & G_{13} \\ G_{21} & G_{22} & G_{23} \end{bmatrix}$, which one of the following expressions correctly represents the generalized linear inverse formula for the model, m, satisfying $d = Gm$?

(A) $(G^T G)^{-1} G^T$

(B) $G^T (G G^T)^{-1}$

(C) $(G G^T)^{-1} G$

(D) $G G^T G^{-1}$

Correct Answer: (B) $G^T (G G^T)^{-1}$

Solution:

Here d is 2×1 (two data values) and G is 2×3 (three model parameters, two equations), so the system $d = Gm$ is UNDERDETERMINED -- fewer equations than unknowns, meaning infinitely many exact solutions m satisfy $Gm = d$. The standard way to pick a unique, well-behaved solution from this infinite family is the minimum-norm solution: among all m satisfying $Gm = d$, choose the one with the smallest $\|m\|^2 = m^T m$.

Set up the constrained minimisation with a Lagrange multiplier vector λ:

$$L = m^T m - \lambda^T (Gm - d)$$

Differentiating with respect to m and setting it to zero:

$$2m - G^T\lambda = 0 \quad \Rightarrow \quad m = \tfrac{1}{2}G^T\lambda$$

Absorbing the factor of 2 into λ, write $m = G^T\lambda$. Substitute into the constraint $Gm = d$:

$$G(G^T\lambda) = d \quad \Rightarrow \quad (GG^T)\lambda = d \quad \Rightarrow \quad \lambda = (GG^T)^{-1}d$$

(this inverse exists because GG^T is a 2×2 matrix, which is invertible whenever G has full row rank -- the underdetermined case). Substituting back:

$$m = G^T(GG^T)^{-1}d$$

so the generalized (minimum-norm) inverse operator that maps data to model is $G^T(GG^T)^{-1}$, matching option (B). This is the standard underdetermined generalized inverse, complementary to the overdetermined least-squares inverse $(G^TG)^{-1}G^T$ used when there are more equations than unknowns (option A, not applicable here since G^TG would be a singular 3×3 matrix for only 2 independent rows).

$$\boxed{G^{-g} = G^T(GG^T)^{-1}}$$

Quick Tip: With 2 data and 3 model parameters the system is underdetermined; minimize ||m||^2 subject to Gm=d using a Lagrange multiplier to get m = G^T(GG^T)^{-1}d.

38. If a seismic wave is travelling along the radial direction, then which of the following statements is/are CORRECT for Rayleigh (R) and Love (L) waves?

(A) Radial component shows the largest amplitude for both R and L waves

(B) Transverse component shows the largest and smallest amplitudes for L and R waves, respectively

(C) Vertical component shows the largest and smallest amplitudes for L and R waves, respectively

(D) Vertical component shows largest amplitude for R wave only

Correct Answer: (B) Transverse component shows the largest and smallest amplitudes for L and R waves, respectively ; (D) Vertical component shows largest amplitude for R wave only

Solution:

Set up local ground-motion axes at a receiver for a wave arriving along the radial (great-circle) direction: radial (R-component, horizontal, along the source-receiver azimuth), transverse (T-component, horizontal, perpendicular to that azimuth) and vertical (Z-component).

Love waves are formed by constructive interference of horizontally-polarised shear (SH) waves trapped in a low-velocity surface layer. Their particle motion is PURELY horizontal and PURELY perpendicular to the propagation direction -- i.e. entirely on the transverse component. A Love wave therefore has essentially ZERO radial and ZERO vertical motion, and its transverse-component amplitude is the largest (in fact the only) component present.

Rayleigh waves arise from coupled P-SV motion at the free surface. Particle motion is retrograde-elliptical and confined entirely to the vertical plane containing the propagation direction -- i.e. it has radial and vertical components only, and ZERO transverse component (SH motion plays no part in Rayleigh waves). Of its two non-zero

components, the vertical component is typically the larger/dominant one (this is why vertical-component seismometers are used to pick Rayleigh-wave dispersion), with the radial component present but generally smaller.

Now check each option:

- (A) False -- radial component is zero for the Love wave, so it cannot be the largest for both wave types.
- (B) Transverse component: the Love wave's motion is entirely transverse, so it is the LARGEST (in fact the only) component for L; the Rayleigh wave has zero transverse motion, so it is the SMALLEST (zero) for R. This is exactly what option (B) states -- CORRECT.
- (C) False -- this reverses the truth. Vertical motion is zero for the Love wave and dominant for the Rayleigh wave, i.e. smallest for L and largest for R -- the opposite of what option (C) claims.
- (D) Vertical component shows the largest amplitude for the R wave only -- correct, since Rayleigh has substantial (dominant) vertical motion while Love has none at all, so R wave only is an accurate qualifier -- CORRECT.

So the correct statements are (B) and (D).

(B) and (D)

Quick Tip: Love waves are purely SH (transverse-only) surface waves; Rayleigh waves are P-SV coupled with dominant vertical and secondary radial motion but zero transverse motion.

39. If a signal is sampled at a sampling rate of 0.2 ms, then its sampling frequency is _____kHz.

Correct Answer: 5

Solution:

The sampling frequency (sampling rate in Hz) is simply the reciprocal of the sampling interval (the time between successive samples):

$$f_s = \frac{1}{\Delta t}$$

Here the sampling interval is given as $\Delta t = 0.2\ \text{ms} = 0.2 \times 10^{-3}\ \text{s} = 2 \times 10^{-4}\ \text{s}$. Substituting:

$$f_s = \frac{1}{2 \times 10^{-4}\ \text{s}} = 5000\ \text{Hz} = 5\ \text{kHz}$$

So a sample is taken every 0.2 milliseconds, which corresponds to exactly 5000 samples being recorded every second, i.e. a sampling frequency of 5 kHz.

$\boxed{f_s = 5\ \text{kHz}}$

Quick Tip: Sampling frequency is the reciprocal of the sampling interval: f_s = 1/delta_t.

40. A P-wave of frequency 20 Hz is travelling through a non-dispersive medium with a velocity of 5 km/s. The amplitude retained at a distance of 10 km from source is _____ % (rounded off to one decimal place). (Use quality factor, $Q = 80$)

Correct Answer: 20 to 21

Solution:

Intrinsic (anelastic) attenuation of a seismic wave's amplitude as it travels a distance x through a medium of quality factor Q is given by:

$$\frac{A}{A_0} = \exp\left(-\frac{\pi f x}{Q v}\right)$$

where f is the frequency, v is the (non-dispersive) wave velocity, and x is the propagation distance. This form follows from writing the travel time $t = x/v$ and using the standard temporal attenuation law $A/A_0 = e^{-\pi f t/Q}$.

Substitute the given values: $f = 20$ Hz, $x = 10$ km, $Q = 80$, $v = 5$ km/s:

$$\frac{\pi f x}{Q v} = \frac{\pi \times 20 \times 10}{80 \times 5} = \frac{200\pi}{400} = \frac{\pi}{2} = 1.5708$$

So:

$$\frac{A}{A_0} = e^{-1.5708} = 0.2079$$

Converting to a percentage:

$$\frac{A}{A_0} \times 100\% = 20.8\%$$

This value of 20.8% lies within the official accepted range of 20 to 21%. Physically, this shows how strongly a 20 Hz P-wave is attenuated by intrinsic damping (Q = 80) over just 10 km -- only about a fifth of the original amplitude survives, which is why higher-frequency seismic signals lose amplitude much faster with distance than lower-frequency ones for the same Q.

$$\boxed{\frac{A}{A_0} \approx 20.8\%}$$

Quick Tip: Use A/A0 = exp(-pi*f*x/(Q*v)); with f=20 Hz, x=10 km, v=5 km/s, Q=80 the exponent is pi/2, giving about 20.8% amplitude retained.

