

Math Review Notes

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1 Introduction

These are notes I've collected on various math topics. I originally created this document to prepare for the GRE Math Subject test. Since then I've expanded it as I've reviewed concepts from past classes and reinforced concepts from new classes. This document is very much a work in progress, with many typos, omissions to be filled in, and probably errors. Nonetheless, I share this document in case it's useful to anyone else as a reference.

I use many sources throughout this document, which I either cite at the beginning of the section (for sources I use broadly) or as I use them (for sources I use for one or two isolated results).

2 Linear Algebra

These are my notes from taking EE 588 at USC, Math 541A at USC, and various other sources which I mostly cite within the text.

2.1 Properties of Projection Matrices

- i. Formula:

$$P = A(A^T A)^{-1} A^T$$

(Note that if A is an invertible (square) matrix, then $P = A(A^T A)^{-1} A^T = AA^{-1}(A^T)^{-1} A^T = I$.)

The projection matrix projects any vector b into the column space of A . In other words, $p = Pb$ is the component of b in the column space, and the error $e = b - Pb$ is the component in the orthogonal complement. ($I - P$ is also a projection matrix. It projects b onto the orthogonal complement, and the projection is $b - Pb = e$.)

(Note that if A is an invertible (square) matrix, then its column space is all of \mathbb{R}^n , so b is already in the column space of A .)

- ii. The projection matrix is **idempotent**: it equals its square— $P^2 = P$.
- iii. The projection matrix is **symmetric**: it equals its transpose— $P^T = P$.
- iv. Conversely, **any symmetric idempotent matrix represents a projection**. P is unique for a given subspace.
- v. If A is an $m \times n$ matrix with rank n , then $\text{rank}(P) = n$. The eigenvalues of P consist of n ones and $m - n$ zeroes. P always contains n independent eigenvectors and is thus diagonalizable.

Suppose A is a square nonsingular matrix and λ is an eigenvalue of A . Then λ^{-1} is an eigenvalue of the matrix A^{-1} .

The trace of an idempotent matrix with rank r is r .

2.2 Eigenvalues, Eigenvectors, Diagonalization, Symmetric Matrices

Notes on Diagonalization

Suppose the $n \times n$ matrix A has n linearly independent eigenvectors. If these eigenvectors are the columns of a matrix S , then $S^{-1}AS$ is a diagonal matrix Λ . The eigenvalues of A are on the diagonal of Λ :

$$S^{-1}AS = \Lambda = \begin{bmatrix} \lambda_1 & 0 & \cdots & 0 \\ 0 & \lambda_2 & \cdots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & \lambda_n \end{bmatrix}$$

We call S the **eigenvector matrix** and Λ the **eigenvalue matrix**.

1. If the matrix A has no repeated eigenvalues, then its n eigenvectors are automatically independent. Therefore **any matrix with n distinct eigenvalues can be diagonalized.**
2. **The diagonalizing matrix S is not unique.** An eigenvector x can be multiplied by a constant and remains an eigenvector. We can multiply the columns of S by any nonzero constants and produce a new diagonalizing S . Repeated eigenvalues leave even more freedom in S (columns with identical eigenvalues can be interchanged).

(Note that for the trivial example $A = I$, any invertible S will do. $S^{-1}IS$ is always diagonal, and Λ is just I . **All vectors are eigenvectors of the identity.**)

3. **Other matrices S will not produce a diagonal Λ .** Since $\Lambda = S^{-1}AS$, S must satisfy $S\Lambda = AS$. Suppose the first column of S is y . Then the first column of $S\Lambda$ is $\lambda_1 y$. If this is to agree with the first column of AS , which by matrix multiplication is Ay , then y must be an eigenvector: $Ay = \lambda_1 y$.

(Note that the *order* of the eigenvectors in S and the eigenvalues in Λ must match.)

4. Not all matrices possess n linearly independent eigenvectors, so **not all matrices are diagonalizable.** **Diagonalizability of A depends on having enough (n) independent eigenvectors. Invertibility of A depends on having nonzero eigenvalues.**

There is no connection between diagonalizability (n independent eigenvectors) and invertibility (no zero eigenvalues). The only indication given by the eigenvalues is that diagonalization can fail only if there are repeated eigenvalues. (But even then, it does not always fail—e.g. I .)

The test is to check, for an eigenvalue that is repeated p times, whether there are p independent eigenvectors—in other words, whether $A - \lambda$ has rank $n - p$.

5. **Projection matrices always contain n independent eigenvectors and thus are always diagonalizable.**

Eigenvalues of Symmetric Matrices: If A is symmetric, then it has the following properties:

1. A has exactly n (not necessarily distinct) eigenvalues
2. There exists a set of n eigenvectors, one for each eigenvalue, that are mutually orthogonal (even if the eigenvalues are not distinct).

Eigenvalues of the Inverse of a Matrix: Suppose A is a square nonsingular matrix and λ is an eigenvalue of A . Then λ^{-1} is an eigenvalue of the matrix A^{-1} . Proof: Note that since A is nonsingular, A^{-1} exists and λ is nonnegative for all eigenvalues of A . Let λ be an eigenvalue of A and let $x \neq 0$ be an eigenvector of A for λ . Suppose A is n by n . Then we have

$$A^{-1}x = A^{-1}\lambda^{-1}\lambda x = \lambda^{-1}A^{-1}\lambda x = \lambda^{-1}A^{-1}Ax = \lambda^{-1}x$$

The inverse of a symmetric matrix is symmetric. Proof: Let A be a symmetric matrix.

$$I = I'$$

$$AA^{-1} = (AA^{-1})'$$

$$A^{-1}A = (A^{-1})'A'$$

$$A^{-1}AA^{-1} = (A^{-1})'AA^{-1}$$

$$A^{-1} = (A^{-1})'$$

2.3 Positive Definite Matrices

For any real invertible matrix A , the product $A'A$ is a positive definite matrix. (Proof: Let z be a non-zero vector. We want $z'A'Az > 0 \forall z$. Note that $z'A'Az = (Az)'(Az)$. Because A is invertible and $z \neq 0$, $Az \neq 0$, so $(Az)'(Az) > 0$.)

Let $A \in \mathbb{R}^{m \times n}$ with $m \geq n$ and let $\text{rank}(A) = n$ (that is, A has full column rank). Then $A'A$ is a positive definite matrix. (Proof: Let z be a non-zero vector. We want $z'A'Az > 0 \forall z$. Note that $z'A'Az = (Az)'(Az)$. Because A has full column rank (and n linearly independent columns) and $z \neq 0$, $Az \neq 0$, so $(Az)'(Az) > 0$.)

Every positive definite matrix is invertible and its inverse is also positive definite.

2.4 Matrix Decompositions

Schur complement, Schur decomposition: For information, see Section 11.2.

QR decompositon

Orthogonal Decomposition

Spectral Decomposition (eigenvalue decomposition)

Generalized eigenvalue decomposition

Singular value decomposition and Pseudo-inverse

Jordan decomposition

Cholesky decomposition

2.5 Other

Frobenius norm

From appendix of Time Series:

Quadratic forms

Special matrices

Difference Equations

2.6 Practice Problems

[The Power Method] This exercise gives an algorithm for finding the eigenvectors and eigenvalues of a symmetric matrix. In modern statistics, this is often a useful thing to do. The Power Method described below is not the best algorithm for this task, but it is perhaps the easiest to describe and analyze.

Let A be an $n \times n$ real symmetric matrix. Let $\lambda_1 \geq \dots \geq \lambda_n$ be the (unknown) eigenvalues of A , and let $v_1, \dots, v_n \in \mathbb{R}^n$ be the corresponding (unknown) eigenvectors of A such that $|v_i| = 1$ and such that $Av_i = \lambda_i v_i$ for all $1 \leq i \leq n$.

Given A , our first goal is to find v_1 and λ_1 . For simplicity, assume that $1/2 < \lambda_1 < 1$, and $0 \leq \lambda_n \leq \dots \leq \lambda_2 < 1/4$. Suppose we have found a vector $v \in \mathbb{R}^n$ such that $|v| = 1$ and $|\langle v, v_1 \rangle| > 1/n$. Let k be a positive integer. Show that

$$A^k v$$

approximates v_1 well as k becomes large. More specifically, show that for all $k \geq 1$,

$$\|A^k v - \langle v, v_1 \rangle \lambda_1^k v_1\|^2 \leq \frac{n-1}{16^k}.$$

(Hint: use the spectral theorem for symmetric matrices.)

Solution. Since the eigenvectors for A are orthogonal, they form a basis for \mathbb{R}^n , so for any $v \in \mathbb{R}^n$ we have $v = \sum_{i=1}^n c_i v_i$ for some $c = (c_1, \dots, c_n) \in \mathbb{R}^n$. It also follows then that $\langle v, v_1 \rangle = \langle \sum_{i=1}^n c_i v_i, v_1 \rangle = c_1 v'_1 v_1 = c_1$. And finally, since $\|v\| = 1$ and $\|v_i\| = 1$ for all i , clearly we have $-1 \leq c_i \leq 1$. Using these facts, we have

$$\begin{aligned} \|A^k v - \langle v, v_1 \rangle \lambda_1^k v_1\|^2 &= \left\| \sum_{i=1}^n \lambda_i^k c_i v_i - \langle v, v_1 \rangle \lambda_1^k v_1 \right\|^2 = \left\| \sum_{i=1}^n \lambda_i^k c_i v_i - \lambda_1^k c_1 v_1 \right\|^2 = \left\| \sum_{i=2}^n \lambda_i^k c_i v_i \right\|^2 \\ &= \sum_{i=2}^n \lambda_i^{2k} c_i^2 v'_i v_i = \sum_{i=2}^n \lambda_i^{2k} c_i^2 \end{aligned}$$

Since by assumption $0 \leq \lambda_n \leq \dots \leq \lambda_2 \leq 1/4$, $\lambda_i^{2k} \leq 1/16^k$ for all i , so we have

$$\|A^k v - \langle v, v_1 \rangle \lambda_1^k v_1\|^2 \leq \frac{1}{16^k} \sum_{i=2}^n c_i^2$$

Since $-1 \leq c_i \leq 1 \implies 0 \leq c_i^2 \leq 1$, we have $\sum_{i=2}^n c_i^2 \leq n-1$, so this can be written as

$$\|A^k v - \langle v, v_1 \rangle \lambda_1^k v_1\|^2 \leq \frac{n-1}{16^k}$$

Remark. Since $|\langle v, v_1 \rangle| \lambda_1^k > 2^{-k}/n$, this inequality implies that $A^k v$ is approximately an eigenvector of A with eigenvalue λ_1 . That is, by the triangle inequality,

$$|A(A^k v) - \lambda_1(A^k v)| \leq |A^{k+1}v - \langle v, v_1 \rangle \lambda_1^{k+1} v_1| + \lambda_1 |\langle v, v_1 \rangle \lambda_1^k v_1 - A^k v| \leq 2 \frac{\sqrt{n-1}}{4^k}.$$

Moreover, by the reverse triangle inequality,

$$|A^k v| = |A^k v - \langle v, v_1 \rangle \lambda_1^k v_1 + \langle v, v_1 \rangle \lambda_1^k v_1| \geq \frac{1}{n} 2^{-k} - \frac{\sqrt{n-1}}{4^k}.$$

If we take k to be large (say $k > 10 \log n$), and if we define $z : equals A^k v$, then z is approximately an eigenvector of A , that is

$$|A \frac{A^k v ||A^k v| - \lambda_1 \frac{A^k v}{|A^k v|}}{|A^k v|} 4n^{3/2} 2^{-k}| \leq 4n^{-4}.$$

And to approximately find the first eigenvalue λ_1 , we simply compute

$$\frac{z^T A z}{z^T z}.$$

That is, we have approximately found the first eigenvector and eigenvalue of A .

To find the second eigenvector and eigenvalue, we can repeat the above procedure, where we start by choosing v such that $\langle v, v_1 \rangle = 0$, $|v| = 1$ and $|\langle v, v_2 \rangle| > 1/(10\sqrt{n})$. To find the third eigenvector and eigenvalue, we can repeat the above procedure, where we start by choosing v such that $\langle v, v_1 \rangle = \langle v, v_2 \rangle = 0$, $|v| = 1$ and $|\langle v, v_3 \rangle| > 1/(10\sqrt{n})$. And so on.

Google's PageRank algorithm uses the power method to rank websites very rapidly. In particular, they let n be the number of websites on the internet (so that n is roughly 10^9). They then define an $n \times n$ matrix C where $C_{ij} = 1$ if there is a hyperlink between websites i and j , and $C_{ij} = 0$ otherwise. Then, they let B be an $n \times n$ matrix such that B_{ij} is 1 divided by the number of 1's in the i^{th} row of C , if $C_{ij} = 1$, and $B_{ij} = 0$ otherwise. Finally, they define

$$A = (.85)B + (.15)D/n$$

where D is an $n \times n$ matrix all of whose entries are 1.

The power method finds the eigenvector v_1 of A , and the size of the i^{th} entry of v_1 is proportional to the "rank" of website i .

12. Let A be a 2×2 matrix for which there is a constant k such that the sum of the entries in each row and each column is k . Which of the following must be an eigenvector of A ?

I. $\begin{pmatrix} 1 \\ 0 \end{pmatrix}$

II. $\begin{pmatrix} 0 \\ 1 \end{pmatrix}$

III. $\begin{pmatrix} 1 \\ 1 \end{pmatrix}$

- (A) I only (B) II only (C) III only (D) I and II only (E) I, II, and III

Solution 12. (C) This condition makes the matrix of the form

$$\begin{pmatrix} a & b \\ b & a \end{pmatrix}.$$

There is no reason that $a = 0$ or $b = 0$, so there is no reason $(1, 0)$ or $(0, 1)$ should be eigenvectors. But it is easy to verify that $(1, 1)$ must be.

24. Consider the system of linear equations

$$\begin{aligned} w + 3x + 2y + 2z &= 0 \\ w + 4x + y &= 0 \\ 3w + 5x + 10y + 14z &= 0 \\ 2w + 5x + 5y + 6z &= 0 \end{aligned}$$

with solutions of the form (w, x, y, z) , where w, x, y , and z are real. Which of the following statements is FALSE?

- (A) The system is consistent.
- (B) The system has infinitely many solutions.
- (C) The sum of any two solutions is a solution.
- (D) $(-5, 1, 1, 0)$ is a solution.
- (E) Every solution is a scalar multiple of $(-5, 1, 1, 0)$.

Solution 24. (E) Looking at our answers, we can verify directly that $(-5, 1, 1, 0)$ is a solution. Any multiple of $(-5, 1, 1, 0)$ is also a solution, which shows that (A), (B), (C), and (D) are all true – leaving only (E). Another solution, for example, is $(0, 2, -8, 5)$

$$A = \begin{pmatrix} 1 & 2 & 3 & 4 & 5 \\ 0 & 2 & 3 & 4 & 5 \\ 0 & 0 & 3 & 4 & 5 \\ 0 & 0 & 0 & 4 & 5 \\ 0 & 0 & 0 & 0 & 5 \end{pmatrix}$$

34. Which of the following statements about the real matrix shown above is FALSE?

- (A) A is invertible.
- (B) If $\mathbf{x} \in \mathbb{R}^5$ and $A\mathbf{x} = \mathbf{x}$, then $\mathbf{x} = \mathbf{0}$.
- (C) The last row of A^2 is $(0 \ 0 \ 0 \ 0 \ 25)$.
- (D) A can be transformed into the 5×5 identity matrix by a sequence of elementary row operations.
- (E) $\det(A) = 120$

Solution 34. (B) An upper triangular matrix is easily verified to be invertible so long as its diagonal entries are all nonzero. Specifically, $\det A$ is still the product of its diagonal entries, so (E) and (D) and (A) are all true. (C) can easily be verified to be true by computing that the bottom-right corner is 25 (the product of upper triangular matrices still being upper triangular). This leaves (B). (B) can be checked directly to be false: if we let $x = (1, 0, 0, 0, 0)$, then $Ax = x$.

37. Let V be a finite-dimensional real vector space and let P be a linear transformation of V such that $P^2 = P$. Which of the following must be true?

- I. P is invertible.
 - II. P is diagonalizable.
 - III. P is either the identity transformation or the zero transformation.
- (A) None (B) I only (C) II only (D) III only (E) II and III

Solution 37. (C) $P^2 = P$ means that P is projection onto some subspace. There is no reason to believe that this should be invertible, but it should definitely be diagonalisable (with eigenbasis some basis of that subspace). III also need not be true if the subspace is anything proper or nontrivial.

50. Let A be a real 2×2 matrix. Which of the following statements must be true?

- I. All of the entries of A^2 are nonnegative.
 - II. The determinant of A^2 is nonnegative.
 - III. If A has two distinct eigenvalues, then A^2 has two distinct eigenvalues.
- (A) I only (B) II only (C) III only (D) II and III only (E) I, II, and III

Solution 50. (B) There is no reason that all the entries of A^2 need to be nonnegative. Its determinant must be nonnegative though: $\det(A^2) = (\det A)^2$. For III, suppose A is the diagonal matrix with entries $\pm\lambda$. Then those are its eigenvalues, and they are distinct so long as $\lambda \neq 0$. But A^2 has only one eigenvalue: λ^2 .

51. Which of the following is an orthonormal basis for the column space of the real matrix $\begin{pmatrix} 1 & -1 & 2 & -3 \\ -1 & 1 & -3 & 2 \\ 2 & -2 & 5 & -5 \end{pmatrix}$?

- (A) $\left\{ \begin{pmatrix} 1 \\ 0 \\ 0 \end{pmatrix}, \begin{pmatrix} 0 \\ 1 \\ 0 \end{pmatrix} \right\}$
- (B) $\left\{ \begin{pmatrix} 1 \\ 0 \\ 0 \end{pmatrix}, \begin{pmatrix} 0 \\ 1 \\ 0 \end{pmatrix}, \begin{pmatrix} 0 \\ 0 \\ 1 \end{pmatrix} \right\}$
- (C) $\left\{ \begin{pmatrix} 1 \\ 0 \\ 0 \end{pmatrix}, \begin{pmatrix} \frac{2}{\sqrt{5}} \\ \frac{1}{\sqrt{5}} \\ 0 \end{pmatrix} \right\}$
- (D) $\left\{ \begin{pmatrix} 1 \\ -1 \\ 2 \end{pmatrix}, \begin{pmatrix} 2 \\ -3 \\ 5 \end{pmatrix} \right\}$
- (E) $\left\{ \begin{pmatrix} \frac{1}{\sqrt{6}} \\ -\frac{1}{\sqrt{6}} \\ \frac{2}{\sqrt{6}} \end{pmatrix}, \begin{pmatrix} \frac{1}{\sqrt{2}} \\ \frac{1}{\sqrt{2}} \\ 0 \end{pmatrix} \right\}$

Solution 51. (E) The basis (C) is not orthogonal and (D) is not normal, so we can rule those out. We can throw out the first column, since it is the negation of the second. A little bit of math shows that the remaining 3×3 matrix has determinant 0, so the rank of our column space is 2. That leaves only (A) and (E), but (A) cannot be correct. Our column space contains vectors that have nonzero third entry, so cannot lie in the span of that basis.

3 Calculus

These notes include some screenshots from Wikipedia as well as from *Calculus* by Gilbert Strang, available at <https://ocw.mit.edu/ans7870/resources/Strang/Edited/Calculus/Calculus.pdf>. I also used parts from some other resources which I mention when they arise.

3.1 List of common derivatives and integrals to know

$$\begin{aligned}\frac{d}{dx}(\sin^{-1} x) &= \frac{1}{\sqrt{1-x^2}} & \frac{d}{dx}(\ln(x)) &= \frac{1}{x}, \quad x > 0 \\ \frac{d}{dx}(\cos^{-1} x) &= -\frac{1}{\sqrt{1-x^2}} & \frac{d}{dx}(\ln|x|) &= \frac{1}{x}, \quad x \neq 0 \\ \frac{d}{dx}(\tan^{-1} x) &= \frac{1}{1+x^2} & \frac{d}{dx}(\log_a(x)) &= \frac{1}{x \ln a}, \quad x > 0\end{aligned}$$

$$\int \tan u \, du = \ln|\sec u| + c$$

$$\int \sec u \, du = \ln|\sec u + \tan u| + c$$

$$\int \frac{1}{a^2+u^2} \, du = \frac{1}{a} \tan^{-1}\left(\frac{u}{a}\right) + c$$

$$\int \frac{1}{\sqrt{a^2-u^2}} \, du = \sin^{-1}\left(\frac{u}{a}\right) + c$$

$$\int \ln u \, du = u \ln(u) - u + c$$

$$\int \sinh x \, dx = \cosh x + C$$

$$\int \cosh x \, dx = \sinh x + C$$

3.2 Matrix Differentiation

Recommended resource: “Matrix Differentiation (and some other stuff)” by Randal J. Barnes (Department of Civil Engineering, University of Minnesota). Available for download at <https://atmos.washington.edu/~dennis/MatrixCalculus.pdf>.

More information not contained in that pdf (from the appendix of *Convex Optimization* by Stephen Boyd and Lieven Vandenberghe, available for free download at <https://web.stanford.edu/~boyd/cvxbook/>):

Chain rule. Suppose $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ is differentiable at $x \in \text{int dom } f$ and $g : \mathbb{R}^m \rightarrow \mathbb{R}^p$ is differentiable at $f(x) \in \text{int dom } g$. Define the composition $h : \mathbb{R}^n \rightarrow \mathbb{R}^p$ by $h(z) = g(f(z))$. Then

$$Dh(x) = Dg(f(x))Df(x)$$

In particular, if $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and $g : \mathbb{R} \rightarrow \mathbb{R}$,

$$\nabla h(x) = g'(f(x))\nabla f(x)$$

Example with an affine function. Suppose $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ is differentiable, $A \in \mathbb{R}^{n \times p}$, and $b \in \mathbb{R}^n$. Define $g : \mathbb{R}^p \rightarrow \mathbb{R}^m$ as $g(x) = f(Ax + b)$ with $\text{dom } g = \{x \mid Ax + b \in \text{dom } f\}$. Then

$$\nabla g(x) = A^T \nabla f(Ax + b)$$

Example 2. Consider $f : \mathbb{R}^n \rightarrow \mathbb{R}$ where

$$f(x) = \log \sum_{i=1}^m \exp(a_i^T x + b_i) =$$

where $a_1, \dots, a_m \in \mathbb{R}^n$ and $b_1, \dots, b_m \in \mathbb{R}$. Note that $f(\cdot)$ can be expressed as a composition of $Ax + b$ (where $A \in \mathbb{R}^{m \times n}$ has rows a_1^T, \dots, a_m^T) and the function $g : \mathbb{R}^m \rightarrow \mathbb{R}$ given by $g(y) = \log(\sum_{i=1}^m \exp(y_i))$. We have

$$\nabla g(y) = \left[\sum_{i=1}^m e^{y_i} \right]^{-1} (\exp(y_1) \dots \exp(y_m))^T$$

so applying the chain rule yields

$$\nabla f(x) = \left[\sum_{i=1}^m \exp(a_i^T x + b_i) \right]^{-1} A^T z$$

where $z_i = \exp(a_i^T x + b_i)$, $i = 1, \dots, m$.

Hessians. The Hessian matrix of $f : \mathbb{R}^n \rightarrow \mathbb{R}$ is denoted by $\nabla^2 f(x)$ and is given by

$$\nabla^2 f(x)_{ij} = \frac{\partial^2 f(x)}{\partial x_i \partial x_j}, \quad i = 1, 2, \dots, n, \quad j = 1, 2, \dots, n$$

The quadratic function

$$f(x_0) + \nabla f(x_0)^T (x - x_0) + \frac{1}{2} (x - x_0)^T \nabla^2 f(x_0) (x - x_0)$$

is called the **second-order approximation of f near x_0** .

Chain rule for second derivative. A chain rule for the second derivative is difficult in general. Here are some special cases.

Composition with scalar function. Suppose $f : \mathbb{R}^n \rightarrow \mathbb{R}$, $g : \mathbb{R} \rightarrow \mathbb{R}$, and $h(x) = g(f(x))$. We have

$$\nabla^2 h(x) = g'(f(x)) \nabla^2 f(x) + g''(f(x)) \nabla f(x) \nabla f(x)^T$$

Composition with affine function. Suppose $f : \mathbb{R} \rightarrow \mathbb{R}$, $a \in \mathbb{R}^m$, $b \in \mathbb{R}$. Define $g : \mathbb{R}^m \rightarrow \mathbb{R}$ by $g(x) = f(a^T x + b)$. Then

$$\nabla^2 g(x) = a^T \nabla^2 f(a^T x + b) a$$

More generally, suppose $f : \mathbb{R}^n \rightarrow \mathbb{R}$, $A \in \mathbb{R}^{n \times m}$, and $b \in \mathbb{R}^n$. Define $g : \mathbb{R}^m \rightarrow \mathbb{R}$ by $g(x) = f(Ax + b)$. Then

$$\nabla^2 g(x) = A^T \nabla^2 f(Ax + b) A$$

3.3 Some theorems in higher dimensions

Taylor's Theorem (first order). (borrowed from <https://www.rose-hulman.edu/~bryan/lottamath/mTaylor.pdf>) Consider a function $f : \mathbb{R}^n \rightarrow \mathbb{R}$. Let $a \in \mathbb{R}^n$ be a fixed point. Then Taylor's Theorem states:

If $f(x)$ is differentiable on an open ball B around a and $x \in B$ then

$$f(x) = f(a) + \nabla f(b)^T (x - a)$$

for some b on the line segment joining a and x .

This can also be expressed as follows. Let $x, y \in \mathbb{R}^n$. If $f(x)$ is continuously differentiable, then

$$f(y) = f(x) + \nabla f(tx + (1-t)y)^T (y - x)$$

for some $t \in [0, 1]$.

Proof. Consider $g(z) = f(zy + (1-z)x)$. If f is differentiable then so is g . Then by the Mean Value Theorem, for some $t \in (0, 1)$ we have $g(1) - g(0) = g'(t)$. By the chain rule,

$$g'(t) = \nabla f(x + t(y - x))^T (y - x)$$

Using $g(1) = f(y)$ and $g(0) = f(x)$, we have

$$\iff \nabla f(tx + (1-t)y)^T (y - x) = g(1) - g(0) = f(y) - f(x)$$

□

Taylor's Theorem (second order). Consider a function $f : \mathbb{R}^n \rightarrow \mathbb{R}$. Let $a \in \mathbb{R}^n$ be a fixed point. Then Taylor's Theorem states:

If $f(x)$ is twice differentiable on an open ball B around a and $x \in B$ then

$$f(x) = f(a) + (x - a)^T \nabla f(a) + \frac{1}{2}(x - a)^T \nabla^2 f(b)(x - a)$$

for some b on the line segment joining a and x .

This can also be expressed as follows. Let $x, y \in \mathbb{R}^n$. If $f(x)$ is twice continuously differentiable, then

$$f(y) = f(x) + \nabla f(x)^T (y - x) + \frac{1}{2}(y - x)^T \nabla^2 f(ty + (1 - t)x)(y - x)$$

for some $t \in [0, 1]$.

Proof. Consider $g(z) = f(zy + (1 - z)x)$. If f is differentiable then so is g . Then by the second order case of Taylor's Theorem in one dimension, for some $t \in (0, 1)$ we have $g(1) = g(0) + g'(0) + (1/2)g''(t)$. By the chain rule,

$$g''(t) = \frac{\partial}{\partial t} \nabla f(x + t(y - x))^T (y - x) = (y - x)^T \nabla^2 f(x + t(y - x))^T (y - x)$$

Using this result along with $g(1) = f(y)$, $g(0) = f(x)$, and $g'(0) = \nabla f(x)^T (y - x)$, we have

$$f(y) = f(x) + \nabla f(x)^T (y - x) + \frac{1}{2}(y - x)^T \nabla^2 f(x + t(y - x))^T (y - x)$$

□

3.4 Optimizing functions of several variables

Functions of two variables [\[edit\]](#)

Suppose that $f(x, y)$ is a differentiable [real function](#) of two variables whose second [partial derivatives](#) exist. The [Hessian matrix](#) H of f is the 2×2 matrix of partial derivatives of f :

$$H(x, y) = \begin{pmatrix} f_{xx}(x, y) & f_{xy}(x, y) \\ f_{yx}(x, y) & f_{yy}(x, y) \end{pmatrix}.$$

Define $D(x, y)$ to be the [determinant](#)

$$D(x, y) = \det(H(x, y)) = f_{xx}(x, y)f_{yy}(x, y) - (f_{xy}(x, y))^2,$$

of H . Finally, suppose that (a, b) is a critical point of f (that is, $f_x(a, b) = f_y(a, b) = 0$). Then the second partial derivative test asserts the following:^[1]

1. If $D(a, b) > 0$ and $f_{xx}(a, b) > 0$ then (a, b) is a local minimum of f .
2. If $D(a, b) > 0$ and $f_{xx}(a, b) < 0$ then (a, b) is a local maximum of f .
3. If $D(a, b) < 0$ then (a, b) is a [saddle point](#) of f .
4. If $D(a, b) = 0$ then the second derivative test is inconclusive, and the point (a, b) could be any of a minimum, maximum or saddle point.

Functions of many variables [\[edit\]](#)

For a function f of two or more variables, there is a generalization of the rule above. In this context, instead of examining the determinant of the Hessian matrix, one must look at the eigenvalues of the Hessian matrix at the critical point. The following test can be applied at any critical point (a, b, \dots) for which the Hessian matrix is invertible:

1. If the Hessian is positive definite (equivalently, has all eigenvalues positive) at (a, b, \dots) , then f attains a local minimum at (a, b, \dots) .
2. If the Hessian is negative definite (equivalently, has all eigenvalues negative) at (a, b, \dots) , then f attains a local maximum at (a, b, \dots) .
3. If the Hessian has both positive and negative eigenvalues then (a, b, \dots) is a saddle point for f (and in fact this is true even if (a, b, \dots) is degenerate).

3.5 Lagrange Multipliers

: to flesh out! <http://tutorial.math.lamar.edu/Classes/CalcIII/LagrangeMultipliers.aspx>

3.6 Line Integrals

(p. 555 of Strang book)

Suppose a force in two-dimensional space is given by $\mathbf{F} = M\mathbf{i} + N\mathbf{j}$. Then the work done by this force on a particle moving along a curve C is given by

$$W = \int_C \mathbf{F} \cdot d\mathbf{R} = \int_C Mdx + Ndy$$

Along a curve in three-dimensional space the work done by a three-dimensional force $\mathbf{F} = M\mathbf{i} + N\mathbf{j} + P\mathbf{k}$ is given by

$$W = \int_C \mathbf{F} \cdot \mathbf{T} ds = \int_C \mathbf{F} \cdot d\mathbf{R} = \int_C Mdx + Ndy + Pdz$$

where the tangent vector \mathbf{T} is given by

$$\mathbf{T} = \frac{d\mathbf{R}}{ds}$$

Green's Theorem: Suppose the region R is bounded by the simple closed piecewise smooth curve C . Then an integral over R equals a line integral around C :

$$\oint_C \mathbf{F} \cdot d\mathbf{R} = \oint_C Mdx + Ndy = \iint_R \left(\frac{\partial N}{\partial x} - \frac{\partial M}{\partial y} \right) dx dy$$

Line integrals chapter! <http://tutorial.math.lamar.edu/Classes/CalcIII/LineIntegralsIntro.aspx>

Surface integrals chapter! <http://tutorial.math.lamar.edu/Classes/CalcIII/SurfaceIntegralsIntro.aspx>

3.7 Miscellaneous

13A The tangent plane at (x_0, y_0, z_0) has the same slopes as the surface $z = f(x, y)$. The equation of the tangent plane (a linear equation) is

$$z - z_0 = \left(\frac{\partial f}{\partial x} \right)_0 (x - x_0) + \left(\frac{\partial f}{\partial y} \right)_0 (y - y_0). \quad (1)$$

The normal vector \mathbf{N} to that plane has components $(\partial f / \partial x)_0, (\partial f / \partial y)_0, -1$.

13B The tangent plane to the surface $F(x, y, z) = c$ has the linear equation

$$\left(\frac{\partial F}{\partial x} \right)_0 (x - x_0) + \left(\frac{\partial F}{\partial y} \right)_0 (y - y_0) + \left(\frac{\partial F}{\partial z} \right)_0 (z - z_0) = 0. \quad (7)$$

The normal vector is $\mathbf{N} = \left(\frac{\partial F}{\partial x} \right)_0 \mathbf{i} + \left(\frac{\partial F}{\partial y} \right)_0 \mathbf{j} + \left(\frac{\partial F}{\partial z} \right)_0 \mathbf{k}$.

$$dz = (\partial z / \partial x)_0 dx + (\partial z / \partial y)_0 dy \quad \text{or} \quad df = f_x dx + f_y dy. \quad (10)$$

This is the **total differential**. All letters dz and df and dw can be used, but ∂z and ∂f are not used. Differentials suggest small movements in x and y ; then dz is the resulting movement in z . On the tangent plane, equation (10) holds exactly.

The **directional derivative**, denoted $D_v f(x, y)$, is a derivative of a multivariable function in the direction of a vector \mathbf{v} . It is the scalar projection of the gradient onto \mathbf{v} .

$$D_v f(x, y) = \text{comp}_v \nabla f(x, y) = \frac{\nabla f(x, y) \cdot \mathbf{v}}{|\mathbf{v}|}$$

3.8 Practice Problems

13F The directional derivative is $D_{\mathbf{u}} f = (\text{grad } f) \cdot \mathbf{u}$. The level direction is perpendicular to $\text{grad } f$, since $D_{\mathbf{u}} f = 0$. **The slope $D_{\mathbf{u}} f$ is largest when \mathbf{u} is parallel to $\text{grad } f$** . That maximum slope is the length $|\text{grad } f| = \sqrt{f_x^2 + f_y^2}$:

$$\text{for } \mathbf{u} = \frac{\text{grad } f}{|\text{grad } f|} \quad \text{the slope is } (\text{grad } f) \cdot \mathbf{u} = \frac{|\text{grad } f|^2}{|\text{grad } f|} = |\text{grad } f|.$$

$$\int_C g(x, y) ds = \text{limit of } \sum_{i=1}^N g(x_i, y_i) \Delta s_i \quad \text{as } (\Delta s)_{\max} \rightarrow 0.$$

The differential ds becomes $(ds/dt)dt$. Everything changes over to t :

$$\int g(x, y) ds = \int_{t=a}^{t=b} g(x(t), y(t)) \sqrt{(dx/dt)^2 + (dy/dt)^2} dt.$$

19. Let f and g be twice-differentiable real-valued functions defined on \mathbb{R} . If $f'(x) > g'(x)$ for all $x > 0$, which of the following inequalities must be true for all $x > 0$?
- (A) $f(x) > g(x)$
 - (B) $f''(x) > g''(x)$
 - (C) $f(x) - f(0) > g(x) - g(0)$
 - (D) $f'(x) - f'(0) > g'(x) - g'(0)$
 - (E) $f''(x) - f''(0) > g''(x) - g''(0)$

Solution 19. (C) There is no reason that $f(x) > g(x)$, or that $f''(x) > g''(x)$. But we do know that

$$\int_0^x f'(t) dt > \int_0^x g'(t) dt \implies f(x) - f(0) > g(x) - g(0).$$

This is precisely an answer.

22. What is the volume of the solid in xyz -space bounded by the surfaces $y = x^2$, $y = 2 - x^2$, $z = 0$, and $z = y + 3$?

- (A) $\frac{8}{3}$
- (B) $\frac{16}{3}$
- (C) $\frac{32}{3}$
- (D) $\frac{104}{105}$
- (E) $\frac{208}{105}$

Solution 22. (C) It looks like our x -coordinates are running over $[-1, 1]$, with y depending on x and z depending on y . To find the volume of the solid, we just need to integrate the constant function 1. We must therefore compute

$$\begin{aligned} \int_{-1}^1 \int_{x^2}^{2-x^2} \int_0^{y+3} 1 dz dy dx &= \int_{-1}^1 \int_{x^2}^{2-x^2} y + 3 dy dx \\ &= \int_{-1}^1 ((2 - x^2)^2/2 + 3(2 - x^2)) - ((x^2)^2/2 + 3(x^2)) dx \\ &= \int_{-1}^1 8 - 8x^2 dx \\ &= 8x - 8x^3/3 \Big|_{-1}^1 = (8 - 8/3) - (-8 + 8/3) = 32/3. \end{aligned}$$

24. Let h be the function defined by $h(x) = \int_0^{x^2} e^{x+t} dt$ for all real numbers x . Then $h'(1) =$

- (A) $e - 1$
- (B) e^2
- (C) $e^2 - e$
- (D) $2e^2$
- (E) $3e^2 - e$

Solution 24. (E) We can actually just integrate this, and not worry about differentiation under the integral.

$$\int_0^{x^2} e^{x+t} dt = e^x \int_0^{x^2} e^t dt = e^x (e^{x^2} - 1) = e^{x^2+x} - e^x.$$

Then deriving that,

$$h'(x) = (2x+1)e^{x^2+x} - e^x,$$

whence our result follows immediately.

26. Let $f(x, y) = x^2 - 2xy + y^3$ for all real x and y . Which of the following is true?

- (A) f has all of its relative extrema on the line $x = y$.
- (B) f has all of its relative extrema on the parabola $x = y^2$.
- (C) f has a relative minimum at $(0, 0)$.
- (D) f has an absolute minimum at $\left(\frac{2}{3}, \frac{2}{3}\right)$.
- (E) f has an absolute minimum at $(1, 1)$.

Solution 26. (A) We are concerned about its extrema, we should find some partial derivatives.

$$f_x = 2x - 2y, \quad f_y = -2x + 3y^2.$$

We would like to know when they are both zero. The first equation gives us $x = y$ and the second gives us $2x = 3y^2$, so that

$$2y = 3y^2 \implies (3y-2)y = 0 \implies y = 0, 2/3.$$

Therefore our solutions are $(0, 0)$ and $(2/3, 2/3)$. Indeed, our relative extrema are all on the line $x = y$. To do some more checking (which you should not do on the actual test),

$$f_{xx} = 2, \quad f_{yy} = 6y, \quad f_{xy} = f_{yx} = -2.$$

Then the determinant of the Hessian is $12y - 4$. This shows that $(0, 0)$ is a saddle point. There is no reason that $(2/3, 2/3)$ is an absolute minimum without further verification, and $(1, 1)$ needn't be an extreme point.

27. Consider the two planes $x + 3y - 2z = 7$ and $2x + y - 3z = 0$ in \mathbb{R}^3 . Which of the following sets is the intersection of these planes?