41. A land magnetic survey was carried out along a profile length of 500 m with an inter-station spacing of 5 m over a buried ore body. The optimum width of the body that can be best resolved is _____ m (answer in integer).

Correct Answer: 10

Solution:

In any profile survey (magnetic, gravity, resistivity, etc.), the station spacing Δx sets a limit on the smallest feature that can be reliably picked up, in the same way a sampling interval limits the highest frequency recoverable in time-series work.

Step 1: Treat the station spacing as a spatial sampling interval.
Here $\Delta x = 5$ m. By the spatial version of the sampling (Nyquist) theorem, at least two independent readings must fall across the width of a target for its anomaly to be resolved without aliasing - a single station only tells you a value exists, not the shape or width of the source causing it.

Step 2: Apply the two-samples-per-target rule.
Optimum (minimum reliably resolvable) width, $w_{opt} = 2\Delta x = 2 \times 5 = 10$ m.

Step 3: Note what the profile length is for.
The 500 m profile length only fixes how many stations exist along the line ($500/5 + 1 = 101$ stations); it does not enter the resolution

calculation, since resolution is governed by the sampling interval, not the total survey length.

So the ore body's optimum resolvable width is $\boxed{10\ \text{m}}$, matching the key.

Quick Tip: Use the spatial sampling (Nyquist) rule - a target needs at least two stations across its width to be properly resolved.

42. If the depth of penetration of an inducing electromagnetic wave of 1 kHz frequency is 400 m, then the resistivity of the subsurface is _____ Ωm (answer in integer).

Correct Answer: 630 to 640

Solution:

The depth of penetration of a plane EM wave into a conducting Earth is its skin depth, δ, given by the standard geophysical rule-of-thumb formula $\delta = 503\sqrt{\rho/f}$, with δ in metres, ρ in Ωm and f in Hz.

Step 1: Write down the knowns.

$\delta = 400$ m, $f = 1000$ Hz.

Step 2: Solve for ρ.

$400 = 503\sqrt{\rho/1000}$

$\sqrt{\rho/1000} = 400/503 = 0.7952$

$\rho/1000 = 0.7952^2 = 0.6324$

$\rho = 632.4\ \Omega$m.

Rounded to the nearest integer, $\rho \approx 632\ \Omega$m, which lies inside the accepted band of $\boxed{630 \text{ to } 640\ \Omega\text{m}}$.

Quick Tip: Use the EM skin-depth (depth of penetration) relation, $\delta = 503\sqrt{\rho/f}$, and solve for ρ.

43. If F represents the Earth's total magnetic field corresponding to its dipole source only, then the increase in F from 0° to 60° N magnetic latitude is _____ % (rounded off to one decimal place).

Correct Answer: 80.0 to 80.5

Solution:

For a pure geocentric dipole, the total field magnitude at magnetic latitude λ is $F(\lambda) = F_0\sqrt{1 + 3\sin^2\lambda}$, where F_0 is the field magnitude at the magnetic equator ($\lambda = 0$).

Step 1: Field at the equator.

At $\lambda = 0°$: $F(0) = F_0$.

Step 2: Field at 60°N.

$$F(60°) = F_0\sqrt{1 + 3\sin^2 60°} = F_0\sqrt{1 + 3(0.75)} = F_0\sqrt{3.25} = 1.8028\,F_0.$$

Step 3: Percentage increase.

$$\frac{F(60°) - F(0)}{F(0)} \times 100 = (1.8028 - 1) \times 100 = 80.28\%.$$

Rounded to one decimal place, the increase is $\boxed{80.3\%}$, inside the accepted band 80.0 to 80.5%.

Quick Tip: Use the geocentric dipole field formula $F(\lambda) = F_0\sqrt{1 + 3\sin^2\lambda}$ and compare its values at 0 and 60 degrees.

44. Considering a surface point source, the horizontal resolution recoverable from the seismic data acquired using a 10 Hz seismic pulse reflected from a depth of 30 km within the crust, with a velocity of 6.5 km/s, is _____ km (rounded off to two decimal places).

Correct Answer: 6.2 to 6.3

Solution:

The lateral (horizontal) resolution limit of unmigrated seismic reflection data is set by the size of the first Fresnel zone - the patch of the reflector from which reflected energy arrives within half a period of the first arrival and therefore interferes constructively.

Step 1: Two-way travel time to the reflector.

$$t_0 = \frac{2z}{v} = \frac{2 \times 30}{6.5} = 9.2308 \text{ s.}$$

Step 2: First Fresnel-zone radius.

$$r_F = \frac{v}{2}\sqrt{\frac{t_0}{f}} = \frac{6.5}{2}\sqrt{\frac{9.2308}{10}} = 3.25 \times 0.9608 = 3.1225 \text{ km.}$$

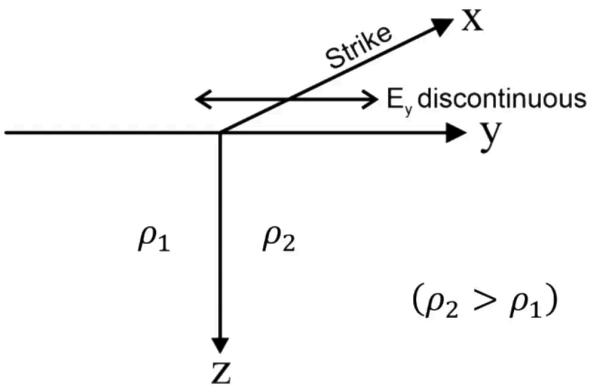
Step 3: Horizontal resolution.

The horizontal resolution - the closest that two reflecting points can be and still be distinguished by an unmigrated, surface-point-source survey - is the Fresnel-zone diameter: $2r_F = 2 \times 3.1225 = 6.245$ km.

Rounded to two decimal places, the horizontal resolution is $\boxed{6.24\text{-}6.25 \text{ km}}$, inside the accepted band 6.20 to 6.30 km.

Quick Tip: Compute the first Fresnel-zone radius, $r_F = \sqrt{vz/(2f)}$, and remember horizontal resolution is its diameter, $2r_F$.

45. A vertical contact separating two resistive domains, ρ_1 and ρ_2 (with $\rho_2 > \rho_1$), is shown in the figure below, with the strike direction along X, the profile direction (perpendicular to strike) along Y - where the electric field component E_y is discontinuous across the contact - and depth along Z.



The magnitude of the discontinuity in the apparent resistivity, when the electric field, E, is perpendicular to the strike, is

(A) $(\rho_1/\rho_2)^2$

(B) $(\rho_2/\rho_1)^2$

(C) $(\rho_1/\rho_2)^{1/2}$

(D) $(\rho_2/\rho_1)^{1/2}$

Correct Answer: (A) $(\rho_1/\rho_2)^2$

Solution:

When the electric field is perpendicular to strike, current flows across the vertical contact - this is the H-polarization (TM) mode of a 2-D resistivity structure, exactly the mode the figure highlights by marking E_y as discontinuous.

Step 1: Boundary conditions across the vertical contact.

The magnetic field component along strike, H_x, is tangential to the contact and must be continuous across it. The current density

perpendicular to the contact, J_y, must also be continuous (charge conservation - no sources or sinks of current at the boundary itself).

Step 2: Relate the electric fields on the two sides.
Since $J_y = E_y/\rho$ and J_y is continuous, $E_{y1}/\rho_1 = E_{y2}/\rho_2 = J_y$, so $E_{y1} = J_y\rho_1$ and $E_{y2} = J_y\rho_2$, giving $E_{y1}/E_{y2} = \rho_1/\rho_2$.

Step 3: Apparent resistivity on each side.
Apparent resistivity from the impedance $Z_{yx} = E_y/H_x$ goes as $\rho_a \propto |E_y/H_x|^2$. Since H_x is the same (continuous) on both sides of the contact, the ratio of the apparent resistivities computed just on the ρ_1 side to that on the ρ_2 side is $\rho_{a1}/\rho_{a2} = (E_{y1}/E_{y2})^2 = (\rho_1/\rho_2)^2$.

So the magnitude of the discontinuity in apparent resistivity across the vertical contact, for E perpendicular to strike, is $\boxed{(\rho_1/\rho_2)^2}$ - option (A).

Quick Tip: E perpendicular to strike is the H-polarization (TM) mode - use continuity of tangential H and of normal current density across the contact.

46. If g_{FA} and g_{BA} respectively denote the free-air and Bouguer gravity anomalies over a fully compensated mountain, then which one of the following is CORRECT in case of Airy-isostacy?

(A) $g_{FA} = 0,\ g_{BA} > 0$

(B) $g_{FA} > 0,\ g_{BA} = 0$

(C) $g_{FA} > 0,\ g_{BA} < 0$

(D) $g_{FA} = 0,\ g_{BA} = 0$

Correct Answer: (C) $g_{FA} > 0,\ g_{BA} < 0$

Solution:

Under Airy's model of isostasy, a mountain is supported by a low-density crustal root that displaces denser mantle material, so that the total mass in every vertical column - from the mountain top down through the compensation depth - is the same everywhere ('fully compensated').

Step 1: What the free-air correction does.

The free-air correction only accounts for the change of gravity with elevation (about 0.3086 mGal/m); it removes the effect of being at a different height, but it does NOT remove the attraction of the topographic mass itself sitting between the station and sea level. Because that mass is physically present and pulling on the gravimeter, the free-air anomaly over a mountain retains a small positive signature, $g_{FA} > 0$, even though the system is isostatically compensated overall.

Step 2: What the Bouguer correction adds.

The Bouguer correction goes a step further and mathematically strips away the gravitational effect of the topographic mass itself (approximated as an infinite slab), as if that rock were not there. Once that mass is removed, all that is left in the anomaly is the effect of the low-density compensating root at depth. A root of density lower than the surrounding mantle is a mass deficit, and a mass deficit produces a negative gravity anomaly, so $g_{BA} < 0$ over the mountain.

Step 3: Combine.

Hence, for a fully (Airy) compensated mountain, $g_{FA} > 0$ and $g_{BA} < 0$ - option $\boxed{(C)}$, the classic diagnostic used to confirm isostatic compensation from real gravity surveys (near-zero-to-small-positive free-air anomaly alongside a strongly negative Bouguer anomaly that deepens with elevation).

Quick Tip: Free-air only corrects for elevation (topographic mass still present, so slightly positive); Bouguer removes that mass and exposes the low-density root (a mass deficit, so negative).

47. Analytic signal, $A(\theta)$, for $f(\theta) = \cos\theta$ is equal to

(A) $A(\theta) = -\cos\theta$

(B) $A(\theta) = -\sin\theta$

(C) $A(\theta) = e^{i\theta}$

(D) $A(\theta) = -e^{i\theta}$

Correct Answer: (C) $A(\theta) = e^{i\theta}$

Solution:

The analytic signal of a real function $f(\theta)$ is built by adding i times its Hilbert transform to itself: $A(\theta) = f(\theta) + i\,\hat{f}(\theta)$, where $\hat{f}$ denotes the Hilbert transform of f.

Step 1: Hilbert transform of $\cos\theta$.

The Hilbert transform acts as a 90-degree phase-delay filter: it turns $\cos\theta$ into $\sin\theta$ (and, consistently, turns $\sin\theta$ into $-\cos\theta$, so that applying it twice gives back $-f(\theta)$, as required of any Hilbert-transform pair). So $\hat{f}(\theta) = \sin\theta$.

Step 2: Build the analytic signal.

$A(\theta) = \cos\theta + i\sin\theta$.

Step 3: Recognize the complex exponential.

By Euler's formula, $\cos\theta + i\sin\theta = e^{i\theta}$.

So $A(\theta) = \boxed{e^{i\theta}}$ - option (C).

Quick Tip: The analytic signal is $f(\theta) + i\hat{f}(\theta)$; the Hilbert transform of $\cos\theta$ is $\sin\theta$, giving Euler's identity.

48. In regions of lateral conductivity contrasts, the time-independent separation between the subsurface apparent resistivity curves of transverse electric (TE) and transverse magnetic (TM) modes relative to each other arises mainly due to

(A) impedance phase change

(B) amplitude magnification

(C) local distortion of electric field

(D) variation in the frequency of the inducing magnetic field

Correct Answer: (C) local distortion of electric field

Solution:

The TE and TM apparent resistivity curves are built from two different combinations of the surface electric and magnetic field components, so anything that distorts the electric field alone will move the two curves apart.

In areas with lateral conductivity contrasts, small near-surface inhomogeneities (bodies much smaller than the skin depth of the sounding) become galvanically charged as current flows through them. The charge that piles up on their boundaries creates a secondary, essentially static electric field that adds to the regional field. This charge build-up is a DC (galvanic) effect - it depends only on the geometry and resistivity contrast of the near-surface body, not on how fast the inducing field is oscillating. Because of this, the

distortion is present at every period equally, i.e. it is time-independent.

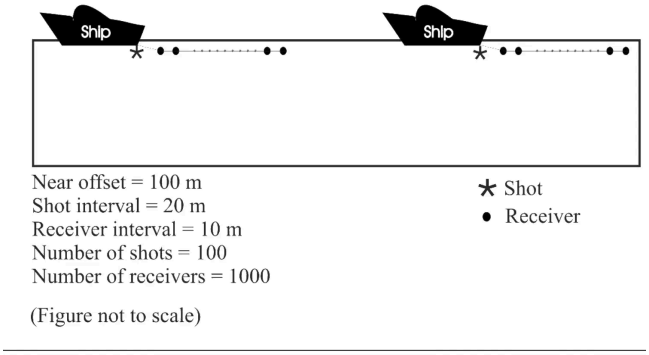
Since apparent resistivity scales as $(E/H)^2$, a constant multiplicative distortion of E shows up as a constant vertical shift of the apparent resistivity curve on a log-log plot, at all periods - the well-known "static shift". Because TE apparent resistivity uses the electric field parallel to strike and TM apparent resistivity uses the field perpendicular to strike, the same near-surface distortion shifts the two curves by different (but each frequency-independent) amounts, so the curves sit apart from one another by a fixed, time-independent separation.

Impedance phase change, amplitude magnification, and variation of the inducing frequency are all inductive, frequency-dependent phenomena - they would produce effects that change with period, not a fixed offset. The observed time-independent separation is therefore caused by local (galvanic) distortion of the electric field around near-surface conductivity heterogeneities.

Hence the correct answer is **local distortion of electric field**, option (C).

Quick Tip: Think about static shift in MT: near-surface charge build-up distorts only the electric field, and this distortion does not change with frequency/period.

49. How many traces will be present in a zero-offset section that is generated from the following marine seismic survey?



Near offset = 100 m

Shot interval = 20 m

Receiver interval = 10 m

Number of shots = 100

Number of receivers = 1000

(Figure not to scale)

(A) 2400

(B) 2414

(C) 2450

(D) 2420

Correct Answer: (B) 2414

Solution:

A single vessel tows one streamer while steaming along the line, firing a shot every 20 m (the shot interval). The streamer carries 1000 receiver channels spaced 10 m apart, with the nearest channel sitting 100 m behind the source (the near offset). Because a fresh streamer record is captured after every shot, the survey ends up illuminating a

stretch of subsurface midpoints that is the SUM of two separate lengths: the physical length of the towed spread, plus the distance the boat itself travels while shooting.