- (A) \emptyset
- (B) $\{(0, 3, 1)\}$
- (C) $\{(x, y, z) : x = t, y = 3t, z = 7 - 2t, t \in \mathbb{R}\}$
- (D) $\{(x, y, z) : x = 7t, y = 3 + t, z = 1 + 5t, t \in \mathbb{R}\}$
- (E) $\{(x, y, z) : x - 2y - z = -7\}$

Solution 27. (D) First, we know that the intersection of two planes in \mathbb{R}^3 should be either a plane or a line. In our case, the two planes are definitely not the same, so we will obtain a line. The slope of the line can be found by taking the cross product of the normal vectors of the two planes in question.

$$(1, 3, -2) \times (2, 1, -3) = \det \begin{bmatrix} i & j & k \\ 1 & 3 & -2 \\ 2 & 1 & -3 \end{bmatrix} = (-7, -1, -5).$$

The only solution corresponding to this slope is (D), as the coefficients of t in (x, y, z) are $(7, 1, 5)$.

32. $\frac{d}{dx} \int_{x^3}^{x^4} e^{t^2} dt =$

- (A) $e^{x^6} (e^{x^8-x^6} - 1)$ (B) $4x^3 e^{x^8}$ (C) $\frac{1}{\sqrt{1-e^{x^2}}}$ (D) $\frac{e^{x^2}}{x^2} - 1$ (E) $x^2 e^{x^6} (4xe^{x^8-x^6} - 3)$

Solution 32. (E) We can sort this out in two steps and apply the fundamental theorem to each.

$$\frac{d}{dx} \left(\int_{x^3}^0 e^{t^2} dx + \int_0^{x^4} e^{t^2} dx \right)$$

For the first,

$$\frac{d}{dx} \int_{x^3}^0 e^{t^2} dx = -\frac{d}{dx} \int_0^{x^3} e^{t^2} dx = -3x^2 e^{x^6}$$

For the second,

$$\frac{d}{dx} \int_0^{x^4} e^{t^2} dx = 4x^3 e^{x^8}.$$

All told, our integral is $x^2 e^{x^6} (4xe^{x^8-x^6} - 3)$.

41. Let ℓ be the line that is the intersection of the planes $x + y + z = 3$ and $x - y + z = 5$ in \mathbb{R}^3 . An equation of the plane that contains $(0, 0, 0)$ and is perpendicular to ℓ is

- (A) $x - z = 0$
 (B) $x + y + z = 0$
 (C) $x - y - z = 0$
 (D) $x + z = 0$
 (E) $x + y - z = 0$

Solution 41. (A) The first plane is determined by the normal vector $(1, 1, 1)$, and the second determined by $(1, -1, 1)$. Therefore the slope of ℓ is determined by a vector perpendicular to those, i.e. the cross product.

$$(1, 1, 1) \times (1, -1, 1) = \det \begin{bmatrix} i & j & k \\ 1 & 1 & 1 \\ 1 & -1 & 1 \end{bmatrix} = (2, 0, -2).$$

41. Let C be the circle $x^2 + y^2 = 1$ oriented counterclockwise in the xy -plane. What is the value of the line integral $\oint_C (2x - y) dx + (x + 3y) dy$?

- (A) 0 (B) 1 (C) $\frac{\pi}{2}$ (D) π (E) 2π

Solution 41. (E) This is a classic Green's theorem problem.

$$\oint_{\partial D} L dx + M dy = \iint_D \left(\frac{\partial M}{\partial x} - \frac{\partial L}{\partial y} \right) dx dy.$$

In our case,

$$\oint_C (2x - y) dx + (x + 3y) dy = \iint_D (1 + 1) dx dy = 2A,$$

where A is the area of the unit circle, i.e. π .

So that is the slope of ℓ . We need this to be the normal vector for the plane in question, so it seems that $(1, 0, -1)$ is our best bet (out of the given options).

$$\begin{aligned} y' + xy &= x \\ y(0) &= -1 \end{aligned}$$

44. If y is a real-valued function defined on the real line and satisfying the initial value problem above, then $\lim_{x \rightarrow -\infty} y(x) =$

- (A) 0 (B) 1 (C) -1 (D) ∞ (E) $-\infty$

Solution 44. (B) Putting it in simpler terms,

$$\frac{dy}{dx} + xy = x \implies \frac{dy}{dx} = x(1 - y) \implies \frac{dy}{1-y} = x dx.$$

Integrating both sides, we obtain

$$-\log(1-y) = x^2/2 + C' \implies 1-y = Ce^{-x^2/2} \implies y = 1 - Ce^{-x^2/2}.$$

Solving the initial value problem gives $C = 2$. Furthermore, as $x \rightarrow -\infty$, the second term above vanishes so we get 1 in the limit.

48. Let g be the function defined by $g(x, y, z) = 3x^2y + z$ for all real x, y , and z . Which of the following is the best approximation of the directional derivative of g at the point $(0, 0, \pi)$ in the direction of the vector $\mathbf{i} + 2\mathbf{j} + 3\mathbf{k}$? (Note: \mathbf{i}, \mathbf{j} , and \mathbf{k} are the standard basis vectors in \mathbb{R}^3 .)
- (A) 0.2 (B) 0.8 (C) 1.4 (D) 2.0 (E) 2.6

Solution 48. (B) It would be good to recall the formula for the directional derivative. We take the gradient of the function then take its scalar product with the normalised vector in the direction we want. To begin,

$$\nabla g = (6xy, 3x^2, 1).$$

At the point $(0, 0, \pi)$, we have $\nabla g = (0, 0, 1)$. That works out pretty well for us. The normalised version of the vector $(1, 2, 3)$ is $(1/\sqrt{14}, 2/\sqrt{14}, 3/\sqrt{14})$. Dotting this with $(0, 0, 1)$ gives $3/\sqrt{14}$, and since $\sqrt{14} = 3.5$ or so our answer should be closer to 0.8 than 0.2.

48. Consider the theorem: If f and f' are both strictly increasing real-valued functions on the interval $(0, \infty)$, then $\lim_{x \rightarrow \infty} f(x) = \infty$. The following argument is suggested as a proof of this theorem.

- (1) By the Mean Value Theorem, there is a c_1 in the interval $(1, 2)$ such that

$$f'(c_1) = \frac{f(2) - f(1)}{2 - 1} = f(2) - f(1) > 0.$$

- (2) For each $x > 2$, there is a c_x in $(2, x)$ such that $\frac{f(x) - f(2)}{x - 2} = f'(c_x)$.

- (3) For each $x > 2$, $\frac{f(x) - f(2)}{x - 2} = f'(c_x) > f'(c_1)$ since f' is strictly increasing.

- (4) For each $x > 2$, $f(x) > f(2) + (x - 2)f'(c_1)$.

- (5) $\lim_{x \rightarrow \infty} f(x) = \infty$

Which of the following statements is true?

- (A) The argument is valid.
 (B) The argument is not valid since the hypotheses of the Mean Value Theorem are not satisfied in (1) and (2).
 (C) The argument is not valid since (3) is not valid.
 (D) The argument is not valid since (4) cannot be deduced from the previous steps.
 (E) The argument is not valid since (4) does not imply (5).

Solution 48. (A) The only issue here seems to be that (4) implies that $f(x)$ gets very large so long as $f'(c_1)$ is positive. But we know that it is, since f is a strictly increasing function. Therefore everything is satisfactory.

4 Differential Equations

61. A tank initially contains a salt solution of 3 grams of salt dissolved in 100 liters of water. A salt solution containing 0.02 grams of salt per liter of water is sprayed into the tank at a rate of 4 liters per minute. The sprayed solution is continually mixed with the salt solution in the tank, and the mixture flows out of the tank at a rate of 4 liters per minute. If the mixing is instantaneous, how many grams of salt are in the tank after 100 minutes have elapsed?

(A) 2 (B) $2 - e^{-2}$ (C) $2 + e^{-2}$ (D) $2 - e^{-4}$ (E) $2 + e^{-4}$

Solution 61. (E) We can set this up as a differential equation. Let s denote the amount of salt in the tank, and let t denote time. We have the initial condition of $s(0) = 3$. $s'(t)$ depends on two factors: the salt flowing in and the salt flowing out. The salt flows in constantly at a rate of 0.08 grams per minute, and the salt flows out at a rate of $4 \cdot (s/100) = s/25$ grams per minute. Therefore

$$s'(t) = \frac{ds}{dt} = 0.08 - s(t)/25 \implies \frac{ds}{dt} = 0.04(2 - s) \implies \frac{ds}{2 - s} = 0.04 dt.$$

Doing the usual calculus,

$$-\log(2 - s) = 0.04t + C' \implies 2 - s = Ce^{-0.04t} \implies s(t) = 2 - Ce^{-0.04t}.$$

The initial condition tells us that $C = -1$, so $s(t) = 2 + e^{-0.04t}$. Plugging in $t = 100$ gives our answer.

5 Real Analysis

These are my notes from Math 4650: Analysis I at Cal State LA as well as Prof. Steven Heilman's notes from Math 541A at USC.

5.1 Midterm 1

5.1.1 Homework 1

Definition: Let $S \subseteq \mathbb{R}$. We say that S is **bounded from above** if $\exists b \in \mathbb{R}$ where

$$s \leq b \quad \forall s \in S$$

If this is the case, we call b an **upper bound** of S .

If $b \leq c$ for all upper bounds c of S , we call b the **supremum** of S : $b = \sup(S)$.

We say that S is **bounded from below** if $\exists a \in \mathbb{R}$ where

$$s \geq a \quad \forall s \in S$$

If this is the case, we call a a **lower bound** of S .

If $a \geq d$ for all lower bounds d of S , we call a the **infimum** of S : $a = \inf(S)$.

Useful Sup/Inf Fact: Let $S \in \mathbb{R}$, $S \neq \emptyset$.

(1) Suppose S is bounded from above by an element b . Then $b = \sup(S) \iff \forall \epsilon > 0 \exists x \in S$ with

$$b - \epsilon < x \leq b$$

(2) Suppose S is bounded from below by an element a . Then $a = \inf(S) \iff \forall \epsilon > 0 \exists x \in S$ with

$$a \leq x < a + \epsilon$$

Completeness Axiom: Let S be a nonempty subset of \mathbb{R} . If S is bounded from above, then $\sup(S)$ exists. If S is bounded from below, then $\inf(S)$ exists.

Facts about absolute value:

(1) $|x - y| < \epsilon \iff y - \epsilon < x < y + \epsilon$ (proof: in notes 08/23)

(2) $|ab| = |a||b|$ (proof: 7(c) in Homework 1)

(3) Let $\epsilon > 0$. Then $|a| < \epsilon \iff -\epsilon < a < \epsilon$. (Proof: follows from (1) if $x = a$, $y = 0$.)

(4) $-|a| \leq a \leq |a|$ (proof: Follows from (1) if $x = a$, $y = 0$, $\epsilon = |a|$.)

(5) **Triangle Inequality:** $|a + b| \leq |a| + |b|$ (Proof in notes 08/23)

(6) $||a| - |b|| \leq |a - b|$ (Proof: 7(d) in Homework 1)

- (7) **Triangle Inequality:** $|a - b| \leq |a| + |b|$ (Proof: follows from (5), let $b = -b$.)
- (8) If $a < x < b$ and $a < y < b$ then $|x - y| < b - a$. (Proof: 7(a) in Homework 1)
- (9) $|a - b| = |b - a|$ (Proof: 7(b) in Homework 1.)

5.1.2 Homework 2

Definition: A sequence (a_n) of real numbers is said to **converge** to a **limit** $L \in \mathbb{R}$ if $\forall \epsilon > 0 \exists N > 0$ where

$$n \geq N \implies |a_n - L| < \epsilon$$

We say that (a_n) **diverges** if it does not converge.

Definition: A sequence (a_n) of real numbers is **bounded** if $\exists M > 0$ where $\forall n \in \mathbb{N}$

$$|a_n| \leq M$$

Theorem. If (a_n) converges then (a_n) is bounded.

Definition: Let (a_n) be a sequence of real numbers. We say that (a_n) is a **Cauchy sequence** if $\forall \epsilon > 0 \exists N$ where

$$n, m \geq N \implies |a_n - a_m| < \epsilon$$

Theorem. (a_n) is Cauchy if and only if (a_n) converges.

Corollary. If (a_n) is Cauchy then (a_n) is bounded.

5.2 Midterm 2

5.2.1 Homework 3

Limits of functions at infinity. Let f be a real-valued function defined on some set D where D contains an interval of the form (a, ∞) . Let $L \in \mathbb{R}$. We say

$$\lim_{x \rightarrow \infty} f(x) = L$$

if $\forall \epsilon > 0 \exists N \in \mathbb{R}$ where

$$x \geq N \implies |f(x) - L| < \epsilon$$

Definition: Let $D \subseteq \mathbb{R}$. Let $a \in \mathbb{R}$. We say that a is a **limit point** (or “cluster point,” or “accumulation point”) of D if $\forall \delta > 0 \exists x \in D$ where

$$x \neq a \text{ and } |x - a| < \delta$$

(Note that a may or may not be contained in D .)

Limit of a function at a : Let $D \subseteq \mathbb{R}$ and $f : d \rightarrow \mathbb{R}$. Let a be a limit point of D . Let $x \in D$. We say that f has a *limit as x tends to a* if $\exists L \in \mathbb{R}$ where $\forall \epsilon > 0 \exists \delta > 0$ such that

$$0 < |x - a| < \delta \implies |f(x) - L| < \epsilon$$

and we write

$$\lim_{x \rightarrow a} f(x) = L$$

Properties of Limits: Let $D \in \mathbb{R}$ and let a be a limit point of D . Suppose $f : D \rightarrow \mathbb{R}$ and $g : D \rightarrow \mathbb{R}$. Let $\alpha \in \mathbb{R}$.

(1) If $\lim_{x \rightarrow a} f(x) = L$ and $\lim_{x \rightarrow a} g(x) = M$ then

(a)

$$\lim_{x \rightarrow a} \alpha = \alpha$$

(b)

$$\lim_{x \rightarrow a} [f(x) + g(x)] = L + M$$

(c)

$$\lim_{x \rightarrow a} [f(x) - g(x)] = L - M$$

(d)

$$\lim_{x \rightarrow a} [f(x) \cdot g(x)] = L \cdot M$$

(e)

$$\lim_{x \rightarrow a} [\alpha \cdot f(x)] = \alpha \cdot L$$

(2) If $h : D \rightarrow \mathbb{R}$ and $h(x) \neq 0 \forall x \in D$ and $\lim_{x \rightarrow a} h(x) = H \neq 0$, then

$$\lim_{x \rightarrow a} \frac{1}{h(x)} = \frac{1}{H}$$

Note that properties (2) and (1)(d) combined imply

$$\lim_{x \rightarrow a} \frac{f(x)}{h(x)} = \frac{L}{H}$$

5.2.2 Homework 4

Continuity: Let $D \subseteq \mathbb{R}$ and $f : D \rightarrow \mathbb{R}$ and $a \in D$. Then f is **continuous** at a if $\lim_{x \rightarrow a} f(x)$ exists and

$$\lim_{x \rightarrow a} f(x) = f(a)$$

(Note: if f is continuous at a , then we can say $\forall \epsilon > 0 \exists \delta > 0$ such that

$$|x - a| < \delta \implies |f(x) - L| < \epsilon$$

that is, we don't need to say $0 < |x - a| < \delta$).

If $B \subseteq D$, then f is **continuous on B** if f is continuous at every $b \in B$.

Intermediate Value Theorem: Let f be continuous on $[a, b]$ and suppose $f(a) < f(b)$. $\forall d$ such that

$$f(a) < d < f(b)$$

$\exists c \in \mathbb{R}$ where

$$a < c < b, f(c) = d$$

5.3 Final

5.3.1 Homework 5

Definition: Let $S \subseteq \mathbb{R}$. We say $x \in \mathbb{R}$ is an **interior point** of S if there exists an open interval (a, b) where

$$x \in (a, b) \text{ and } (a, b) \subseteq S$$

Open sets: Let $S \subseteq \mathbb{R}$. We say S is **open** if every $x \in S$ is an interior point of S .

Closed sets: Let $S \subseteq \mathbb{R}$. We say S is **closed** if $\mathbb{R} \setminus S$ is open.

Theorem. A set is closed if and only if it contains all of its limit points.

Facts about open and closed sets: Suppose $a, b \in \mathbb{R}$. Then

- (a, ∞) is open (Proof: Homework 5 problem 5b).
- $(-\infty, b)$ is open (Proof: Homework 5 problem 5a).
- (a, b) is open (Proof: class notes).
- If $a < b$, then $[a, b]$ is closed (Proof: Homework 5 problem 5c).
- If A and B are open, then $A \cup B$ and $A \cap B$ are open (Proof: Homework 5 problem 3).
- If A and B are closed, then $A \cup B$ and $A \cap B$ are closed (Proof: Homework 5 problem 4).
- \mathbb{R} is open (Proof: Homework 5 problem 1) and closed (Proof: $\mathbb{R} \setminus \mathbb{R} = \emptyset$ is open).
- \emptyset is open (Proof: Homework 5 problem 2) and closed (Proof: $\mathbb{R} \setminus \emptyset = \mathbb{R}$ is open).

Definition: Let $S \subseteq \mathbb{R}$. An **open cover** of S is a collection $X = \{\mathcal{O}_\alpha \mid \alpha \in I\}$ where each set \mathcal{O}_α is an open subset of \mathbb{R} such that

$$S \subseteq \bigcup_{\alpha \in I} \mathcal{O}_\alpha$$

(Here I is some set that indexes the \mathcal{O}_α).

If $X' \subseteq X$ such that

$$S \subseteq \bigcup_{\mathcal{O}_\alpha \in X'} \mathcal{O}_\alpha$$

then X' is called a **subcover** of S contained in X . In addition, if X' is finite then we call X' a **finite subcover** of S contained in X .

Compactness: Let $S \subseteq \mathbb{R}$. We say that S is **compact** if every open cover of S contains a finite subcover.

Definition: Let $S \subseteq \mathbb{R}$. We say that S is **bounded** if $\exists M > 0$ where $S \subseteq [-M, M]$.

Note: S is bounded if and only if $|s| \leq M \forall s \in S$.

Heine-Borel Theorem. Let $S \subseteq \mathbb{R}$. S is compact if and only if S is closed and bounded.

Theorem. Let $f : D \rightarrow \mathbb{R}$ be continuous on D . If $X \subseteq D$ and X is compact (closed and bounded), then

$$f(\bar{x}) = \{f(x) \mid x \in X\}$$

is compact (closed and bounded).

Corollary: Suppose $f : D \rightarrow \mathbb{R}$ where D is closed and bounded. Then there exists $a, b \in D$ where $f(a)$ is the min of f on D and $f(b)$ is the max of f on D .

5.3.2 Homework 6

Uniform Continuity: Let $D \subseteq \mathbb{R}$ and let $f : D \rightarrow \mathbb{R}$. We say that f is **uniformly continuous** on D if $\forall \epsilon > 0 \exists \delta > 0$ where

$$x, y \in D \text{ and } 0 < |x - y| < \delta \implies |f(x) - f(y)| < \epsilon$$

Uniform continuity implies continuity. Suppose $f : D \rightarrow \mathbb{R}$ where $D \subseteq \mathbb{R}$. If f is uniformly continuous on D , then f is continuous at every $a \in D$.

5.4 More Theorems

Theorem 5.1. Fubini's Theorem. Let $h : \mathbb{R}^2 \rightarrow \mathbb{R}$ be a continuous function such that $\int \int_{\mathbb{R}^2} |h(x, y)| dx dy < \infty$. Then

$$\int \int_{\mathbb{R}^2} h(x, y) dx dy = \int_{\mathbb{R}} \left(\int_{\mathbb{R}} h(x, y) dx \right) dy = \int_{\mathbb{R}} \left(\int_{\mathbb{R}} h(x, y) dy \right) dx$$

5.5 Problems from Practice Math GRE Subject Tests

38. Let A and B be nonempty subsets of \mathbb{R} and let $f : A \rightarrow B$ be a function. If $C \subseteq A$ and $D \subseteq B$, which of the following must be true?

(A) $C \subseteq f^{-1}(f(C))$

(B) $D \subseteq f(f^{-1}(D))$

(C) $f^{-1}(f(C)) \subseteq C$

Solution 38. (A) Neither of the equalities should hold – these are in fact nonsense statements, as one side lies in A and the other in B . To unravel the remaining two sets,

$$f^{-1}(f(C)) = \{x \in A : f(x) \in f(C)\}, \quad f(f^{-1}(D)) = f(\{y \in A : f(y) \in D\})$$

Clearly the second set must always be contained in D , but not the other way around. Similarly the first set certainly contains all $c \in C$ (as $f(c) \in f(C)$) but not the other way around.

47. The function $f : \mathbb{R} \rightarrow \mathbb{R}$ is defined as follows.

$$f(x) = \begin{cases} 3x^2 & \text{if } x \in \mathbb{Q} \\ -5x^2 & \text{if } x \notin \mathbb{Q} \end{cases}$$

Which of the following is true?

- (A) f is discontinuous at all $x \in \mathbb{R}$.
- (B) f is continuous only at $x = 0$ and differentiable only at $x = 0$.
- (C) f is continuous only at $x = 0$ and nondifferentiable at all $x \in \mathbb{R}$.
- (D) f is continuous at all $x \in \mathbb{Q}$ and nondifferentiable at all $x \in \mathbb{R}$.
- (E) f is continuous at all $x \notin \mathbb{Q}$ and nondifferentiable at all $x \in \mathbb{R}$.

Solution 47. (B) A classic kind of problem. We are clearly continuous and differentiable at 0. Anywhere else, near a rational number there is an irrational number and vice versa. Therefore there can be no continuity anywhere but at 0, and hence no differentiability either.

57. For each positive integer n , let x_n be a real number in the open interval $\left(0, \frac{1}{n}\right)$. Which of the following statements must be true?

- I. $\lim_{n \rightarrow \infty} x_n = 0$
 - II. If f is a continuous real-valued function defined on $(0, 1)$, then $\{f(x_n)\}_{n=1}^{\infty}$ is a Cauchy sequence.
 - III. If g is a uniformly continuous real-valued function defined on $(0, 1)$, then $\lim_{n \rightarrow \infty} g(x_n)$ exists.
- (A) I only (B) I and II only (C) I and III only (D) II and III only (E) I, II, and III

Solution 57. (C) I is true, since $\lim_{n \rightarrow \infty} x_n$ must be bounded between 0 and $\lim_{n \rightarrow \infty} 1/n = 0$. Unfortunately, x_n does not converge inside $(0, 1)$. There is no reason therefore that $f(x_n)$ should be a convergent sequence – suppose that $f(x) = 1/x$, so that $f(x_n)$ is certainly not Cauchy. However, if g is uniformly continuous, then g extends to a continuous function on $[0, 1]$. Now x_n is a convergent sequence, so $\lim_{n \rightarrow \infty} g(x_n) = g(\lim_{n \rightarrow \infty} x_n) = g(0)$ exists.

60. A real-valued function f defined on \mathbb{R} has the following property.

For every positive number ϵ , there exists a positive number δ such that

$$|f(x) - f(1)| \geq \epsilon \text{ whenever } |x - 1| \geq \delta.$$

This property is equivalent to which of the following statements about f ?

- (A) f is continuous at $x = 1$.
- (B) f is discontinuous at $x = 1$.
- (C) f is unbounded.
- (D) $\lim_{|x| \rightarrow \infty} |f(x)| = \infty$
- (E) $\int_0^{\infty} |f(x)| dx = \infty$

Solution 60. (D) While it looks like this is the opposite of continuity, that should read ‘there exists $\epsilon > 0$ ’. What the statement says is that we not only get arbitrarily far away from $f(1)$, but we must for all x sufficiently far away from 1. So as $|x|$ gets very large, so does $|f(x)|$.

63. For any nonempty sets A and B of real numbers, let $A \cdot B$ be the set defined by

$$A \cdot B = \{xy : x \in A \text{ and } y \in B\}.$$

If A and B are nonempty bounded sets of real numbers and if $\sup(A) > \sup(B)$, then $\sup(A \cdot B) =$

- (A) $\sup(A) \sup(B)$
- (B) $\sup(A) \inf(B)$
- (C) $\max\{\sup(A) \sup(B), \inf(A) \inf(B)\}$
- (D) $\max\{\sup(A) \sup(B), \sup(A) \inf(B)\}$
- (E) $\max\{\sup(A) \sup(B), \inf(A) \sup(B), \inf(A) \inf(B)\}$

Solution 63. (E) The supremum is either going to be the product of the two largest positive numbers in A and B or the product of the two smallest negative numbers in A and B . That means we should look for $\sup \cdot \sup$ or $\inf \cdot \inf$. However, it might be the case that B contains only negative numbers and A contains only positive numbers. Then the largest value in $A \cdot B$ will be attained by the smallest positive element of A and the largest negative element of B , giving us our third option: $\inf A \cdot \sup B$.

6 Probability

These are my notes from taking Math 505A at USC, Math 541A at USC, ISE 620 at USC and the textbook *Probability and Random Processes* (Grimmett and Stirzaker) 3rd edition.

6.1 To Know for Math 505A Midterm 1 (Discrete Random Variables)

6.1.1 Definitions

Definition 6.1. The **probability mass function** of a discrete random variable X is the function $f : \mathbb{R} \rightarrow [0, 1]$ given by $f(x) = \Pr(X = x)$.

Definition 6.2. The **(cumulative) distribution function** of a discrete random variable F is given by

$$F(x) = \sum_{i:x_i \leq x} f(x_i)$$

Definition 6.3. The **joint probability mass function** $f : \mathbb{R}^2 \rightarrow [0, 1]$ of two discrete random variables X and Y is given by

$$f(x, y) = \Pr(X = x \cap Y = y)$$

Definition 6.4. The **joint distribution function** $F : \mathbb{R}^2 \rightarrow [0, 1]$ is given by

$$F(x, y) = \Pr(X \leq x \cap Y \leq y)$$

Definition 6.5. If $\Pr(B) > 0$ then the **conditional probability** that A occurs given that B occurs is defined to be

$$\Pr(A | B) = \frac{\Pr(A \cap B)}{\Pr(B)}$$

Definition 6.6. (Independent sets.) Let A_1, A_2, \dots be subsets of a sample space Ω , and let \mathbb{P} be a probability law on Ω . We say that A_1, A_2, \dots are **independent** if for any finite subset S of $\{1, 2, \dots\}$, we have

$$\mathbb{P}\left(\bigcap_{i \in S} A_i\right) = \prod_{i \in S} \mathbb{P}(A_i)$$

Definition 6.7. (Independence of random variables.) Random variables X_1, X_2, \dots are **independent** if for every $B_1, B_2, \dots \subseteq \mathbb{R}$, the events $\{X_1 \in B_1\}, \{X_2 \in B_2\}, \dots$ are independent; that is,

$$\mathbb{P}\left(\bigcap_{i=1}^n \{X_i \in B_i\}\right) = \prod_{i=1}^n \mathbb{P}(\{X_i \in B_i\})$$

Remark. A more informal definition is as follows: Two random variables X and Y are **independent** if and only if $\Pr(X \cap Y) = \Pr(X)\Pr(Y)$.

Theorem 6.1. (Law of total probability). If X is a random variable and Y is a discrete random variable taking on values y_1, y_2, \dots, y_n , then $\Pr(X) = \sum_i \Pr(X | Y = y_i) \cdot \Pr(Y = y_i)$. (Can be used to prove independence.)

Definition 6.8. Two random variables X and Y are **uncorrelated** if $\mathbb{E}(XY) = \mathbb{E}(X)\mathbb{E}(Y)$.

Proposition 6.2. (a) Two random variables are uncorrelated if and only if their covariance $\text{Cov}(X, Y) = \mathbb{E}[(X - \mathbb{E}(X))(Y - \mathbb{E}(Y))] = \mathbb{E}(XY) - \mathbb{E}(X)\mathbb{E}(Y)$ equals 0.

(b) If X and Y are independent then they are uncorrelated.

Theorem 6.3. If X and Y are independent and $g, h : \mathbb{R} \rightarrow \mathbb{R}$, then $g(X)$ and $h(Y)$ are also independent.

6.1.2 Conditioning

Definition 6.9. The **conditional distribution function** of Y given $X = x$, written $F_{Y|X}(\cdot | x)$, is defined by

$$F_{Y|X}(y | x) = \Pr(Y \leq y | X = x)$$

Definition 6.10. The **conditional probability mass function** of Y given $X = x$, written $f_{Y|X}(\cdot | x)$, is defined by

$$f_{Y|X}(y | x) = \Pr(Y = y | X = x)$$

Theorem 6.4. Iterated expectations:

- (i) $\mathbb{E}[\mathbb{E}(X | Y)] = \mathbb{E}(X)$ (**Law of Total Expectation**)
- (ii) $\mathbb{E}[(X | Y) | Z] = \mathbb{E}(X | Y)$
- (iii) $\mathbb{E}(E(XY | Y)) = \mathbb{E}(Y\mathbb{E}(X | Y))$

Proof. (i) Discrete case:

$$\begin{aligned} \mathbb{E}[\mathbb{E}(X | Y)] &= \sum_y \mathbb{E}(X | Y = y) \Pr(Y = y) = \sum_y \sum_x x \Pr(X = x | Y = y) \Pr(Y = y) \\ &= \sum_y \sum_x x \Pr(X = x \cap Y = y) = \sum_x x \sum_y \Pr(X = x \cap Y = y) = \sum_x x \Pr(X = x) = \mathbb{E}(X) \end{aligned}$$

Continuous case:

$$\mathbb{E}[\mathbb{E}(X | Y)] = \int_{-\infty}^{\infty} \mathbb{E}(X | Y = y) f_Y(y) dy = \text{(by definition 1.75)} \int_{-\infty}^{\infty} \left(\int_{-\infty}^{\infty} x f_{X|Y}(x | y) dx \right) f_Y(y) dy$$

$$\text{(by Fubini's Theorem, Theorem 5.1)} \quad \int_{-\infty}^{\infty} \left(\int_{-\infty}^{\infty} xf_{X|Y}(x | y) f_Y(y) dy \right) dx = \int_{-\infty}^{\infty} xf_X(x) dx = \mathbb{E}(X)$$

□

Definition 6.11. Conditional Variance: $\text{Var}(X | Y) = \mathbb{E}[(X - \mathbb{E}(X | Y))^2 | Y]$

6.1.3 Odds and Ends

Proposition 6.5. Inclusion-Exclusion Principle:

(a)

$$\Pr \left(\bigcup_{i=1}^n A_i \right) = \sum_{k=1}^n (-1)^{k+1} \left(\sum_{1 \leq i_1 < \dots < i_k \leq m} \Pr(A_{i1} \cap \dots \cap A_{ik}) \right)$$

(b)

$$\left| \bigcup_{i=1}^n A_i \right| = \sum_{k=1}^n (-1)^{k+1} \left(\sum_{1 \leq i_1 < \dots < i_k \leq m} |A_{i1} \cap \dots \cap A_{ik}| \right)$$

To prove this, we will first prove the Multi-Binomial Theorem.

Lemma 6.6. (Multi-Binomial Theorem)

$$\prod_{i=1}^d (x_i + y_i)^{n_i} = \sum_{k_1=0}^{n_1} \sum_{k_2=0}^{n_2} \dots \sum_{k_d=0}^{n_d} \binom{n_1}{k_1} x_1^{k_1} y_1^{n_1-k_1} \binom{n_2}{k_2} x_2^{k_2} y_2^{n_2-k_2} \dots \binom{n_d}{k_d} x_d^{k_d} y_d^{n_d-k_d}$$

Proof.

□

We are now ready to prove the Inclusion-Exclusion Principle.

Proof. (Proof of (a).) Begin by noting that

$$\mathbf{1}_{\{\bigcup_{i=1}^n A_i\}} = 1 - \prod_{i=1}^n (1 - \mathbf{1}_{\{A_i\}}) \tag{1}$$

because the expression on the right will equal 1 if at least one term in the product equals 0 (that is, if $\mathbf{1}_{\{A_i\}} = 1$ for some $i \in 1, \dots, n$) and will equal 0 if every term in the product equals 1 (if $\mathbf{1}_{\{A_i\}} = 0$ for every $i \in 1, \dots, n$), which is exactly what we want. Expanding the right side of (1) using the Multi-Binomial Theorem (Lemma 6.6), we have

$$= 1 - \prod_{i=1}^n (1 - \mathbf{1}_{\{A_i\}}) = 1 - (1 - \mathbf{1}_{\{A_1\}})(1 - \mathbf{1}_{\{A_2\}}) \cdots (1 - \mathbf{1}_{\{A_n\}})$$

$$\begin{aligned}
&= 1 - \left[1 + \sum_{k=1}^n (-1)^k \left(\sum_{1 \leq i_1 < \dots < i_k \leq k} \mathbf{1}_{\{A_{i_1}\}} \cdots \mathbf{1}_{\{A_{i_k}\}} \right) \right] = -1 \cdot \sum_{k=1}^n (-1)^k \left(\sum_{1 \leq i_1 < \dots < i_k \leq k} \mathbf{1}_{\{A_{i_1} \cap \dots \cap A_{i_k}\}} \right) \\
&= \sum_{k=1}^n (-1)^{k+1} \left(\sum_{1 \leq i_1 < \dots < i_k \leq k} \mathbf{1}_{\{A_{i_1} \cap \dots \cap A_{i_k}\}} \right) \\
\implies \mathbf{1}_{\{\cup_{i=1}^n A_i\}} &= \sum_{k=1}^n (-1)^{k+1} \left(\sum_{1 \leq i_1 < \dots < i_k \leq k} \mathbf{1}_{\{A_{i_1} \cap \dots \cap A_{i_k}\}} \right) \tag{2}
\end{aligned}$$

Taking expectations of both sides of (2) yields

$$\begin{aligned}
\Pr(\cup_{i=1}^n A_i) &= \mathbb{E} \left[\sum_{k=1}^n (-1)^{k+1} \left(\sum_{1 \leq i_1 < \dots < i_k \leq k} \mathbf{1}_{\{A_{i_1} \cap \dots \cap A_{i_k}\}} \right) \right] \\
&= \sum_{k=1}^n (-1)^{k+1} \left(\sum_{1 \leq i_1 < \dots < i_k \leq k} \mathbb{E}[\mathbf{1}_{\{A_{i_1} \cap \dots \cap A_{i_k}\}}] \right) \\
&= \sum_{k=1}^n (-1)^{k+1} \left(\sum_{1 \leq i_1 < \dots < i_k \leq k} \Pr(A_{i_1} \cap \dots \cap A_{i_k}) \right)
\end{aligned}$$

□

Theorem 6.7. Sums of random variables. If X and Y are independent then

$$\Pr(X + Y = z) = f_{X+Y}(z) = \sum_x f_X(x) f_Y(z - x) = \sum_y f_X(z - y) f_Y(y)$$

Proposition 6.8. Variance-Covariance Expansion. Let X_1, \dots, X_n be random variables. If $\mathbb{E}|X_k|^2 < \infty$, then

$$\text{Var}(X_1 + \dots + X_n) = \sum_k \text{Var}(X_k) + \sum_{k \neq m} \sum_m \text{Cov}(X_k, X_m)$$

Definition 6.12. Let $\{X_i\}$ be i.i.d. random variables. Let N be a random variable taking on positive integer values. Let

$$S = \sum_{i=1}^N X_i$$

Then S is a **compound random variable**.

Proposition 6.9. (Wald's Equation.) Let $\{X_i\}$ be i.i.d. random variables with mean $\mathbb{E}(X)$. Let N be a random variable taking on positive integer values, and let $S = \sum_{i=1}^N X_i$. Then $\mathbb{E}(S) = \mathbb{E}(N)\mathbb{E}(X)$.

Proof.

$$\mathbb{E}(S \mid N) = \mathbb{E}\left(\sum_{i=1}^N X_i\right) = \sum_{i=1}^N \mathbb{E}(X_i) = N\mathbb{E}(X)$$

$$\mathbb{E}(S) = \mathbb{E}(\mathbb{E}[S \mid N]) = \mathbb{E}(N\mathbb{E}(X)) = \mathbb{E}(N)\mathbb{E}(X)$$

□

Proposition 6.10. (**Proposition 1.6.1 in Sheldon Ross *A First Course in Probability*.**) There are $\binom{n-1}{r-1}$ distinct positive integer-valued vectors $(x_1, x_2, \dots, x_r), x_i > 0 \forall i$ satisfying the equation $x_1 + x_2 + \dots + x_r = n$.

Proof. (Not rigorous, but a justification.) Imagine we have n indistinguishable objects to allocate to r people. We lay out the n objects and take $r - 1$ sticks to place in the $n - 1$ spaces between them. The first person gets all the objects to the left of the leftmost stick, the second person gets the objects between the leftmost and second leftmost stick, and so on, until the last person gets all the objects to the right of the rightmost stick. The constraint that x_i be positive is equivalent to saying that each person must receive at least one object. Therefore we must place each stick in a different place. There are $\binom{n-1}{r-1}$ ways to do this.

□

Proposition 6.11. (**Proposition 1.6.2 in Sheldon Ross *A First Course in Probability*.**) There are $\binom{n+r-1}{r-1}$ distinct nonnegative integer-valued vectors $(x_1, x_2, \dots, x_r), x_i \geq 0 \forall i$ satisfying the equation $x_1 + x_2 + \dots + x_r = n$.

Proof. We would like to solve the problem

$$x_1 + x_2 + \dots + x_r = n, x_i \geq 0 \forall i$$

Note that we can transform this problem in the following way:

$$x_1 + 1 + x_2 + 1 + \dots + x_r + 1 = n + 1 \cdot r, x_i + 1 \geq 1 \forall i$$

Letting $y_i = x_i + 1$, we have the equivalent system

$$y_1 + y_2 + \dots + y_r = n + r, y_i \geq 1 \forall i$$

Since $y_i \geq 1 \iff y_i > 0$, by Proposition 6.10, the number of distinct solutions to this equation is $\binom{n+r-1}{r-1}$.

□

Proposition 6.12. More generally, if we desire solutions to $x_1 + x_2 + \dots + x_r = n$ such that $x_i \geq k \in \mathbb{N}$, then $\binom{n+r \cdot (1-k)-1}{r-1} = \binom{n+r-1-rk}{r-1}$ solutions are possible.

Proof. We can construct a similar argument to that used in the proof of Proposition 6.11 by adding $r \cdot (1 - k)$ to each side of $x_1 + x_2 + \dots + x_r = n$:

$$x_1 + 1 - k + x_2 + 1 - k + \dots + x_r + 1 - k = n + r \cdot (1 - k), \quad x_i \geq k \quad \forall i$$

Then substitute $y_i = x_i + 1 - k$ to yield

$$y_1 + y_2 + \dots + y_r = n + r \cdot (1 - k), \quad y_i + k - 1 \geq k \quad \forall i$$

then apply Proposition 6.10 noting that $y_i + k - 1 \geq k \iff y_i \geq 1 \iff x_i \geq k$ to yield the result. \square

Proposition 6.13. Even more generally, suppose we desire solutions to $x_1 + x_2 + \dots + x_r = n$ such that $x_1 \geq k_1 \in \mathbb{N}, x_2 \geq k_2 \in \mathbb{N}, \dots, x_r \geq k_r \in \mathbb{N}$. Then

$$\binom{n + \sum_{i=1}^r (1 - k_i) - 1}{r - 1} = \binom{n + r - 1 - \sum_{i=1}^r k_i}{r - 1}$$

solutions are possible.