Length of the towed spread, from the source out to the farthest (last) channel:

$L_{spread} = \text{near offset} + (N_{receivers} - 1) \times RI = 100 + (1000 - 1) \times 10 = 100 + 9990 = 10090 \text{ m}$

Distance the vessel sails between the first shot and the last shot:

$L_{sail} = (N_{shots} - 1) \times SI = (100 - 1) \times 20 = 1980 \text{ m}$

Total subsurface length covered along the profile:

$L_{total} = L_{spread} + L_{sail} = 10090 + 1980 = 12070 \text{ m}$

Because the shot interval (20 m) is exactly twice the receiver interval (10 m), this acquisition geometry gives smooth, uniform-fold coverage when the common-midpoint (CMP) bin is taken as half of the receiver interval:

$$\text{CMP interval} = \frac{RI}{2} = \frac{10}{2} = 5 \text{ m}$$

The number of distinct zero-offset (CMP) traces that this profile can generate is the total covered length divided by the CMP bin size:

$$N_{traces} = \frac{L_{total}}{\text{CMP interval}} = \frac{12070}{5} = 2414$$

So the zero-offset section produced from this survey will contain $\boxed{2414}$ traces, option (B).

Quick Tip: Total CMP-covered length = spread length + sail-line distance; divide by the CMP bin size (half the receiver interval, since shot interval = 2 x

receiver interval here).

50. $\Delta T = [\Delta T_1, \Delta T_2, \ldots, \Delta T_{19}, \Delta T_{20}]$, and $\Delta T_u = [\Delta T_{u1}, \Delta T_{u2}, \ldots, \Delta T_{u19}, \Delta T_{u20}]$ denote the magnetic data observed at heights 0 km and 5 km, respectively, along a profile of length 100 km. What is the maximum attenuation at a height of 5 km?
[Use wave number in radian/km].

(A) $e^{-\pi/2}$

(B) $e^{-\pi}$

(C) $e^{-2\pi}$

(D) $e^{-3\pi}$

Correct Answer: (B) $e^{-\pi}$

Solution:

Upward continuation of a potential field is a simple multiplication in the wavenumber domain: each spatial Fourier component of the field is attenuated according to

$$F_u(k) = F(k)\, e^{-kh}$$

where k is the angular wavenumber (in radian/km, as specified) of that component and h is the height gained (here 5 km). Since the attenuation factor e^{-kh} decreases as k grows, the STRONGEST (maximum) attenuation of the dataset happens for the shortest-wavelength (highest wavenumber) component that the sampled profile can actually carry - the Nyquist component.

The profile is 100 km long and is represented by 20 discrete samples $\Delta T_1, \ldots, \Delta T_{20}$, so the station (sample) spacing is

$$\Delta x = \frac{100\ \text{km}}{20} = 5\ \text{km}$$

The highest (Nyquist) angular wavenumber that this sampling can resolve is

$$k_{max} = \frac{\pi}{\Delta x} = \frac{\pi}{5}\ \text{rad/km}$$

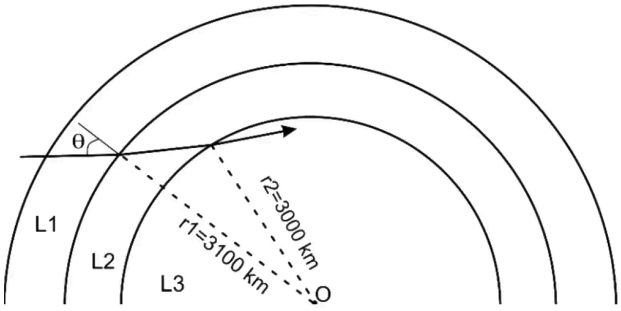
Putting this into the upward-continuation attenuation formula for $h = 5$ km:

$$e^{-k_{max}h} = e^{-(\pi/5)(5)} = e^{-\pi}$$

So the maximum attenuation suffered by the data on being continued upward to 5 km is $\boxed{e^{-\pi}}$, option (B).

Quick Tip: Upward continuation attenuates each Fourier component by e^{-kh}; maximum attenuation occurs at the Nyquist wavenumber set by the 20-sample, 100 km profile spacing.

51. The given figure shows the geometry of a seismic ray path inside the Earth, with the P-wave velocities of 10 km/s, 11 km/s and 12 km/s corresponding to the layers L1, L2 and L3, respectively. If the angle of incidence (θ) at the L1-L2 boundary is $40°$, then what is the angle of refraction at the L2-L3 boundary? (answer in nearest integer)



(In the figure, $r_1 = 3100$ km is the radius, measured from the Earth's centre O, of the L1-L2 boundary, and $r_2 = 3000$ km is the radius of the L2-L3 boundary.)

(A) $49°$

(B) $50°$

(C) $53°$

(D) $55°$

Correct Answer: (C) $53°$

Solution:

This is a spherically layered Earth, so the problem has to be handled in two distinct steps: an ordinary (flat) Snell's law refraction at each velocity boundary, plus a geometrical correction because the ray travels as a straight line between two boundaries that sit at different radii from the Earth's centre.

Step 1 - Refraction at the L1-L2 boundary (radius r_1). Both the

incident and refracted rays touch the boundary at the same radius r_1, so ordinary Snell's law applies exactly as in flat-layered media:

$$\frac{\sin\theta_2}{v_2} = \frac{\sin\theta_1}{v_1} \Rightarrow \sin\theta_2 = \frac{v_2}{v_1}\sin\theta_1 = \frac{11}{10}\sin 40^\circ = 1.1 \times 0.6428 = 0.7071$$

$$\theta_2 = \sin^{-1}(0.7071) \approx 45^\circ$$

Step 2 - Straight-line travel through L2, from radius r_1 to radius r_2. Inside layer L2 the velocity is constant, so the ray is a straight chord. For any straight line, the perpendicular distance from the centre O to the line - equal to $r\sin\theta(r)$ at any point on the line - never changes. Applying this at the two ends of the chord (radius r_1 where the angle is θ_2, and radius r_2 where the angle is θ_2'):

$$r_1\sin\theta_2 = r_2\sin\theta_2' \Rightarrow \sin\theta_2' = \frac{r_1}{r_2}\sin\theta_2 = \frac{3100}{3000} \times 0.7071 = 0.7308$$

$$\theta_2' = \sin^{-1}(0.7308) \approx 46.9^\circ$$

Step 3 - Refraction at the L2-L3 boundary (radius r_2). Again both rays share the same radius, so ordinary Snell's law applies:

$$\frac{\sin\theta_3}{v_3} = \frac{\sin\theta_2'}{v_2} \Rightarrow \sin\theta_3 = \frac{v_3}{v_2}\sin\theta_2' = \frac{12}{11} \times 0.7308 = 0.7972$$

$$\theta_3 = \sin^{-1}(0.7972) \approx 52.85^\circ$$

Rounded to the nearest integer, the angle of refraction at the L2-L3 boundary is $\boxed{53^\circ}$, option (C).

Quick Tip: Apply ordinary Snell's law at each boundary (same radius on both sides), and separately use $r \sin\theta$ =constant for the straight ray segment travelling between the two different radii inside L2.

52. Which of the following statements about digital filters is/are CORRECT?

(A) Non-recursive filters do not have poles

(B) Recursive filters can have only poles OR both poles and zeros

(C) The order of the digital filter defines the presence of minimum number of previous inputs only

(D) The ratio of output to input defines the transfer function of the digital filter

Correct Answer: (A) Non-recursive filters do not have poles ; (B) Recursive filters can have only poles OR both poles and zeros ; (D) The ratio of output to input defines the transfer function of the digital filter

Solution:

Go through each statement using the standard definitions of digital filter structures.

(A) Non-recursive filters do not have poles - CORRECT. A non-recursive (FIR) filter computes its output purely from the current and past INPUT samples:

$$y_n = \sum_{k=0}^{M} b_k \, x_{n-k}$$

Its transfer function $H(z) = \sum b_k z^{-k}$ is a polynomial in z^{-1} with only zeros (roots of the numerator); there is no feedback denominator, so there are no non-trivial poles. This statement is true.

(B) Recursive filters can have only poles OR both poles and zeros - CORRECT. A recursive (IIR) filter feeds back past OUTPUT values as well:

$$y_n = \sum_k b_k x_{n-k} + \sum_k a_k y_{n-k}$$

giving a transfer function $H(z) = \dfrac{\sum b_k z^{-k}}{1 - \sum a_k z^{-k}}$. The denominator always contributes poles. If the numerator is just a constant (an all-pole filter), the filter has poles only; if the numerator also has multiple terms, the filter has both poles and zeros. Either situation is possible for a recursive filter, so this statement is true.

(C) The order of the digital filter defines the presence of minimum number of previous inputs only - INCORRECT. The order of a filter is set by the number of delay elements needed, which for a recursive filter includes both previous INPUT samples and previous OUTPUT samples (the feedback terms), not inputs alone. Restricting the definition to "previous inputs only" is therefore wrong.

(D) The ratio of output to input defines the transfer function of the digital filter - CORRECT. By definition, the (z-

domain) transfer function is exactly $H(z) = Y(z)/X(z)$, the ratio of the z-transform of the output to that of the input. This is textbook-correct.

So the correct statements are (A), (B) and (D), i.e. options $\boxed{1, 2, 4}$.

Quick Tip: Write the general recursive-filter transfer function $H(z) = Y(z)/X(z)$ and check each claim about poles, zeros, and filter order against it.

53. ΔB_z represents the maximum vertical magnetic anomaly along a profile due to a horizontal cylinder with susceptibility contrast (Δk), and radius (r) at a depth (z) below the Earth's surface. Which combination(s) of Δk, r, and z labelled as P, Q, R and S given below, produces/produce the same ΔB_z?

(P) $\Delta k = 0.02$, $r = 100\ m$, and $z = 200\ m$

(Q) $\Delta k = 0.01$, $r = 80\ m$, and $z = 160\ m$

(R) $\Delta k = 0.005$, $r = 200\ m$, and $z = 400\ m$

(S) $\Delta k = 0.02$, $r = 150\ m$, and $z = 300\ m$

(A) P, S

(B) P, Q

(C) R, S

(D) Q, R

Correct Answer: (A) P, S

Solution:

The maximum vertical magnetic anomaly of a long horizontal cylinder is proportional to the susceptibility contrast times a shape factor that

depends only on the ratio of the radius to the depth:

$$\Delta B_z \propto \Delta k \cdot r^2/z^2$$

(the anomaly of a magnetized cylinder falls off with depth the way a line of dipoles does, and the size term always enters as the dimensionless combination $(r/z)^2$, scaled by Δk).

The key observation is to check the ratio z/r for each of the four models, because if that ratio is the same for all of them, the shape factor $(r/z)^2$ becomes a common constant and ΔB_z then depends on Δk alone:
P: $z/r = 200/100 = 2$
Q: $z/r = 160/80 = 2$
R: $z/r = 400/200 = 2$
S: $z/r = 300/150 = 2$

All four models share exactly the same $z/r = 2$, so $(r/z)^2 = 1/4$ is identical in every case. This means ΔB_z for each model is controlled purely by its Δk value:
P: $\Delta k = 0.02$
Q: $\Delta k = 0.01$
R: $\Delta k = 0.005$
S: $\Delta k = 0.02$

P and S both have $\Delta k = 0.02$, so they produce the same ΔB_z; Q (0.01) and R (0.005) are each different from the rest and from each other. Hence the pair that cannot be distinguished by their maximum vertical anomaly is P and S, option $\boxed{\text{(A) P, S}}$.

Quick Tip: Note that $z/r = 2$ for all four models, so the shape factor cancels and ΔB_z depends only on Δk; look for the pair with equal Δk.

54. If $f(t) = [-1, 2, 1]$ and $g(t) = [0, -1, 2]$ are two wavelets and $f(\tau)$ and $g(\tau)$ represent flipped versions of $f(t)$ and $g(t)$, respectively, then which of the following expressions is/are CORRECT?
[* and $\otimes$ denote convolution and cross-correlation operations, respectively]

(A) $f(t) * g(t) = f(t) \otimes g(t)$

(B) $f(t) * g(t) = f(t) \otimes g(\tau)$

(C) $f(t) * g(t) = f(\tau) \otimes g(t)$

(D) $f(t) * g(t) = f(\tau) \otimes g(\tau)$

Correct Answer: (B) $f(t) * g(t) = f(t) \otimes g(\tau)$; (C) $f(t) * g(t) = f(\tau) \otimes g(t)$

Solution:

Convolution reverses (flips) one of the two sequences before sliding and multiplying, whereas cross-correlation slides the two sequences past each other WITHOUT reversing either one. Comparing the two definitions directly gives the general identity: convolving f and g is the same as cross-correlating either signal, PROVIDED the OTHER signal is flipped first - flipping just one of the two signals before cross-correlating reproduces the convolution.

Let us check this with the numbers given. First compute the full linear convolution:
$f(t) = [-1, 2, 1]$, $g(t) = [0, -1, 2]$

$$f * g = [0,\ 1,\ -4,\ 3,\ 2]$$

(term by term: $(-1)(0) = 0$; $(-1)(-1) + (2)(0) = 1$; $(-1)(2) + (2)(-1) + (1)(0) = -4$; $(2)(2) + (1)(-1) = 3$; $(1)(2) = 2$).

Now flip f: $f(\tau) = [1, 2, -1]$. Cross-correlating $f(\tau)$ with the unflipped $g(t)$ (sliding g past $f(\tau)$) reproduces exactly:

$$f(\tau) \otimes g(t) = [0,\ 1,\ -4,\ 3,\ 2]$$

which matches $f * g$ term for term - so option (C) is correct.

Similarly, flip g instead: $g(\tau) = [2, -1, 0]$. Cross-correlating the unflipped $f(t)$ with $g(\tau)$ also reproduces the same convolution sequence $[0, 1, -4, 3, 2]$ - so option (B) is correct as well.

By contrast, cross-correlating f and g with NEITHER of them flipped gives $[0, -1, 0, 5, -2]$, which is a completely different sequence from $f * g$, so option (A) is wrong. And cross-correlating $f(\tau)$ with $g(\tau)$ (BOTH flipped) reproduces $[2, 3, -4, 1, 0]$, which is just the time-reversed version of $f * g$ rather than $f * g$ itself, so option (D) is also wrong.

Hence the correct expressions are $f(t) * g(t) = f(t) \otimes g(\tau)$ and $f(t) * g(t) = f(\tau) \otimes g(t)$, i.e. options $\boxed{(B) \text{ and } (C)}$, answer 2,3.

Quick Tip: Convolution equals cross-correlation once EXACTLY ONE of the two signals is time-reversed (flipped); flipping both or neither does not reproduce the convolution.

55. Which of the following assumptions is/are valid for electromagnetic induction to happen in the Earth?

(A) Plane-wave approximation of the inducing field

(B) Displacement currents are not neglected

(C) Earth behaves as Ohmic conductor due to conservation of charge

(D) Earth does not generate electromagnetic energy, but only absorbs or dissipates it

Correct Answer: (A) Plane-wave approximation of the inducing field ; (C) Earth behaves as Ohmic conductor due to conservation of charge ; (D) Earth does not generate electromagnetic energy, but only absorbs or dissipates it

Solution:

In magnetotelluric (MT) and geomagnetic depth-sounding theory, EM induction in the Earth rests on a set of standard simplifying assumptions. Check each option against them.

Step 1 - Plane-wave approximation: The natural source currents (ionospheric and magnetospheric current systems) sit far above the Earth's surface compared with the induction scale lengths involved, so at the surface the primary field can be treated as a uniform, vertically incident plane wave. This is a foundational assumption of MT theory, so (A) is correct.