Proof. Very similar to the proof of Proposition 6.12. Add $\sum_{i=1}^r (1 - k_i)$ to each side of $x_1 + x_2 + \dots + x_r = n$:

$$x_1 + 1 - k_1 + x_2 + 1 - k_2 + \dots + x_r + 1 - k_r = n + \sum_{i=1}^r (1 - k_i), \quad x_i \geq k_i \quad \forall i$$

Then substitute $y_i = x_i + 1 - k_i$ to yield

$$y_1 + y_2 + \dots + y_r = n + \sum_{i=1}^r (1 - k_i), \quad y_i + k_i - 1 \geq k_i \quad \forall i$$

Finally, apply Proposition 6.10 noting that $y_i + k_i - 1 \geq k_i \iff y_i \geq 1 \iff x_i \geq k_i$ to yield the result. \square

Proposition 6.14. Suppose we desire solutions to $x_1 + x_2 + \dots + x_r = n$ such that $\tau \leq r$ of the $\{x_i\}$ exceed some threshold k . For example, if we have $x_1 \geq k, x_2 \geq k, \dots, x_\tau \geq k$, with $x_{\tau+1}, \dots, x_r$ taking on arbitrary values, then the condition is satisfied. (The particular x_i that exceed k does not matter, as long as τ of them exceed k). Then

$$\binom{r}{\tau} \binom{n + r - 1 - k\tau}{r - 1}$$

solutions are possible.

Proof. By Proposition 6.13, the number of ways this condition can be met for a particular set of τ variables x_i is $\binom{n+r-1-k\tau}{r-1}$. Since there are $\binom{r}{\tau}$ ways to choose which τ variables will exceed k , the result follows. \square

Remark. Convolution on the integers. Let X, Y be independent integer-valued random variables. Let $t \in \mathbb{Z}$.

$$\begin{aligned}\Pr(X + Y = t) &= \sum_{j,k \in \mathbb{Z}: j+k=t} \Pr(X = j, Y = k) = \sum_{j \in F} \Pr(X = j, Y = t-j) = \sum_{j \in \mathbb{Z}} \Pr(X = j) \Pr(Y = t-j) \\ &= \sum_{j \in \mathbb{Z}} p_X(j) p_Y(t-j)\end{aligned}$$

Definition 6.13. Let $g, h : \mathbb{Z} \rightarrow \mathbb{R}$ be functions. The **convolution** of g and h , denoted by $g * h$, is a function $g * h : \mathbb{Z} \rightarrow \mathbb{R}$ defined by

$$(g * h)(t) = \sum_{j \in \mathbb{Z}} g(j)h(t-j) \quad \forall t \in \mathbb{Z}$$

Definition 6.14. (Convolution on the real line.) Let $g, h : \mathbb{R} \rightarrow \mathbb{R}$ be functions. The **convolution** of g and h , denoted by $g * h$, is a function $g * h : \mathbb{R} \rightarrow \mathbb{R}$ defined by

$$(g * h)(t) = \int_{-\infty}^{\infty} g(x)h(t-x)dx, \quad \forall t \in \mathbb{R}$$

Proposition 6.15. Let X, Y be two continuous independent random variables such that $\Pr(X + Y \leq t)$ is differentiable with respect to $t \in \mathbb{R}$. Then

$$f_{X+Y}(t) = (f_X * f_Y)(t), \quad \forall t \in \mathbb{R}$$

Proof.

$$\Pr(X + Y \leq t) = \int_{\{(x,y) \in \mathbb{R}^2: x+y \leq t\}} f_{X,Y}(x,y)dxdy = \int_{x=-\infty}^{x=\infty} \int_{y=-\infty}^{y=t-x} f_X(x)f_Y(y)dydx$$

Then, since $\Pr(X + Y \leq t)$ is differentiable with respect to t , we have by the Fundamental Theorem of Calculus

$$f_{X+Y}(t) = \frac{d}{dt} \Pr(X + Y \leq t) = \int_{x=-\infty}^{x=\infty} f_X(x) \frac{d}{dt} \int_{y=-\infty}^{y=t-x} f_Y(y)dydx = \int_{x=-\infty}^{x=\infty} f_X(x)f_Y(t-x)dx$$

□

6.1.4 Methods for Calculating Quantities

- Expectation

—

Definition 6.15. $\mathbb{E}(X) = \sum_x x \Pr(X = x)$

—

Theorem 6.16. (a) $\mathbb{E}(aX + bY) = a\mathbb{E}(X) + b\mathbb{E}(Y)$

(b) If $X \geq 0$ then $\mathbb{E}(X) \geq 0$

—

Theorem 6.17. Law of the Unconscious Statistician: If X has mass function f , and $g : \mathbb{R} \rightarrow \mathbb{R}$, then

$$\mathbb{E}(g(X)) = \sum_x g(x)f(x)$$

—

Proposition 6.18. Expectation is a linear operator: $\mathbb{E}(\sum_i X_i) = \sum_i \mathbb{E}(X_i)$

- Variance

—

Definition 6.16. $\text{Var}(X) = \mathbb{E}(X - \mathbb{E}(X))^2$

—

Proposition 6.19. (Useful reformulation:) $\text{Var}(X) = \mathbb{E}(X^2) - \mathbb{E}(X)^2$

—

Theorem 6.20. (Some useful results):

- (a) $\text{Var}(aX) = a^2 \text{Var}(X)$
- (b) $\text{Var}(X + Y) = \text{Var}(X) + \text{Var}(Y) + 2\text{Cov}(X, Y)$
- (c) $\text{Var}(aX \pm bY) = a^2 \text{Var}(X) + b^2 \text{Var}(Y) \pm 2ab\text{Cov}(X, Y)$
- (d) $\text{Var}\left(\sum_{i=1}^n X_i\right) = \sum_{i=1}^n \text{Var}(X_i) + 2 \sum_{1 \leq i < j \leq n} \text{Cov}(X_i, X_j)$

—

Definition 6.17. Conditional variance:

$$\text{Var}(X | Y) = \mathbb{E}[(X - \mathbb{E}(X | Y))^2 | Y]$$

—

Theorem 6.21. Law of Total Variance: $\text{Var}(X) = \text{Var}(\mathbb{E}(X | Y)) + \mathbb{E}(\text{Var}(X | Y))$

Proof.

$$\text{Var}(\mathbb{E}(X | Y)) = \mathbb{E}[(\mathbb{E}(X | Y))^2] - [\mathbb{E}(\mathbb{E}(X | Y))]^2 = \mathbb{E}[(\mathbb{E}(X | Y))^2] - \mathbb{E}(X)^2 \quad (3)$$

$$\text{Var}(X | Y) = \mathbb{E}(X^2 | Y) - (\mathbb{E}(X | Y))^2 \implies \mathbb{E}[\text{Var}(X | Y)] = \mathbb{E}(X^2) - \mathbb{E}[(\mathbb{E}(X | Y))^2] \quad (4)$$

Adding together (3) and (4) yields

$$\begin{aligned} \text{Var}(\mathbb{E}(X | Y)) + \mathbb{E}(\text{Var}(X | Y)) &= \mathbb{E}[(\mathbb{E}(X | Y))^2] - \mathbb{E}(X)^2 + \mathbb{E}(X^2) - \mathbb{E}[(\mathbb{E}(X | Y))^2] \\ &= \mathbb{E}(X^2) - \mathbb{E}(X)^2 = \text{Var}(X) \end{aligned}$$

□

Corollary 6.21.1. (Rao-Blackwell Theorem.) $\text{Var}(X) \geq \text{Var}(\mathbb{E}(X | Y))$

Proof. Follows immediately from Theorem 6.21 by noting that since the variance is nonnegative, $\mathbb{E}(\text{Var}(X | Y)) \geq 0$.

□

—

Proposition 6.22. If $c \in \mathbb{R}$, then $\text{Var}(c) = 0$.

- Covariance

—

Definition 6.18. $\text{Cov}(X, Y) = \mathbb{E}[(X - \mathbb{E}(X))(Y - \mathbb{E}(Y))]$

—

Proposition 6.23. (Useful reformulation): $\text{Cov}(X, Y) = \mathbb{E}(XY) - \mathbb{E}(X)\mathbb{E}(Y)$

—

Theorem 6.24. (Some useful results):

- (a) $\text{Cov}(aX, BY) = ab\text{Cov}(X, Y)$
- (b) $\text{Cov}(X, X) = \text{Var}(X)$
- (c) $\text{Cov}(aX + bY) = ac\text{Var}(X) + bd\text{Var}(Y) + (ad + bc)\text{Cov}(X, Y)$
- (d) $\text{Cov}(X_1 + X_2, Y) = \text{Cov}(X_1, Y) + \text{Cov}(X_2, Y)$
- (e) $\text{Cov}\left(\sum_{i=1}^n X_i, Y\right) = \sum_{i=1}^n \text{Cov}(X_i, Y)$
- (f) $\text{Cov}\left(\sum_{i=1}^n X_i, \sum_{j=1}^m Y_j\right) = \sum_{i=1}^n \sum_{j=1}^m \text{Cov}(X_i, Y_j)$

—

Definition 6.19. Conditional covariance:

$$\text{Cov}(X, Y | Z) = \mathbb{E}(XY | Z) - \mathbb{E}(X | Z)\mathbb{E}(Y | Z) = \mathbb{E}[(X - \mathbb{E}(X | Z))(Y - \mathbb{E}(Y | Z)) | Z]$$

—

Theorem 6.25. Law of Total Covariance:

$$\text{Cov}(X, Y) = \mathbb{E}(\text{Cov}(X, Y | Z)) + \text{Cov}(\mathbb{E}(X | Z), \mathbb{E}(Y | Z))$$

6.1.5 Discrete Random Variable Distributions

Binomial: Binomial(n, p) (sum of n Bernoulli random variables)

- Mass function: $\Pr(X = k) = \binom{n}{k} p^k (1-p)^{n-k}$
- Distribution: $\Pr(X \leq k) = \sum_{i=0}^k \binom{n}{i} p^i (1-p)^{n-i}$
- Expectation: $\mathbb{E}(X) = np$
- Variance: $\text{Var}(X) = np(1-p)$

Poisson: Poisson(λ): an approximation of the binomial distribution for n very large, p very small, $np \rightarrow \lambda \in (0, \infty)$.

- Mass function:

$$\Pr(X = k) = \frac{e^{-\lambda} \lambda^k}{k!}$$

- Distribution: $\Pr(X \leq k) = \sum_{i=0}^k \frac{e^{-\lambda} \lambda^i}{i!}$
- Expectation: $\mathbb{E}(X) = \lambda$ (derive from basic definitions)
- Variance: $\text{Var}(X) = \lambda$

Proposition 6.26. Let $X \sim \text{Binomial}(n, p)$. Then

$$\lim_{n \rightarrow \infty, p \rightarrow 0, np \rightarrow \lambda} X \sim \text{Poisson}(np).$$

Proof.

$$\begin{aligned} & \lim_{n \rightarrow \infty, p \rightarrow 0, np \rightarrow \lambda} \binom{n}{k} p^k (1-p)^{n-k} = \lim_{n \rightarrow \infty, p \rightarrow 0, np \rightarrow \lambda} \frac{n!}{(n-k)! k!} p^k (1-p)^{n-k} \\ &= \frac{1}{k!} \cdot \lim_{n \rightarrow \infty, p \rightarrow 0, np \rightarrow \lambda} (1-p)^{n-k} p^k \prod_{i=0}^{k-1} (n-i) = \frac{1}{k!} \cdot \lim_{n \rightarrow \infty, p \rightarrow 0, np \rightarrow \lambda} \left(1 - \frac{np}{n}\right)^{n-k} p^k \prod_{i=0}^{k-1} (n-i) \end{aligned}$$

Using $\lim_{n \rightarrow \infty} (1 - \lambda/n)^n = \exp(\lambda)$, and letting $\lambda = np$, we have

$$= \frac{\exp(-np)(np)^k}{k!} = \boxed{\frac{\exp(-\lambda)\lambda^k}{k!}} \sim \text{Poisson}(np)$$

□

Proposition 6.27. Let $X \sim \text{Poisson}(\lambda)$. If λ is sufficiently large (say $\lambda > 20$, then we can use the approximation

$$X \sim \mathcal{N}(\lambda, \lambda)$$

Proof. (Informal justification.) By Proposition 6.26, a Poisson distribution can be thought of as a close approximation of a binomial distribution. Since a binomial distribution can be approximated by a normal distribution for n large and np not too large, the same is true of a Poisson distribution. □

Geometric: $G_1(p)$: the number of Bernoulli trials before the first success.

- Mass function: $\Pr(X = k) = p(1-p)^{k-1}$
- Distribution: $\Pr(X \leq k) = \sum_{i=1}^k p(1-p)^{k-1}$
- Expectation: $\mathbb{E}(X) = 1/p$

- Variance: $\text{Var}(X) = (1-p)/p^2$

Negative binomial: $\text{NB}(r, p)$: The number of Bernoulli trials required for r successes. (Can be derived as the sum of r identically distributed geometric random variables.)

- Mass function: $\Pr(X = k) = \binom{k-1}{r-1} p^r (1-p)^{k-r}$
- Distribution: $\Pr(X \leq k) = \sum_{i=r}^k \binom{i-1}{r-1} p^r (1-p)^{i-r}$
- Expectation: $\mathbb{E}(X) = \frac{pr}{1-p}$
- Variance: $\text{Var}(X) = \frac{pr}{(1-p)^2}$

Hypergeometric: $\text{Hypergeometric}(N, M, K)$: When drawing a sample of size K from a group of N items, M of which are special, X is the number of special items retrieved.

- Mass function:

$$\Pr(X = k) = \frac{\binom{M}{k} \binom{N-M}{K-k}}{\binom{N}{K}}$$

- Distribution:

$$\Pr(X \leq k) = \sum_{i=0}^k \frac{\binom{M}{i} \binom{N-M}{K-i}}{\binom{N}{K}}$$

- Expectation: $\mathbb{E}(X) = \frac{MK}{N}$

Proof. Let Y_j be an indicator variable for special item j being selected. Note that $X = \sum_{j=1}^M Y_j$ and that

$$\mathbb{E}(Y_j) = \frac{\binom{1}{1} \binom{N-1}{K-1}}{\binom{N}{K}} = \frac{1 \cdot \frac{(N-1)!}{(N-K)!(K-1)!}}{\frac{N!}{(N-K)!K!}} = \frac{K}{N},$$

so we have

$$\mathbb{E}(X) = \mathbb{E}\left(\sum_{j=1}^M Y_j\right) = \sum_{j=1}^M \mathbb{E}(Y_j) = \frac{MK}{N}.$$

Alternative proof:

Let Z_i be an indicator variable for the i th selected item to be special. Then $X = \sum_{i=1}^K Z_i$ and $\mathbb{E}(Z_i) = \frac{M}{N}$, so we have

$$\mathbb{E}(X) = \mathbb{E}\left(\sum_{i=1}^K Z_i\right) = \sum_{i=1}^K \mathbb{E}(Z_i) = \frac{MK}{N}.$$

Variance: $\text{Var}(X) = (\text{find by indicator method. Proof from notes, I think this may have errors:})$

Σ_{random}

$X \sim \mathcal{H}(M, m, n)$

$X = \sum_{k=1}^n X_k$ $X_k = \begin{cases} 1, & \text{if element is "special"} \\ 0, & \text{if not.} \end{cases}$

$E(X_k) = \frac{m}{M}$ $\Rightarrow E(X) = n \cdot \frac{m}{M}$

$V_{\text{var}}(X) = \sum_{k=1}^n V_{\text{var}}(X_k) + \sum_{k \neq m} \sum_{m} \text{cov}(X_k, X_m)$

$= n \rho(1-\rho) - 2 \binom{n}{2} \left[\frac{\frac{m(m+1)}{M(M-1)}}{n(n-1)} - \left(\frac{m}{M}\right)^2 \right]$

□

Proposition 6.28. Let $X \sim \text{Hypergeometric}(N, M, K)$. Then

$$\lim_{M, N \rightarrow \infty, M/N \rightarrow p} \Pr(X = k) \sim \text{Binomial}(K, p)$$

Proof.

$$\begin{aligned}
 \lim_{M, N \rightarrow \infty, M/N \rightarrow p} p_k(M, N, K) &= \lim_{M, N \rightarrow \infty, M/N \rightarrow p} \frac{\binom{M}{k} \binom{N-M}{K-k}}{\binom{N}{K}} \\
 &= \lim_{M, N \rightarrow \infty, M/N \rightarrow p} \frac{M!(N-M)!/[k!(M-k)!(K-k)!(N-M-K+k)!]}{N!/[K!(N-K)!]} \\
 &= \lim_{M, N \rightarrow \infty, M/N \rightarrow p} \frac{K!}{(K-k)!k!} \cdot \frac{M!(N-M)!(N-K)!}{N!(M-k)!(N-M-K+k)!} \\
 &= \lim_{M, N \rightarrow \infty, M/N \rightarrow p} \binom{K}{k} \cdot \frac{M!/(M-k)!}{N!/(N-k)!} \cdot \frac{(N-M)!(N-K)!}{(N-k)!(N-M-(K-k))!} \\
 &= \binom{K}{k} \lim_{M, N \rightarrow \infty, M/N \rightarrow p} \frac{M!/(M-k)!}{N!/(N-k)!} \cdot \frac{(N-M)!(N-M-(K-k))!}{(N-K+(K-k))!/(N-K)!} \\
 &= \binom{K}{k} \lim_{M, N \rightarrow \infty, M/N \rightarrow p} \prod_{i=0}^{k-1} \frac{M-i}{N-i} \cdot \prod_{j=0}^{K-k-1} \frac{N-M-j}{N-K+1+j} \\
 &= \binom{K}{k} \left(\frac{M}{N}\right)^k \left(\frac{N-M}{N}\right)^{K-k} \\
 &= \binom{K}{k} \left(\frac{M}{N}\right)^k \left(1 - \frac{M}{N}\right)^{K-k} = \binom{K}{k} p^k (1-p)^{K-k}
 \end{aligned}$$

□

6.1.6 Indicator Method

Proposition 6.29. If $\mathbf{1}_{A_k}$ is an indicator then

(a)

$$\text{Cov}(\mathbf{1}_{A_k}, \mathbf{1}_{A_m}) = \mathbb{E}(\mathbf{1}_{A_k} \mathbf{1}_{A_m}) - \mathbb{E}(\mathbf{1}_{A_k})\mathbb{E}(\mathbf{1}_{A_m}) = \Pr(A_k \cap A_m) - \Pr(A_k)\Pr(A_m)$$

(b)

$$\text{Var}(\mathbf{1}_{A_k}) = \mathbb{E}(\mathbf{1}_{A_k}^2) = \mathbb{E}(\mathbf{1}_{A_k})^2 = \Pr(A_k) - (\Pr(A_k))^2$$

Theorem 6.30. X is independent of Y if and only if X is independent of $\mathbf{1}_A$, $A \in Y$.

Example problems: 505A Homework 3 problem 9(a)

Worked examples in p. 56 - 59 of Grimmett and Stirzaker 3rd edition.

6.1.7 Linear transformations of random variables

6.1.8 Poisson Paradigm (Poisson approximation for indicator method)

Theorem 6.31. (Theorem 4.12.9, p. 129 of Grimmett and Stirzaker.) Let A_i be an event. If $X = \sum_{i=1}^m \mathbf{1}_{A_i}$ where $\mathbf{1}_{A_i}$ is an indicator variable for A_i , and the A_i are only weakly dependent on each other, then

$$\text{As } m \rightarrow \infty, \quad X \sim \text{Poisson}(\mathbb{E}(X))$$

More specifically, let B_i be n independent Bernoulli random variables with probabilities p_i . If $Y = \sum_{i=1}^n B_i$ then

$$\text{As } n \rightarrow \infty, \quad Y \sim \text{Poisson} \left(\mathbb{E} \left(\sum_i B_i \right) \right) = \text{Poisson} \left(\sum_i \mathbb{E} B_i \right) = \text{Poisson} \left(\sum_i p_i \right)$$

Proof. Full proof available in Grimmett and Stirzaker, section 4.12, page 129. A justification of the first claim is as follows: if the A_i are independent and $\Pr(A_i) = p \forall i$, then $X \sim \text{Binomial}(m, p)$. Then by Proposition 6.26, the result follows. It turns out that this result holds up if the probabilities are not necessarily identical (but all small) and the variables are not necessarily independent (but only weakly dependent). \square

6.1.9 Asymptotic Distributions

Proposition 6.32.

$$e^x = \lim_{n \rightarrow \infty} \left(1 + \frac{x}{n} \right)^n$$

Theorem 6.33. Stirling's Formula:

$$n! \approx n^n e^{-n} \sqrt{2\pi n}$$

6.2 Worked problems

6.2.1 Example Problems That Will Likely Appear on Midterm (and Final)

- (1) (a) **Fall 2011 Problem 1 (same as HW1 problem 5; similar to HW3 problem 2(5).)** Let A and B be events such that $0 < \Pr(A) < 1$. Show that if $\Pr(B | A) = \Pr(B | A^c)$, then A and B are independent.
- (b) Let X and Y be two discrete random variables, each taking only two possible values. Show that if $\mathbb{E}(XY) = \mathbb{E}(X)\mathbb{E}(Y)$ then X and Y are independent.

Solution.

- (a) A and B are independent if and only if

$$\Pr(A \cap B) = \Pr(A) \cdot \Pr(B)$$

We know that

$$\Pr(B) = \Pr(B|A) \cdot \Pr(A) + \Pr(B|A^c) \cdot \Pr(A^c)$$

$$= \Pr(B|A) \cdot \Pr(A) + \Pr(B|A) \cdot (1 - \Pr(A)) = \Pr(B|A) \cdot \Pr(A) + \Pr(B|A) - \Pr(B|A) \cdot \Pr(A)$$

$$= \Pr(B|A)$$

Also, we know that since $\Pr(A) \neq 0$,

$$\Pr(B|A) = \frac{\Pr(A \cap B)}{\Pr(A)}$$

Per above $\Pr(B|A) = \Pr(B)$, so we have

$$\Pr(B) = \frac{\Pr(A \cap B)}{\Pr(A)}$$

$$\Pr(A \cap B) = \Pr(A) \cdot \Pr(B)$$

which is what we were trying to prove. So the answer is true.

- (b) Without loss of generality, let X and Y have mass functions

$$X = \begin{cases} x_1 & \text{with probability } \Pr(A) \\ x_2 & \text{with probability } \Pr(A^c) \end{cases}$$

$$Y = \begin{cases} y_1 & \text{with probability } \Pr(B) \\ y_2 & \text{with probability } \Pr(B^c) \end{cases}$$

Then $X \perp\!\!\!\perp Y \iff \Pr(A \cap B) = \Pr(A) \Pr(B)$. Let $\alpha = X - x_2$, $\beta = Y - y_2$; that is,

$$\alpha = \begin{cases} x_1 - x_2 & \text{with probability } \Pr(A) \\ 0 & \text{with probability } \Pr(A^c) \end{cases}$$

$$\beta = \begin{cases} y_1 - y_2 & \text{with probability } \Pr(B) \\ 0 & \text{with probability } \Pr(B^c) \end{cases}$$

Then we have

- $\mathbb{E}(\alpha) = (x_1 - x_2) \Pr(A)$
- $\mathbb{E}(\beta) = (y_1 - y_2) \Pr(B)$
- $\mathbb{E}(\alpha\beta) = (x_1 - x_2)(y_1 - y_2) \Pr(A \cap B)$

which we can use to obtain

$$\begin{aligned}\mathbb{E}(XY) &= \mathbb{E}[(\alpha + x_2)(\beta + y_2)] = \mathbb{E}(\alpha\beta) + y_2\mathbb{E}(\alpha) + x_2\mathbb{E}(\beta) + x_2y_2 \\ &= (x_1 - x_2)(y_1 - y_2) \Pr(A \cap B) + y_2(x_1 - x_2) \Pr(A) + x_2(y_1 - y_2) \Pr(B) + x_2y_2\end{aligned}\quad (5)$$

$$\begin{aligned}\mathbb{E}(X)\mathbb{E}(Y) &= \mathbb{E}(\alpha + x_2)\mathbb{E}(\beta + y_2) = [x_2 + (x_1 - x_2) \Pr(A)][y_2 + (y_1 - y_2) \Pr(B)] \\ &= x_2y_2 + x_2(y_1 - y_2) \Pr(B) + y_2(x_1 - x_2) \Pr(A) + (x_1 - x_2)(y_1 - y_2) \Pr(A) \Pr(B)\end{aligned}\quad (6)$$

Using $\mathbb{E}(XY) = \mathbb{E}(X)\mathbb{E}(Y)$, we set (5) and (6) equal to each other. Canceling terms appearing in both yields

$$(x_1 - x_2)(y_1 - y_2) \Pr(A \cap B) = (x_1 - x_2)(y_1 - y_2) \Pr(A) \Pr(B) \iff \Pr(A \cap B) = \Pr(A) \Pr(B)$$

which proves the independence of X and Y if $\mathbb{E}(XY) = \mathbb{E}(X)\mathbb{E}(Y)$.

- (2) **Fall 2013 Qual Problem 1.** Consider a sequence of independent tosses of a pair of fair dice. Compute the probability that the sum 4 will occur before the sum 5.

Solution. Let Y_k be the outcome of the k th toss. Let X_{k1} be the number of the first die on the k th toss and X_{k2} be the outcome of the second die. Note that

$$\Pr(Y_k = 4) = \Pr(X_{k1} = 1 \cap X_{k2} = 3) + \Pr(X_{k1} = 2 \cap X_{k2} = 2) + \Pr(X_{k1} = 3 \cap X_{k2} = 1) = 3 \cdot \frac{1}{36} = \frac{1}{12}$$

$$\Pr(Y_k = 5) = \Pr(X_{k1} = 1 \cap X_{k2} = 4) + \Pr(X_{k1} = 2 \cap X_{k2} = 3) + \Pr(X_{k1} = 3 \cap X_{k2} = 2)$$

$$+ \Pr(X_{k1} = 4 \cap X_{k2} = 1) = 4 \cdot \frac{1}{36} = \frac{1}{9}$$

Let A_k be the event that $Y_k = 4$ and $Y_j \neq 4$ or 5, $j = 1, \dots, k-1$. Note that all A_k are mutually exclusive and

$$\Pr(A_k) = \frac{1}{12} \cdot \left(1 - \frac{3+4}{36}\right)^{k-1} = \frac{1}{12} \cdot \left(\frac{29}{36}\right)^{k-1}.$$

Then

$$\begin{aligned}\Pr(\{\text{roll a 4 before a 5}\}) &= \Pr\left(\bigcup_{k=1}^{\infty} A_k\right) = \sum_{k=1}^{\infty} \Pr(A_k) = \frac{1}{12} \sum_{k=1}^{\infty} \left(\frac{29}{36}\right)^{k-1} = \frac{1/12}{1 - 29/36} \\ &= \frac{3}{36 - 29} = \boxed{\frac{3}{7}}\end{aligned}$$

- (3) **Spring 2013 Problem 1, in notes from 09/21.** Let X and Y be random variables such that $\mathbb{E}(X | Y) = Y, \mathbb{E}(Y | X) = X, \mathbb{E}(X^2) < \infty, \mathbb{E}(Y^2) < \infty$. Show that $\mathbb{E}(X - Y)^2 = 0$ (or equivalently, show $\Pr(X = Y) = 1$).

Solution.

$$\mathbb{E}(X - Y)^2 = \mathbb{E}(X^2 - 2XY + Y^2) = \mathbb{E}(X^2) - 2\mathbb{E}(XY) + \mathbb{E}(Y^2)$$

$$\mathbb{E}(XY) = \mathbb{E}(\mathbb{E}(XY | Y)) = \mathbb{E}(Y\mathbb{E}(X | Y)) = \mathbb{E}(Y \cdot Y) = \mathbb{E}(Y^2)$$

Also,

$$\mathbb{E}(XY) = \mathbb{E}(\mathbb{E}(XY | X)) = \mathbb{E}(X\mathbb{E}(Y | X)) = \mathbb{E}(X \cdot X) = \mathbb{E}(X^2)$$

Therefore

$$\mathbb{E}(X - Y)^2 = 0$$

- (4) **Fall 2016 Problem 4.** Consider a group of $n \geq 4$ people, among whom are Alice, Bob, Charles, and Diana, standing in a row. Assume that all possible orderings of the n people are equally likely.
- (a) Compute the probability that Charles stands somewhere between Alice and Bob. (Note: this does not mean that the three are necessarily adjacent; there can be other people between Alice and Bob.)
 - (b) Compute the probability that Diana stands somewhere between Alice and Bob given that Charles stands somewhere between Alice and Bob.
 - (c) Let X be the number of people who stand between Alice and Bob. Compute the expected value and the variance of X . (Note: Alice and Bob themselves are not counted in this number.)

Solution.

- (a) For this part, n does not make a difference; all we need to know is the ordering of A, B , and C . This is because conditional on a specific ordering of A, B , and C , all arrangements of everyone else are equally likely; and conversely, the a particular ordering of A, B , and C is independent of which three particular slots are made available to them. Examining the permutations of A, B , and C , two of them have C in the middle, so the answer is $2/6 = \boxed{1/3}$.
- (b) Similarly, the answer is independent of n , so we work with $n = 4$. All possible orderings with Charles between Alice and Bob are as follows:

$$ACDB, ADCB, BCDA, BDCA, ACBD, BCAD, DACB, DBCA$$

The first four of these have Diana between Alice and Bob, so the answer is $4/8 = \boxed{1/2}$.

- (c) Let I_k be an indicator variable for the event that person k is between A and B . By the result from part (a), $\mathbb{E}(I_k) = 1/3$. Then we have

$$\text{Var}(I_k) = \mathbb{E}(I_k^2) - \mathbb{E}(I_k)^2 = 1^2 \cdot \Pr(I_k = 1) - \frac{1}{9} = \frac{1}{3} - \frac{1}{9} = \frac{2}{9}$$

Noting that the four arrangements above with Charles and Diana in between Alice and Bob are the only ones where this will be the case of the $4! = 24$ possible orderings, we have

$$\mathbb{E}(I_k I_j) = \frac{4}{24} = \frac{1}{6}$$

so

$$\text{Cov}(I_j, I_k) = \mathbb{E}(I_k I_j) - \mathbb{E}(I_k)\mathbb{E}(I_j) = \frac{1}{6} - \frac{1}{9} = \frac{1}{18}$$

Therefore

$$\mathbb{E}(X) = \mathbb{E}\left(\sum_{i=1}^{n-2} I_k\right) = \sum_{i=1}^{n-2} \frac{1}{3} = \frac{n-2}{3}$$

$$\text{Var}(X) = \text{Var}\left(\sum_{i=1}^{n-2} I_k\right) = \sum_{i=1}^{n-2} \text{Var}(I_k) + 2 \sum_{1 \leq j < k \leq n-2} \text{Cov}(I_j, I_k) = \frac{2(n-2)}{9} + [(n-2)^2 - (n-2)] \cdot \frac{1}{18}$$

$$= \frac{4(n-2)}{18} + \frac{(n-2)(n-3)}{18} = \boxed{\frac{(n-2)(n+1)}{18}}$$

- (5) **Spring 2015 Problem 2.** A deck of 52 cards is shuffled thoroughly. Someone goes through all 52 cards, scoring 1 point each time 2 cards of the same value are consecutive (that is, when two consecutive cards have the same rank but different suits). For example, the sequence 9H, 8H, 7D, 6C, 7S, 7H, 7C scores 2 points, one for the 7H following the 7S, one for the 7C following the 7H. Let X be the total score.

- (a) Compute $\mathbb{E}(X)$.
- (b) Compute $\Pr(X = 39)$. (Note that there are 13 different ranks and you cannot score more than 3 per rank.)
- (c) In the line below, circle the number that you think is the closest to the value $\Pr(X = 0)$ and briefly explain your choice:

$$\frac{1}{1000}, \quad \frac{1}{500}, \quad \frac{1}{100}, \quad \frac{1}{50}, \quad \frac{1}{20}, \quad \frac{1}{10}, \quad \frac{1}{5}, \quad \frac{1}{2}$$

Solution.

- (a) Start by assuming the permutation is cyclic (that is, after the last card you go back to the beginning). Let Y be the number of matches in this situation. Let A_i be the event that the i th card is followed by a match. Then $\Pr(A_i) = 3/51 = 1/17$, so

$$Y = \sum_{k=1}^{52} \mathbf{1}_{\{A_i\}} \implies \mathbb{E}(Y) = \sum_{i=1}^{52} \mathbb{E}(\mathbf{1}_{\{A_i\}}) = \sum_{i=1}^{52} \Pr(A_i) = 52 \cdot 1/17 = 52/17$$

Note that $\mathbb{E}(Y) = \mathbb{E}(X) + \mathbb{E}(\mathbf{1}_{\{A_i\}})$ because if the permutation is cyclic then you have one extra opportunity to match at the end.

$$\implies \mathbb{E}(X) = \frac{52}{17} - \frac{1}{17} = \frac{51}{17} = \boxed{3}$$

(b) $\Pr(X = 39) = \Pr(\{\text{all possible matches occur}\})$, so in this event all cards of the same rank are clustered together. There are 13 clusters of 4 cards, so there are $13!$ ways to order the clusters and $4!$ ways to order the cards within each cluster. Therefore

$$\boxed{\Pr(X = 39) = \frac{13!(4!)^{13}}{52!}}$$

(c) Because A_i are only weakly dependent and $\Pr(A_i)$ is small for all A_i , we can use the Poisson approximation (see Section 6.1.8); that is, $X \sim \text{Poisson}(\mathbb{E}(X)) = \text{Poisson}(3)$. Therefore

$$\Pr(X = 0) \approx \frac{e^{-3} \cdot 3^0}{0!} = \frac{1}{e^3} \approx \frac{1}{2.8^3} \approx \boxed{\frac{1}{20}}$$

6.2.2 Problems we did in class that professor mentioned

(Fall 2014 Problem 1) (Variation of Midterm problem 1 above) Let A and B be two events with $0 < \Pr(A) < 1$, $0 < \Pr(B) < 1$. Define the random variables $\xi = \xi(\omega)$ and $\eta = \eta(\omega)$ by

$$\xi(\omega) = \begin{cases} 5 & \text{if } \omega \in A \\ -7 & \text{if } \omega \notin A \end{cases}, \quad \eta(\omega) = \begin{cases} 2 & \text{if } \omega \in B \\ 3 & \text{if } \omega \notin B \end{cases}$$

True or false: the events A and B are independent if and only if the random variables ξ and η are uncorrelated?

Solution. (\implies) Suppose A and B are independent. Then ξ and η are uncorrelated if and only if $\mathbb{E}(\xi\eta) = \mathbb{E}(\xi)\mathbb{E}(\eta)$. We can write $\xi = 5 \cdot \mathbf{1}_A - 7 \cdot \mathbf{1}_{A^c}$ and $\eta = 2 \cdot \mathbf{1}_B + 3 \cdot \mathbf{1}_{B^c}$. So we have

$$\xi\eta = (5 \cdot \mathbf{1}_A - 7 \cdot \mathbf{1}_{A^c})(2 \cdot \mathbf{1}_B + 3 \cdot \mathbf{1}_{B^c}) = 10 \cdot \mathbf{1}_{A \cap B} + 15 \cdot \mathbf{1}_{A \cap B^c} - 14 \cdot \mathbf{1}_{A^c \cap B} - 21 \cdot \mathbf{1}_{A^c \cap B^c}$$

$$\implies \mathbb{E}(\xi\eta) = 10 \Pr(A \cap B) + 15 \Pr(A \cap B^c) - 14 \Pr(A^c \cap B) - 21 \Pr(A^c \cap B^c)$$

Then

$$\mathbb{E}(\xi)\mathbb{E}(\eta) = (5 \Pr(A) - 7 \Pr(A^c))(2 \Pr(B) + 3 \Pr(B^c))$$

$$= 10 \Pr(A \cap B) + 15 \Pr(A \cap B^c) - 14 \Pr(A^c \cap B) - 21 \Pr(A^c \cap B^c) = \mathbb{E}(\xi\eta)$$

where the second-to-last step follows from the independence of A and B . Therefore η and ξ are uncorrelated.

(\impliedby) Now suppose η and ξ are uncorrelated. Then ξ and η are independent if and only if $\Pr(\xi \cap \eta) = \Pr(\xi)\Pr(\eta)$. Define

$$\alpha(\omega) = \xi(\omega) + 7 = \begin{cases} 12 & \text{if } \omega \in A \\ 0 & \text{if } \omega \notin A \end{cases}, \quad \beta(\omega) = \eta(\omega) - 3 = \begin{cases} -1 & \text{if } \omega \in B \\ 0 & \text{if } \omega \notin B \end{cases}$$

Then we have

$$(\alpha\beta)(\omega) = \begin{cases} -12 & \text{if } \omega \in A \cap B \\ 0 & \text{otherwise} \end{cases}$$

Then

$$\mathbb{E}(\xi\eta) = \mathbb{E}[(\alpha - 7)(\beta + 3)] = \mathbb{E}(\alpha\beta) + 3\mathbb{E}(\alpha) - 7\mathbb{E}(\beta) - 21$$

$$\mathbb{E}(\xi)\mathbb{E}(\eta) = (\mathbb{E}(\alpha) - 7)(\mathbb{E}(\beta) + 3) = \mathbb{E}(\alpha)\mathbb{E}(\beta) - 7\mathbb{E}(\beta) + 3\mathbb{E}(\alpha) - 21$$

Since by assumption $\mathbb{E}(\xi\eta) = \mathbb{E}(\xi)\mathbb{E}(\eta)$, this yields $\mathbb{E}(\alpha\beta) = \mathbb{E}(\alpha)\mathbb{E}(\beta)$. But

$$\mathbb{E}(\alpha\beta) = -12 \Pr(A \cap B), \quad \mathbb{E}(\alpha)\mathbb{E}(\beta) = 12 \Pr(A)(-1) \Pr(B) = -12 \Pr(A) \Pr(B)$$

Therefore $\Pr(\xi \cap \eta) = \Pr(\xi) \Pr(\eta)$ and ξ and η are independent.

Example (Letter/envelope matching problem; sometimes referred to as Montmort's matching problem). An assistant brings n sandwiches for n employees at a company. Each employee ordered a unique sandwich, but unfortunately the assistant forgot to ask that the sandwiches be labeled, so they are all indistinguishable, wrapped in the same paper. The assistant plans to distribute one sandwich to each employee and hope for the best. Let X be the number of sandwiches that are delivered to the correct person.