Step 2 - Displacement currents: MT induction operates in the quasi-static (diffusive) regime because Earth materials are relatively good conductors and the frequencies used are low. Here the conduction current density σE dominates the displacement current density $\epsilon \partial E/\partial t$ by many orders of magnitude, so displacement

currents ARE neglected, not retained. The statement "displacement currents are not neglected" is therefore FALSE, so (B) is incorrect.

Step 3 - Ohmic behaviour: Applying Ohm's law $\mathbf{J} = \sigma\mathbf{E}$ together with charge conservation (the continuity equation $\nabla \cdot \mathbf{J} = -\partial\rho/\partial t$) shows that any free charge in a conductor relaxes away almost instantaneously, so the Earth behaves as a pure Ohmic conductor. This matches (C), which is correct.

Step 4 - Passive medium: The Earth has no internal EM source in this problem; it only responds to the externally imposed inducing field, dissipating the energy resistively (Joule heating) rather than generating any of its own. This matches (D), which is correct.

The valid assumptions are (A), (C) and (D). $\boxed{A,\ C,\ D}$

Quick Tip: MT/EM induction theory uses the quasi-static (diffusion) approximation, which means displacement currents are dropped, not retained.

56. The model regularization in a damped least square problem is/are used to

(A) introduce additional constraints to the ill-posed problems

(B) add stability to the inversion

(C) convert well-posed problems into ill-posed problems

(D) control the trade-off between the misfit and the model variance

Correct Answer: (A) introduce additional constraints to the ill-posed problems ; (B) add stability to the inversion ; (D) control the trade-off between the misfit and the model variance

Solution:

The damped least-squares (Marquardt-Levenberg / Tikhonov-style) inversion minimizes an objective function of the form

$$\Phi(\mathbf{m}) = \|\mathbf{d} - \mathbf{G}\mathbf{m}\|^2 + \lambda^2\|\mathbf{m} - \mathbf{m}_0\|^2$$

where the first term is the data misfit and the second is the model regularization (damping) term, weighted by the trade-off parameter λ. Check each statement against this.

Step 1: Most geophysical inverse problems (gravity, magnetic, EM, seismic) are ill-posed - a unique, stable model cannot be recovered from noisy, finite data alone. Adding the regularization term introduces an extra piece of information (e.g. a preference for a smooth or small model) that helps select one acceptable solution among the many that fit the data almost equally well. This matches (A).

Step 2: Without regularization, the plain least-squares solution $(\mathbf{G}^T\mathbf{G})^{-1}\mathbf{G}^T\mathbf{d}$ can become wildly unstable whenever $\mathbf{G}^T\mathbf{G}$ is near-singular (small eigenvalues amplify noise enormously). Adding $\lambda^2\mathbf{I}$ (or another regularization operator) before inverting keeps the matrix well-conditioned, stabilizing the inversion. This matches (B).

Step 3: Regularization does the OPPOSITE of what (C) states - it is applied precisely because the problem is already ill-posed, and its entire purpose is to make the solution well-behaved, not to make a well-posed problem worse. So (C) is FALSE.

Step 4: The damping parameter λ directly sets how much weight is given to fitting the data versus keeping the model close to the reference/simple - i.e. it controls the trade-off between the data misfit and the model variance (or model norm). This matches (D).

The correct statements are (A), (B) and (D). $\boxed{A, B, D}$

Quick Tip: Regularization stabilizes ill-posed inversions and never turns a well-posed problem into an ill-posed one.

57. The vector field $\mathbf{F}$ for a region is described by $\mathbf{F} = ar^n\hat{r}$ $(r \neq 0)$, where a is a non-zero constant and r is the radial distance from the source. For what value(s) of n, $\mathbf{F}$ becomes both solenoidal and irrotational?

(A) 2

(B) -2

(C) 1

(D) -1

Correct Answer: (B) -2

Solution:

We are given the radial vector field $\mathbf{F} = ar^n\hat{r}$, $r \neq 0$.

Step 1 - Irrotational check: For any purely radial vector field that depends only on r (no angular dependence), the curl vanishes identically: $\nabla \times (f(r)\hat{r}) = 0$ for any function $f(r)$, because none of the angular derivatives that would produce a non-zero curl component are present. So $\mathbf{F}$ is irrotational for EVERY value of n - this condition alone places no restriction on n.

Step 2 - Solenoidal check: The divergence of a radial field in spherical coordinates is

$$\nabla \cdot \mathbf{F} = \frac{1}{r^2}\frac{d}{dr}\left(r^2 \cdot ar^n\right) = \frac{a}{r^2}\frac{d}{dr}\left(r^{n+2}\right) = a(n+2)r^{n-1}$$

For $\mathbf{F}$ to be solenoidal, $\nabla \cdot \mathbf{F} = 0$ for all $r \neq 0$, which (since $a \neq 0$) requires

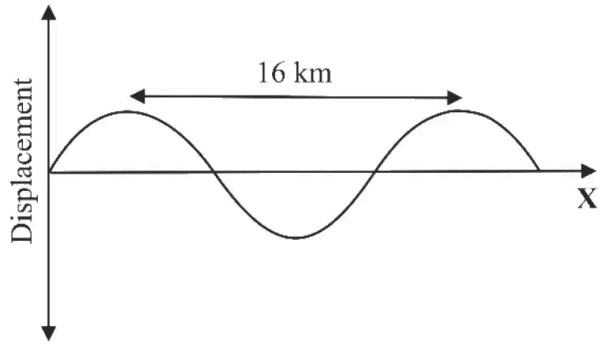
$$n + 2 = 0 \implies n = -2$$

Step 3: Since the irrotational condition holds for every n, the binding constraint is the solenoidal one, giving $n = -2$. This is exactly the exponent of the familiar inverse-square-law fields (gravitational field of a point mass, Coulomb field of a point charge), which are known to be divergence-free away from their source.

$\boxed{n = -2}$, option (B).

Quick Tip: The curl of any purely radial field is automatically zero for any n; the real constraint comes from setting the divergence to zero.

58. The maximum strain for the plane wave at $t = 0$, having a wavelength of 16 km and unit amplitude, travelling along the X-direction, as shown in the figure, is _____ (rounded off to three decimal places).



Correct Answer: 0.385 to 0.405

Solution:

A plane wave travelling along the X-direction with unit amplitude and wavelength λ can be written, at $t = 0$, as a displacement field

$$u(x, 0) = A \sin\left(\frac{2\pi x}{\lambda}\right), \qquad A = 1$$

Step 1: In a 1-D wave, strain is the spatial gradient of displacement:

$$\varepsilon(x) = \frac{\partial u}{\partial x} = A\frac{2\pi}{\lambda}\cos\left(\frac{2\pi x}{\lambda}\right)$$

Step 2: The strain is maximum where the cosine term equals 1 (at the zero-crossings of the displacement, where the wave's local slope is steepest), giving

$$\varepsilon_{max} = \frac{2\pi A}{\lambda}$$

Step 3: Substituting $A = 1$ and $\lambda = 16$ km,

$$\varepsilon_{max} = \frac{2\pi(1)}{16} = \frac{\pi}{8} \approx 0.3927$$

Rounded to three decimal places, $\varepsilon_{max} \approx 0.393$, which lies within the accepted range (0.385 to 0.405).

$\boxed{\varepsilon_{max} \approx 0.393}$

Quick Tip: Strain is the spatial derivative of displacement; for a sinusoid the peak strain equals amplitude times wavenumber (2 pi / lambda).

59. A causal recursive filter is given by $y_n = x_n - 2x_{n-1} + 3y_{n-1}$. If the input values, x_0, x_1 and x_2 are 1.2, -0.8 and 2.3, respectively, the output, y_2, of the filter is _____ (rounded off to one decimal place).