- (a) What is the probability of at least one match; that is, $\Pr(X \geq 1)$?
- (b) What is the probability of r correct matches?
- (c) What $\mathbb{E}(X)$?
- (d) What is $\text{Var}(X)$?

Solution.

- (a) Let A_k be an indicator variable for the event that sandwich k is matched to the correct employee.
Then

$$\Pr(X \geq 1) = \Pr\left(\bigcup_{k=1}^n A_k\right)$$

Consider that if there are k correct matches, there are $\binom{n}{k}$ sets of k sandwiches that could be correctly distributed. Also, the probability of a particular set of k sandwiches being correctly distributed is $(n - k)!/n!$. So we have

$$\Pr(X = k) = \binom{n}{k} \frac{(n - k)!}{n!}$$

Therefore by the Inclusion-Exclusion Principle (Proposition 6.5),

$$\begin{aligned} \Pr\left(\bigcup_{k=1}^n A_k\right) &= \sum_{k=1}^n (-1)^{k-1} \Pr(X = k) = \sum_{k=1}^n (-1)^{k-1} \binom{n}{k} \frac{(n - k)!}{n!} = \sum_{k=1}^n (-1)^{k-1} \frac{n!}{(n - k)!k!} \frac{(n - k)!}{n!} \\ &= \sum_{k=1}^n \frac{(-1)^{k-1}}{k!} = \frac{(-1)^0}{0!} - \sum_{k=0}^n \frac{(-1)^k}{k!} = \boxed{1 - \sum_{k=0}^n \frac{(-1)^k}{k!}} \end{aligned}$$

As $n \rightarrow \infty$, we have

$$1 - \sum_{k=0}^n \frac{(-1)^k}{k!} \rightarrow 1 - e^{-1} = \boxed{1 - \frac{1}{e}}$$

- (b) Clearly there is only one way to match all the sandwiches correctly, so $\Pr(X = r \mid r = n) = 1/n!$. Also, note that it is impossible to match all but one sandwich, so $\Pr(X = r \mid r = n - 1) = 0$. Only the cases for $r \leq n - 2$ are nontrivial. Using a similar argument as part (a), we see that for any set of m sandwiches, the probability that at least one was correctly distributed is

$$\Pr\left(\bigcup_{k=1}^m A_k\right) = \sum_{k=1}^m (-1)^{k-1} \frac{(n - k)!}{n!}$$

and that the probability that *any* set of m sandwiches contained at least one correct match is

$$\begin{aligned} \sum_{k=1}^m (-1)^{k-1} \binom{n}{k} \frac{(n - k)!}{n!} &= \sum_{k=1}^m (-1)^{k-1} \frac{n!}{(n - k)!k!} \frac{(n - k)!}{n!} \\ &= \sum_{k=1}^m \frac{(-1)^{k-1}}{k!} = 1 + \sum_{k=2}^m \frac{(-1)^{k-1}}{k!} = 1 - \sum_{k=2}^m \frac{(-1)^k}{k!} \end{aligned}$$

So for $m \geq 2$, the probability of *no* correct matches is $\sum_{k=2}^m \frac{(-1)^k}{k!}$ if $n \geq 2$, and of course 0 if $n = 1$. Therefore the probability of r matches is the probability of any one set of r sandwiches all matching and none of the remaining $n - r$ sandwiches matching times the number of sets of r sandwiches; that is,

$$\begin{aligned} \Pr(X = r \mid r \leq n - 2) &= \binom{n}{r} \cdot \frac{(n - r)!}{n!} \cdot \left(\sum_{k=2}^{n-r} \frac{(-1)^k}{k!} \right) = \frac{r!}{(n - r)!r!} \cdot \frac{(n - r)!}{n!} \sum_{k=2}^{n-r} \frac{(-1)^k}{k!} \\ &= \frac{1}{r!} \sum_{k=2}^{n-r} \frac{(-1)^k}{k!} \end{aligned}$$

Therefore we have

$$\Pr(X = r) = \begin{cases} \frac{1}{r!} \sum_{k=2}^{n-r} \frac{(-1)^k}{k!} & r \leq n-2 \\ 0 & r = n-1 \\ \frac{1}{r!} & r = n \end{cases}$$

(c)

$$\mathbb{E}(X) = \mathbb{E}\left(\sum_{k=1}^n A_k\right) = \sum_{k=1}^n \mathbb{E}(A_k) = \sum_{k=1}^n \Pr(A_k = 1) = n \cdot \frac{1}{n} = \boxed{1}$$

(d)

$$\text{Var}(X) = \mathbb{E}(X^2) - \mathbb{E}(X)^2$$

$$\mathbb{E}(X^2) = \mathbb{E}\left(\sum_{k=1}^n A_k\right)^2 = \mathbb{E}\left(\sum_{k=1}^n A_k^2 + 2 \sum_{1 \leq i < j \leq n} A_i A_j\right) = \sum_{k=1}^n \mathbb{E}(A_k^2) + 2 \sum_{1 \leq i < j \leq n} \mathbb{E}(A_i A_j)$$

Because

$$\mathbb{E}(A_k^2) = 1^2 \cdot \Pr(A_k = 1) = \frac{1}{n}$$

$$\mathbb{E}(A_i A_j) = \Pr(A_i = 1 \cap A_j = 1) = \frac{(n-2)!}{n!} = \frac{1}{n(n-1)} = \frac{1}{2} \cdot \binom{n}{2}^{-1}$$

we have

$$\mathbb{E}(X^2) = \sum_{k=1}^n \frac{1}{n} + 2 \sum_{1 \leq i < j \leq n} \frac{1}{n(n-1)} = 1 + 2 \cdot \binom{n}{2} \cdot \frac{1}{2} \cdot \binom{n}{2}^{-1} = 2$$

$$\implies \text{Var}(X) = \mathbb{E}(X^2) - \mathbb{E}(X)^2 = 2 - 1 = \boxed{1}$$

HW3 Problem 2(5). Verify: $\mathbb{E}(X | Y) = \mathbb{E}(X)$ if X and Y are independent.**Solution.** X and Y are independent if and only if

$$\Pr(X \cap Y) = \Pr(X) \cdot \Pr(Y) \iff \Pr(X = x \cap Y = y) = \Pr(X = x) \Pr(Y = y)$$

$$\iff \Pr(X = x | Y = y) \cdot \Pr(Y = y) = \Pr(X = x) \Pr(Y = y) \iff \Pr(X = x | Y = y) = \Pr(X = x)$$

$$\implies \mathbb{E}(X | Y) = \sum_x x \cdot \Pr(X = x | Y = y) = \sum_x x \cdot \Pr(X = x) = \mathbb{E}(X)$$

HW3 Problem 2 (parts 1 - 4). Verify:

$$(1) \quad \mathbb{E}(\mathbb{E}(X | Y)) = \mathbb{E}(X)$$

- (2) $\mathbb{E}(g(Y)X | Y) = g(Y)\mathbb{E}(X | Y)$
 (3) $\text{Cov}(\mathbb{E}(X | Y), Y) = \text{Cov}(X, Y)$
 (4) Y and $X - \mathbb{E}(X | Y)$ are uncorrelated.

Solution.

(1)

$$\begin{aligned}\mathbb{E}(\mathbb{E}(X | Y)) &= \sum_y \mathbb{E}(X | Y) \Pr(Y = y) = \sum_y \left[\sum_x x \cdot \Pr(X = x | Y = y) \Pr(Y = y) \right] \\ &= \sum_y \left[\sum_x x \cdot \Pr(X = x \cap Y = y) \right] = \sum_y \left[\sum_x x \cdot \Pr(Y = y | X = x) \cdot \Pr(X = x) \right] \\ &= \sum_x \left[x \cdot \Pr(X = x) \cdot \sum_y (\Pr(Y = y | X = x)) \right] = \sum_x \left[x \cdot \Pr(X = x) \cdot 1 \right] \\ &= \mathbb{E}(X)\end{aligned}$$

(2) 2

(3)

$$\begin{aligned}\text{Cov}(\mathbb{E}(X | Y), Y) &= \mathbb{E} \left(\left[\mathbb{E}(X | Y) - \mathbb{E}(\mathbb{E}(X | Y)) \right] \left[Y - \mathbb{E}(Y) \right] \right) \\ &= \mathbb{E} \left(\left[\mathbb{E}(X | Y) - \mathbb{E}(X) \right] \left[Y - \mathbb{E}(Y) \right] \right) = \mathbb{E} \left(\mathbb{E}(X | Y)Y - \mathbb{E}(X)Y - \mathbb{E}(X | Y)\mathbb{E}(Y) + \mathbb{E}(X)\mathbb{E}(Y) \right) \\ &= \mathbb{E}(\mathbb{E}(X | Y)Y) - \mathbb{E}(X)\mathbb{E}(Y) - \mathbb{E}(Y)\mathbb{E}(\mathbb{E}(X | Y)) + \mathbb{E}(X)\mathbb{E}(Y) = \mathbb{E}(\mathbb{E}(X | Y)Y) - \mathbb{E}(Y)\mathbb{E}(X) \\ &= \mathbb{E}(XY) - \mathbb{E}(X)\mathbb{E}(Y) = \text{Cov}(X, Y)\end{aligned}$$

(4) Y and $X - \mathbb{E}(X | Y)$ are uncorrelated if and only if $\text{Cov}(Y, X - \mathbb{E}(X | Y)) = 0 \iff \mathbb{E}(Y \cdot [X - \mathbb{E}(X | Y)]) - \mathbb{E}(Y)\mathbb{E}(X - \mathbb{E}(X | Y)) = 0$.

$$\begin{aligned}\mathbb{E}(Y \cdot [X - \mathbb{E}(X | Y)]) - \mathbb{E}(Y)\mathbb{E}(X - \mathbb{E}(X | Y)) &= \mathbb{E}(YX - Y\mathbb{E}(X | Y)) - \mathbb{E}(Y)\mathbb{E}(X) + \mathbb{E}(Y)\mathbb{E}(\mathbb{E}(X | Y)) \\ &= \mathbb{E}(YX) - \mathbb{E}(Y\mathbb{E}(X | Y)) - \mathbb{E}(Y)\mathbb{E}(X) + \mathbb{E}(Y)\mathbb{E}(X) = \mathbb{E}(YX) - \mathbb{E}(YX) = 0\end{aligned}$$

Spring 2018 Problem 2 (did not complete)

2. Consider positions 1 to n arranged in a circle, so that 2 comes after 1, 3 comes after 2, ..., n comes after $n - 1$, and 1 comes after n . Similarly, take 1 to n as values, with cyclic order, and consider all $n!$ ways to assign values to positions, bijectively, with all $n!$ possibilities equally likely. For $i = 1$ to n , let X_i be the indicator that position i and the one following are filled in with two consecutive values in increasing order, and define

$$S_n = \sum_{i=1}^n X_i, \quad T_n = \sum_{i=1}^n iX_i$$

For example, with $n = 6$ and the circular arrangement 314562, we get $X_3 = 1$ since 45 are consecutive in increasing order, and similarly $X_4 = X_6 = 1$, so that $S_6 = 3, T_6 = 13$.

- a) Compute the mean and the variance of S_n .
- b) Compute the mean and the variance of T_n .

Fall 2008 Problem 2 (HW1 Problem 10). Consider a lottery with n^2 tickets, of which only n tickets win prizes. Let p_n be the probability that, out of n randomly selected tickets, at least one wins a prize. Compute $\lim_{n \rightarrow \infty} p_n$.

Solution. There are $\binom{n^2}{n}$ possible sets of n tickets. The number of these sets that do not contain at least one winner (that is, they only contain members of the $n^2 - n$ losing tickets) is $\binom{n^2 - n}{n}$. Therefore the probability of selecting a set of n tickets that contains at least one winner is

$$\begin{aligned} p_n &= 1 - \binom{n^2 - n}{n} / \binom{n^2}{n} = 1 - \frac{(n^2 - n)!}{n!(n^2 - n - n)!} / \frac{(n^2)!}{(n^2 - n)!n!} = 1 - \frac{(n^2 - n)!}{n!(n^2 - 2n)!} \cdot \frac{(n^2 - n)!n!}{(n^2)!} \\ &= 1 - \frac{(n^2 - n)!}{(n^2 - 2n)!} \cdot \frac{(n^2 - n)!}{(n^2)!} = 1 - \prod_{i=0}^{n-1} (n^2 - n - i) / \prod_{i=0}^{n-1} (n^2 - i) = 1 - \prod_{i=0}^{n-1} \frac{n^2 - n - i}{n^2 - i} \\ &= 1 - \prod_{i=0}^{n-1} \left(\frac{n^2 - i}{n^2 - i} - \frac{n}{n^2 - i} \right) = 1 - \prod_{i=0}^{n-1} \left(1 - \frac{n}{n^2 - i} \right) \end{aligned}$$

Therefore

$$\begin{aligned} \lim_{n \rightarrow \infty} p_n &= \lim_{n \rightarrow \infty} \left[1 - \prod_{i=0}^{n-1} \left(1 - \frac{n}{n^2 - i} \right) \right] = 1 - \lim_{n \rightarrow \infty} \prod_{i=0}^n \left(1 - \frac{n}{n^2 - i} \right) = 1 - \lim_{n \rightarrow \infty} \prod_{i=0}^n \left(1 - \frac{n \cdot \frac{1}{n}}{\frac{n^2}{n} - \frac{i}{n}} \right) \\ &= 1 - \lim_{n \rightarrow \infty} \prod_{i=0}^n \left(1 - \frac{1}{n - \frac{i}{n}} \right) = 1 - \lim_{n \rightarrow \infty} \prod_{i=0}^n \left(1 - \frac{1}{n} \right) = 1 - \lim_{n \rightarrow \infty} \left(1 - \frac{1}{n} \right)^n = \boxed{1 - \exp(-1)} \end{aligned}$$

6.2.3 Problems we did on homework

Fall 2017 Problem 2 (Homework 3 Problem 6). An urn contains $2n$ balls, coming in pairs: two balls are labeled “1”, two balls are labeled “2”, ..., two balls are labeled “ n ”. A sample of size n is taken without replacement. Denote by N the number of pairs in the sample. Compute the expected value and the variance of N . You do not need to simplify the expression for the variance.

Solution. Let X_k be an indicator variable for both balls labeled k being in the sample. Note that

$$\mathbb{E}(X_k) = \Pr(X_k = 1) = \frac{\binom{2n-2}{n-2}}{\binom{2n}{n}} = \frac{(2n-2)!}{(n-2)!n!} / \frac{(2n)!}{n!n!} = \frac{(2n-2)!n!}{(2n)!(n-2)!} = \frac{n(n-1)}{2n(2n-1)} = \frac{n-1}{2(2n-1)}$$

Now since $N = \sum_{k=1}^n X_k$, we have

$$\mathbb{E}(N) = \mathbb{E}\left(\sum_{k=1}^n X_k\right) = \sum_{k=1}^n \mathbb{E}(X_k) = \boxed{\frac{n(n-1)}{2(2n-1)}}$$

To obtain the variance, note that

$$\mathbb{E}(N^2) = \mathbb{E}\left(\sum_{k=1}^n X_k\right)^2 = \mathbb{E}\left(\sum_{k=1}^n X_k^2 + 2 \sum_{1 \leq i < j \leq n} X_i X_j\right) = \sum_{k=1}^n \mathbb{E}(X_k^2) + 2 \sum_{1 \leq i < j \leq n} \mathbb{E}(X_i X_j)$$

Because

$$\mathbb{E}(X_k^2) = 1^2 \cdot \Pr(X_k = 1) = \mathbb{E}(X_k) = \frac{n-1}{2(2n-1)}$$

$$\mathbb{E}(X_i X_j) = \Pr(X_i = 1 \cap X_j = 1) = \frac{\binom{2n-4}{n-4}}{\binom{2n}{n}} = \frac{(2n-4)!}{(n-4)!n!} / \frac{(2n)!}{n!n!} = \frac{(2n-4)!n!}{(2n)!(n-4)!}$$

$$= \frac{n(n-1)(n-2)(n-3)}{2n(2n-1)(2n-2)(2n-3)} = \frac{(n-1)(n-2)(n-3)}{2(2n-1)(2n-2)(2n-3)}$$

we have

$$\begin{aligned} \mathbb{E}(N^2) &= \sum_{k=1}^n \frac{n-1}{2(2n-1)} + 2 \sum_{1 \leq i < j \leq n} \frac{(n-1)(n-2)(n-3)}{2(2n-1)(2n-2)(2n-3)} = \frac{n(n-1)}{2(2n-1)} + 2 \binom{n}{2} \frac{(2n-4)!n!}{(2n)!(n-4)!} \\ &= \frac{n(n-1)}{2(2n-1)} + \frac{n!}{(n-2)!} \cdot \frac{(2n-4)!n!}{(2n)!(n-4)!} = \frac{n(n-1)}{2(2n-1)} + n(n-1) \cdot \frac{(n-1)(n-2)(n-3)}{2(2n-1)(2n-2)(2n-3)} \\ &= \frac{n(n-1)}{2(2n-1)} + \frac{n(n-1)^2(n-2)(n-3)}{2(2n-1)(2n-2)(2n-3)} \\ \implies \text{Var}(N) &= \mathbb{E}(N^2) - \mathbb{E}(N)^2 = \boxed{\frac{n(n-1)}{2(2n-1)} + \frac{n(n-1)^2(n-2)(n-3)}{2(2n-1)(2n-2)(2n-3)} - \frac{n^2(n-1)^2}{4(2n-1)^2}} \end{aligned}$$

Fall 2017 Problem 3 (HW3 Problem 8—almost full solution)

Let U_1, U_2, \dots be iid random variables, uniformly distributed on $[0, 1]$, and let N be a Poisson random variable with mean value equal to 1. Assume that N is independent of U_1, U_2, \dots and define

$$Y = \begin{cases} 0 & \text{if } N = 0 \\ \max_{1 \leq i \leq N} U_i & \text{if } N > 0 \end{cases}$$

Compute the expected value of Y .

Solution. Since Y is a function of N , let $Y = y(N)$. By the Law of the Unconscious Statistician,

$$\mathbb{E}(Y) = \mathbb{E}(\mathbb{E}(Y | N)) = \mathbb{E}(\mathbb{E}(\max_{1 \leq i \leq N} U_i | N = n))$$

Let $Z_n = \max_{1 \leq i \leq n} U_i$. The cdf of Z_n can be calculated as follows:

$$\Pr(Z_n \leq x) = \Pr(\max_{1 \leq i \leq n} U_i \leq x) = \Pr(U_1 \leq x \cap U_2 \leq x \cap \dots \cap U_n \leq x) = x^n$$

for $x \in [0, 1]$. Therefore the pdf of Z_n is its derivative, nx^{n-1} . So we have

$$\mathbb{E}(\max_{1 \leq i \leq N} U_i | N = n) = \mathbb{E}(Z_n) = \int_0^1 x n x^{n-1} dx = n \int_0^1 x^n dx = n \frac{x^{n+1}}{n+1} \Big|_0^1 = \frac{n}{n+1}$$

Plugging this into the expression for $\mathbb{E}(Y)$ yields

$$\begin{aligned} \mathbb{E}(Y) &= \mathbb{E}(\mathbb{E}(Y | N)) = \sum_{n=0}^{\infty} \frac{n}{n+1} \Pr(N = n) = \sum_{n=1}^{\infty} \frac{n}{n+1} \frac{\exp(-1) 1^n}{n!} \\ &= \frac{1}{e} \sum_{n=1}^{\infty} \frac{n+1-1}{(n+1)!} = \frac{1}{e} \left(\sum_{n=1}^{\infty} \frac{n+1}{(n+1)!} - \sum_{n=1}^{\infty} \frac{1}{(n+1)!} \right) = \frac{1}{e} \left(\sum_{n=1}^{\infty} \frac{1}{n!} - \sum_{m=2}^{\infty} \frac{1}{m!} \right) \\ &= \frac{1}{e} [e - 1 - (e - 1 - 1)] = \boxed{\frac{1}{e}} \end{aligned}$$

Fall 2013 Problem 3/Spring 2011 Problem 2 (HW3 Problem 9; coupon collector problem)
Only parts I didn't do: Let D be the event that no box receives more than 1 ball. Fix $a \in (0, 1)$. If both $n, d \rightarrow \infty$ together, what relation must they satisfy in order to have $\Pr(D) \rightarrow a$?

HW3 Problem 9. Consider n (different) balls placed at random in m boxes so that each of m^n configurations is equally likely.

- (a) Compute the expected value and the variance of the number of empty boxes.

- (b) Show that if $\lim_{m,n \rightarrow \infty} m \exp(-n/m) = \lambda \in (0, \infty)$, then, in the same limit, the number of empty boxes has Poisson distribution with parameter λ .
- (c) For $k \geq 1$ such that $k+3 \leq m$, define the event A_k that the boxes $k, k+1, k+2, k+3$ are empty. Assuming that $m > 8$, compute $\Pr(A_1 \cup A_3 \cup A_5)$. How will the answer change if $m = 8$?
- (d) Now imagine that the balls are dropped one-by-one (with each ball equally likely to go into any of the m boxes, independent of all other balls), and denote by N_m the minimal number of balls required to fill all the boxes. Compute $\mathbb{E}(N_m)$, $\text{Var}(N_m)$ and

$$\lim_{m \rightarrow \infty} \Pr\left(\frac{N_m - m \log m}{m} \leq x\right)$$

- (e) Suppose we instead place an unlimited number of balls into the m boxes until we have k consecutive balls land in the same box (it doesn't matter which box). What is the expected number of balls we will drop until this happens?

Solution.

- (a) Let A_i be the event that the i th box is empty. Let $\mathbf{1}_{A_i}$ be the indicator for A_i . Then $X = \sum_{i=1}^m \mathbf{1}_{A_i}$.

$$\mathbb{E}(X) = \mathbb{E}\left(\sum_{i=1}^m \mathbf{1}_{A_i}\right) = \sum_{i=1}^m (\mathbb{E}\mathbf{1}_{A_i}) = \sum_{i=1}^m \Pr(A_i) = \sum_{i=1}^m \left(\frac{m-1}{m}\right)^n = \boxed{\frac{(m-1)^n}{m^{n-1}}}$$

$$\text{Var}(X) = \text{Var}\left(\sum_{i=1}^m \mathbf{1}_{A_i}\right) = \sum_{i=1}^m \text{Var}(\mathbf{1}_{A_i}) + 2 \sum_{1 \leq i < j \leq m} \text{Cov}(\mathbf{1}_{A_i}, \mathbf{1}_{A_j})$$

$$\text{Var}(\mathbf{1}_{A_i}, \mathbf{1}_{A_j}) = \mathbb{E}(\mathbf{1}_{A_i} \mathbf{1}_{A_j}) - \mathbb{E}(\mathbf{1}_{A_i})^2 = \Pr(A_i \cap A_j) - \Pr(A_i)^2 = \left(\frac{m-1}{m}\right)^n - \left(\frac{m-1}{m}\right)^{2n}$$

$$\text{Cov}(\mathbf{1}_{A_i}, \mathbf{1}_{A_j}) = \mathbb{E}(\mathbf{1}_{A_i} \mathbf{1}_{A_j}) - \mathbb{E}(\mathbf{1}_{A_i})\mathbb{E}(\mathbf{1}_{A_j}) = \Pr(A_i \cap A_j) - \Pr(A_i)\Pr(A_j) = \left(\frac{m-2}{m}\right)^n - \left(\frac{m-1}{m}\right)^{2n}$$

$$\begin{aligned} \implies \text{Var}(X) &= m \cdot \left[\left(\frac{m-1}{m}\right)^n - \left(\frac{m-1}{m}\right)^{2n} \right] + \frac{m!}{(m-2)!} \left[\left(\frac{m-2}{m}\right)^n - \left(\frac{m-1}{m}\right)^{2n} \right] \\ &= \frac{(m-1)^n}{m^{n-1}} - \frac{(m-1)^{2n}}{m^{2n-1}} + (m^2 - m) \left[\left(\frac{m-2}{m}\right)^n - \left(\frac{m-1}{m}\right)^{2n} \right] \end{aligned}$$

$$\boxed{\text{Var}(X) = \frac{(m-1)^n}{m^{n-1}} - \frac{(m-1)^{2n}}{m^{2n-1}} + (m-1) \left[\frac{(m-2)^n}{m^{n-1}} - \frac{(m-1)^{2n}}{m^{2n-1}} \right]}$$

- (b) Note that

$$X = \sum_{i=1}^m \mathbf{1}_{A_i}$$

and that the A_i are only weakly dependent on each other, especially as m and n increase. Therefore as $m, n \rightarrow \infty$, the Poisson paradigm (see Section 6.1.8) suggests $X \sim \text{Poisson}(\mathbb{E}(X))$. We have

$$\mathbb{E}(X) = \frac{(m-1)^n}{m^{n-1}}$$

so

$$\begin{aligned} \lim_{n,m \rightarrow \infty} \mathbb{E}(X) &= \lim_{n,m \rightarrow \infty} m \cdot \left(\frac{m-1}{m} \right)^n = \lim_{n,m \rightarrow \infty} m \cdot \left(1 - \frac{1}{m} \right)^n = \lim_{n,m \rightarrow \infty} m \cdot \left[\left(1 - \frac{1}{m} \right)^m \right]^{n/m} \\ &\approx \lim_{n,m \rightarrow \infty} m \cdot [e^{-1}]^{n/m} = \lim_{n,m \rightarrow \infty} m e^{-n/m} \end{aligned}$$

Using

$$\lim_{m,n \rightarrow \infty} m \exp(-n/m) = \lambda \in (0, \infty)$$

we have $\boxed{X \sim \text{Poisson}(\lambda) \text{ as } m, n \rightarrow \infty}$.

(c)

$$\Pr(A_1 \cup A_3 \cup A_5) = \Pr(A_1) + \Pr(A_3) + \Pr(A_5) - \Pr(A_1 \cap A_3) - \Pr(A_1 \cap A_5) - \Pr(A_3 \cap A_5) + \Pr(A_1 \cap A_3 \cap A_5)$$

We have

$$\Pr(A_1) = \Pr(A_3) = \Pr(A_5) = \left(\frac{m-4}{m} \right)^n$$

$$\Pr(A_1 \cap A_3) = \Pr(A_3 \cap A_5) = \left(\frac{m-6}{m} \right)^n$$

$$\Pr(A_1 \cap A_5) = \Pr(A_1 \cap A_3 \cap A_5) = \left(\frac{m-8}{m} \right)^n$$

Therefore

$$\Pr(A_1 \cup A_3 \cup A_5) = 3 \left(\frac{m-4}{m} \right)^n - 2 \left(\frac{m-6}{m} \right)^n = \boxed{\frac{3(m-4)^n - 2(m-6)^n}{m^n}}$$

- (d) N_m is the minimal number of balls required to fill all the boxes. Let T_i be the number of balls that have to be dropped to fill the i th box after $i-1$ boxes have been filled. The probability of filling a new box after $i-1$ boxes have been filled is $\frac{m-(i-1)}{m}$. Therefore T_i has a geometric distribution with $E(T_i) = \frac{m}{m-(i-1)}$. Since $N_m = \sum_{i=1}^m T_i$, we have

$$\mathbb{E}(N_m) = \mathbb{E} \left(\sum_{i=1}^m T_i \right) = \sum_{i=1}^m \mathbb{E}(T_i) = \sum_{i=1}^m \frac{m}{m-(i-1)} = \boxed{m \sum_{i=1}^m \frac{1}{i}}$$

Because the T_i are independent, we have

$$\text{Var}(N_m) = \text{Var} \left(\sum_{i=1}^m T_i \right) = \sum_{i=1}^m \text{Var}(T_i) = \sum_{i=1}^m \left(1 - \frac{m-(i-1)}{m} \right) \left/ \left(\frac{m-(i-1)}{m} \right)^2 \right.$$

$$= \sum_{i=1}^m \frac{i-1}{m} \cdot \left(\frac{m}{m-(i-1)} \right)^2 = \boxed{m \sum_{i=1}^m \frac{i-1}{[m-(i-1)]^2}}$$

Finally, to find

$$\lim_{m \rightarrow \infty} \Pr \left(\frac{N_m - m \log m}{m} \leq x \right)$$

begin by noting that we can also express N_m as

$$\Pr(N_m \leq k) = \Pr(X_{m,k} = 0)$$

where $X_{m,k}$ is defined as X is in part (b) with k being the number of balls that have been dropped so far, $k \in \mathbb{N} \geq m$. (For $k < m$, $\Pr(N_m \leq k) = 0$.)

Again, let $A_{i,k}$ be the event that the i th box is empty after dropping k balls. Then because $X_{m,k} = \sum_{i=1}^m \mathbf{1}_{A_{i,k}}$ and the $A_{i,k}$ are only weakly dependent on each other (especially as m becomes large), the Poisson paradigm (see Section 6.1.8) again suggests that as $m \rightarrow \infty$, $X_{m,k} \sim \text{Poisson}(\lambda_k)$ where $\lambda_k = \mathbb{E}(X_{m,k})$ is defined as above. Therefore we have

$$\begin{aligned} \lim_{m \rightarrow \infty} \Pr \left(\frac{N_m - m \log m}{m} \leq x \right) &= \lim_{m \rightarrow \infty} \Pr(N_m \leq xm + m \log m) = \lim_{m \rightarrow \infty} \Pr(X_{m,xm+m \log m} \\ &= 0) \approx \frac{\exp(-\lambda_{xm+m \log m}) \cdot \lambda_{xm+m \log m}^0}{0!} = \exp(-\lambda_{xm+m \log m}) \end{aligned}$$

And we have

$$\begin{aligned} \lambda_{xm+m \log m} &= \lim_{m \rightarrow \infty} m \exp \left(-\frac{xm + m \log m}{m} \right) = \lim_{m \rightarrow \infty} m \exp(-x - \log m) = \lim_{m \rightarrow \infty} m/m \exp(-x) \\ &= \exp(-x) \end{aligned}$$

which yields

$$\boxed{\lim_{m \rightarrow \infty} \Pr \left(\frac{N_m - m \log m}{m} \leq x \right) = \exp(\exp(-x))}$$

- (e) Let $N = N_k$ be the number of balls that are dropped until k consecutive balls land in the same box, and likewise for N_{k-1} . Suppose we have already observed $k-1$ consecutive outcomes (of any kind) in N_{k-1} trials. Then we finish on the next term (by having another consecutive outcome) with probability $1/m$. Otherwise we have a different outcome and then repeat the same process again. So we have

$$\mathbb{E}(N_k | N_{k-1}) = N_{k-1} + 1 \cdot \frac{1}{m} + \mathbb{E}(N_k) \cdot \left(1 - \frac{1}{m} \right)$$

Therefore

$$\mathbb{E}(N) = \mathbb{E}(N_k) = \mathbb{E}[\mathbb{E}(N_k \mid N_{k-1})] = \mathbb{E}(N_{k-1}) + \frac{1}{m} + \left(1 - \frac{1}{m}\right)\mathbb{E}(N_k)$$

$$\iff \frac{1}{m}\mathbb{E}(N_k) = \mathbb{E}(N_{k-1}) + \frac{1}{m} \iff \mathbb{E}(N_k) = m\mathbb{E}(N_{k-1}) + 1$$

We have a recursive formula. Note that $\mathbb{E}(N_1) = 1$ because the number of trials until there is 1 consecutive outcome of any kind is simply 1. We can then calculate as follows:

$$\mathbb{E}(N_2) = m\mathbb{E}(N_{2-1}) + 1 = m + 1$$

$$\mathbb{E}(N_3) = m\mathbb{E}(N_{3-1}) + 1 = m(m + 1) + 1 = 1 + m + m^2$$

$$\mathbb{E}(N_4) = m\mathbb{E}(N_{4-1}) + 1 = m(1 + m + m^2) + 1 = 1 + m + m^2 + m^3$$

⋮

$$\mathbb{E}(N_k) = \sum_{i=0}^{k-1} m^i = \frac{1 \cdot (1 - m^k)}{1 - m} = \boxed{\frac{m^k - 1}{m - 1}}$$

Fall 2012 Problem 1 (HW2 Problem 10/HW 1 Problem 9) Only part I didn't do: Find the mean and variance of $S_n = X_1 + \dots + X_n$, the total number of white balls added to the urn up to time n .

HW1 Problem 9. An urn contains b black and w white balls. At each step, a ball is removed from the urn at random and then put back together with one more ball of the same color. Compute the probability p_n to get a black ball on step n , $n \geq 1$.

Solution. Step 1:

$$p_1 = \frac{b}{b+w}$$

Step 2: We need to separately consider the cases where a black ball was selected on step 1 (with probability p_1) or a white ball (with probability $1 - p_1$).

$$\begin{aligned} p_2 &= p_1 \cdot \frac{b+1}{b+w+1} + (1-p_1) \cdot \frac{b}{b+w+1} = p_1 \left(\frac{b+1}{b+w+1} - \frac{b}{b+w+1} \right) + \frac{b}{b+w+1} \\ &= p_1 \left(\frac{1}{b+w+1} + \frac{1}{p_1} \frac{b}{b+w+1} \right) = p_1 \left(\frac{1}{b+w+1} + \frac{b+w}{b} \frac{b}{b+w+1} \right) \\ &= p_1 \left(\frac{b+w+1}{b+w+1} \right) = p_1 \end{aligned}$$

$$\implies p_2 = p_1 = \frac{b}{b+w}$$

Step 3: Regardless of the previous steps, there are now $b + w + 2$ balls in the urn. Since we know that $p_1 = p_2$, the probability that we have selected k black balls so far (and thus, the probability that there are currently $b + k$ black balls in the urn) is given by

$$\begin{aligned} \Pr(k \text{ balls chosen in first 2 rounds}) &= \binom{2}{k} p_1^k (1 - p_1)^{2-k} = \binom{2}{k} \left(\frac{b}{b+w}\right)^k \left(\frac{w}{b+w}\right)^{2-k} \\ &= \binom{2}{k} \frac{b^k w^{2-k}}{(b+w)^2} \end{aligned}$$

for $k \in \{0, 1, 2\}$. Given that we have selected k black balls so far, the probability of selecting a black ball this time is $\frac{b+k}{b+w+2}$. Therefore the probability of selecting a black ball this round is

$$\begin{aligned} p_3 &= \sum_{k=0}^2 \binom{2}{k} \frac{b^k w^{2-k}}{(b+w)^2} \frac{b+k}{b+w+2} = \frac{1}{(b+w+2)(b+w)^2} \sum_{k=0}^2 \binom{2}{k} (b+k) b^k w^{2-k} \\ &= \frac{1}{(b+w+2)(b+w)^2} \left(\binom{2}{0} b w^2 + \binom{2}{1} (b+1) b w + \binom{2}{2} (b+2) b^2 \right) \\ &= \frac{bw^2 + 2(b+1)bw + (b+2)b^2}{(b+w+2)(b+w)^2} = \frac{b}{b+w} \left(\frac{w^2 + 2bw + 2w + b^2 + 2b}{b^2 + bw + 2b + wb + w^2 + 2w} \right) \\ &= \frac{b}{b+w} \left(\frac{w^2 + 2bw + 2w + b^2 + 2b}{b^2 + 2bw + 2b + w^2 + 2w} \right) = \frac{b}{b+w} = p_1 \end{aligned}$$

There seems to be a clear pattern here. Let's find the general formula by induction.

Step $n+1$: Assume that the probability of choosing a black ball on steps $1, 2, \dots, n$ was $\frac{b}{b+w}$ each time.
(a bunch of boring stuff, then it worked.)

HW2 Problem 10. Random variables (X_1, \dots, X_n) are called *exchangeable* if $\Pr(X_1 = x_1, \dots, X_n = x_n) = \Pr(X_{\tau(1)} = x_1, \dots, X_{\tau(n)} = x_n)$ for all real numbers x_1, \dots, x_n and every permutation τ of the set $\{1, \dots, n\}$. In the setting of Problem 9 from Homework 1, let $X_k = 1$ if a white ball is drawn on step k , and $X_k = 0$ otherwise. Show that the random variables X_1, \dots, X_n are exchangeable for every $n \geq 2$.

Solution. For $n = 2$: There are two cases which we must show are equal to show exchangeability:

$$\Pr(X_1 = 0, X_2 = 1) = \Pr(X_1 = 1, X_2 = 0)$$

First,

$$\Pr(X_1 = 0, X_2 = 1) = \Pr(\text{black first}) \Pr(\text{white second} \mid \text{black first}) = \left(\frac{b}{b+w} \right) \left(\frac{w}{b+w+1} \right)$$

$$\left(\frac{w}{b+w} \right) \left(\frac{b}{b+w+1} \right) = \Pr(X_1 = 1, X_2 = 0)$$

which proves exchangeability for $n = 2$. In the general case, we seek to show that X_1, \dots, X_n are exchangeable. That is, in all $n+1$ unordered sets $\mathbb{X}_k = \{x_{1k}, x_{2k}, \dots, x_{nk} \mid x_{ik} \in \{0, 1\}, \sum_i x_{ik} = k\}$, in all $\binom{n}{k}$ permutations of \mathbb{X}_k ,

$$\Pr(\mathbb{X}_{kj} = \Pr(\mathbb{X}_{kj'})$$

where j and j' denote different permutations of \mathbb{X}_k . That is,

$$\Pr(X_1 = x_{1k}, X_2 = x_{2k}, \dots, X_n = x_{nk}) = \Pr(X_{j_1} = x_{1k}, X_{j_2} = x_{2k}, \dots, X_{j_n} = x_{nk})$$

where j_1, j_2, \dots, j_n index the permuted variables. Consider \mathbb{X}_{kj^*} where all k white balls are chosen first and all $n-k$ black balls are chosen last. We have

$$\Pr(\mathbb{X}_{kj^*}) = \prod_{i=1}^k \left(\frac{w+i-1}{b+w+i-1} \right) \cdot \prod_{i=k+1}^n \left(\frac{b+i-k-1}{b+w+i-1} \right)$$

$$= \prod_{i=1}^n \left(\frac{1}{b+w+i-1} \right) \cdot \left[\prod_{i=1}^k (w+i-1) \prod_{i=k+1}^n (b+i-k-1) \right] = \prod_{i=1}^n \left(\frac{1}{b+w+i-1} \right) \cdot \left[\prod_{i=1}^k (w+i-1) \prod_{i'=1}^{n-k} (b+i'-1) \right]$$

It is easy to see that the leftmost product will always equal the product of the denominators, regardless of the permutation, since one ball is added to the urn after every draw. Similarly, regardless of permutation, the numerator of the probability of drawing the i th white ball will always equal $w+i-1$, the number of white balls already in the urn. Likewise, the numerator of the probability of drawing the i' th black ball is always $b+i'-1$. Because multiplication is commutative, all permutations of these numbers will have equal products. Therefore $\Pr(\mathbb{X}_{kj^*}) = \Pr(\mathbb{X}_{kj})$ for all k . That is,

$$\Pr(X_1 = x_1, \dots, X_n = x_n) = \Pr(X_{\tau(1)} = x_1, \dots, X_{\tau(n)} = x_n)$$

for all $(x_1, \dots, x_n) \in \mathbb{R}^n$, all $n \in \mathbb{Z}$ such that $n \geq 2$, all permutations τ .

Homework 2 Problem 2. Consider the function

$$f(x) = \begin{cases} C(2x - x^2) & 0 < x < 2 \\ 0 & \text{otherwise} \end{cases}$$

- (a) Could f be a distribution function? If so, determine C .
 (b) Could f be a probability density function? If so, determine C .

Solution.

(a) If f is a distribution function, $\lim_{x \rightarrow -\infty} f(x) = 0$, $\lim_{x \rightarrow \infty} f(x) = 1$, and $f'(x) \geq 0 \forall x \in \mathbb{R}$. f clearly does not meet the second or third conditions and is therefore not a distribution function.

(b) If f is a density function then $\int_{-\infty}^{\infty} f(x)dx = 1$ and $f(x) \geq 0 \forall x \in \mathbb{R}$.

$$\begin{aligned} \int_{-\infty}^{\infty} f(x)dx &= \int_0^2 C(2x - x^2)dx = C \left[x^2 - \frac{x^3}{3} \right]_0^2 = C \left(4 - \frac{8}{3} - 0 \right) = C \cdot \frac{4}{3} \\ &= 1 \iff C = \frac{3}{4} \end{aligned}$$

Next we check that f is always nonnegative. It equals zero except on $(0, 2)$.

$$\frac{3}{4}(2x - x^2) \geq 0 \iff x(2 - x) \geq 0 \iff x \in (0, 2)$$

Therefore f is nonnegative $\forall x \in \mathbb{R}$, so f is a probability density function if $C = \frac{3}{4}$.

HW1 Problem 8. Two people, A and B , are involved in a duel. The rules are simple: shoot at each other once; if at least one is hit, the duel is over, if both miss, repeat (go to the next round), and so on. Denote by p_A and p_B the probabilities that A hits B and B hits A with one shot, and assume that hitting/missing is independent from round to round. Compute the probabilities of the following events:
 (a) the duel ends and A is not hit; (b) the duel ends and both are hit; (c) the duel ends after round number n ; (d) the duel ends after round number n GIVEN that A is not hit; (e) the duel ends after n rounds GIVEN that both are hit; (f) the duel goes on forever.

Solution.

(a) Let A_k denote the event that the duel is ended by A shooting B in the k th round (with neither person being shot in the first $k - 1$ rounds). Note that $\{A_k | k = 1, 2, \dots\}$ are all mutually exclusive. Therefore the probability of the duel ending without A being hit is $\sum_{k=1}^{\infty} A_k$. Because the probabilities in each round are constant and independent,

$$A_k = (1 - p_A)^{k-1} p_A (1 - p_B)^k$$

So the probability that the duel ends and A is not hit is

$$\sum_{k=1}^{\infty} A_k = \sum_{k=1}^{\infty} (1 - p_A)^{k-1} p_A (1 - p_B)^k = p_A (1 - p_B) \sum_{k=1}^{\infty} (1 - p_A)^{k-1} (1 - p_B)^{k-1}$$

This is an infinite geometric series. Since the ratio $(1 - p_A)(1 - p_B)$ has absolute value less than 1, the sum can be calculated.

$$\sum_{k=1}^{\infty} A_k = p_A (1 - p_B) \cdot \frac{1}{1 - (1 - p_A)(1 - p_B)} = \frac{p_A (1 - p_B)}{p_A + p_B - p_A p_B} = \boxed{\frac{p_A (1 - p_B)}{p_A (1 - p_B) + p_B}}$$

- (b) Similar to part (a). Let C_k denote the event that the duel is ended with both players being shot in the k th round (with neither person being shot in the first $k - 1$ rounds). Again, $\{C_k|k = 1, 2, \dots\}$ are all mutually exclusive, so the probability of the duel ending in these circumstances is $\sum_{k=1}^{\infty} C_k$. We have

$$\begin{aligned} C_k &= (1 - p_A)^{k-1} p_A (1 - p_B)^{k-1} p_B \\ \sum_{k=1}^{\infty} C_k &= \sum_{k=1}^{\infty} (1 - p_A)^{k-1} p_A (1 - p_B)^{k-1} p_B = p_A p_B \sum_{k=1}^{\infty} (1 - p_A)^{k-1} (1 - p_B)^{k-1} \\ &= p_A p_B \cdot \frac{1}{1 - (1 - p_A)(1 - p_B)} = \boxed{\frac{p_A p_B}{p_A + p_B - p_A p_B}} \end{aligned}$$

Note that this value is less than the answer from part (a) if $p_B < \frac{1}{2}$ and greater if $p_B > \frac{1}{2}$

- (c) Let B_k denote the event that the duel is ended by B shooting A in the k th round (with neither person being shot in the first $k - 1$ rounds), with

$$B_k = (1 - p_A)^k p_B (1 - p_B)^{k-1}$$

Let A_k and C_k be defined as above. Note that $\{A_k|k = 1, 2, \dots\}$, $\{B_k|k = 1, 2, \dots\}$, $\{C_k|k = 1, 2, \dots\}$ are all mutually exclusive, and that the event that the duel ends in round n is $\{A_n \cup B_n \cup C_n\}$. So the probability of the duel ending in round n is

$$\begin{aligned} \Pr(A_n \cup B_n \cup C_n) &= \Pr(A_n) + \Pr(B_n) + \Pr(C_n) \\ &= (1 - p_A)^{n-1} p_A (1 - p_B)^n + (1 - p_A)^n p_B (1 - p_B)^{n-1} + (1 - p_A)^{n-1} p_A (1 - p_B)^{n-1} p_B \\ &= (1 - p_A)^{n-1} (1 - p_B)^{n-1} [p_A (1 - p_B) + (1 - p_A) p_B + p_A p_B] \\ &= \boxed{(1 - p_A)^{n-1} (1 - p_B)^{n-1} (p_A + p_B - p_A p_B)} \end{aligned}$$

- (d) Let A_k , B_k , C_k be defined as above. The event that the duel ends at round n without A being hit is given by $\{A_n\}$.

$$\Pr(A_n) = \boxed{(1 - p_A)^{n-1} p_A (1 - p_B)^n}$$

- (e) Let A_k , B_k , C_k be defined as above. The event that the duel ends at round n with both players being hit is given by $\{C_n\}$.

$$\Pr(C_n) = \boxed{(1 - p_A)^{n-1} p_A (1 - p_B)^{n-1} p_B}$$

- (f) Let A_k , B_k , C_k be defined as above. The probability that the duel never ends is equal to 1 - the probability that the duel ends at some point, which is $\{A_k|k = 1, 2, \dots\} \cup \{B_k|k = 1, 2, \dots\} \cup \{C_k|k = 1, 2, \dots\}$. Since all of these events are mutually exclusive, we have

$$1 - \Pr(\{A_k|k = 1, 2, \dots\} \cup \{B_k|k = 1, 2, \dots\} \cup \{C_k|k = 1, 2, \dots\}) = 1 - \sum_{k=1}^{\infty} (A_k + B_k + C_k)$$

$$\begin{aligned}
&= 1 - \sum_{k=1}^{\infty} ((1-p_A)^{k-1} p_A (1-p_B)^k + (1-p_A)^k p_B (1-p_B)^{k-1} + (1-p_A)^{k-1} p_A (1-p_B)^{k-1} p_B) \\
&= 1 - [p_A(1-p_B) + (1-p_A)p_B + p_A p_B] \sum_{k=1}^{\infty} (1-p_A)^{k-1} (1-p_B)^{k-1} \\
&= 1 - [p_A(1-p_A p_B) + p_B - p_A)p_B + p_A p_B] \cdot \frac{1}{1 - (1-p_A)(1-p_B)} \\
&= 1 - \frac{p_A - p_A p_B + p_B - p_A p_B + p_A p_B}{p_A + p_B - p_A p_B} = 1 - \frac{p_A + p_B - p_A p_B}{p_A + p_B - p_A p_B} = \boxed{0}
\end{aligned}$$

Homework 1 Problem 1.

- (I) Seven different gifts are distributed among 10 children. How many different outcomes are possible if every child can receive (a) at most one gift, (b) at most two gifts, (c) any number of gifts?
- (II) Answer the same questions if the gifts are identical (but the children are still different).

Solution.

(I) (a) $\binom{10}{7} 7! = \boxed{604,800}$

(b) Clearly all outcomes that satisfy part (I)(a) also satisfy these conditions, so we start with $\binom{10}{7} 7! = 604,800$ possible outcomes. In addition, the following outcomes are possible:

- (i) **A set of 6 children receive gifts; one child receives two gifts.** There are $\binom{10}{6}$ ways to pick a group of 6 children to receive the gifts. Next, there are $\binom{6}{1} = 6$ ways to choose which child receives two gifts. Finally, there are $7!/2!$ unique ways to distribute the gifts among the children once a particular partition is chosen (since order matters for all of the gifts except for the two that are received by the same child).

- (ii) **A set of 5 children receive gifts; two children receive two gifts.** There are $\binom{10}{5}$ ways to pick a group of 5 children to receive the gifts. Next, there are $\binom{5}{2}$ ways to choose which of these children receive one gift and which receive two. Finally, there are $7!/(2!2!)$ unique ways to distribute the gifts among the children once a particular partition is chosen (since order matters for all of the gifts except for the two batches of two gifts that are received by the same child).

(Note that without the restriction that a child can receive at most two gifts, another possibility is that 1 child could receive 3 gifts, but that wouldn't work in this case.)

- (iii) **A set of 4 children receive gifts; three children each receive two gifts.** There are $\binom{10}{4}$ ways to pick a group of 4 children to receive the gifts. Next, there are $\binom{4}{3} = 4$ ways to choose which of these children receive one gift and which receive two. Finally, there are $7!/(2!2!2!)$ unique ways to distribute the gifts among the children once a particular partition is chosen (since order matters for all of the gifts except for the three batches of two gifts that are received by the same child).

(Again, there are other possibilities for 4 children to receive 7 gifts, but none that satisfy the condition that no child receives more than 2 gifts.)

Clearly each of these outcomes are mutually exclusive. Therefore the answer is

$$\begin{aligned}
 & \binom{10}{7} 7! + \binom{10}{6} \cdot \binom{6}{1} \cdot \frac{7!}{2!} + \binom{10}{5} \cdot \binom{5}{2} \cdot \frac{7!}{2!2!} + \binom{10}{4} \cdot \binom{4}{3} \cdot \frac{7!}{2!2!2!} \\
 &= 7! \cdot \left(\frac{10!}{3!} + \frac{10!}{6!4!} \cdot 6 \cdot \frac{1}{2} + \frac{10!}{5!5!} \cdot \frac{5!}{3!2!} \cdot \frac{1}{4} + \frac{10!}{4!6!} \cdot \frac{4!}{3!} \cdot \frac{1}{8} \right) \\
 &= 7!10! \cdot \left(\frac{1}{3!} + \frac{1}{6!4!} \cdot \frac{6}{2} + \frac{1}{5!} \cdot \frac{1}{3!2!} \cdot \frac{1}{4} + \frac{1}{6!3!} \cdot \frac{1}{8} \right) \\
 &= \boxed{7,484,400}
 \end{aligned}$$

(c) $10^7 = \boxed{10,000,000}$

(II) (a) $\binom{10}{7} = \boxed{120}$

(b) Clearly all outcomes that satisfy part (I)(a) also satisfy these conditions, so we start with $\binom{10}{7} = 120$ possible outcomes. In addition, the following outcomes are possible:

- (i) A set of 6 children receive gifts; one child receives two gifts (6 distinct ways this could happen for each set of 6 children).
- (ii) A set of 5 children receive gifts; two children receive two gifts ($\binom{5}{2}$ distinct ways this could happen for each set of 5 children).
- (iii) A set of 4 children receive gifts; three children each receive two gifts (4 distinct ways this could happen for each set of 4 children).

Clearly each of these outcomes are mutually exclusive. Therefore the answer is

$$\binom{10}{7} + \binom{10}{6} \cdot \binom{6}{7-6} + \binom{10}{5} \cdot \binom{5}{7-5} + \binom{10}{4} \cdot \binom{4}{7-4} = \boxed{4,740}$$

- (c) By Proposition 6.11, the number of nonnegative integer-valued vectors (x_1, x_2, \dots, x_r) satisfying the equation

$$x_1 + x_2 + \dots + x_r = n$$

is equal to $\binom{n+r-1}{r-1} = \binom{7+10-1}{10-1} = \boxed{11,440}$.

Homework 1 Problem 2.

- (I) 20 different gifts are distributed among seven children. How many different outcomes are possible if every child can receive (a) at least one gift, (b) at least two gifts, (c) any number of gifts?
- (II) Answer the same questions if the gifts are identical (but the children are still different).
- (III) Now try to generalize problems (1) and (2).

Solution.

(I) (a) There are 7^{20} possible allocations of gifts if we have no restrictions. If one child doesn't get a gift, there are $\binom{7}{1}$ ways to choose which child that is and 6^{20} subsequent allocations of gifts. Likewise, there are $\binom{7}{2} \cdot (7-2)^{20}$ ways to allocate the gifts if two children don't receive gifts, $\binom{7}{3} \cdot (7-3)^{20}$ ways if three children don't receive gifts, $\binom{7}{4} \cdot (7-4)^{20}$ ways if four children don't receive gifts, $\binom{7}{5} \cdot (7-5)^{20}$ ways if five children don't receive gifts, and $\binom{7}{6} \cdot (7-6)^{20}$ ways if six children don't receive gifts.

Let A_i denote the number of allocations in which i children do not receive gifts. In order to make the calculation, we must use the Inclusion-Exclusion principle (Proposition 6.5) (because, for example, some of the allocations in which three children don't receive gifts include allocations where four or more children don't receive gifts, and we don't want to double-count). Therefore the number of ways that at least one child can not receive a gift (i.e. the complement of every child receiving at least one gift) is

$$\begin{aligned} \left| \bigcup_{i=1}^6 A_i \right| &= \sum_{i=1}^6 |A_i| - \sum_{1 \leq i < j \leq 6} |A_i \cap A_j| + \sum_{1 \leq i < j < k \leq 6} |A_i \cap A_j \cap A_k| - \dots \\ &\quad + (-1)^{6-1} |A_1 \cap A_2 \cap A_3 \cap \dots \cap A_6| \end{aligned}$$

Fortunately, these allocations are nested in the sense that all the allocations where e.g. 5 children do not receive gifts are a subset of all the allocations where 4 children do not receive gifts; that is

$$A_6 \subset A_5 \subset A_4 \subset A_3 \subset A_2 \subset A_1$$

which implies e.g.

$$A_1 \cap A_2 \cap A_3 \cap \dots \cap A_6 = A_6,$$

$$\sum_{1 \leq i < j \leq 6} |A_i \cap A_j| = 5|A_6| + 4|A_5| + 3|A_4| + 2|A_3| + |A_2|$$

So we have

$$\begin{aligned} \left| \bigcup_{i=1}^6 A_i \right| &= |A_6| + |A_5| + |A_4| + |A_3| + |A_2| + |A_1| - (5|A_6| + 4|A_5| + 3|A_4| + 2|A_3| + |A_2|) \\ &\quad + (4|A_6| + 3|A_5| + 2|A_4| + |A_3|) - (3|A_6| + 2|A_5| + |A_4|) + \dots - |A_6| \\ &= |A_1| - |A_2| + |A_3| - |A_4| + |A_5| - |A_6| \\ &= \binom{7}{1} \cdot 6^{20} - \binom{7}{2} \cdot (7-2)^{20} + \binom{7}{3} \cdot (7-3)^{20} - \binom{7}{4} \cdot (7-4)^{20} \\ &\quad + \binom{7}{5} \cdot (7-5)^{20} - \binom{7}{6} \cdot (7-6)^{20} \end{aligned}$$

The final answer is

$$7^{20} - \left| \bigcup_{i=1}^6 A_i \right| = 7^{20} - \binom{7}{1} \cdot 6^{20} + \binom{7}{2} \cdot (7-2)^{20} - \binom{7}{3} \cdot (7-3)^{20} + \binom{7}{4} \cdot (7-4)^{20} \\ - \binom{7}{5} \cdot (7-5)^{20} + \binom{7}{6} \cdot (7-6)^{20} \approx [5.616 \cdot 10^{16}]$$

- (b) Similar to above, but more complicated. The complement of every child receiving at least two gifts is that at least one child doesn't receive a gift (same as above) or at least one child only receives one gift. So we start from the baseline answer above, and subtract out all the possible allocations in which at least one child receives one gift.

If one child only receives one gift (and the rest receive more than one), there are $\binom{7}{1}$ ways to choose which child that is, $\binom{20}{1}$ ways to choose which gift that child receives, and 6^{20-1} allocations of the remaining gifts. If two children receive only one gift, there are $\binom{7}{2}$ ways to choose which children those are, $\binom{20}{2} \cdot 2!$ ways to choose which gifts those children get and distribute them among those children, and $(7-2)^{20-2}$ ways to allocate the remaining gifts. Likewise, if three children receive only one gift there are $\binom{7}{3} \binom{20}{3} \cdot 3! \cdot (7-3)^{20-3}$ ways to allocate the gifts, $\binom{7}{4} \binom{20}{4} \cdot 4! \cdot (7-4)^{20-4}$ ways if four children receive only one gift, $\binom{7}{5} \binom{20}{5} \cdot 5! \cdot (7-5)^{20-5}$ ways if five children receive only one gift, and $\binom{7}{6} \binom{20}{6} \cdot 6! \cdot (7-6)^{20-6}$ ways if six children don't receive gifts.

Let B_j be the event that j children receive only one gift. Note that $B_1 \cap A_i$ is nonempty $\forall i < 7-1$, $B_2 \cap A_i$ is nonempty $\forall i < 7-2$, and in general, $B_j \cap A_i$ is nonempty $\forall i < 7-j, j \in \{1, 2, \dots, 6\}$. Applying the Inclusion-Exclusion Principle (Proposition 6.5) in a similar way as in part (I)(a), the answer is

$$7^{20} - \left| \bigcup_{i=1}^6 A_i \right| - \left| \bigcup_{j=1}^6 B_j \right| + \sum_{i \in \{1, \dots, 6\}, j \in \{1, \dots, 6\}} \left| A_i \cap B_j \right|$$

Per part (I)(a), the first two terms approximately equal $5.616 \cdot 10^{16}$. Clearly

$$\bigcup_{i \in \{1, \dots, 6\}, j \in \{1, \dots, 6\}} \left(A_i \cap B_j \right) \subset \bigcup_{j=1}^6 B_j$$

which implies

$$-\left| \bigcup_{j=1}^6 B_j \right| + \left| \bigcap_{i \in \{1, \dots, 6\}, j \in \{1, \dots, 6\}} B_j \right| < 0$$

so the answer to this part will be less than $5.616 \cdot 10^{16}$, which makes sense.

Calculating $\left| \bigcup_{j=1}^6 B_j \right|$ is not too difficult using Inclusion-Exclusion:

$$\left| \bigcup_{j=1}^6 B_j \right| = \sum_{j=1}^6 |B_j| - \sum_{1 \leq j < k \leq 6} |B_j \cap B_k| + \sum_{1 \leq j < k < \ell \leq 6} |B_j \cap B_k \cap B_\ell| - \dots \\ + (-1)^{6-1} |B_1 \cap B_2 \cap B_3 \cap \dots \cap B_6|$$

where since

$$B_6 \subset B_5 \subset B_4 \subset B_3 \subset B_2 \subset B_1$$

which implies e.g.

$$B_1 \cap B_2 \cap B_3 \cap \dots \cap B_6 = B_6,$$

$$\sum_{1 \leq j < k \leq 6} |B_j \cap B_k| = 5|B_6| + 4|B_5| + 3|B_4| + 2|B_3| + |B_2|$$

we have

$$\begin{aligned} \left| \bigcup_{j=1}^6 B_j \right| &= |B_6| + |B_5| + |B_4| + |B_3| + |B_2| + |B_1| - (5|B_6| + 4|B_5| + 3|B_4| + 2|B_3| + |B_2|) \\ &\quad + (4|B_6| + 3|B_5| + 2|B_4| + |B_3|) - (3|B_6| + 2|B_5| + |B_4|) + \dots - |B_6| \\ &= |B_1| - |B_2| + |B_3| - |B_4| + |B_5| - |B_6| \\ &= \binom{7}{1} \binom{20}{1} \cdot (7-1)^{20-1} - \binom{7}{2} \binom{20}{2} \cdot 2! \cdot (7-2)^{20-2} + \binom{7}{3} \binom{20}{3} \cdot 3! \cdot (7-3)^{20-3} - \binom{7}{4} \binom{20}{4} \cdot 4! \cdot (7-4)^{20-4} \\ &\quad + \binom{7}{5} \binom{20}{5} \cdot 5! \cdot (7-5)^{20-5} - \binom{7}{6} \binom{20}{6} \cdot 6! \cdot (7-6)^{20-6} \\ &\approx 5.846 \cdot 10^{16} \end{aligned}$$

However, calculating

$$\sum_{i \in \{1, \dots, 6\}, j \in \{1, \dots, 6\}} |A_i \cap B_j|$$

is very difficult because, for example, $B_2 \cap A_3$ is nonempty but $B_2 \not\subset A_3$ and $A_3 \not\subset B_2$.

(c) $7^{20} \approx 7.979 \cdot 10^{16}$

(II) (a) By Proposition 6.10, there are $\binom{19}{6} = [27, 132]$ ways to do this.

(b) Similar to Problem 1 part (II)(c), if the vector (x_1, x_2, \dots, x_7) represents the number of gifts given to each child, we would like a solution such that

$$x_1 + x_2 + \dots + x_7 = 20, x_i \geq 2 \forall i$$

By Proposition 6.12, the number of possible allocations under these conditions, is $\binom{20+7 \cdot (1-2)-1}{7-1} = \binom{12}{6} = [924]$.

(c) By Proposition 6.11, the number of nonnegative integer-valued vectors (x_1, x_2, \dots, x_r) satisfying the equation

$$x_1 + x_2 + \dots + x_r = n$$

is equal to $\binom{n+r-1}{r-1}$. In distributing 20 identical gifts to 7 different children, we can imagine the vector $(x_1, x_2, \dots, x_{10})$ represents the number of gifts given to each child (where x_i is a nonnegative integer for all i). So we have $n = 20$ and $r = 7$. Therefore the number of possible allocations is

$$\binom{20+7-1}{7-1} = \boxed{165,765,600}$$

(III) Generalization of 1(I): If there are g distinguishable gifts and $c \geq g$ children, the number of distinct allocations if each child can receive

- (a) at most one gift is $\binom{c}{g}g!$.
- (b) at most two gifts is

$$\sum_{i=c-g+1}^g \binom{c}{i} \cdot \binom{i}{g-i} \cdot \frac{g!}{(2!)^{g-i}}$$

- (c) any number of gifts is c^g .

Generalization of 1(II): If there are g identical gifts and $c \geq g$ children, the number of distinct allocations if each child can receive

- (a) at most one gift is $\binom{c}{g}$.
- (b) at most two gifts is

$$\sum_{i=c-g+1}^g \binom{c}{i} \cdot \binom{i}{g-i}$$

- (c) any number of gifts is $\binom{g+c-1}{c-1}$

Generalization of 2(I): If there are g distinguishable gifts and $c \leq g$ children, the number of distinct allocations if each child must receive

- (a) at least one gift is

$$c^g - \sum_{i=1}^{c-1} (-1)^{i+1} \binom{c}{i} \cdot (c-i)^g$$

- (b) at least two gifts is

$$c^g - \sum_{i=1}^{c-1} (-1)^{i+1} \binom{c}{i} \cdot (c-i)^g - \sum_{i=1}^{c-1} (-1)^{i+1} \binom{c}{i} \binom{g}{i} \cdot i! \cdot (c-i)^{g-i}$$

- (c) any number of gifts is c^g

Generalization of 2(II): If there are g identical gifts and $c \leq g$ children, the number of distinct allocations if each child must receive

- (a) at least one gift is

$$\binom{g-1}{c-1}$$

(b) at least two gifts is

$$\binom{g - c - 1}{c - 1}$$

(c) any number of gifts is

$$\binom{g + c - 1}{c - 1}$$

Homework 1 Problem 4. You have \$20K to invest, and have a choice of stocks, bonds, mutual funds, or a CD. Investments must be made in multiples of \$1K, and there are minimal amounts to be invested: \$2K in stocks, \$2K in bonds, \$3K in mutual funds, and \$4K in the CD. Count the number of choices in each situation: (a) You want to invest in all four, (b) you want to invest in at least three out of four.

Solution.

- (a) If the vector $(x_S, x_B, x_{MF}, x_{CD})$ represents the amount of money (in thousands of dollars) invested in each instrument, we would like a solution such that

$$x_S + x_B + x_{MF} + x_{CD} = 20$$

where

$$x_S \geq 2, x_B \geq 2, x_{MF} \geq 3, x_{CD} \geq 4$$

In a way similar to the proof for Proposition 6.12, note that we can transform this problem in the following way:

$$x_S - 1 + x_B - 1 + x_{MF} - 2 + x_{CD} - 3 = 20 - (1 + 1 + 2 + 3)$$

where

$$x_S - 1 \geq 1, x_B - 1 \geq 1, x_{MF} - 2 \geq 1, x_{CD} - 3 \geq 1$$

Letting $y_S = x_S - 1, y_B = x_B - 1, y_{MF} = x_{MF} - 2, y_{CD} = x_{CD} - 3$, we have the equivalent system

$$y_S + y_B + y_{MF} + y_{CD} = 13, y \geq 1 \forall y$$

By Proposition 6.10, the number of distinct solutions to this equation, and therefore the number of possible allocations under these conditions, is $\binom{13-1}{4-1} = \boxed{220}$.

- (b) Enumerate the $\binom{4}{3} = 4$ possibilities.

(i) **Invest in stocks, bonds, and mutual funds.**

$$x_S + x_B + x_{MF} = 202$$

where

$$x_S \geq 2, x_B \geq 2, x_{MF} \geq 3$$

Note that we can transform this problem in the following way:

$$x_S - 1 + x_B - 1 + x_{MF} - 2 = 20 - (1 + 1 + 2)$$

where

$$x_S - 1 \geq 1, x_B - 1 \geq 1, x_{MF} - 2 \geq 1$$

Letting $y_S = x_S - 1, y_B = x_B - 1, y_{MF} = x_{MF} - 2$, we have the equivalent system

$$y_S + y_B + y_{MF} = 16, y \geq 1 \forall y$$

Therefore the number of possible allocations under these conditions is $\binom{16-1}{3-1} = [105]$.

(ii) **Invest in stocks, bonds, and CDs.**

$$x_S + x_B + x_{CD} = 20$$

where

$$x_S \geq 2, x_B \geq 2, x_{CD} \geq 4$$

Note that we can transform this problem in the following way:

$$x_S - 1 + x_B - 1 + x_{CD} - 3 = 20 - (1 + 1 + 3)$$

where

$$x_S - 1 \geq 1, x_B - 1 \geq 1, x_{CD} - 3 \geq 1$$

Letting $y_S = x_S - 1, y_B = x_B - 1, y_{CD} = x_{CD} - 3$, we have the equivalent system

$$y_S + y_B + y_{CD} = 15, y \geq 1 \forall y$$

Therefore the number of possible allocations under these conditions is $\binom{15-1}{3-1} = [91]$.

(iii) **Invest in stocks, mutual funds, and CDs.**

$$x_S + x_{MF} + x_{CD} = 2$$

where

$$x_S \geq 2, x_{MF} \geq 3, x_{CD} \geq 4$$

Note that we can transform this problem in the following way:

$$x_S - 1 + x_{MF} - 2 + x_{CD} - 3 = 20 - (1 + 2 + 3)$$

where

$$x_S - 1 \geq 1, x_{MF} - 2 \geq 1, x_{CD} - 3 \geq 1$$

Letting $y_S = x_S - 1, y_{MF} = x_{MF} - 2, y_{CD} = x_{CD} - 3$, we have the equivalent system

$$y_S + y_{MF} + y_{CD} = 14, y \geq 1 \forall y$$

Therefore the number of possible allocations under these conditions is $\binom{14-1}{3-1} = [78]$.

(iv) **Invest in bonds, mutual funds, and CDs.**

$$x_B + x_{MF} + x_{CD} = 2$$

where

$$x_B \geq 2, x_{MF} \geq 3, x_{CD} \geq 4$$

Note that we can transform this problem in the following way:

$$x_B - 1 + x_{MF} - 2 + x_{CD} - 3 = 20 - (1 + 2 + 3)$$

where

$$x_B - 1 \geq 1, x_{MF} - 2 \geq 1, x_{CD} - 3 \geq 1$$

Letting $y_B = x_B - 1, y_{MF} = x_{MF} - 2, y_{CD} = x_{CD} - 3$, we have the equivalent system

$$y_B + y_{MF} + y_{CD} = 14, y \geq 1 \quad \forall y$$

therefore the number of possible allocations under these conditions is $\binom{14-1}{3-1} = \boxed{78}$.

(v) **Invest in all four:** per part 4(a), there are $\boxed{220}$ ways to do this.

Note that all of these possibilities are mutually exclusive. Therefore the total number is

$$\binom{16-1}{3-1} + \binom{15-1}{3-1} + \binom{14-1}{3-1} + \binom{14-1}{3-1} + \binom{13-1}{4-1} = 105 + 91 + 78 + 78 + 220 = \boxed{572}$$

6.3 To Know for Math 505A Midterm 2

6.3.1 Definitions

Definition 6.20. A random variable X is **continuous** if its distribution function $F(x) = \Pr(X \leq x)$ can be written as

$$F(x) = \int_{-\infty}^x f(u)du$$

for some integrable $f : \mathbb{R} \rightarrow [0, \infty)$.

Definition 6.21. The function f is called the **(probability) density function** of the continuous random variable X .

Proposition 6.34. If X has pdf $f_X(x)$, then for $\mu \in \mathbb{R}, \sigma > 0$,

$$h(x) = \frac{1}{\sigma} f_X\left(\frac{x-\mu}{\sigma}\right)$$

is a pdf. In this setting μ is sometimes called a “location parameter” and σ is called a “scale parameter.”

Definition 6.22. The **joint distribution function** of X and Y is the function $F : \mathbb{R}^2 \rightarrow [0, 1]$ given by

$$F(x, y) = \Pr(X \leq x \cap Y \leq y)$$

Definition 6.23. The random variables X and Y are **jointly continuous** with **joint (probability) density function** $f : \mathbb{R}^2 \rightarrow [0, \infty)$ if

$$F(x, y) = \int_{v=-\infty}^y \int_{u=-\infty}^x f(u, v)dudv \text{ for each } x, y \in \mathbb{R}$$

Definition 6.24. Two continuous random variables are **independent** if and only if $\{X \leq x\}$ and $\{Y \leq y\}$ are independent events for all $x, y \in \mathbb{R}$.

Ways to show independence:

- Use Definition 6.24: show that $\Pr(X \leq x \cap Y \leq y) = \Pr(X \leq x) \Pr(Y \leq y)$ for all $x, y \in \mathbb{R}$.

•

Theorem 6.35. The random variables X and Y are independent if and only if $F(x, y) = F_X(x)F_Y(y)$ for all $x, y \in \mathbb{R}$.

•

Proposition 6.36. For continuous random variables, the previous condition is equivalent to requiring $f(x, y) = f_X(x)f_Y(y)$.

•

Theorem 6.37. If two variables are bivariate normal, they are independent if and only if their covariance

$$\text{Cov}(X, Y) = \mathbb{E}(XY) - \mathbb{E}(X)\mathbb{E}(Y) = \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} xyf(x, y)dxdy$$

is equal to 0.

Theorem 6.38. (Change of variables.) Let X_1, Y_1 be random variables with joint PDF f_{X_1, Y_1} . Let X_2, Y_2 be random variables with joint PDF f_{X_2, Y_2} . Let $T : \mathbb{R}^2 \rightarrow \mathbb{R}^2$ and let $S : \mathbb{R}^2 \rightarrow \mathbb{R}^2$ so that $ST(x, y) = (x, y)$ and $TS(x, y) = (x, y)$ for every $(x, y) \in \mathbb{R}^2$. Let $J(x, y)$ denote the determinant of the Jacobian of S at (x, y) . Then

$$f_{X_2, Y_2}(x, y) = f_{X_1, Y_1}(S(x, y)) |J(x, y)|.$$

Proof. If the transformation from X_1, Y_1 to X_2, Y_2 is given by $S(X_1, Y_1) = (X_2, Y_2)$, then the change of variables formula from calculus is as follows:

$$\int \int_A f_{X_1, Y_1}(x, y) dxdy = \int \int_B f_{X_1, Y_1}(S(x, y)) |J(x, y)| dxdy$$

where $A \subseteq \text{domain}(f_{X_1, Y_1}(\cdot))$, B is the transformation of the region A under S , and $|J(x, y)|$ is the Jacobian of S at (x, y) . It follows from the definition of joint pdfs that the integrand on the right is the joint pdf of (X_2, Y_2) ; that is,

$$f_{X_2, Y_2}(x, y) = f_{X_1, Y_1}(S(x, y)) |J(x, y)|.$$

□

- Characteristic functions:

Theorem 6.39. X and Y are independent if and only if $\phi_{X,Y}(s, t) = \phi_X(s)\phi_Y(t)$.

Theorem 6.40. (Theorem 4.2.3, Grimmett and Stirzaker.) Let X and Y be random variables, and let $g, h : \mathbb{R} \rightarrow \mathbb{R}$. If X and Y are independent, then so are $g(X)$ and $h(Y)$.

Definition 6.25. The **correlation coefficient** between random variables X and Y is given by

$$\rho = \frac{\text{Cov}(X, Y)}{\sqrt{\text{Var}(X)\text{Var}(Y)}}$$

Theorem 6.41. The correlation coefficient satisfies $|\rho| \leq 1$.

Proof. Apply the Cauchy-Schwarz Inequality (Theorem 8.6) to $X - \mathbb{E}(X)$ and $Y - \mathbb{E}(Y)$:

$$\begin{aligned} \text{Cov}(X, Y)^2 &= (\mathbb{E}[(X - \mathbb{E}(X))(Y - \mathbb{E}(Y))]^2)^2 \leq \mathbb{E}[(X - \mathbb{E}(X))^2]\mathbb{E}[(Y - \mathbb{E}(Y))^2] = \text{Var}(X)\text{Var}(Y) \\ &\iff \frac{\text{Cov}(X, Y)^2}{\text{Var}(X)\text{Var}(Y)} \leq 1 \iff \rho^2 \leq 1 \iff |\rho| \leq 1 \end{aligned}$$

□

Theorem 6.42. (Stein identity.) Let X be a standard Gaussian random variable, so that X has density $x \rightarrow e^{-x^2/2}/\sqrt{2\pi}$, $\forall x \in \mathbb{R}$. Let $g : \mathbb{R} \rightarrow \mathbb{R}$ be a continuously differentiable function such that g and g' have polynomial volume growth. That is, $\exists a, b > 0$ such that $|g(x)|, |g'(x)| \leq a(1 + |x|)^b$, $\forall x \in \mathbb{R}$. Then

$$\mathbb{E}Xg(X) = \mathbb{E}g'(X).$$

Proof. Examining the left side, we have

$$\mathbb{E}(Xg(X)) = \frac{1}{\sqrt{2\pi}} \int_{-\infty}^{\infty} xg(x)e^{-x^2/2}dx$$

We will use integration by parts with $u = g(x) \implies du = g'(x)dx$, $dv = xe^{-x^2/2}dx \implies v = -e^{-x^2/2}$ to yield the result:

$$\mathbb{E}(Xg(X)) = -g(x)e^{-x^2/2} + \int_{-\infty}^{\infty} g'(x)e^{-x^2/2}dx = \mathbb{E}(g'(X)).$$

□

Example. Using Theorem 6.42, recursively compute $\mathbb{E}X^k$ for any positive integer k . Alternatively, for any $t > 0$, show that $\mathbb{E}e^{tX} = e^{t^2/2}$, i.e. compute the **moment generating function** of X . Then, using $\frac{d^k}{dt^k}|_{t=0}\mathbb{E}e^{tX} = \mathbb{E}X^k$ and using the power series expansion of the exponential, compute $\mathbb{E}X^k$ directly from the identity $\mathbb{E}e^{tX} = e^{t^2/2}$.

Solution. We can use this result to recursively calculate $\mathbb{E}(X^k)$ for any positive integer k . Suppose we have $\mathbb{E}(X^{k-1})$. Letting $g(X) = X^{k-1} \implies g'(X) = (k-1)X^{k-2}$, we have

$$\mathbb{E}(X^k) = \mathbb{E}(X \cdot X^{k-1}) = \mathbb{E}(Xg(X)) = \mathbb{E}(g'(X)) \iff \boxed{\mathbb{E}(X^k) = (k-1)\mathbb{E}(X^{k-2})}.$$

Since $X \sim \mathcal{N}(0, 1)$, we have $\mathbb{E}(X) = 0$, $\mathbb{E}(X^2) = \text{Var}(X) + \mathbb{E}(X)^2 = 1 + 0 = 1$. Therefore we have

$$\mathbb{E}(X^k) = \begin{cases} \prod_{i=1}^{(k-1)/2} (k - (2i-1)) \mathbb{E}(X) & k \text{ is odd} \\ \prod_{i=1}^{k/2} (k - (2i-1)) \mathbb{E}(X^2) & k \text{ is even} \end{cases}$$

$$\boxed{\mathbb{E}(X^k) = \begin{cases} 0 & k \text{ is odd} \\ \prod_{i=1}^{k/2} (k - (2i-1)) & k \text{ is even} \end{cases}}$$

6.3.2 Probability-Generating Functions

Definition 6.26.

$$G_X(s) = \mathbb{E}(s^X)$$

Theorem 6.43. Some useful properties:

- (a) $\mathbb{E}(X) = G'_X(1)$, $\mathbb{E}[X(X-1)\cdots(X-k+1)] = G^{(k)}(1)$
- (b) If X and Y are independent then $G_{X+Y}(s) = G_X(s)G_Y(s)$.

6.3.3 Moment-Generating Functions

Definition 6.27.

$$M_X(t) = \mathbb{E}(e^{tX})$$

Theorem 6.44. Some useful properties:

- (a) $\mathbb{E}(X) = M'_X(0)$, $\mathbb{E}(X^k) = M^{(k)}(0)$
- (b) If X and Y are independent then $M_{X+Y}(t) = M_X(t)M_Y(t)$.

6.3.4 Characteristic Functions

Definition 6.28.

$$\phi_X(t) = \mathbb{E}(e^{itX})$$

Proposition 6.45. Necessary and sufficient conditions for a function to be a characteristic function:

- (a) $\phi_X(0) = 1$
- (b) $|\phi(t)| \leq 1 \forall t$
- (c) ϕ is uniformly continuous on \mathbb{R}

(d) ϕ is positive semidefinite; that is,

$$\sum_{i,j} \phi(t_j - t_k) z_j \bar{z}_k \geq 0 \text{ for all real } t_1, t_2, \dots, t_n \text{ and complex } z_1, z_2, \dots, z_n$$

Or, equivalently, or every set of real numbers t_1, t_2, \dots, t_n , the matrix $\phi(t_i - t_j), i, j \in \{1, 2, \dots, n\}$ is Hermitian and nonnegative definite.