Correct Answer: 5.1

Solution:

The filter is causal and recursive, so the outputs are computed sequentially starting from $n = 0$, taking all values before the start of the record (x_{-1} and y_{-1}) to be zero (the filter starts at rest).

Step 1 (n = 0):

$$y_0 = x_0 - 2x_{-1} + 3y_{-1} = 1.2 - 2(0) + 3(0) = 1.2$$

Step 2 (n = 1):

$$y_1 = x_1 - 2x_0 + 3y_0 = -0.8 - 2(1.2) + 3(1.2) = -0.8 - 2.4 + 3.6 = 0.4$$

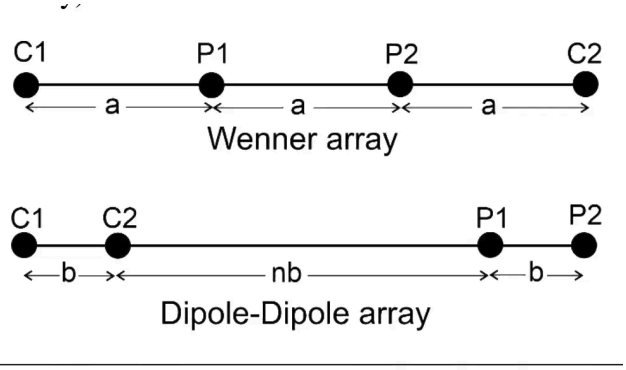
Step 3 (n = 2):

$$y_2 = x_2 - 2x_1 + 3y_1 = 2.3 - 2(-0.8) + 3(0.4) = 2.3 + 1.6 + 1.2 = 5.1$$

$\boxed{y_2 = 5.1}$

Quick Tip: Evolve the recursion one sample at a time from rest (x(-1) = y(-1) = 0); you need y0 and y1 before you can get y2.

60. If C1 & C2 and P1 & P2 are the pairs of current and potential electrodes in the Wenner (W) and Dipole-Dipole (DD) array configurations as shown in the figure below, then the fraction of the geometric factor for DD array that will be equal to half of that of W array is _____ (rounded off to three decimal places). (Use n = 1 in DD array)



Correct Answer: 0.166 to 0.167

Solution:

Step 1 - Geometric factor of the Wenner array: With equal electrode spacing a between C1-P1, P1-P2 and P2-C2, the Wenner geometric factor is

$$K_W = 2\pi a$$

Step 2 - Geometric factor of the Dipole-Dipole array: With current-dipole spacing a, potential-dipole spacing a, and dipole separation factor n (the gap between the two dipoles is na), the standard Dipole-Dipole geometric factor is

$$K_{DD} = \pi n(n+1)(n+2)a$$

For $n = 1$:

$$K_{DD} = \pi(1)(2)(3)a = 6\pi a$$

Step 3 - Set up the required fraction: We need the fraction x of K_{DD} that equals half of K_W:

$$x \cdot K_{DD} = \frac{1}{2} K_W$$

$$x = \frac{K_W/2}{K_{DD}} = \frac{2\pi a/2}{6\pi a} = \frac{\pi a}{6\pi a} = \frac{1}{6}$$

Step 4: Evaluating,

$$x = \frac{1}{6} = 0.1\overline{6} \approx 0.167$$

This lies within the accepted range (0.166 to 0.167).

$\boxed{x \approx 0.167}$

Quick Tip: Write K_W = 2*pi*a and K_DD = pi*n(n+1)(n+2)*a; with n = 1, K_DD = 6*pi*a, so the required fraction is (K_W/2)/K_DD = 1/6.

61. A scintillometer records 300 counts per second (cps) in a radiometric survey. If the background radiation and dead-time of the instrument are 100 cps and 250 μs, respectively, then the true net count rate is _____ cps (rounded off to two decimals).

Correct Answer: 220 to 223

Solution:

A scintillometer always under-counts at high count rates because the detector needs a finite dead-time τ to recover after each pulse before it can register the next one. For a non-paralysable (Type I) counter, the observed (recorded) count rate N_{obs} and the true count rate N_{true} are related by:

$$N_{true} = \frac{N_{obs}}{1 - N_{obs}\tau}$$

Step 1: Correct the gross (signal + background) rate.

Here $N_{obs} = 300$ cps and $\tau = 250\ \mu s = 2.5 \times 10^{-4}\ s$.

$N_{obs}\tau = 300 \times 2.5 \times 10^{-4} = 0.075$

$$N_{true,\,gross} = \frac{300}{1 - 0.075} = \frac{300}{0.925} = 324.32 \text{ cps}$$

Step 2: Correct the background rate the same way.

The quoted background of 100 cps is itself a value recorded on the same dead-time-limited instrument, so it must also be corrected before being subtracted:

$100 \times 2.5 \times 10^{-4} = 0.025$

$$N_{true,\,bg} = \frac{100}{1 - 0.025} = \frac{100}{0.975} = 102.56 \text{ cps}$$

Step 3: Net (source-only) true count rate.

$N_{net} = N_{true,\,gross} - N_{true,\,bg} = 324.32 - 102.56 = 221.76$ cps

$$\boxed{N_{net} \approx 221.76 \text{ cps}}$$

This lies inside the accepted 220–223 cps band.

Quick Tip: Apply the non-paralysable dead-time correction $N_{true} = N_{obs}/(1 - N_{obs}\tau)$ to BOTH the gross count rate and the background count rate before subtracting them.

62. Gamma-ray (GR) log values for clean sandstone and shale are 15 API and 115 API, respectively. If GR records 35 API for shaly-sandstone, then the GR index is _____ (rounded off to one decimal place).

Correct Answer: 0.2

Solution:

The Gamma Ray Index (also called the shale index, I_{GR}) is a normalised measure of how 'shaly' a formation reading is between the two log end-members — the clean (shale-free) sand baseline GR_{min} and the pure shale baseline GR_{max}:

$$I_{GR} = \frac{GR_{log} - GR_{min}}{GR_{max} - GR_{min}}$$

Given: $GR_{log} = 35$ API (the shaly-sandstone reading), $GR_{min} = 15$ API (clean sand), $GR_{max} = 115$ API (shale).

$$I_{GR} = \frac{35 - 15}{115 - 15} = \frac{20}{100} = 0.2$$

$$\boxed{I_{GR} = 0.2}$$

An index of 0.2 means the interval is only lightly shaly — about 20% of the way from clean sand toward pure shale on the linear GR scale — consistent with the log describing it as 'shaly sandstone' rather than shale.

Quick Tip: GR index = (measured GR – clean-sand GR) / (shale GR – clean-sand GR); it is a simple linear normalisation between the two end-member log readings.

63. If the inclination of the remanent magnetic field of a 140 million years old crustal block, now located at the equator, is 50°, then its drift-rate is _____ cm/yr (rounded off to two decimal places). [Use 1° = 111 km]

Correct Answer: 2.3 to 2.5

Solution:

For a geocentric axial dipole field, the inclination I frozen into a rock's remanent magnetisation is related to the (paleo)magnetic latitude λ at which the rock acquired that magnetisation by the dipole formula:

$$\tan I = 2\tan\lambda$$

Step 1: Solve for the paleolatitude.

$$\tan 50^\circ = 2\tan\lambda \Rightarrow \tan\lambda = \frac{\tan 50^\circ}{2} = \frac{1.19175}{2} = 0.59588$$

$$\lambda = \tan^{-1}(0.59588) \approx 30.8^\circ$$

Step 2: Interpret the geometry.

The block currently sits AT the equator (present-day latitude 0°), but its remanent inclination shows it actually acquired its magnetisation while sitting at a paleolatitude of about 30.8°. So, over the last 140 million years, the block has physically drifted from ~30.8° latitude down to 0° (the equator) — a net displacement of 30.8° of latitude.

Step 3: Convert the angular drift into a real distance.