Remark. Relationship between characteristic functions and probability and moment generating functions:

$$\phi_X(t) = M_X(it) = G_X(e^{it})$$

Theorem 6.46. Some useful properties:

- (a) $X \perp Y \implies \phi_{X+Y}(t) = \phi_X(t)\phi_Y(t)$
- (b) $Y = aX + b \implies \phi_Y(t) = e^{itb}\phi_X(at)$
- (c) $\phi_X^{(k)}(0) = i^k \mathbb{E}(X^k)$
- (d) $\phi_{X,Y}(s,t) = \mathbb{E}(e^{isX}e^{itY})$
- (e) $X \perp Y \iff \phi_{X,Y}(s,t) = \phi_X(s)\phi_Y(t)$

Theorem 6.47. Other facts from notes on course website

- (a) If $\phi(t)$ is even, $\phi(0) = 1$, ϕ is convex for $t > 0$, and $\lim_{t \rightarrow \infty} \phi(t) = 0$, then ϕ is a characteristic function of an absolutely continuous random variable.
- (b) If ϕ is a characteristic function and $\phi(t) = 1 + o(t^2), t \rightarrow 0$, then $\phi(t) = 1$ for all t . The random variable with such a characteristic function must have zero mean and zero variance. In particular, if $r > 2$, then $\exp(-|t|^r)$ is not a characteristic function.
- (c) If $\phi(t) = e^{p(t)}$ is a characteristic function and $p = p(t)$ is a polynomial, then the degree of p is at most 2. For example, $e^{t^2-t^4}$ is not a characteristic function.
- (d) If ξ is absolutely continuous, then $\lim_{|t| \rightarrow \infty} |\phi_\xi(t)| = 0$ (Riemann-Lebesgue).
- (e) If $\int_{-\infty}^{\infty} |\phi_\xi(t)| dt < \infty$, then ξ is absolutely continuous with pdf

$$f(x) = \frac{1}{2\pi} \int_{-\infty}^{\infty} \exp(-itx) \phi(t) dt$$

6.3.5 Continuous Random Variable Distributions

Uniform: $U(a, b)$

- Probability density function:

$$f(x) = \begin{cases} \frac{1}{b-a} & a \leq x \leq b \\ 0 & \text{otherwise} \end{cases}$$

- Cumulative distribution function:

$$F(x) = \Pr(X \leq x) = \begin{cases} 0 & x \leq a \\ \frac{x-a}{b-a} & a < x \leq b \\ 1 & x > b \end{cases}$$

- Probability-generating function:

- Moment-generating function:

$$M_X(t) = \frac{1}{(b-a)t} [\exp(bt) - \exp(at)]$$

Proof.

$$M_X(t) = \mathbb{E}(\exp(tx)) = \int_a^b \frac{1}{b-a} \cdot \exp(tx) dx = \frac{1}{b-a} \left[\frac{1}{t} \exp(tx) \right]_a^b = \frac{1}{(b-a)t} [\exp(bt) - \exp(at)]$$

□

- Characteristic function:

$$\frac{2}{(b-a)t} \sin\left(\frac{1}{2}(b-a)t\right) \exp\left(i(a+b)\frac{t}{2}\right)$$

- Expectation: $\mathbb{E}(X) = (b-a)/2$
- Variance: $\text{Var}(X) = (b-a)^2/12$

Proposition 6.48. If $X \sim U(0, 1)$, then $Y = -\log(X) \sim \text{Exponential}(1)$.

Proof.

$$\Pr(Y \leq y) = \Pr(-\log(X) \leq y) = \Pr(\log(X) \geq -y) = \Pr(X \geq e^{-y}) = \int_{\exp(-y)}^{\infty} dt = \int_{\exp(-y)}^1 dt$$

Substituting $t = e^{-u}$ (so that we have $u = -\log(t)$, $dt = -e^{-u}du$, we have

$$\Pr(Y \leq y) = - \int_y^0 e^{-u} du = [e^{-u}]_y^0 = 1 - e^{-y}$$

which is the cdf for an exponential distribution with mean 1. □

Normal: $\mathcal{N}(\mu, \sigma^2)$

- Probability density function:

$$f_X(x) = \frac{1}{\sqrt{2\pi\sigma^2}} \exp\left(-\frac{(x-\mu)^2}{2\sigma^2}\right)$$

- Cumulative distribution function: $F(x) = \Pr(X \leq x) =$

- Probability-generating function:
- Moment-generating function:
- Characteristic function: $\phi(t) = \exp(i\mu t - (1/2)\sigma^2 t^2)$. Standard normal: $\phi(t) = \exp((-1/2)t^2)$.
- Expectation: $\mathbb{E}(X) = \mu$
- Variance: $\text{Var}(X) = \sigma^2$

Gamma: $\Gamma(\alpha, \beta)$

- Probability density function:

$$f(x) = \frac{1}{\beta^\alpha \Gamma(\alpha)} x^{\alpha-1} e^{-x/\beta} = \frac{1}{\Gamma(\alpha, \beta)} x^{\alpha-1} e^{-x/\beta}$$

- Cumulative distribution function: $F(x) = \Pr(X \leq x) =$
- Probability-generating function:
- Moment-generating function:
- Characteristic function:
- Expectation: $\mathbb{E}(X) = \alpha\beta$
- Variance: $\text{Var}(X) = \alpha\beta^2$

Proposition 6.49. Let $X \sim \text{Gamma}(\alpha_1, \beta)$ and $Y \sim \text{Gamma}(\alpha_2, \beta)$. Then $X + Y \sim \text{Gamma}(\alpha_1 + \alpha_2, \beta)$.

Proposition 6.50. Let $X \sim \text{Gamma}(\alpha, \beta)$. Then as $\beta \rightarrow \infty$, $X \xrightarrow{d} \mathcal{N}(\alpha\beta, \alpha\beta^2)$.

Proof. See <http://www.math.wm.edu/~leemis/chart/UDR/PDFs/GammaNormal1.pdf>. □

χ_n^2 : special case of a gamma distribution: $\Gamma(n/2, 2)$. Also the sum of n independent standard normally distributed variables.

- Probability density function:

$$f(x) = \frac{1}{2^{n/2} \Gamma(n/2)} x^{n/2-1} e^{-x/2} = \frac{1}{\Gamma(n/2, 2)} x^{n/2-1} e^{-x/2}$$

- Cumulative distribution function: $F(x) = \Pr(X \leq x) =$
- Probability-generating function:
- Moment-generating function:
- Characteristic function:
- Expectation: $\mathbb{E}(X) = n/2 \cdot 2 = n$
- Variance: $\text{Var}(X) = n/2 \cdot 2^2 = 2n$

Exponential: (special case of a gamma distribution: $\Gamma(1, \beta)$. Also a special case of a Weibull distribution with $\beta = 1$.)

- Probability density function: $f(x) = \frac{1}{\beta} \exp(-x/\beta) = \lambda e^{-\lambda x}$
- Cumulative distribution function: $F(x) = \Pr(X \leq x) = 1 - e^{-\lambda x}$
- Probability-generating function:
- Moment-generating function: $\frac{\lambda}{\lambda-t}$
- Characteristic function:
- Expectation: $\mathbb{E}(X) = \beta = \lambda^{-1}$
- Variance: $\text{Var}(X) = \beta^2 = \lambda^{-2}$

Remark. Recall Proposition 6.48: If $X \sim U(0, 1)$, then $Y = -\log(X) \sim \text{Exponential}(1)$.

Cauchy:

- Probability density function:

$$f(x) = \frac{1}{\pi(1+x^2)} \text{ (standard Cauchy) , } f(x) = \frac{1}{\pi\sigma(1+(x-\mu)^2/\sigma^2)} \text{ (general)}$$

- Cumulative distribution function: $F(x) = \Pr(X \leq x) =$
- Probability-generating function:
- Moment-generating function:
- Characteristic function:
- Expectation: does not exist
- Variance: does not exist (Cauchy distribution has no moments.)

Beta: Recall:

$$B(\alpha, \beta) = \frac{\Gamma(\alpha)\Gamma(\beta)}{\Gamma(\alpha+\beta)}$$

$$\implies \frac{B(\alpha+1, \beta)}{B(\alpha, \beta)} = \frac{\Gamma(\alpha+1)\Gamma(\beta)}{\Gamma(\alpha+1+\beta)} \frac{\Gamma(\alpha+\beta)}{\Gamma(\alpha)\Gamma(\beta)} = \frac{\alpha}{\alpha+\beta}$$

- Probability density function: $f(x) =$
- Cumulative distribution function: $F(x) = \Pr(X \leq x) =$
- Probability-generating function:
- Moment-generating function:

- Characteristic function:
- Expectation: $\mathbb{E}(X) =$
- Variance: $\text{Var}(X) =$

t_n :

- Probability density function:

$$f(x) = \frac{\Gamma((n+1)/2)}{\sqrt{n\pi} \cdot \Gamma(n/2)} \left(1 + \frac{x^2}{n}\right)^{-(n+1)/2}$$

- Cumulative distribution function: $F(x) = \Pr(X \leq x) =$
- Probability-generating function:
- Moment-generating function:
- Characteristic function:
- Expectation: $\mathbb{E}(X) = 0$
- Variance: $\text{Var}(X) = n/(n-2)$

Weibull:

- Probability density function: $f(x) = \alpha\beta x^{\beta-1} \exp(-\alpha x^\beta)$
- Cumulative distribution function: $F(x) = \Pr(X \leq x) = 1 - \exp(-\alpha x^\beta)$
- Probability-generating function:
- Moment-generating function:
- Characteristic function:
- Expectation: $\mathbb{E}(X) =$
- Variance: $\text{Var}(X) =$

Pareto: (parameters α, x_m)

- Probability density function: $f(x) = \alpha x_m^\alpha / x^{\alpha+1}$ for $x \geq 1, 0$ otherwise.
- Cumulative distribution function: $F(x) = 1 - (x_m/x)^\alpha$ for $x \geq 1, 0$ otherwise.
- Probability-generating function:
- Moment-generating function:
- Characteristic function:
- Expectation: $\mathbb{E}(X) = \alpha x_m / (\alpha - 1)$ for $\alpha > 1, \infty$ otherwise.
- Variance: $\text{Var}(X) = \frac{x_m^2 \alpha}{(\alpha-1)^2 (\alpha-2)}$ for $\alpha > 2, \infty$ otherwise.

6.3.6 Multivariate Gaussian (Normal) Distributions

Definition 6.29. From <http://pluto.huji.ac.il/~pchiga/teaching/MathStat/SIAnotes2013.pdf> (definition 2b6): A random vector $X = (X_1, X_2)$ is Gaussian with mean $\mu = (\mu_1, \mu_2)$ and the covariance matrix

$$\Sigma = \begin{bmatrix} \sigma_1^2 & \rho\sigma_1\sigma_2 \\ \rho\sigma_1\sigma_2 & \sigma_2^2 \end{bmatrix}$$

if it has a joint pdf of the form

$$f_X(x) = \frac{1}{2\pi\sigma_2\sigma_2\sqrt{1-\rho^2}} \exp \left[-\frac{1}{2} \frac{1}{1-\rho^2} \left(\frac{(x_1 - \mu_1)^2}{\sigma_1^2} - \frac{2\rho(x_1 - \mu_1)(x_2 - \mu_2)}{\sigma_1\sigma_2} + \frac{(x_2 - \mu_2)^2}{\sigma_2^2} \right) \right]$$

for $x \in \mathbb{R}^2$.

Proposition 6.51. From <http://pluto.huji.ac.il/~pchiga/teaching/MathStat/SIAnotes2013.pdf> (Proposition 3c1): Let X be a Gaussian random variable in \mathbb{R}^2 such that

$$\begin{bmatrix} X_1 \\ X_2 \end{bmatrix} \sim \mathcal{N} \left(\begin{bmatrix} \mu_1 \\ \mu_2 \end{bmatrix}, \begin{bmatrix} \sigma_1^2 & \rho\sigma_1\sigma_2 \\ \rho\sigma_1\sigma_2 & \sigma_2^2 \end{bmatrix} \right)$$

Then $f_{X_1|X_2}(x_1; x_2)$ is Gaussian with the (conditional) mean

$$\mathbb{E}(X_1 | X_2 = x_2) = \mu_1 + \frac{\rho\sigma_1}{\sigma_2}(x_2 - \mu_2)$$

and the (conditional) variance

$$\text{Var}(X_1 | X_2 = x_2) = \sigma_1^2(1 - \rho^2)$$

That is, the conditional distribution of X_1 given X_2 is

$$X_1 | X_2 = x_2 \sim \mathcal{N} \left(\mu_1 + \rho \frac{\sigma_1}{\sigma_2} (x_2 - \mu_2), (1 - \rho^2)\sigma_1^2 \right)$$

Remark. Note that this matches the OLS coefficients in the univariate case. In other words, the univariate OLS formula can be derived using only this fact.

Recall Theorem 6.37: if two variables are bivariate normal, they are independent if and only if their covariance

$$\text{Cov}(X, Y) = \mathbb{E}(XY) - \mathbb{E}(X)\mathbb{E}(Y) = \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} xyf(x, y)dxdy$$

equals 0.

6.4 Worked problems

6.4.1 Example Problems That Will Likely Appear on Midterm (and Final)

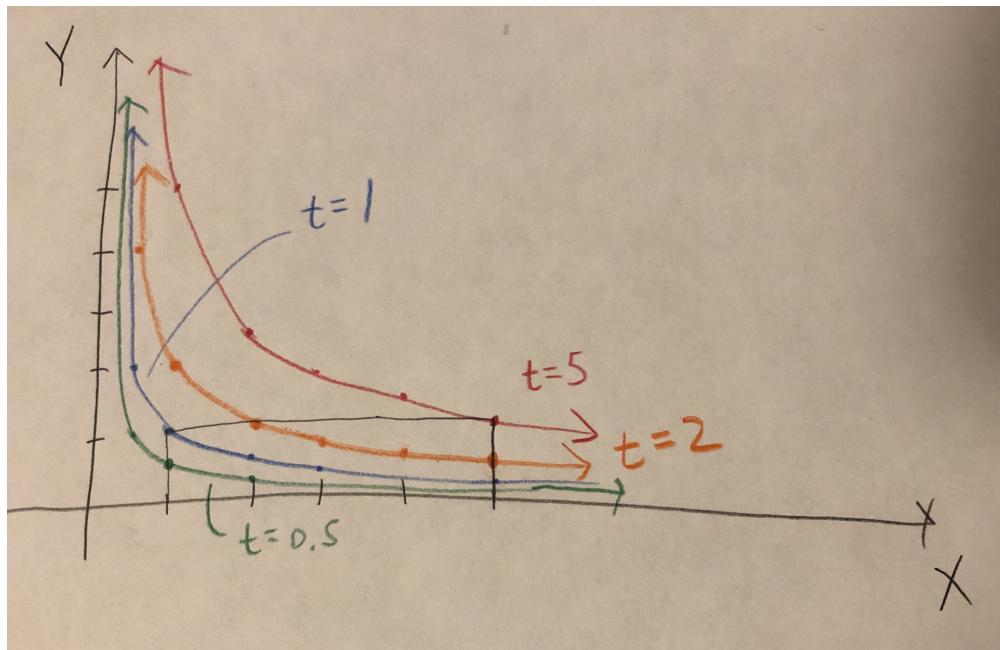
- (1) Let X be uniform on $[1, 5]$, let Y be uniform on $[0, 1]$, and assume that X and Y are independent.
- Compute the probability density function of the product XY .
 - Only part included on midterm.** Compute the cumulative distribution function of the ratio X/Y .
 - Compute the characteristic function of the sum $X + Y$.
 - Compute the moment-generating function of the random variable $X - \ln(Y)$.

Solution.

- (a) We will find the cdf and then differentiate to yield the pdf. Observe that

$$F_{XY}(t) = \Pr(XY \leq t) = \Pr(Y \leq tX^{-1})$$

Plotting the density function of XY along with plots of $F_{XY}(t)$ as a function of X for various values of t , we have the following:



Now since both X and Y are distributed uniformly, for a given t , $\Pr(Y \leq tX^{-1})$ is the area under the curve and in the rectangle, weighted by $1/4$ since the rectangle has total area 4 but total probability 1. It is clear that the four regimes we need to consider are (1) $t < 0$, (2) $0 \leq t < 1$, (3) $1 \leq t < 5$, and (4) $t \geq 5$.

- (1) $t < 0$: The curve lies below the rectangle, so there is no area below the curve and in the rectangle. Therefore $\boxed{\Pr(Y \leq tX^{-1} \mid t < 0) = 0}$. (This is also clear since tX^{-1} would be a negative number and Y is nonnegative.)

(2) $0 \leq t < 1$: Integrating the relevant area, we have

$$\Pr(Y \leq tX^{-1} \mid 0 \leq t < 1) = \frac{1}{4} \int_1^5 \frac{t}{x} dx = \frac{t}{4} [\log(x)]_1^5 = \boxed{\frac{t}{4} \log(5)}$$

(3) $1 \leq t < 5$: In this case, the area is a rectangle of height 1 and width $t - 1$ plus the area under the curve from t to 5.

$$\Pr(Y \leq tX^{-1} \mid 1 \leq t < 5) = \frac{1}{4} \left(1 \cdot (t-1) + \int_t^5 \frac{t}{x} dx \right) = \frac{1}{4} \left(t-1 + t [\log(x)]_t^5 \right) = \boxed{\frac{1}{4} [t(1 + \log(5/t)) - 1]}$$

(4) $t \geq 5$: In this case, the entire rectangle lies below the curve. Therefore $\boxed{\Pr(Y \leq tX^{-1} \mid t \geq 5) = 1}$.

So we have

$$F_{XY}(t) = \begin{cases} 0 & t < 0 \\ \frac{t}{4} \log(5) & 0 \leq t < 1 \\ \frac{1}{4} [t(1 + \log(5/t)) - 1] & 1 \leq t < 5 \\ 1 & t \geq 5 \end{cases}$$

Finally, differentiating yields

$$f_{XY}(t) = \begin{cases} 0 & t < 0 \\ \frac{1}{4} \log(5) & 0 \leq t < 1 \\ \frac{1}{4} \log\left(\frac{5}{t}\right) & 1 \leq t < 5 \\ 0 & t \geq 5 \end{cases}$$

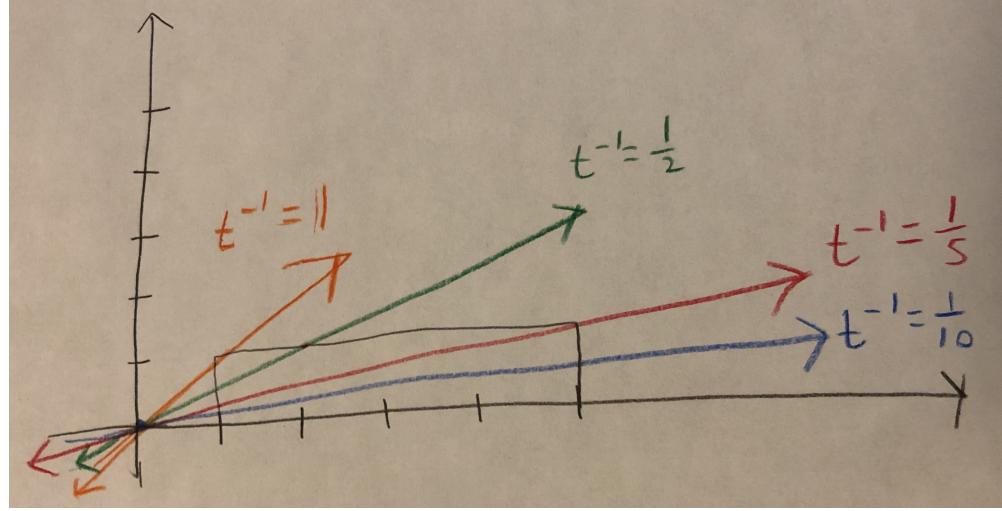
since

$$\begin{aligned} \frac{d}{dt} \left(\frac{1}{4} [t(1 + \log(5/t)) - 1] \right) &= \frac{1}{4} \left[1 + \log(5/t) + t \left(\frac{1}{5/t} - 5 \cdot t^{-2} \right) \right] = \frac{1}{4} \left[1 + \log(5/t) + t \left(t - 1 \cdot t^{-2} \right) \right] \\ &= \frac{1}{4} \log\left(\frac{5}{t}\right) \end{aligned}$$

(b) **Only part included on midterm.** We will proceed in a similar way as part (a). Observe that

$$F_{X/Y}(t) = \Pr\left(\frac{X}{Y} \leq t\right) = \Pr(Y \geq X/t)$$

Plotting the density function of X/Y along with plots of $F_{X/Y}(t)$ as a function of X for various values of t , we have the following:



Now since both X and Y are distributed uniformly, for a given t , $\Pr(Y \geq X/t)$ is the area above the curve and in the rectangle, weighted by $1/4$ since the rectangle has total area 4 but total probability 1. It is clear that the three regimes we need to consider are (1) $t^{-1} \geq 1 \iff t \leq 1$, (2) $1/5 \leq t^{-1} < 1 \iff 1 < t \leq 5$, and (3) $0 < t^{-1} < 1/5 \iff t > 5$.

- (1) $t \leq 1$: The curve lies above the rectangle, so there is no area above the curve and in the rectangle. Therefore $\Pr(Y \geq X/t \mid t \leq 1) = 0$. (This is also clear since X/t would have to be greater than 1 and Y is less than or equal to 1.)
- (2) $1 < t \leq 5$: The relevant area is the triangle above the green line in the rectangle. Note that it intersects the vertical line at $Y = 1/t$ and the horizontal line at $X = t$.

$$\Pr(Y \geq X/t \mid 1 < t \leq 5) = \frac{1}{4} \cdot \frac{1}{2} \left(1 - \frac{1}{t}\right)(t-1) = \frac{1}{8} \left(t - 1 - 1 + \frac{1}{t}\right) = \frac{1}{8} \left(t - 2 + \frac{1}{t}\right)$$

- (3) $t > 5$: In this case, the area is a trapezoid above the blue line and in the rectangle. Note that the blue line intersects the left vertical line at $Y = 1/t$ and the right vertical line at $Y = 5/t$.

$$\Pr(Y \geq X/t \mid t > 5) = \frac{1}{4} \cdot \frac{1}{2} \cdot \left(1 - \frac{1}{t} + 1 - \frac{5}{t}\right) \cdot 4 = \frac{1}{2} \cdot \left(2 - \frac{6}{t}\right) = 1 - \frac{3}{t}$$

So we have

$$F_{XY}(t) = \begin{cases} 0 & t \leq 1 \\ \frac{1}{8} \left(t - 2 + \frac{1}{t}\right) & 1 < t \leq 5 \\ 1 - \frac{3}{t} & t > 5 \end{cases}$$

- (c) The characteristic function for a uniform distribution on $[a, b]$ is

$$\frac{2}{(b-a)t} \sin\left(\frac{1}{2}(b-a)t\right) \exp\left(i(a+b)\frac{t}{2}\right).$$

Using the fact that $X \perp\!\!\!\perp Y \implies \phi_{X+Y}(t) = \phi_X(t)\phi_Y(t)$, we have

$$\phi_{X+Y}(t) = \phi_X(t)\phi_Y(t) = \frac{2}{(5-1)t} \sin\left(\frac{1}{2}(5-1)t\right) \exp\left(i(5+1)\frac{t}{2}\right) \cdot \frac{2}{t} \sin\left(\frac{1}{2}t\right) \exp\left(i\frac{t}{2}\right)$$

$$= \frac{1}{2t} \sin(2t) \exp(3it) \cdot \frac{2}{t} \sin\left(\frac{1}{2}t\right) \exp\left(i\frac{t}{2}\right) = \boxed{\frac{1}{t^2} \exp\left(\frac{7}{2}it\right) \cdot \sin(2t) \sin\left(\frac{1}{2}t\right)}$$

(d) The moment-generating function for a uniform distribution on $[a, b]$ is

$$M_X(t) = \mathbb{E}(\exp(tx)) = \int_a^b \frac{1}{b-a} \cdot \exp(tx) dx = \frac{1}{b-a} \left[\frac{1}{t} \exp(tx) \right]_a^b = \frac{1}{(b-a)t} [\exp(bt) - \exp(at)]$$

Therefore the moment-generating function for X is $t^{-1}[\exp(t) - 1]$. Note that

$$\Pr(Y \leq y) = \Pr(-\log(X) \leq y) = \Pr(\log(X) \geq -y) = \Pr(X \geq e^{-y}) = \int_{\exp(-y)}^{\infty} dt = \int_{\exp(-y)}^1 dt$$

Substituting $t = e^{-u}$ (so that we have $u = -\log(t)$, $dt = -e^{-u}du$, we have

$$\Pr(Y \leq y) = - \int_y^0 e^{-u} du = [e^{-u}]_y^0 = 1 - e^{-y}$$

which is the cdf for an exponential distribution with mean 1. Therefore $Y = -\log(X) \sim \text{Exponential}(1)$, so

$$M_Y(t) = \frac{1}{1-t}$$

Using the fact that if X and Y are independent then $M_{X+Y}(t) = M_X(t)M_Y(t)$, we have

$$M_{X+Y}(t) = M_X(t)M_Y(t) = \frac{\exp(t) - 1}{t} \cdot \frac{1}{1-t} = \boxed{\frac{\exp(t) - 1}{t - t^2}}$$

- (2) **Fall 2016 Problem 2.** Let X and Y be i.i.d. exponential with mean 1. Show that for every $t > 0$ the events $\{\omega : \min\{X, Y\} > t\}$ and $\{\omega : X < Y\}$ are independent.

Solution. Note that $\min\{X, Y\} > t \iff X > t \cap Y > t$.

- $\Pr(\min\{X, Y\} > t) = \Pr(X > t \cap Y > t) = \Pr(X > t) \Pr(Y > t)$

$$= \int_t^{\infty} e^{-x} dx \int_t^{\infty} e^{-y} dy = -e^{-x}|_t^{\infty} - e^{-y}|_t^{\infty} = \boxed{e^{-2t}}$$

- $\Pr(X < Y)$: Note that in Figure 2, the region that satisfies this condition is region G_1 plus G_2 (note that X and Y are nonnegative). Therefore we can find this probability by integrating the joint pdf over that region.

$$\Pr(X < Y) = \iint_{\{G_1+G_2\}} f_{X,Y}(x,y) dx dy$$

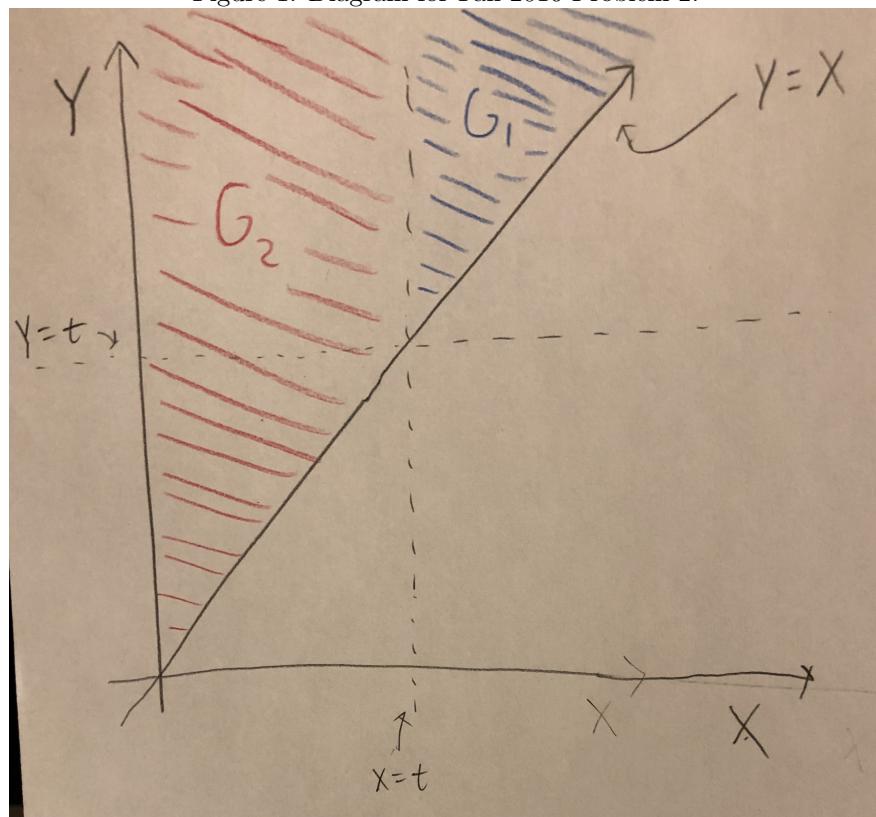
Note that the joint pdf is the probability of the marginal pdfs since X and Y are independent.

$$= \int_0^{\infty} \int_x^{\infty} e^{-x-y} dy dx = \int_0^{\infty} e^{-x} \int_x^{\infty} e^{-y} dy dx$$

$$\int_x^{\infty} e^{-y} dy = -e^{-y}|_x^{\infty} = e^{-x}$$

$$\implies \Pr(X < Y) = \int_0^{\infty} e^{-2x} dx = -\frac{1}{2} e^{-2x}|_0^{\infty} = \boxed{\frac{1}{2}}$$

Figure 1: Diagram for Fall 2016 Problem 2.



- $\Pr(X < Y \cap \min\{X, Y\} > t)$: Note that in Figure 2, the region that satisfies this condition is region G_1 . Therefore we can find this probability by integrating the joint pdf over that region.

$$\begin{aligned}\Pr(X < Y \cap \min\{X, Y\} > t) &= \int_{G_1} \int f_{X,Y}(x, y) dx dy = \int_t^\infty \int_t^y e^{-x-y} dx dy = \int_t^\infty e^{-y} \int_t^y e^{-x} dx dy \\ &\quad \int_t^y e^{-x} dx = -e^{-x} \Big|_t^y = -e^{-y} + e^{-t} \\ \implies \Pr(X < Y \cap \min\{X, Y\} > t) &= \int_t^\infty e^{-y} (-e^{-y} + e^{-t}) dy = \int_t^\infty (e^{-t-y} - e^{-2y}) dy \\ &= \frac{1}{2} e^{-2y} - e^{-t} e^{-y} \Big|_t^\infty = -\frac{1}{2} e^{-2t} - e^{-2t} = \boxed{\frac{1}{2} e^{-2t}}\end{aligned}$$

Note that

$$\Pr(X < Y \cap \min\{X, Y\} > t) = \frac{1}{2} \cdot e^{-2t} = \Pr(X < Y) \Pr(\min\{X, Y\} > t)$$

Therefore the events $\{\omega : \min\{X, Y\} > t\}$ and $\{\omega : X < Y\}$ are independent for every $t > 0$.

- (3) In a certain area, earthquakes happen at a frequency of one every four days. What is the probability that more than 100 earthquakes will occur in this area in one year (365 days)?

Solution. We can think of this as a Poisson process (see section 7.1) with $\lambda = 1/4$. Then there are two ways to obtain the answer: we can either examine the number of earthquakes in a 365 day period $N(365)$ and find the probability that $N(365) > 100$, or we can examine the number of days until the 101st earthquake T_{101} and find the probability that $T_{101} < 365$.

- (i) **Number of earthquakes in 365 days:** Let $N(t)$ be the number of earthquakes that occur in t days after the start of this process. By Theorem 7.1, $N(t) \sim \text{Poisson}(t \cdot 1/4)$. Then

$$\Pr(N(t) > 100) = \sum_{j=101}^{\infty} \frac{(365 \cdot 1/4)^j \exp(-365 \cdot 1/4)}{j!}$$

To obtain an answer for this, we can use the normal approximation to a Poisson distribution (Proposition 6.27):

$$\begin{aligned}N(t) \sim \mathcal{N}(t/4, t/4) \implies \Pr(N(365) > 100) &\approx \Pr\left(\mathcal{N}(0, 1) > \frac{100.5 - 365/4}{\sqrt{365/4}}\right) \\ &= \Pr\left(\mathcal{N}(0, 1) > \frac{100.5 - 91.25}{\sqrt{91.25}}\right) \approx \Pr\left(\mathcal{N}(0, 1) > \frac{9.25}{9.1}\right) \approx \boxed{0.1664}\end{aligned}$$

- (ii) **Number of days before 100th earthquake:** Let T_n be the number of days until the n th earthquake happens. By Corollary 7.2.1, $T_n \sim \text{Gamma}(n, 4)$. Then

$$\Pr(T_{101} < 365) = \int_0^{365} \frac{1}{\Gamma(101, 4)} x^{101-1} e^{-x/4} dx$$

To obtain an answer for this, we can use the normal approximation to a Gamma distribution (Proposition 6.50):

$$T_n \sim \mathcal{N}(404, 1616) \implies \Pr(T_{101} < 365) \approx \Pr\left(\mathcal{N}(0, 1) < \frac{365 - 404}{\sqrt{1616}}\right) \approx \Pr\left(\mathcal{N}(0, 1) < \frac{-40}{40}\right)$$

$$= \Pr(\mathcal{N}(0, 1) < -1) \approx [0.1660]$$

6.4.2 More Problems From Homework

Homework 5 Problem 4.

Let X_1, X_2, \dots be i.i.d. having moment-generating functions $M_X = M_X(t), t \in (-\infty, \infty)$. Let N be an integer-valued random variable with moment-generating function $M_N = M_N(t), t \in (-\infty, \infty)$. Assume that N is independent of all X_k and define $S = \sum_{k=1}^N X_k$. Confirm that the random variable S has the moment-generating function $M_S = M_S(t)$ defined for all $t \in (-\infty, \infty)$ and

$$M_S(t) = M_N(M_X(t))$$

Then use the result to derive the formulae

$$\mathbb{E}(S) = \mu_N \mu_X, \text{Var}(S) = (\sigma_N^2 - \mu_N) \mu_X^2 + \mu_N \sigma_X^2$$

where $\mu_N = \mathbb{E}(N)$, $\mu_X = \mathbb{E}(X_1)$, $\sigma_N^2 = \text{Var}(N)$, and $\sigma_X^2 = \text{Var}(X_1)$. How will the above computations change if we use the characteristic function ϕ_X instead of the moment-generating function M_X ?

Solution.

$$\begin{aligned} M_S(t) &= \mathbb{E}(e^{tS}) = \mathbb{E}[\mathbb{E}(e^{tS} \mid N)] = \sum_{n=0}^{\infty} \mathbb{E}(e^{tS} \mid N = n) \Pr(N = n) = \sum_{n=0}^{\infty} \mathbb{E}(e^{t(X_1+X_2+\dots+X_n)} \mid N = n) \Pr(N = n) \\ &= \sum_{n=0}^{\infty} \mathbb{E}(e^{tX_1} e^{tX_2} \cdots e^{tX_n}) \Pr(N = n) \end{aligned}$$

By independence of the X_i we have

$$= \sum_{n=0}^{\infty} \mathbb{E}(e^{tX_1}) \mathbb{E}(e^{tX_2}) \cdots \mathbb{E}(e^{tX_n}) \Pr(N = n)$$

which, since the X_i are i.i.d., can be written as

$$= \sum_{n=0}^{\infty} \mathbb{E}(e^{tX_1})^n \Pr(N = n) = \sum_{n=0}^{\infty} (M_X(t))^n \Pr(N = n)$$

But since $G_N(s) = \mathbb{E}(s^N) = \sum_{n=0}^{\infty} s^n \Pr(N = n)$, this can be written as

$$M_S(t) = G_N(M_X(t))$$

as desired. Note that

$$M'_S(t) = G'_N(M_X(t))M'_X(t)$$

$$M''_S(t) = G''_N(M_X(t))(M'_X(t))^2 + G'_N(M_X(t))M''_X(t)$$

So we have

- $\mathbb{E}(S) = M'_S(0) = G'_N(M_X(0))M'_X(0) = G'_N(1)\mathbb{E}(X_1) = \mathbb{E}(N)\mathbb{E}(X_1) = \mu_N\mu_X$
- $\text{Var}(S) = \mathbb{E}(S^2) - \mathbb{E}(S)^2 = M''_S(0) - (M'_S(0))^2$

$$\begin{aligned} &= G''_N(M_X(0))(M'_X(0))^2 + G'_N(M_X(0))M''_X(0) - \mu_N^2\mu_X^2 = G''_N(1)\mathbb{E}(X_1)^2 + G'_N(1)\text{Var}(X_1) - \mu_N^2\mu_X^2 \\ &= \mathbb{E}[N(N-1)]\mathbb{E}(X_1)^2 + \mathbb{E}(N)\text{Var}(X_1) - \mu_N^2\mu_X^2 = \mathbb{E}[N^2 - N]\mathbb{E}(X_1)^2 + \mathbb{E}(N)\text{Var}(X_1) - \mu_N^2\mu_X^2 \\ &= [\mathbb{E}(N^2) - \mathbb{E}(N)^2 + \mathbb{E}(N)^2 - \mathbb{E}(N)]\mathbb{E}(X_1)^2 + \mathbb{E}(N)\text{Var}(X_1) - \mu_N^2\mu_X^2 = \\ &= [\text{Var}(N) + \mathbb{E}(N)^2 - \mathbb{E}(N)]\mathbb{E}(X_1)^2 + \mathbb{E}(N)\text{Var}(X_1) - \mu_N^2\mu_X^2 = (\sigma_N^2 + \mu_N^2 - \mu_N)\mu_X^2 + \mu_N\sigma_X^2 - \mu_N^2\mu_X^2 \\ &= \boxed{(\sigma_N^2 - \mu_N)\mu_X^2 + \mu_N\sigma_X^2} \end{aligned}$$

To use the characteristic function ϕ_X instead of the moment generating function M_X , we would do the following:

$$\begin{aligned} \phi_S(t) &= \mathbb{E}(e^{itS}) = \mathbb{E}[\mathbb{E}(e^{itS} \mid N)] = \sum_{n=0}^{\infty} \mathbb{E}(e^{itS} \mid N = n) \Pr(N = n) = \sum_{n=0}^{\infty} \mathbb{E}(e^{it(X_1+X_2+\dots+X_n)} \mid N = n) \Pr(N = n) \\ &= \sum_{n=0}^{\infty} \mathbb{E}(e^{itX_1} e^{tX_2} \dots e^{itX_n}) \Pr(N = n) \end{aligned}$$

By independence of the X_i we have

$$= \sum_{n=0}^{\infty} \mathbb{E}(e^{itX_1}) \mathbb{E}(e^{itX_2}) \cdots \mathbb{E}(e^{itX_n}) \Pr(N = n)$$

which, since the X_i are i.i.d., can be written as

$$= \sum_{n=0}^{\infty} \mathbb{E}(e^{itX_1})^n \Pr(N = n) = \sum_{n=0}^{\infty} (\phi_X(t))^n \Pr(N = n)$$

But since $G_N(s) = \mathbb{E}(s^N) = \sum_{n=0}^{\infty} s^n \Pr(N = n)$, this can be written as

$$\phi_S(t) = G_N(\phi_X(t))$$

Homework 5 Problem 7.

- (a) Let X_1, X_2, \dots, X_n be independent with mean zero and finite third moment. Prove that

$$\mathbb{E}(X_1 + \dots + X_n)^3 = \mathbb{E}X_1^3 + \dots + \mathbb{E}X_n^3$$

Solution.

- (a) Let $\mathbb{E}(\exp(itX_i)) = \phi_{X_i}(t_i)$. Let $S_n = \sum_{i=1}^n X_i$. Then by independence the characteristic function for S_n is

$$\mathbb{E}(\exp(itS_n)) = \phi_{S_n}(t) = \prod_{i=1}^n \phi_{X_i}(t)$$

Then

$$\mathbb{E}(X_1 + X_2 + \dots + X_n)^3 = \mathbb{E}(S_n^3) = \phi_{S_n}^{(3)}(0)$$

$$= \sum_{i=1}^n \phi_{X_i}^{(3)}(0) \cdot \left(\prod_{j \in \{1, \dots, n\}, j \neq i} \phi_{X_j}(0) \right) + C \left[\sum_{i=1}^n \left(\sum_{j \in \{1, \dots, n\}, j \neq i} \phi_{X_i}^{(2)}(0) \phi_{X_j}^{(1)}(0) \right) \cdot \left(\prod_{k \in \{1, \dots, n\}, k \neq i, j} \phi_{X_k}(0) \right) \right]$$

where C is some coefficient resulting from the multinomial expansion of S_n after repeated differentiation product rules. But because $\mathbb{E}(X_i) = 0$, $\phi_{X_i}^{(1)}(0) = 0 \forall i$, so the second term goes to 0. Therefore we have

$$\mathbb{E}(X_1 + X_2 + \dots + X_n)^3 = \sum_{i=1}^n \phi_{X_i}^{(3)}(0) \cdot \left(\prod_{j \in \{1, \dots, n\}, j \neq i} \phi_{X_j}(0) \right) = \sum_{i=1}^n \mathbb{E}(X_i^3) \cdot 1^{n-1} = \sum_{i=1}^n \mathbb{E}(X_i^3)$$

as desired.

Homework 6 Problem 10.

- (a) For $p \in (0, 1)$, let $x(p)$ be the smallest number of people so that there is a better than $100 \cdot p\%$ chance to have at least two born on the same day. Find an approximate expression for $x(p)$, and sketch the graph of the function $x = x(p)$.
- (b) Repeat part (a) when you want at least three people to share a birthday.

Solution.

- (a) Let $f(x)$ be the probability of no matches in birthdays in a group of x people; that is,

$$f(x) = \frac{365 \cdot 364 \cdot 363 \cdots (365 - x + 1)}{365^x} = \frac{1}{365^x} \cdot \frac{365!}{(365 - x)!} = \left(1 - \frac{1}{365}\right) \left(1 - \frac{2}{365}\right) \cdots \left(1 - \frac{x-1}{365}\right)$$

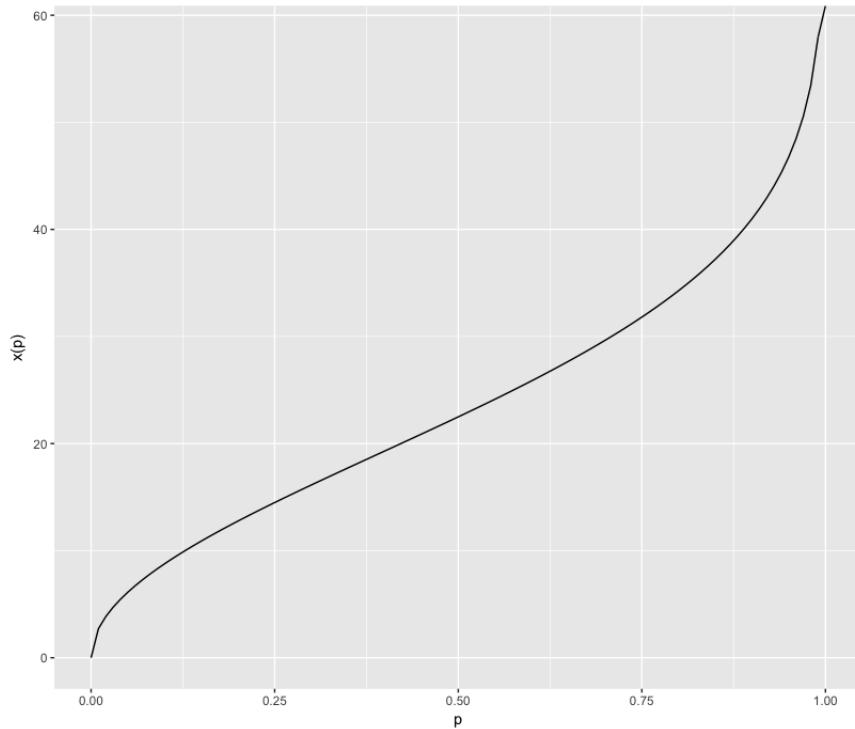
Using the first order Taylor approximation $\exp(-k/x) \approx 1 - k/x$, we have

$$\begin{aligned} f(x) &= \left(1 - \frac{1}{365}\right) \left(1 - \frac{2}{365}\right) \cdots \left(1 - \frac{x-1}{365}\right) \approx \exp(-1/365) \exp(-2/365) \cdots \exp(-(x-1)/365) \\ &= e^{-(x^2-x)/(2 \cdot 365)} \end{aligned}$$

We want the probability of a match to be at least p ; that is, $f(x) \leq 1 - p$. Setting this equal to $q = 1 - p$, we have

$$\begin{aligned} e^{-(x^2-x)/(2 \cdot 365)} = q &\iff -\frac{x^2 - x}{730} = \log(q) \iff x^2 - x + 730 \log(q) = 0 \\ &\implies x = 0.5 + \sqrt{1/4 + 730 \log(1/q)} \approx \boxed{\sqrt{2 \cdot 365 \log(1/q)}} \end{aligned}$$

where we discard the negative root because we have to have a nonnegative number of people, and we don't worry about the decimals since this is an approximation and we have to round up to the nearest whole person anyway.



- (b) For a group of three people, the Poisson approximation (see Section 6.1.8) is more convenient. The number of groups of 3 people in a room of x people is $\binom{x}{3}$. For a group of three people, the probability that all three have the same birthday is $1 \cdot 1/365 \cdot 1/365 = 365^{-2}$. Therefore we can think of the number of matches of three people as distributed Poisson with expectation $\binom{x}{3} \cdot 365^{-2}$. Then we have the probability of at least one “success” (triplet with three matched birthdays) is

$$1 - \frac{\exp(-\lambda)\lambda^0}{0!} = 1 - \exp\left(-\binom{x}{3} \cdot 365^{-2}\right)$$

We set this equal to p and solve:

$$\begin{aligned} p = 1 - \exp\left(-\binom{x}{3} \cdot 365^{-2}\right) &\iff -\binom{x}{3} \cdot 365^{-2} = \log(1-p) \iff \frac{x!}{(x-3)!3!} = 365^2 \cdot \log\left(\frac{1}{1-p}\right) \\ &\iff x(x-1)(x-2) = 6 \cdot 365^2 \cdot \log\left(\frac{1}{1-p}\right) \iff (x^2-x)(x-2) = x^3 - 3x^2 + 2x = 6 \cdot 365^2 \cdot \log\left(\frac{1}{1-p}\right) \end{aligned}$$

This has a unique real solution, but it is hard to find.

Question:

Let X, Y, Z be independent uniform on $(0, 1)$. Compute the cdfs of XY , X/Y , and XY/Z .

Solution.

Using the information from part (a), and the fact that $f_X(x) = 1$ (for $x \in [0, 1]$) and likewise for $f_Y(y)$:

- XY :

$$\begin{aligned}
F_{XY}(z) &= \int_0^\infty f_X(x) \int_{-\infty}^{z/x} f_Y(y) dy dx - \int_{-\infty}^0 f_X(x) \int_\infty^{z/x} f_Y(y) dy dx \\
&= \int_0^1 [(z/x)\mathbf{1}_{\{0 < z/x \leq 1\}} + \mathbf{1}_{\{z/x > 1\}}] dx = \int_0^1 [(z/x)\mathbf{1}_{\{z \leq x\}} + \mathbf{1}_{\{z > x\}}] dx = \int_0^z dx + \int_z^1 (z/x) dx \\
&= z + z \log(x) \Big|_z^1 = z + z \log(1) - z \log(z) = z(1 - \log(z))
\end{aligned}$$

$$\implies F_{XY}(z) = \begin{cases} 0 & z \leq 0 \\ z(1 - \log(z)) & 0 < z \leq 1 \\ 1 & z > 1 \end{cases}$$

- X/Y :

$$\begin{aligned}
F_{X/Y}(z) &= \int_0^\infty f_Y(y) \int_{-\infty}^{zy} f_X(x) dx dy - \int_{-\infty}^0 f_Y(y) \int_\infty^{zy} f_X(x) dx dy \\
&= \int_0^1 [zy\mathbf{1}_{\{0 < zy \leq 1\}} + \mathbf{1}_{\{zy > 1\}}] dy = \int_0^1 [zy\mathbf{1}_{\{y > 0 \cap y \leq 1/z\}} + \mathbf{1}_{\{y > 1/z\}}] dy = \int_0^{1/z} zy \cdot dy + \int_{1/z}^1 dy \\
&= \frac{zy^2}{2} \Big|_0^{1/z} + (1 - 1/z) = \frac{z}{2z^2} + 1 - \frac{2}{2z} = 1 - \frac{1}{2z} \\
\implies F_{X/Y}(z) &= \begin{cases} 0 & z \leq 0 \\ 1 - \frac{1}{2z} & 0 < z \leq 1 \\ \frac{z}{2z^2} & z > 1 \end{cases}
\end{aligned}$$

- XY/Z : Consider this the cdf of the quotient of $W = XY$ and Z .

$$\begin{aligned}
F_U(u) &= \int_0^\infty f_Z(z) \int_{-\infty}^{uz} f_W(w) dw dz - \int_{-\infty}^0 f_Z(z) \int_\infty^{uz} f_W(w) dw dz \\
&= \int_0^1 \int_0^{uz} -\log(w)\mathbf{1}_{\{0 < uz \leq 1\}} dw dz = \int_0^1 -[w \log(w) - w]_0^{uz} \mathbf{1}_{\{0 < z \leq 1/u\}} dz \\
&= \int_0^{1/u} uz [1 - \log(uz)] dz = \frac{u}{4} z^2 (3 - 2 \log(uz)) \Big|_0^{1/u} = \frac{u}{4u^2} (3 - 2 \log(1)) - 0 = \frac{3}{4u} \\
\implies F_{XY/Z}(u) &= \begin{cases} 0 & u \leq 0 \\ \frac{3}{4u} & 0 < u \leq 3/4 \\ 1 & u > 3/4 \end{cases}
\end{aligned}$$

7 Stochastic Processes

These notes are based on my notes from *Time Series and Panel Data Econometrics* (1st edition) by M. Hashem Pesaran and coursework for Economics 613: Economic and Financial Time Series I at USC, as well as notes from *Probability and Random Processes* by Grimmett and Stirzaker.

7.1 Poisson Processes

Definition 7.1. A **Poisson process with intensity** λ is a process $N = \{N(t) : t \geq 0\}$ taking values in $S = \{0, 1, 2, \dots\}$ such that

(a) $N(0) = 0$; if $s < t$ then $N(s) \leq N(t)$.

$$(b) \Pr(N(t+h) = n+m \mid N(t) = n) = \begin{cases} \lambda h + o(h) & \text{if } m = 1, \\ o(h) & \text{if } m > 1; \\ 1 - \lambda h + o(h) & \text{if } m = 0 \end{cases}$$

(c) If $s < t$, the number $N(t) - N(s)$ of emissions in the interval $(s, t]$ is independent of the times of emissions during $[0, s]$.

Remark. λ can be interpreted as the average or long-run frequency of the Poisson process.

Theorem 7.1. (Grimmett and Stirzaker theorem 6.8.2.) Let $N(t)$ be a Poisson process with intensity λ . Then $N(t)$ has the Poisson distribution with parameter λt ; that is,

$$\Pr(N(t) = j) = \frac{(\lambda t)^j \exp(-\lambda t)}{j!}, \quad j = 0, 1, 2, \dots$$

Proof. See Grimmett and Stirzaker section 6.8.2, page 247. □

Definition 7.2. Let $N(t)$ be a Poisson process with intensity λ . Let T_0, T_1, \dots be given by

$$T_0 = 0, T_n = \inf\{t : N(t) = n\} \tag{7}$$

so that T_n is the time of the n th arrival. The **interarrival times** are the random variables X_1, X_2, \dots given by

$$X_n = T_n - T_{n-1}. \tag{8}$$

Remark. From knowledge of N , we can find the values of X_1, X_2, \dots by (7) and (8). Conversely, we can construct N from knowledge of the X_i by

$$T_n = \sum_{i=1}^n X_i, \quad N(t) = \max\{n : T_n \leq t\} \tag{9}$$

Theorem 7.2. (Grimmett and Stirzaker theorem 6.8.10.) Let $N(t)$ be a Poisson process with intensity λ . Let T_0, T_1, \dots be given by (7) and let X_n be given by (8). Then the random variables $\{X_n\}$ are independent, each having an exponential distribution with parameter λ .

Proof. See Grimmett and Stirzaker section 6.8.2, page 249. \square

Corollary 7.2.1. Let $N(t)$ be a Poisson process with intensity λ . Let T_0, T_1, \dots be given by (7). Then $T_n \sim \text{Gamma}(n, \lambda^{-1})$.

Proof. By (9), $T_n = \sum_{i=1}^n X_i$. $X_i \sim \text{Exponential}(\lambda)$ by Theorem 7.2, which means $X_i \sim \text{Gamma}(1, \lambda^{-1})$. Then by Proposition 6.49, $T_n \sim \text{Gamma}(n, \lambda^{-1})$. \square

7.2 Martingales

Definition. Let $\{y_t\}_{t=0}^\infty$ be a sequence of random variables, and let Ω_t denote the information set available at date t , which at least contains $\{y_t, y_{t-1}, y_{t-2}, \dots\}$. If $\mathbb{E}(y_t \mid \Omega_{t-1}) = y_{t-1}$ holds then $\{y_t\}$ is a martingale process with respect to Ω_t .

Definition. Let $\{y_t\}_{t=1}^\infty$ be a sequence of random variables, and let Ω_t denote the information set available at date t , which at least contains $\{y_t, y_{t-1}, y_{t-2}, \dots\}$. If $\mathbb{E}(y_t \mid \Omega_{t-1}) = 0$, then $\{y_t\}$ is a martingale difference process with respect to Ω_t .

7.3 Brownian Motion

Appendix B.13, Brownian motion. A standard Brownian motion $b(\cdot)$ is a continuous-time stochastic process associating each date $a \in [0, 1]$ with the scalar $b(a)$ such that

- (i) $b(0) = 0$
- (ii) For any dates $0 \leq a_1 \leq a_2 \leq \dots \leq a_k \leq 1$ the changes $[b(a_2) - b(a_1)], [b(a_3) - b(a_2)], \dots, [b(a_k) - b(a_{k-1})]$ are independent multivariate Gaussian with $b(a) - b(s) \sim \mathcal{N}(0, a - s)$.
- (iii) For any given realization, $b(a)$ is continuous in a with probability 1.

Other continuous time processes can be generated from the standard Brownian motion. For example, a Brownian motion with variance σ^2 can be obtained as

$$w(a) = \sigma b(a)$$

where $b(a)$ is a standard Brownian motion.

The continuous time process

$$\mathbf{w}(a) = \boldsymbol{\Sigma}^{1/2} \mathbf{b}(a)$$

is a Brownian motion with covariance matrix Σ .

Definition 26 (Wiener process). Let $\Delta w(t)$ be the change in $w(t)$ during the time interval dt . Then $w(t)$ is said to follow a Wiener process if

$$\Delta w(t) = \epsilon_t \sqrt{dt}, \quad \epsilon_t \sim IID(0, 1)$$

and $w(t)$ denotes the value of the $w(\cdot)$ at date t . Clearly,

$$\mathbb{E}[\Delta w(t)] = 0, \text{ and } \text{Var}[\Delta w(t)] = dt$$

Theorem 7.3. Donsker's Theorem, Theorem 43, p.335, Section 15.6.3. Let $a \in [0, 1)$, $t \in [0, T]$, and suppose $(J - 1)/T \leq a < J/T$, $J = 1, 2, \dots, T$. Define

$$R_T(a) = \frac{1}{\sqrt{T}} s_{[Ta]}$$

where

$$s_{[Ta]} = \epsilon_1 + \epsilon_2 + \dots + \epsilon_{[Ta]}$$

$[Ta]$ denotes the largest integer part of Ta and $s_{[Ta]} = 0$ if $[Ta] = 0$. Then $R_T(a)$ weakly converges to $w(a)$, i.e.,

$$R_T(a) \rightarrow w(a)$$

where $w(a)$ is a Wiener process. Note that when $a = 1$, $R_T(1) = 1/\sqrt{T} \cdot S_{[T]} = 1/\sqrt{T} \cdot (\epsilon_1 + \epsilon_2 + \dots + \epsilon_T)$. Since ϵ_t 's are IID, by the central limit theorem, $R_T(1) \rightarrow \mathcal{N}(0, 1)$.

Similar (Theorem 2.1 in Phillips and Durlaf (1986)): Let $\{u_t\}$ be a sequence satisfying $\mathbb{E}(u_t) = 0$, $\gamma(0) = \mathbb{E}(T^{-1}S_t^2) \rightarrow \sigma^2 < \infty$ as $T \rightarrow \infty$, $\{u_t\}$ is square summable, $\sup_t \{\mathbb{E}(|u_t|^\beta)\} < \infty$ for some $2 \leq \beta < \infty$ and all t , $\gamma(h) = \mathbb{E}(T^{-1}(y_t - y_{t-h})^2) \rightarrow K_h < \infty$ as $\min\{h, T\} \rightarrow \infty$. Then $X_T(t) \Rightarrow W(t)$ as $T \rightarrow \infty$, where $W(t)$ is a Wiener process.

Theorem 7.4. Continuous Mapping Theorem (Theorem 44 of Pesaran in 15.6.3). Let $a \in [0, 1)$, $i \in [0, n]$, and suppose $(J - 1)/n \leq a < J/n$, $J = 1, 2, \dots, n$. Define $R_n(a) = n^{-1/2} S_{[n \cdot a]}$. If $f(\cdot)$ is continuous over $[0, 1]$, then

$$f[R_n(a)] \xrightarrow{d} f[w(a)]$$

8 Asymptotics and Convergence

These notes are based on my notes from chapter 8 of *Time Series and Panel Data Econometrics* (1st edition) by M. Hashem Pesaran and coursework for Economics 613: Economic and Financial Time Series I at USC, as well as Math 505A and Math 541A at USC and chapter 7 from *Probability and Random Processes* (Grimmett and Stirzaker) 3rd edition.

8.1 Preliminaries (5.9 and 7.1, Grimmett and Stirzaker)

Definition 8.1. Definition 7.1.4, Grimmett and Stirzaker. If for all $x \in [0, 1]$ the sequence $\{f_n(x)\}$ of real numbers satisfies $f_n(x) \rightarrow f(x)$ as $n \rightarrow \infty$ then we say $f_n \rightarrow f$ **pointwise**.

Remark. In practice pointwise convergence is often not useful for functions because a sequence of functions may be continuous while its limit is not. For instance, consider $\{f_n : f_n = x^n \forall x \in [0, 1]\}$. Then f_n is continuous for all n but

$$\lim_{n \rightarrow \infty} f_n = \begin{cases} 0 & x \leq 1 \\ 1 & x = 1 \end{cases}$$

Instead, the following definition is often more useful.

Definition 8.2. (from class notes.) We say that f_n **uniformly converges to f on $[a, b]$** if for every $\epsilon > 0$ there exists N such that for every $n > N$,

$$\forall x \in [a, b] \quad |f_n(x) - f(x)| < \epsilon$$

Definition 8.3. (Definition 7.1.5, Grimmett and Stirzaker.) Let V be a collection of functions mapping $[0, 1]$ into \mathbb{R} and assume V is endowed with a function $\|\cdot\| : V \rightarrow \mathbb{R}$ satisfying

- (a) $\|f\| \geq 0$ for all $f \in V$
- (b) $\|f\| = 0$ if and only if f is the zero function (or equivalent to it)
- (c) $\|af\| = |a| \cdot \|f\|$ for all $a \in \mathbb{R}$, $f \in V$
- (d) $\|f + g\| \leq \|f\| + \|g\|$ (triangle inequality)

The function $\|\cdot\|$ is called a **norm**. If $\{f_n\}$ is a sequence of members of V then we say that $f_n \rightarrow f$ **with respect to the norm $\|\cdot\|$** if $\|f_n - f\| \rightarrow 0$ as $n \rightarrow \infty$.

Definition 8.4. (Definition 7.16, Grimmett and Stirzaker.) Let $\epsilon > 0$ be prescribed, and define the distance between two functions $g, h : [0, 1] \rightarrow \mathbb{R}$ by

$$d_\epsilon(g, h) = \int_E dx$$

where $E = \{u \in [0, 1] : |g(u) - h(u)| > \epsilon\}$. We say that $f_n \rightarrow f$ **in measure** if

$$d_\epsilon(f_n, f) \rightarrow 0 \text{ as } n \rightarrow \infty \text{ for all } \epsilon > 0$$

Theorem 8.1. Inversion Theorem (Theorem 5.9.2, Grimmett and Stirzaker). Let X have distribution function F and characteristic function ϕ . Define $\bar{F} : \mathbb{R} \rightarrow [0, 1]$ by

$$\bar{F}(x) = \frac{1}{2} [F(x) + \lim_{y \rightarrow x^-} F(y)]$$

Then

$$\bar{F}(b) - \bar{F}(a) = \lim_{N \rightarrow \infty} \int_{-N}^N \frac{\exp(-iat) - \exp(-ibt)}{2\pi i t} \cdot \phi(t) dt$$

Proof. See Kingman and Taylor (1966). \square

Corollary 8.1.1. Corollary 5.9.3. Random variables X and Y have the same characteristic function if and only if they have the same distribution function.

Proof. Available in Grimmett and Stirzaker section 5.9, pp. 189 - 190. \square

Definition 8.5. (Definition 5.9.4, Grimmett and Stirzaker.) We say that the sequence F_1, F_2, \dots of distribution functions **converges** to the distribution function F (written $F_n \rightarrow F$) if $F(x) = \lim_{n \rightarrow \infty} F_n(x)$ at each point x where F is continuous.

Theorem 8.2. Continuity theorem (Thereom 5.9.5; in notes from Friday 10/26, Lecture 28). Suppose that F_1, F_2, \dots is a sequence of distribution functions with corresponding characteristic functions ϕ_1, ϕ_2, \dots

- (a) If $F_n(x) \rightarrow F(x)$ for some distribution function F with characteristic function ϕ (at x where F is continuous), then $\phi_n(t) \rightarrow \phi(t)$ for all t .
- (b) Conversely, if $\phi(t) = \lim_{n \rightarrow \infty} \phi_n(t)$ exists and $\phi(t)$ is continuous at $t = 0$, then ϕ is the characteristic function of some distribution function F , and $F_n \rightarrow F$.

Proof. See Kingman and Taylor (1966). \square

8.2 Inequalities (8.6 of Pesaran)

Inequalities

- Probabilities

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Lemma 8.3. Markov's Inequality (Grimmett and Stirzaker p. 311, 319) : Let $X : \Omega \rightarrow [-\infty, \infty]$ be a random variable. Then for all $a > 0$,

$$\Pr(|X| \geq a) \leq \frac{\mathbb{E}(|X|)}{a}$$

Proof. Note $a \cdot \mathbf{1}_{\{|X| \geq a\}} \leq |X|$, where $\mathbf{1}$ is the indicator function. Dividing both sides by a and taking expectations yields the result. \square

Corollary 8.3.1. If n is a positive integer, then

$$\Pr(|X| \geq t) \leq \frac{\mathbb{E}(|X|^n)}{t^n} \quad \forall t > 0$$

Proof. By Markov's Inequality (Theorem 8.3),

$$\Pr(|X| \geq t) = \Pr(|X|^n \geq t^n) \leq \frac{\mathbb{E}(|X|^n)}{t^n}$$

\square

Theorem 8.4. Chebyshev's Inequality: (probability p. 319) Let $X : \Omega \rightarrow [-\infty, \infty]$ be an (integrable) random variable with $\mathbb{E}(X^2) < \infty$. Then for any real number $k > 0$

$$\Pr(|X - \mathbb{E}(X)| \geq k\sqrt{\text{Var}(X)}) \leq \frac{1}{k^2}$$

This can also be written as

$$\Pr(|X - \mathbb{E}(X)| \geq k) \leq \frac{\text{Var}(X)}{k^2}$$

(Can be used to demonstrate consistency of estimators: if we can show that as $T \rightarrow \infty$ $\text{Var}(X) = \sigma^2 \rightarrow 0$, then this implies $\Pr(|X - \mu| \geq k\sigma) \rightarrow 0$ as $T \rightarrow \infty$, showing consistency.)

Theorem 8.5. Chernoff For $x \geq 0$, $a > 0$, $\forall t > 0$,

$$\Pr(X \geq a) = \Pr(e^{tX} \geq e^{ta}) \leq \frac{\mathbb{E}(e^{tX})}{e^{ta}}$$

• Moments

Theorem 8.6. Cauchy-Schwarz (and Bunyakovsky). If X and Y are random variables with finite variance then

$$\mathbb{E}(XY)^2 \leq \mathbb{E}(X^2)\mathbb{E}(Y^2)$$

Note that this can be a corollary of Theorem 8.10 with $p = q = 2$. We can also prove this theorem on its own in a different one. We first prove a useful result.

Lemma 8.7. If $\text{Var}(X) = 0$ then X is almost surely constant; that is, $\Pr(X = a) = 1$ for some $a \in \mathbb{R}$.

Proof. Note that because $\text{Var}(X) = 0 < \infty$, we know that $\mathbb{E}(X)$ and $\mathbb{E}(X^2)$ exist. We have

$$\text{Var}(X) = \mathbb{E}[(X - \mathbb{E}(X))^2] = 0$$

Let $Y = (X - \mathbb{E}(X))^2$. Note that $Y = (X - \mathbb{E}(X))^2 \geq 0$ and that $\mathbb{E}(Y) = \text{Var}(X) = 0$. Therefore $\Pr(Y = 0) = 1$, so $\Pr(Y \neq 0) = 0$. To see why, in the case that X is discrete,

$$\mathbb{E}(Y) = \sum_{k=0}^{\infty} k \cdot \Pr(Y = k) = \text{Var}(X) = 0$$

which is true if and only if $\Pr(Y = k) = 0$ for all $k > 0$. Since we already showed that $\Pr(Y < 0) = 0$, it follows that $\Pr(Y = 0) = 1$. In the continuous case,

$$\mathbb{E}(Y) = \int_0^{\infty} y \cdot f_Y(y) dy = \text{Var}(X) = 0$$

which implies that $f_Y(x) = 0$ for all $x > 0$. Again, since $\Pr(Y < 0) = 0$, we have $\Pr(Y \neq 0) = 0$. But $Y = 0 \iff X = \mathbb{E}(X)$ so we have $\Pr(X = \mathbb{E}(X)) = 1$. \square

Remark. Note that Lemma 8.7 along with Proposition 6.22 imply that X has variance 0 if and only if it is (almost surely) constant.

We are now ready to prove the Cauchy-Schwarz Inequality.

Proof. if $\mathbb{E}(X^2) = 0$ or $\mathbb{E}(Y^2) = 0$, the Cauchy-Schwarz Inequality follows immediately. To see why, suppose without loss of generality that $\mathbb{E}(X^2) = 0$. Then the right side is 0. Also, $0 \leq \text{Var}(X) = \mathbb{E}(X^2) - \mathbb{E}(X)^2 = -\mathbb{E}(X)^2$. Since $\mathbb{E}(X)^2 \geq 0$, we must have $\mathbb{E}(X)^2 = 0$ and therefore $\text{Var}(X) = 0$. Therefore by Lemma 8.7, X is almost surely constant, which means that $\text{Cov}(X, Y) = 0$.

In the case that $\mathbb{E}(X^2) > 0$ and $\mathbb{E}(Y^2) > 0$, for $a, b \in \mathbb{R}$, let $Z = aX - bY$. Then

$$0 \leq \mathbb{E}(Z^2) = a^2 \mathbb{E}(X^2) - 2ab \mathbb{E}(XY) + b^2 \mathbb{E}(Y^2) \quad (10)$$

The right side of (10) is quadratic in a . Because it is greater than or equal to zero, it has at most one real root, which means its discriminant must be non-positive. That is, if $b \neq 0$,

$$(-2b \mathbb{E}(XY))^2 - 4b^2 \mathbb{E}(X^2) \mathbb{E}(Y^2) \leq 0 \iff \mathbb{E}(XY)^2 - \mathbb{E}(X^2) \mathbb{E}(Y^2) \leq 0$$

which yields the result. Note that equality holds if and only if $\Pr(aX = bY) = 1$ because the discriminant is zero if and only if the quadratic has a real root, which occurs if and only if

$$\mathbb{E}[(aX - bY)^2] = 0$$

which is true if and only if $\Pr(aX = bY) = 1$ by Lemma 8.7 and Proposition 6.22. \square

– Krylov

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Definition 8.6. Let $\phi : \mathbb{R} \rightarrow \mathbb{R}$. We say that ϕ is **convex** if for any $x, y \in \mathbb{R}$ and for any $t \in [0, 1]$, we have

$$\phi(tx + (1 - t)y) \leq t\phi(x) + (1 - t)\phi(y)$$

Proposition 8.8. Let $\phi : \mathbb{R} \rightarrow \mathbb{R}$. Then ϕ is convex if and only if for any $y \in \mathbb{R}$ there exists a constant a and there exists a function $L : \mathbb{R} \rightarrow \mathbb{R}$ defined by $L(x) = a(x - y) + \phi(y)$, $x \in \mathbb{R}$ such that $L(y) = \phi(y)$ and such that $L(x) \leq \phi(x)$ for all $x \in \mathbb{R}$.

Proof. Hint: Suppose ϕ is convex. If x is fixed and y varies, show that $\frac{\phi(y) - \phi(x)}{y - x}$ increases as y increases. Draw a picture. What slope a should L have at x ? \square

Theorem 8.9. Jensen's Inequality (Grimmett and Stirzaker p.181, 349) Let $X : \Omega \rightarrow [-\infty, \infty]$ be a random variable. Let $u : \mathbb{R} \rightarrow \mathbb{R}$ be convex. If $\mathbb{E}|X| < \infty$ and $\mathbb{E}|u(X)| < \infty$, then

$$u(\mathbb{E}(X)) \leq \mathbb{E}(u(X))$$

Proof. Hint: use Proposition 8.8 with $y := \mathbb{E}(X)$. \square

Corollary 8.9.1. Triangle Inequality. Let $X : \Omega \rightarrow [-\infty, \infty]$ be a random variable. If $\mathbb{E}|X| < \infty$, then $|\mathbb{E}(X)| \leq \mathbb{E}(|X|)$.

Proof. Follows immediately from Jensen's Inequality using $u(X) = |X|$. \square

Corollary 8.9.2. If $\mathbb{E}(X^2) < \infty$, then $\mathbb{E}(|X|) < \infty$, so $\mathbb{E}(X) \in \mathbb{R}$.

Proof. Follows immediately from Jensen's Inequality using $u(X) = X^2$. \square

Theorem 8.10. Holder (Grimmett and Stirzaker p. p. 143, 319) Generalization of Cauchy-Schwarz. For $p, q > 1$ satisfying $1/p + 1/q = 1$ we have

$$\mathbb{E}(|XY|) \leq (\mathbb{E}(|X^p|))^{1/p} (\mathbb{E}(|X^q|))^{1/q}$$

Proof. Assume without loss of generality that $\|X\|_p = \|Y\|_q = 1$. Also, the case $p = 1, q = \infty$ follows from the triangle inequality, so we assume $1 < p < \infty$. From concavity of the log function, we have

$$\begin{aligned} \log((x^p)^{1/p}(y^q)^{1/q}) &= (1/p)\log(x^p) + (1/q)\log(y^q) \\ &\leq \log\left(\frac{1}{p}x^p + \frac{1}{q}y^q\right) \\ \implies (x^p)^{1/p}(y^q)^{1/q} &\leq \frac{1}{p}x^p + \frac{1}{q}y^q \end{aligned}$$

Fixing an $\omega \in \Omega$, we have

$$|X(\omega)Y(\omega)| = (|X(\omega)|^p)^{1/p} (|Y(\omega)|^q)^{1/q} \leq \frac{1}{p}|X(\omega)|^p + \frac{1}{q}|Y(\omega)|^q$$

Integrating we have... \square

Theorem 8.11. Minkowski (Grimmett and Stirzaker p. p. 143) For $p \geq 1$,

$$[\mathbb{E}(|X + Y|^p)]^{1/p} \leq (\mathbb{E}|X^p|)^{1/p} + (\mathbb{E}|Y^p|)^{1/p}$$

- Useful for showing lower order moments are finite (e.g. finite variance implies finite mean).

Lemma 8.12. Lyapunov's Inequality (Grimmett and Stirzaker p. 143). For $0 < r \leq s < \infty$,

$$\mathbb{E}(|X|^r)^{1/r} \leq \mathbb{E}(|X|^s)^{1/s}$$

—

Theorem 8.13. Triangle Inequality: Let $X, Y : \Omega \rightarrow \mathbb{R}$ be random variables. Let $1 \leq p \leq \infty$. Then

$$\|X + Y\|_p \leq \|X\|_p + \|Y\|_p, 1 \leq p \leq \infty$$

Proof. The case $p = \infty$ follows from the scalar triangle inequality, so assume $1 \leq p < \infty$. By scaling, we may assume $\|X\|_p = 1 - t$, $\|Y\|_p = t$, for some $t \in (0, 1)$ (zeroes and infinities being trivial). Define $V := X/(1 - t)$, $W := Y/t$. Then by convexity of $x \rightarrow |x|^p$ on \mathbb{R} ,

$$|(1 - t)V(\omega) + t(W(\omega))|^p \leq (1 - t)|V(\omega)|^p + t|W(\omega)|^p$$

Take expectation of both sides:

$$\mathbb{E}|X + Y|^p \leq (1 - t)^{1-p}\mathbb{E}(|X|^p) + t^{1-p}\mathbb{E}(|Y|^p)$$

Since $\|X\|_p = t$, $\|Y\|_p = 1 - t$, we have that the right side is $(1 - t)^{1-p}t^p + t^{1-p}(1 - t)^p = 1$. (Note: $\|Y\|_p = t$, $\mathbb{E}|Y|^p = t^p$, $\|X\|_p = 1 - t$). Therefore

$$(\mathbb{E}|X + Y|^p)^{1/p} = \|X + Y\|_p \leq 1$$

□

Monotone convergence theorem.

Dominated convergence theorem.

8.3 Modes of Convergence (7.2 of Grimmett and Stirzaker, 8.2 and 8.4 of Pesaran)

Let $\{X_n\} = \{X_1, X_2, \dots\}$ and X be random variables defined on a probability space $(\Omega, \mathcal{F}, \mathbb{P})$.

Definition 8.7. Convergence in probability. $\{X_n\}$ is said to **converge in probability** to X if

- Grimmett and Strizaker definition:

$$\lim_{n \rightarrow \infty} \Pr(|X_n - X| > \epsilon) = 0, \text{ for every } \epsilon > 0$$

- Pesaran definition:

$$\lim_{n \rightarrow \infty} \Pr(|X_n - X| < \epsilon) = 1, \text{ for every } \epsilon > 0$$

This mode of convergence is also often denoted by $X_n \xrightarrow{p} X$ and when X is a fixed constant it is referred to as the **probability limit of X_n** , written as $\text{Plim}(X_n) = x$, as $n \rightarrow \infty$.

The above concept is readily extended to multivariate cases where $\{\mathbf{X}_n, n = 1, 2, \dots\}$ denote m -dimensional vectors of random variables. Then the condition is

$$\lim_{n \rightarrow \infty} \Pr(\|\mathbf{X}_n - \mathbf{X}\| < \epsilon) = 1, \text{ for every } \epsilon > 0$$

where $\|\cdot\|$ denotes an appropriate norm (say ℓ_2). Convergence in probability is often referred to as "weak convergence" (in contrast to convergence with probability 1, below).

Definition 8.8. Convergence with probability 1 or almost surely. The sequence of random variables $\{X_n\}$ is said to **converge with probability 1** (or **almost surely**) to X if

- (505A class notes definition)

$$\Pr\left(\{\omega \in \Omega : \lim_{n \rightarrow \infty} X_n(\omega) = X(\omega)\}\right) = 1$$

(Note: pointwise convergence can hardly ever be shown here and is not useful.)

- Grimmett and Stirzaker textbook definition:

$$\Pr\left(\{\omega \in \Omega : X_n(\omega) \rightarrow X(\omega) \text{ as } n \rightarrow \infty\}\right) = 1$$

- Pesaran textbook definition:

$$\Pr\left(\lim_{n \rightarrow \infty} X_n = X\right) = 1$$

This is often written as $X_n \xrightarrow{w.p.1} X$ or $X_n \xrightarrow{a.s.} X$. An equivalent condition for convergence with probability 1 is given by

$$\lim_{n \rightarrow \infty} \Pr(|X_m - X| < \epsilon, \text{ for all } m \geq n) = 1, \text{ for every } \epsilon > 0$$

which shows that convergence in probability is a special case of convergence with probability 1 (obtained by setting $m = n$). Convergence with probability 1 is stronger than convergence in probability and is often referred to as "strong convergence."

Definition 8.9. Convergence in r -th mean. $X_n \rightarrow X$ in **r th mean** where $r \geq 1$ if $\mathbb{E}|X_n|^r < \infty$ for all n and

$$\lim_{n \rightarrow \infty} \mathbb{E}(|X_n - X|^r) = 0$$

Convergence in r th mean is often written $X_n \xrightarrow{r} X$.