Using $1^\circ = 111$ km:

$$\text{Distance} = 30.8^\circ \times 111\ \text{km}/^\circ \approx 3418\ \text{km} = 3.418 \times 10^8\ \text{cm}$$

Step 4: Divide by the elapsed time.

$$\text{Time} = 140\ \text{million years} = 1.4 \times 10^8\ \text{yr}$$

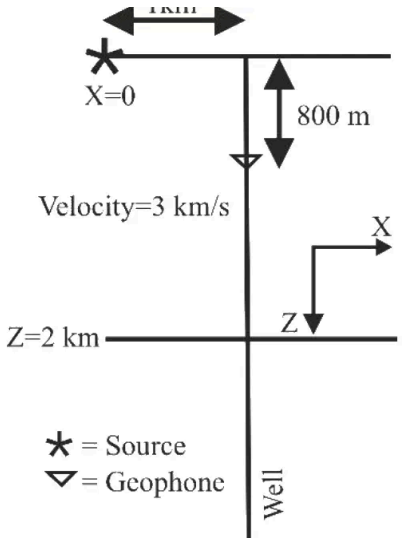
$$\text{Drift rate} = \frac{3.418 \times 10^8\ \text{cm}}{1.4 \times 10^8\ \text{yr}} \approx 2.44\ \text{cm/yr}$$

$$\boxed{\approx 2.44\ \text{cm/yr}}$$

This falls inside the accepted 2.3–2.5 cm/yr band — an ordinary plate-drift speed of a few cm/yr, comparable to real present-day plate velocities.

Quick Tip: Use the axial-dipole relation tan I = 2 tan(paleolatitude) to find how far (in degrees of latitude) the block has drifted from its magnetisation latitude to its present equatorial position, then convert to cm/yr.

64. Consider a seismic source located at the Earth's surface. A vertical well is located at a distance of 1 km away from the seismic source, and a geophone is suspended inside the well at a depth of 800 m, as shown in the figure below.



The travel time of the primary reflected wave from a horizontal reflector located at a depth of 2 km is _____ s (rounded off to three decimal places).

Correct Answer: 1.116 to 1.120

Solution:

The reflector is horizontal and the medium above it has a single constant velocity, $V = 3$ km/s (read off the figure). For this geometry the fastest way to get the travel time of the primary (single-bounce) reflection to an offset receiver is the classical image-source (mirror-source) method used throughout reflection seismology.

Step 1: Set up coordinates.

Let the source S sit at the surface at $x = 0,\ z = 0$. The reflector is horizontal at depth $z = 2$ km. The well is 1 km away (horizontally) from the source, and the geophone inside it sits at depth 800 m = 0.8 km. So the receiver R is at $(x, z) = (1,\ 0.8)$ km.

Step 2: Construct the image source.

For a flat horizontal reflector, the reflected ray path from S to any receiver has exactly the same length as a straight (direct) ray from an 'image' source S^*, obtained by mirroring S across the reflector plane: $z_{S^*} = 2z_{reflector} - z_S = 2(2) - 0 = 4$ km, at the same horizontal position, so $S^* = (0, 4)$ km.

Step 3: Compute the straight-line distance from the image source to the receiver.

$$L = \sqrt{(x_R - x_{S^*})^2 + (z_R - z_{S^*})^2} = \sqrt{(1-0)^2 + (0.8-4)^2} = \sqrt{1 + 10.24} = \sqrt{11.24} \approx 3.3526 \text{ km}$$

Step 4: Divide by velocity to get the travel time.

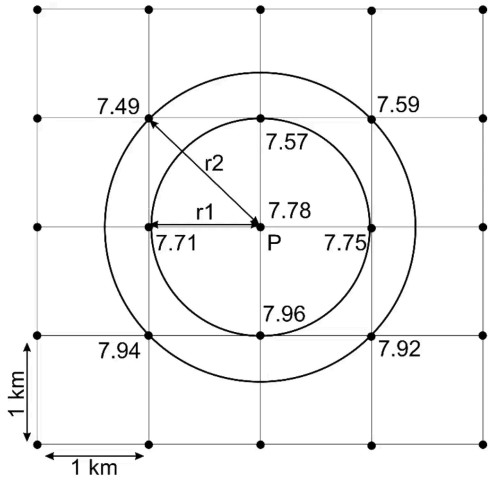
$$t = \frac{L}{V} = \frac{3.3526}{3} \approx 1.1175 \text{ s}$$

$$\boxed{t \approx 1.118 \text{ s}}$$

This lies inside the official 1.116–1.120 s window.

Quick Tip: Use the image-source (mirror) method for a horizontal reflector: reflect the source (or receiver) across the reflector depth and take a straight-line path to the actual receiver (or source), then divide by velocity.

65. The figure below shows Bouguer gravity anomaly values (in mGal) at a grid interval of 1 km, centred on point P (r1 and r2 are radial distances from P).



r1 and r2 are radial distances from P

The second vertical derivative at P is _____ mGal/km^2 (rounded off to two decimals).

Correct Answer: 0.16 to 0.18

Solution:

On a horizontal survey grid, the gravity field satisfies Laplace's equation in the source-free region:

$$\frac{\partial^2 g}{\partial x^2} + \frac{\partial^2 g}{\partial y^2} + \frac{\partial^2 g}{\partial z^2} = 0 \Rightarrow g_{zz} = -(g_{xx} + g_{yy})$$

So the second VERTICAL derivative can be estimated purely from the horizontal (map-view) gravity values around P, using finite differences over the surrounding grid.

From the figure (grid spacing $s = 1$ km), $P = 7.78$ mGal, and its two surrounding rings:
Ring 1 — direct N, S, E, W neighbours, radius $r_1 = s = 1$ km:
7.57, 7.96, 7.71, 7.75 → sum = 30.99

Ring 2 — diagonal NE, NW, SE, SW neighbours, radius $r_2 = s\sqrt{2} \approx 1.414$ km: 7.59, 7.49, 7.92, 7.94 → sum = 30.94

Step 1: Horizontal-Laplacian estimate from Ring 1 alone (ordinary 5-point second difference, spacing s):

$$L_1 = \frac{\Sigma(\text{Ring1}) - 4P}{s^2} = \frac{30.99 - 4(7.78)}{1} = \frac{30.99 - 31.12}{1} = -0.13$$

Step 2: Horizontal-Laplacian estimate from Ring 2 alone (same operator, but the diagonal points sit on a rotated set of axes at spacing $s\sqrt{2}$, so the effective spacing-squared is $2s^2$):

$$L_2 = \frac{\Sigma(\text{Ring2}) - 4P}{2s^2} = \frac{30.94 - 31.12}{2} = \frac{-0.18}{2} = -0.09$$

Step 3: L_1 and L_2 are two independent finite-difference estimates of the SAME Laplacian at different effective grid spacings (1 and 2, in units of s^2). Combining them by Richardson extrapolation cancels the leading truncation error — exactly how the widely used combined-ring second-vertical-derivative operator (Rosenbach/Elkins-type SVD template) is built:

$$L_{true} = 2L_1 - L_2 = 2(-0.13) - (-0.09) = -0.26 + 0.09 = -0.17$$

Step 4: Apply Laplace's equation:

$$g_{zz} = -L_{true} = -(-0.17) = 0.17 \text{ mGal/km}^2$$

$$\boxed{g_{zz} \approx 0.17 \text{ mGal/km}^2}$$

This lies inside the accepted 0.16–0.18 mGal/km^2 band. The positive value indicates P sits over (or very near) the edge of a shallow, dense body — the classic SVD 'sharpening' response used to pinpoint compact source edges.

Quick Tip: Use Laplace's equation (g_zz = -(g_xx+g_yy)) and estimate the horizontal Laplacian from both the near ring (r1) and far diagonal ring (r2) of grid values around P, then combine the two estimates to cancel truncation error.