Definition 8.10. Convergence in Distribution. Let X_1, X_2, \dots have distribution functions $F_1(\cdot), F_2(\cdot), \dots$ respectively. Then X_n is said to **converge in distribution to X** if

$$\lim_{n \rightarrow \infty} \Pr(X_n \leq u) = \Pr(X \leq u)$$

for all u at which $F_X(x) = \Pr(X \leq x)$ is continuous. This can also be written

$$\lim_{n \rightarrow \infty} F_n(u) = F(u)$$

for all u at which F is continuous. Convergence in distribution is usually denoted by $X_n \xrightarrow{d} X$, $X_n \xrightarrow{L} X$, or $F_n \Rightarrow F$. By the Continuity Theorem (section 8.1), this is equivalent to

$$\lim_{n \rightarrow \infty} \phi_{X_n}(t) = \phi_X(t), \quad t \in \mathbb{R}$$

Theorem 8.14. (Theorem 7.2.3, Grimmett and Stirzaker.) The following implications hold:

- $(X_n \xrightarrow{a.s.} X) \implies (X_n \xrightarrow{p} X)$
- $(X_n \xrightarrow{r} X) \implies (X_n \xrightarrow{p} X)$ for any $r \geq 1$
- $(X_n \xrightarrow{p} X) \implies (X_n \xrightarrow{d} X)$

Also, if $r > s \geq 1$, then $(X_n \xrightarrow{r} X) \implies (X_n \xrightarrow{s} X)$. No other implications hold in general.

Theorem 8.15. Some exceptions (Theorem 7.2.4).

- If $X_n \xrightarrow{d} c$ where c is constant, then $X_n \xrightarrow{p} c$.
- If $X_n \xrightarrow{p} X$ and $\Pr(|X_n| \leq k) = 0$ for all n and some k , then $X_n \xrightarrow{r} X$ for all $r \geq 1$.
- If $P_n(\epsilon) = \Pr(|X_n - X| > \epsilon)$ satisfies $\sum_n P_n(\epsilon) < \infty$ for all $\epsilon > 0$, then $X_n \xrightarrow{a.s.} X$.

Proof. (Part (c).) Let $A_n(\epsilon) = \{|X_n - X| > \epsilon\}$ (so that $P_n(\epsilon) = \Pr[A_n(\epsilon)]$), and let $B_m(\epsilon) = \bigcup_{n \geq m} A_n(\epsilon)$. Then

$$\Pr(B_m(\epsilon)) \leq \sum_{n=m}^{\infty} \Pr(A_n(\epsilon))$$

so $\lim_{m \rightarrow \infty} \Pr(B_m(\epsilon)) = 0$ whenever $\sum_n \Pr(A_n(\epsilon)) < \infty$. See also Lemma 8.17 part (b). \square

8.4 More on convergence (7.2 of Grimmett and Stirzaker)

Other theorems to include: Fatou's Lemma, Fubini's Theorem, Kolmogorov's Maximal Inequality, Kolmogorov Three-Series Test, Lindeberg Feller Central Limit Theorem, **this and more at beginning of Mike's 505A qual solutions.**

Definition 8.11. Cauchy Convergence. We say that the sequence $\{X_n : n \geq 1\}$ of random variables on the probability space $(\Omega, \mathcal{F}, \mathbb{P})$ is **almost surely Cauchy convergent** if

$$\Pr(\{\omega \in \Omega : X_m(\omega) - X_n(\omega) \rightarrow 0 \text{ as } m, n \rightarrow \infty\}) = 1$$

That is, the set of points ω of the sample space for which the real sequence $\{X_n(\omega) : n \geq 1\}$ is Cauchy convergent is an event having probability 1.

Lemma 8.16. (Lemma 7.2.6 from Grimmett and Stirzaker)

- (a) If $r > s \geq 1$ and $X_n \xrightarrow{r} X$, then $X_n \xrightarrow{s} X$.
- (b) If $X_n \xrightarrow{1} X$ then $X_n \xrightarrow{p} X$.

The converse assertions fail in general.

Proof. (a) Using Lyapunov's Inequality (Lemma 8.12), if $r > s \geq 1$

$$[\mathbb{E}(|X_n - X|^s)]^{1/s} \leq [\mathbb{E}(|X_n - X|^r)]^{1/r}$$

Therefore if $X_n \xrightarrow{r} X$ (meaning $\lim_{n \rightarrow \infty} \mathbb{E}(|X_n - X|^r) = 0$), (then $\lim_{n \rightarrow \infty} \mathbb{E}(|X_n - X|^s) = 0$, so $X_n \xrightarrow{s} X$. We show the converse fails by counterexample:

$$X_n = \begin{cases} n & \text{with probability } n^{(-1/2)(r+s)} \\ 0 & \text{with probability } 1 - n^{(-1/2)(r+s)} \end{cases}$$

Then $\mathbb{E}|X_n^s| = n^{(1/2)(s-r)} \rightarrow 0$ and $\mathbb{E}|X_n^r| = n^{(1/2)(r-s)} \rightarrow \infty$.

- (b) By Markov's Inequality (Lemma 8.3),

$$\Pr(|X_n - X| > \epsilon) \leq \frac{\mathbb{E}|X_n - X|}{\epsilon} \quad \text{for all } \epsilon > 0$$

Therefore if $X_n \xrightarrow{1} X$; that is, $\lim_{n \rightarrow \infty} \mathbb{E}(|X_n - X|) = 0$, then $\lim_{n \rightarrow \infty} \Pr(|X_n - X| > \epsilon) = 0$ for every $\epsilon > 0$, so $X_n \xrightarrow{p} X$.

To see the converse fails, define an independent sequence $\{X_n\}$ by

$$X_n = \begin{cases} n^3 & \text{with probability } n^{-2} \\ 0 & \text{with probability } 1 - n^{-2} \end{cases}$$

Then $\Pr(|X| > \epsilon) = n^{-2}$ for all large n , and so $X_n \xrightarrow{p} 0$. However, $\mathbb{E}|X_n| = n \rightarrow \infty$.

□

Lemma 8.17. (Lemma 7.2.10, Grimmett and Stirzaker.) Let $A_n(\epsilon) = \{|X_n - X| > \epsilon\}$ and $B_m(\epsilon) = \cup_{n \geq m} A_n(\epsilon)$. Then:

- (a) $X_n \xrightarrow{a.s.} X$ if and only if $\Pr(B_m(\epsilon)) \rightarrow 0$ as $m \rightarrow \infty$ for all $\epsilon > 0$.

- (b) $X_n \xrightarrow{a.s.} X$ if $\sum_n \Pr(A_n(\epsilon)) < \infty$ for all $\epsilon > 0$.
- (c) If $X_n \xrightarrow{a.s.} X$ then $X_n \xrightarrow{p} X$, but the converse fails in general.

Proof. (a)

- (b) As for Theorem 8.15 part (c).
- (c) To see the converse fails, define an independent sequence $\{X_n\}$ by

$$X_n = \begin{cases} 1 & \text{with probability } n^{-1} \\ 0 & \text{with probability } 1 - n^{-1} \end{cases}$$

Clearly $X_n \xrightarrow{p} 0$. However, if $0 < \epsilon < 1$,

$$\begin{aligned} \Pr(B_m(\epsilon)) &= 1 - \lim_{r \rightarrow \infty} \Pr(X_n = 0 \text{ for all } n \text{ such that } m \leq n \leq r) \text{ (by Lemma 1.3.5)} \\ &= 1 - \left(1 - \frac{1}{m}\right) \left(1 - \frac{1}{m+1}\right) \cdots \text{ (by independence)} \\ &= 1 - \lim_{M \rightarrow \infty} \left(\frac{m-1}{m} \cdot \frac{m}{m+1} \cdot \frac{m+1}{m+2} \cdots \frac{M}{M+1}\right) \\ &= 1 - \lim_{M \rightarrow \infty} \frac{m-1}{M+1} = 1 \end{aligned}$$

and so $\{X_n\}$ does not converge almost surely.

□

Lemma 8.18. (Lemma 7.2.12, Grimmett and Stirzaker.) There exist sequences which

- (a) converge almost surely but not in mean,
- (b) converge in mean but not almost surely.

Proof. (a) As for Lemma 8.16 part (b).

□

Theorem 8.19. (Theorem 7.2.13, Grimmett and Stirzaker.) If $X_n \xrightarrow{p} X$, there exists a non-random increasing sequence of integers n_1, n_2, \dots such that $X_{n_i} \xrightarrow{a.s.} X$ as $i \rightarrow \infty$.

Theorem 8.20. Skorokhod's representation theorem (Theorem 7.2.14, Grimmett and Stirzaker). If $\{X_n\}$ and X with distribution functions $\{F_n\}$ and F are such that $X_n \xrightarrow{d} X$ (or equivalently, $F_n \rightarrow F$) as $n \rightarrow \infty$, then there exists a probability space $(\Omega', \mathcal{F}', \mathbb{P}')$ and random variables $\{Y_n\}$ and Y mapping Ω' into \mathbb{R} such that

- (a) $\{Y_n\}$ and Y have distribution functions $\{F_n\}$ and F
- (b) $Y_n \xrightarrow{a.s.} Y$ as $n \rightarrow \infty$

Therefore, although X_n may fail to converge to X in any mode other than in distribution, there exists a sequence $\{Y_n\}$ such that Y_n is distributed identically to X_n for every n , which converges almost surely to a copy of X .

Theorem 8.21. (Theorem 7.2.19, Grimmett and Stirzaker; same as Portmanteau Theorem?)

The following three statements are equivalent:

- (a) $X_n \xrightarrow{d} X$
- (b) $\mathbb{E}[g(X_n)] \rightarrow \mathbb{E}[g(X)]$ for all bounded continuous functions g .
- (c) $\mathbb{E}[g(X_n)] \rightarrow \mathbb{E}[g(X)]$ for all functions g of the form $g(x) = f(x)\mathbf{1}_{[a,b]}(x)$ where f is continuous on $[a, b]$ and a and b are points of continuity of the distribution function of the random variable X .

Theorem 8.22. (Grimmett and Stirzaker Theorem 7.3.9.)

- (a) If $X_n \xrightarrow{a.s.} X$ and $Y_n \xrightarrow{a.s.} Y$ then $X_n + Y_n \xrightarrow{a.s.} X + Y$.
- (b) If $X_n \xrightarrow{r} X$ and $Y_n \xrightarrow{r} Y$ then $X_n + Y_n \xrightarrow{r} X + Y$.
- (c) If $X_n \xrightarrow{p} X$ and $Y_n \xrightarrow{p} Y$ then $X_n + Y_n \xrightarrow{p} X + Y$.
- (d) It is not in general true that $X_n + Y_n \xrightarrow{d} X + Y$ whenever $X_n \xrightarrow{d} X$ and $Y_n \xrightarrow{d} Y$.

Theorem 8.23. Borel-Cantelli lemmas (Grimmett and Stirzaker Theorem 7.3.10.) Let $\{A_n\}$ be an infinite sequence of events from some probability space $(\Omega, \mathcal{F}, \mathbb{P})$. Let $A = \bigcap_n \bigcup_{m=n}^{\infty} A_m = \limsup_{n \rightarrow \infty} A_n = \{A_n \text{ i.o.}\}$ be the event that infinitely many of the A_n occur. Then:

- (a) $\Pr(A) = 0$ if $\sum_n \Pr(A_n) < \infty$
- (b) $\Pr(A) = 1$ if $\sum_n \Pr(A_n) = \infty$ and A_1, A_2, \dots are independent events.

Proof. (a) We have that $A \subseteq \bigcup_{m=n}^{\infty} A_m$ for all n , so

$$\Pr(A) \leq \sum_{m=n}^{\infty} \Pr(A_m) \rightarrow 0 \text{ as } n \rightarrow \infty$$

whenever $\sum_n \Pr(A_n) < \infty$.

- (b) One can confirm that

$$A^c = \bigcup_n \bigcap_{m=n}^{\infty} A_m^c$$

But

$$\begin{aligned} \Pr\left(\bigcap_{m=n}^{\infty} A_m^c\right) &= \lim_{r \rightarrow \infty} \Pr\left(\bigcap_{m=n}^r A_m^c\right) = \prod_{m=n}^{\infty} [1 - \Pr(A_m)] \text{ (by independence)} \leq \prod_{m=n}^{\infty} \exp(-\Pr(A_m)) \\ &= \exp\left(-\sum_{m=n}^{\infty} \Pr(A_m)\right) = 0 \end{aligned}$$

whenever $\sum_n \Pr(A_n) = \infty$, where the fourth step follows since $1 - x \leq e^{-x}$ if $x \geq 0$. Thus

$$\Pr(A^c) = \lim_{n \rightarrow \infty} \Pr\left(\bigcap_{m=n}^{\infty} A_m^c\right) = 0$$

so $\Pr(A) = 1$.

□

Theorem 8.24. Kolmogorov's Two-Series Theorem. Let X_1, X_2, \dots be independent random variables with $\mathbb{E}(X_n) = \mu_n$ and $\text{Var}(X_n) = \sigma_n^2$ such that $\sum_{n=1}^{\infty} \mu_n < \infty$ and $\sum_{n=1}^{\infty} \sigma_n^2 < \infty$. Then $\sum_{n=1}^{\infty} X_n$ converges in \mathbb{R} almost surely.

Proof. Available on wikipedia, https://en.wikipedia.org/wiki/Kolmogorov%27s_two-series_theorem. □

8.4.1 Slutsky's Convergence Theorems (8.4.1 of Pesaran, 7.3 of Grimmett and Stirzaker)

Theorem 8.25. Theorem 6 of Pesaran, Section 8.4.1, p. 173. Let $\{x_t, y_t\}, t = 1, 2, \dots$ be a sequence of pairs of random variables with $y_t \xrightarrow{d} y$ and $|y_t - x_t| \xrightarrow{p} 0$. Then $x_t \xrightarrow{d} y$.

Theorem 8.26. Theorem 7 in Pesaran, on p.318 (section 7.3) of Grimmett and Stirzaker. (Section 8.4.1, p. 174) If $x_t \xrightarrow{d} x$ and $y_t \xrightarrow{p} c$ where c is a finite constant, then

- (i) $x_t + y_t \xrightarrow{d} x + c$
- (ii) $y_t x_t \xrightarrow{d} cx$
- (iii) $x_t/y_t \xrightarrow{d} x/c$, if $c \neq 0$.

Theorem 8.27. on p.318 (section 7.3) of Grimmett and Stirzaker. Suppose that $X_n \xrightarrow{d} 0$ and $Y_n \xrightarrow{p} Y$, and let $g : \mathbb{R}^2 \rightarrow \mathbb{R}$ be such that $g(x, y)$ is a continuous function of y for all x , and $g(x, y)$ is continuous at $x = 0$ for all y . Then $g(X_n, Y_n) \xrightarrow{p} g(0, Y)$.

Theorem 8.28. Continuous Mapping Theorem (Theorem 9 of Pesaran, Section 8.4.1, p. 176: convergence properties of transformed sequences.) Suppose $\{\mathbf{x}_t\}, \{\mathbf{y}_t\}$, \mathbf{x} , and \mathbf{y} are $m \times 1$ vectors of random variables on a probability space, and let $\mathbf{g}(\cdot)$ be a continuous vector-valued function. (Alternatively, suppose g has the set of discontinuity points D_g such that $\Pr(X \in D_g) = 0$.) Then

- (i) $\mathbf{x}_t \xrightarrow{a.s.} \mathbf{x} \implies \mathbf{g}(\mathbf{x}_t) \xrightarrow{a.s.} \mathbf{g}(\mathbf{x})$
- (ii) $\mathbf{x}_t \xrightarrow{p} \mathbf{x} \implies \mathbf{g}(\mathbf{x}_t) \xrightarrow{p} \mathbf{g}(\mathbf{x})$
- (iii) $\mathbf{x}_t \xrightarrow{d} \mathbf{x} \implies \mathbf{g}(\mathbf{x}_t) \xrightarrow{d} \mathbf{g}(\mathbf{x})$
- (iv) $\mathbf{x}_t - \mathbf{y}_t \xrightarrow{p} \mathbf{0}$ and $\mathbf{y}_t \xrightarrow{d} \mathbf{y} \implies \mathbf{g}(\mathbf{x}_t) - \mathbf{g}(\mathbf{y}_t) \xrightarrow{d} \mathbf{0}(\mathbf{x})$

Proof. See Serfling (1980) or Rao (1973). □

See also:

Theorem 8.29. (Theorem 7.2.18, Grimmett and Stirzaker.) If $X_n \xrightarrow{d} X$ and $g : \mathbb{R} \rightarrow \mathbb{R}$ is continuous, then $g(X_n) \xrightarrow{d} g(X)$.

8.5 Stochastic orders $\mathcal{O}_p(\cdot)$ and $o_p(\cdot)$ (Pesaran 8.5)

Definition 8.12. (Pesaran 8.5 Definition 6.) Let $\{a_t\}$ be a sequence of positive numbers and $\{x_t\}$ be a sequence of random variables. Then

- (i) $x_t = \mathcal{O}_p(a_t)$, or x_t/a_t is bounded in probability, if for every $\epsilon > 0$ there exist real numbers M_ϵ and N_ϵ such that

$$\Pr\left(\frac{|x_t|}{a_t} > M_\epsilon\right) < \epsilon, \quad \text{for } t > N_\epsilon$$

- (ii) $x_t = o_p(a_t)$ if

$$\frac{x_t}{a_t} \xrightarrow{p} 0$$

8.6 Laws of Large Numbers and Central Limit Theorems (Pesaran 8.6; Grimmett and Stirzaker 7.4, 7.5)

Theorem 8.30. Weak Law of Large Numbers (Khintchine) (Pesaran 8.6 Theorem 10, Grimmett and Stirzaker Theorem 7.4.7). Suppose that $\{X_k\}$ is a sequence of (i) IID random variables with (ii) constant means, i.e., $\mathbb{E}(X_k) = \mu < \infty$. Then

$$\bar{X}_k = \frac{1}{n} \sum_{k=1}^n X_k \xrightarrow{p} \mu$$

Theorem 8.31. Weak Law of Large Numbers (Chebyshev) (Pesaran Section 8.6, p. 178, Theorem 11.) Let $\{X_k\}$ be a sequence of random variables. If (i) $\mathbb{E}(X_k) = \mu_k$, (ii) $\text{Var}(X_k) = \sigma_k^2$, and (iii) $\text{Cov}(X_k, X_j) = 0$, $k \neq j$, and (iv)

$$\lim_{n \rightarrow \infty} \frac{1}{n} \sum_{k=1}^n \sigma_k^2 < \infty$$

then we have $\bar{X}_n - \bar{\mu}_n \xrightarrow{p} 0$, where $\bar{\mu}_n = n^{-1} \sum_{k=1}^n \mu_k$.

Theorem 8.32. Strong Law of Large Numbers (Grimmett and Stirzaker Theorem 7.4.3). Let $\{X_k\}$ be a sequence of (i) independent (ii) identically distributed random variables with (iii) $\mathbb{E}(X_k) = \mu$ and (iv) $\mathbb{E}(X_k^2) < \infty$. Then

$$\frac{1}{n} \sum_{k=1}^n X_k \rightarrow \mu \text{ almost surely and in mean square.}$$

Theorem 8.33. Strong Law of Large Numbers (Grimmett and Stirzakker Theorem 7.5.1). Let $\{X_k\}$ be a sequence of (i) independent (ii) identically distributed random variables. Then if and only if (iii) $\mathbb{E}|X_k| < \infty$,

$$\frac{1}{n} \sum_{k=1}^n X_k \xrightarrow{a.s.} \mu$$

Theorem 8.34. Strong Law of Large Numbers 1 (Kolmogorov) (Pesaran 8.8 Theorem 12).

Let $\{X_k\}$ be a sequence of (i) independent random variables with (ii) $\mathbb{E}(X_k) = \mu_k < \infty$ and (ii) $\text{Var}(X_k) = \sigma_k^2$ such that (iii)

$$\sum_{k=1}^{\infty} \frac{\sigma_k^2}{k^2} < \infty$$

Then $\overline{X}_n - \bar{\mu}_n \xrightarrow{w.p. 1} 0$. If the independence assumption (i) is replaced by a lack of correlation (i.e. $\text{Cov}(X_k, X_j) = 0, k \neq j$), the convergence of $\overline{X}_n - \bar{\mu}_n$ with probability one requires the stronger condition

$$\sum_{k=1}^{\infty} \frac{\sigma_k^2 (\log k)^2}{k^2} < \infty$$

Theorem 8.35. Strong Law of Large Numbers 2 (Pesaran 8.8 Theorem 13) Suppose that X_1, X_2, \dots are (i) independent random variables, and that (ii) $\mathbb{E}(X_k) = 0$, (iii) $\mathbb{E}(X_k^4) \leq M \forall k$ where M is an arbitrary positive constant. Then

$$\overline{X}_n = \frac{1}{n} \sum_{k=1}^n X_k \xrightarrow{a.s.} 0$$

Theorem 8.36. Central Limit Theorem (Grimmett and Stirzaker theorem 5.10.4.) Let X_1, X_2, \dots be a sequence of independent identically distributed random variables with finite mean μ and finite non-zero variance σ^2 , and let $S_n = \sum_{i=1}^n X_i$. Then

$$\frac{S_n - n\mu}{\sqrt{n\sigma^2}} \xrightarrow{d} \mathcal{N}(0, 1)$$

Theorem 8.37. (Grimmett and Stirzaker theorem 5.10.5.) Let X_1, X_2, \dots be independent random variables satisfying $\mathbb{E}(X_j) = 0$, $\text{Var}(X_j) = \sigma_j^2$, $\mathbb{E}|X_j^3| < \infty$ such that

$$\lim_{n \rightarrow \infty} \frac{1}{\sigma(n)^3} \sum_{j=1}^n \mathbb{E}|X_j^3| = 0$$

where $\sigma(n)^2 = \text{Var}(\sum_{j=1}^n X_j) = \sum_{j=1}^n \sigma_j^2$. Then

$$\frac{1}{\sigma(n)} \sum_{j=1}^n X_j \xrightarrow{d} \mathcal{N}(0, 1)$$

Proof. See Loeve (1977, p. 287) and Grimmett and Stirzaker Problem 5.12.40. \square

Lemma 8.38. Lindeberg's Condition: Let $\{X_k\}$ be a sequence of independent (not necessarily identically distributed) random variables with expectations μ_k and finite variances σ_k^2 . Let $s_n^2 = \sum_{k=1}^n \sigma_k^2$. If such a sequence of independent random variables X_k satisfies the condition

$$\lim_{n \rightarrow \infty} \frac{1}{s_n^2} \sum_{k=1}^n \mathbb{E}[(X_k - \mu_k)^2] \cdot \mathbf{1}_{\{|X_k - \mu_k| > \epsilon s_n\}} = 0$$

for all $\epsilon > 0$ then the central limit theorem holds; that is, the random variables

$$Z_n = \frac{1}{s_n} \sum_{k=1}^n (X_k - \mu_k)$$

converge in distribution to $\mathcal{N}(0, 1)$ as $n \rightarrow \infty$.

8.7 The case of dependent and heterogeneously distributed observations (Pesaran 8.8)

Theorem 8.39. Central limit theorem for martingale difference sequences (Pesaran 8.8 Theorem 28). Let $\{x_t\}$ be a martingale difference sequence with respect to the information set Ω_t . Let $\bar{\sigma}_T^2 = \text{Var}(\sqrt{T}\bar{x}_T) = T^{-1} \sum_{t=1}^T \sigma_t^2$. If $\mathbb{E}(|x_t|^r) < K < \infty$ for any $r > 2$ and for all t , and

$$\frac{1}{T} \sum_{t=1}^T x_t^2 - \bar{\sigma}_T^2 \xrightarrow{p} 0$$

then $\sqrt{T}\bar{x}_T / \bar{\sigma}_T \xrightarrow{d} \mathcal{N}(0, 1)$.

8.8 Worked Examples from Math 505A Midterm 2

- (1) (a) **Fall 2010 Problem 1.** Let X_k , $k \geq 1$, be i.i.d. random variables with mean 1 and variance 1. Show that the limit

$$\lim_{n \rightarrow \infty} \frac{\sum_{k=1}^n X_k}{\sum_{k=1}^n X_k^2}$$

exists in an appropriate sense, and identify the limit.

- (b) **Not included on midterm or final.** Let $(X_j)_{j \geq 1}$ be i.i.d. uniform on $(-1, 1)$. Let

$$Y_n = \frac{\sum_{j=1}^n X_j}{\sum_{j=1}^n X_j^2 + \sum_{j=1}^n X_j^3}$$

Prove that $\lim_{n \rightarrow \infty} \sqrt{n}Y_n$ exists in an appropriate sense, and identify the limit.

Solution.

(a)

$$\lim_{n \rightarrow \infty} \frac{\sum_{k=1}^n X_k}{\sum_{k=1}^n X_k^2} = \lim_{n \rightarrow \infty} \frac{n^{-1} \sum_{k=1}^n X_k}{n^{-1} \sum_{k=1}^n X_k^2}$$

Since X_1, X_2, \dots are i.i.d., $E(X_1^2) = \text{Var}(X_1) + (\mathbb{E}(X_1))^2 = 2 < \infty$, we have

$$n^{-1} \sum_{k=1}^n X_k \xrightarrow{a.s.} \mathbb{E}(X_1) = 1 \text{ as } n \rightarrow \infty$$

by Theorem 8.32 (Strong Law of Large Numbers). Also, X_1^2, X_2^2, \dots are clearly identically distributed, and are independent by Theorem 4.2.3 (“If X and Y are independent, then so are $g(X)$ and $g(Y)$.”). It is clear also that $\mathbb{E}(|X_1^2|) = \mathbb{E}(X_1^2) = \text{Var}(X_1) + \mathbb{E}(X_1)^2 = 1 + 1 = 2 < \infty$. Therefore by Theorem 8.33 (Strong Law of Large Numbers),

$$n^{-1} \sum_{k=1}^n X_k^2 \xrightarrow{a.s.} \mathbb{E}(X_1^2) = 2 \text{ as } n \rightarrow \infty$$

(From here I had two different ways of finishing the problem.)

- Because we have almost sure convergence in the numerator and denominator, by the Continuous Mapping Theorem (Theorem 8.28),

$$\lim_{n \rightarrow \infty} \frac{n^{-1} \sum_{k=1}^n X_k}{n^{-1} \sum_{k=1}^n X_k^2} = \frac{\lim_{n \rightarrow \infty} n^{-1} \sum_{k=1}^n X_k}{\lim_{n \rightarrow \infty} n^{-1} \sum_{k=1}^n X_k^2} \xrightarrow{a.s.} \boxed{\frac{1}{2}}$$

- Then, using one of Slutsky’s convergence theorems (Theorem 8.26: “If $x_t \xrightarrow{d} x$ and $y_t \xrightarrow{p} c$ where c is a finite constant, then $x_t/y_t \xrightarrow{d} x/c$, if $c \neq 0$.”), we have

$$\frac{n^{-1} \sum_{k=1}^n X_k}{n^{-1} \sum_{k=1}^n X_k^2} \xrightarrow{d} \frac{\mathbb{E}(X_1)}{\mathbb{E}(X_1^2)} = \frac{\mathbb{E}(X_1)}{\text{Var}(X_1) + \mathbb{E}(X_1)^2} = \frac{1}{1+1} = \frac{1}{2}$$

But then, by Theorem 8.15 (Theorem 7.2.4(a) in Grimmett and Stirzaker: “If $X_n \xrightarrow{d} c$ where c is constant, then $X_n \xrightarrow{p} c$ ”), we have $\frac{n^{-1} \sum_{k=1}^n X_k}{n^{-1} \sum_{k=1}^n X_k^2} \xrightarrow{p} 1/2$.

(b) (Not included on midterm or final.)

$$Y_n = \frac{\sum_{j=1}^n X_j}{\sum_{j=1}^n X_j^2 + \sum_{j=1}^n X_j^3} = \frac{n^{-1} \sum_{j=1}^n X_j}{n^{-1} \sum_{j=1}^n X_j^2 + n^{-1} \sum_{j=1}^n X_j^3}$$

Note that $\mathbb{E}(X_1) = 0$, $\mathbb{E}(X_1^2) = \text{Var}(X_1) + \mathbb{E}(X_1)^2 = (1 - (-1))^2/12 + 0^2 = 1/3$, $\mathbb{E}(X_1^3) = (1/2) \int_{-1}^1 x^3 dx = 0$. (We derived the formulae for the first three moments of a uniform distribution on Homework 4 problem 2(2).)

$$\implies \sqrt{n} Y_n = \frac{\sqrt{1/3} (\sum_{j=1}^n X_j - n\mathbb{E}(X_1)) / \sqrt{n \cdot 1/3}}{n^{-1} \sum_{j=1}^n X_j^2 + n^{-1} \sum_{j=1}^n X_j^3}$$

By the Central Limit Theorem (Theorem 8.36),

$$\frac{\sum_{j=1}^n X_j - n\mathbb{E}(X_1)}{\sqrt{n \cdot 1/3}} \xrightarrow{d} \mathcal{N}(0, 1)$$

By the Law of Large Numbers (Theorem 8.33), since $\mathbb{E}(|X_1^2|) = \mathbb{E}(X_1^2) = 1/3 < \infty$,

$$\frac{1}{n} \sum_{j=1}^n X_j^2 \xrightarrow{a.s.} \mathbb{E}(X_1^2) = 1/3$$

By the Law of Large Numbers (Theorem 8.33), since $\mathbb{E}(|X_1^3|) = (1/2) \int_{-1}^1 |x^3| dx = \int_0^1 x^3 dx = 1/4 < \infty$,

$$\frac{1}{n} \sum_{j=1}^n X_j^3 \xrightarrow{a.s.} \mathbb{E}(X_1^3) = 0$$

In the denominator, since we have almost sure convergence, the regular rules of calculus/real analysis apply. That is, using the above results,

$$n^{-1} \sum_{j=1}^n X_j^2 + n^{-1} \sum_{j=1}^n X_j^3 \xrightarrow{a.s.} 1/3$$

Therefore

$$\sqrt{n} Y_n = \frac{\sqrt{1/3} (\sum_{j=1}^n X_j - n\mathbb{E}(X_1)) / \sqrt{n \cdot 1/3}}{n^{-1} \sum_{j=1}^n X_j^2 + n^{-1} \sum_{j=1}^n X_j^3} \xrightarrow{d} \frac{\sqrt{1/3}}{1/3} \mathcal{N}(0, 1) = \boxed{\mathcal{N}(0, 3)}$$

- (2) **Fall 2010 Problem 2.** Fix $p \in (0, 1)$ and consider independent Poisson random variables $X_k, k \geq 1$ with

$$\mathbb{E}X_k = \frac{p^k}{k}$$

Verify that the sum $\sum_{k=1}^{\infty} kX_k$ converges with probability one and determine the distribution of the random variable $Y = \sum_{k=1}^{\infty} kX_k$.

Solution. Melike's solution (use for midterm): We have $\mathbb{E}[kX_k] = p^k$ and $\sum_{k=1}^{\infty} p^k = p/(1-p) < \infty$, and $\text{Var}(kX_k) = kp^k$ and

$$\sum_{k=1}^{\infty} kp^k = p \sum_{k=1}^{\infty} kp^{k-1} = p \frac{d}{dp} \sum_{k=1}^{\infty} p^k = p \frac{d}{dp} \frac{p}{1-p} = p \cdot \frac{(1-p) - p(-1)}{(1-p)^2} = \frac{p}{(1-p)^2} < \infty$$

Since the sequence $\{Y_k\}_{k \geq 1}$ is independent, by Kolmogorov's Two Series Theorem (Theorem 8.24: "Let X_1, X_2, \dots be independent random variables with $\mathbb{E}(X_n) = \mu_n$ and $\text{Var}(X_n) = \sigma_n^2$ such that $\sum_{n=1}^{\infty} \mu_n < \infty$ and $\sum_{n=1}^{\infty} \sigma_n^2 < \infty$. Then $\sum_{n=1}^{\infty} X_n$ converges in \mathbb{R} almost surely."), we conclude that $\sum_{k=1}^{\infty} kX_k$ converges almost surely.

To find the distribution of Y , let X be a Poisson random variable and consider its probability generating function:

$$G_X(s) = \mathbb{E}(s^X) = \sum_{k=0}^{\infty} s^k e^{-\lambda} \frac{\lambda^k}{k!} = e^{-\lambda} \sum_{k=0}^{\infty} \frac{(\lambda s)^k}{k!} = e^{-\lambda} e^{\lambda s} = e^{\lambda(s-1)}$$

So $\mathbb{E}(s^{X_k}) = \exp\left(\frac{p^k}{k}(s-1)\right)$ and $\mathbb{E}(s^{kX_k}) = \mathbb{E}[(s^k)^{X_k}] = \exp\left(\frac{p^k}{k}(s^k-1)\right)$. Then define $Y_n = \sum_{k=1}^n kX_k$ and consider

$$\begin{aligned}
G_{Y_n}(s) &= \mathbb{E}(s^{Y_n}) = \mathbb{E}\left(\prod_{k=1}^n s^{kX_k}\right) = \prod_{k=1}^n \mathbb{E}(s^{kX_k}) = \prod_{k=1}^n \exp\left(\frac{p^k}{k}(s^k - 1)\right) = \exp\left(\sum_{k=1}^n \frac{p^k}{k}(s^k - 1)\right) \\
&= \exp\left(\sum_{k=1}^n \frac{(ps)^k}{k} - \sum_{k=1}^n \frac{p^k}{k}\right)
\end{aligned}$$

Now, by taking limits as $n \rightarrow \infty$ (since we are allowed to take limit inside of expectation here), we get

$$\begin{aligned}
G_Y(s) &= \mathbb{E}(s^Y) = \exp\left(\sum_{k=1}^{\infty} \frac{(ps)^k}{k} - \sum_{k=1}^{\infty} \frac{p^k}{k}\right) = \exp\left(\int \sum_{k=1}^{\infty} (ps)^{k-1} dp - \int \sum_{k=1}^{\infty} p^{k-1} dp\right) \\
&= \exp\left(\int \frac{1}{1-ps} dp - \int \frac{1}{1-p} dp\right) = \exp(-\log(1-ps) + \log(1-p)), \quad -1 \leq ps < 1 \text{ and } -1 \leq p < 1 \\
&= \frac{1-p}{1-ps}, \quad -1 \leq ps < 1
\end{aligned}$$

Since we know $\Pr(X = k) = \frac{G_X^{(k)}(0)}{k!}$, we have

$$\begin{aligned}
G_Y(s) &= \frac{1-p}{1-sp}, \quad G'(s) = \frac{p(1-p)}{(1-sp)^2}, \quad G''(s) = \frac{2p^2(1-p)}{(1-sp)^3}, \quad G^{(3)}(s) = \frac{3 \cdot 2p^3(1-p)}{(1-sp)^3}, \dots \\
G^{(k)}(s) &= \frac{k!p^k(1-p)}{(1-sp)^k} \text{ for } k = 0, 1, 2, \dots
\end{aligned}$$

So we have

$$\Pr(Y = k) = (1-p)p^k, \quad k = 0, 1, 2, \dots$$

$$= \Pr(G_1(1-p) = k+1) = \Pr(G_1(1-p) - 1 = k)$$

which means $Y \sim G_1(1-p) - 1$.

(3) Spring 2017 Problem 3.

- (a) Consider the sequence $\{X_k, k \geq 1\}$ of random variables such that X_1 is uniform on $(0, 1)$ and, given X_k , the distribution of X_{k+1} is uniform on $(0, CX_k)$, where $\sqrt{3} < C < 2$.
 - (i) For $n \geq 1$, compute the conditional expectation $\mathbb{E}(X_{n+1}^r | X_n)$.
 - (ii) For $n \geq 1$, compute $\mathbb{E}(X_n^r)$.
 - (iii) Show that $\lim_{x \rightarrow \infty} X_n = 0$ in ℓ_1 and with probability one, but not in ℓ_2 .
 - (iv) Investigate the same questions for all other values of $C > 0$.
- (b) Let $a > 0$, let $X_n, n \geq 1$ be i.i.d. random variables that are uniform on $(0, a)$, and let $Y_n = \prod_{k=1}^n X_k$. Determine, with a proof, all values of a for which $\lim_{n \rightarrow \infty} Y_n = 0$ with probability one.

Solution.

(a) (i) We have that $X_{n+1} | X_n \sim U(0, CX_n)$. Therefore

$$\begin{aligned}\mathbb{E}(X_{n+1}^r | X_n) &= \frac{1}{CX_n} \int_0^{CX_n} x^r dx = \frac{1}{CX_n} \cdot \frac{x^{r+1}}{r+1} \Big|_0^{CX_n} = \frac{C^r X_n^r}{r+1} \\ \implies \mathbb{E}(X_{n+1}^r) &= \mathbb{E}[\mathbb{E}(X_{n+1}^r | X_n)] = \frac{C^r}{r+1} \cdot \mathbb{E}(X_n^r) \\ \implies \boxed{\mathbb{E}(X_{n+1}^r | X_n) = \frac{C^r}{r+1} X_n^r}\end{aligned}$$

(ii) Note that $E(X_1^r) = \int_0^1 x^r dr = 1/(r+1)$. Therefore

$$\mathbb{E}(X_{n+1}^r) = \frac{C^r}{r+1} \cdot \mathbb{E}(X_n^r) = \left(\frac{C^r}{r+1}\right)^n \cdot \mathbb{E}(X_1^r) = \boxed{\left(\frac{C^r}{r+1}\right)^n \cdot \frac{1}{r+1}}$$

(iii) We would like to show that $X_n \xrightarrow{w.p.1} 0$ and that $X_n \xrightarrow{1} 0$, but that the same result does not follow for the ℓ_2 norm.

- **Convergence with probability one:** We seek to show that $\Pr(\{\omega \in \Omega : \lim_{n \rightarrow \infty} X_n(\omega) = 0\}) = 1$. By Markov's Inequality (Lemma 8.3), we have

$$\begin{aligned}\Pr(|X_n| \geq a) &\leq \frac{\mathbb{E}(X_n)}{a} \quad \forall a > 0 \\ \iff \Pr(|X_n| \geq a) &\leq \left(\frac{C^1}{1+1}\right)^{n-1} \cdot \frac{1}{1+1} \cdot \frac{1}{a} = \left(\frac{C}{2}\right)^{n-1} \cdot \frac{1}{2a} \quad \forall a > 0\end{aligned}$$

Since $\sqrt{3} < C < 2$, $\sqrt{3}/2 < C/2 < 1$. Since $X_n \in [0, CX_{n-1}]$, $X_n \geq 0$, so $|X_n| = X_n$. Therefore we have

$$\Pr(\lim_{n \rightarrow \infty} |X_n| \geq a) = \Pr(\lim_{n \rightarrow \infty} X_n \geq a) \leq \lim_{n \rightarrow \infty} \left(\frac{C}{2}\right)^{n-1} \cdot \frac{1}{2a} = 0 \quad \forall a > 0$$

Since $|X_n| \geq 0$, this implies that $\Pr(\lim_{n \rightarrow \infty} X_n = 0) = \Pr(\{\omega \in \Omega : \lim_{n \rightarrow \infty} X_n(\omega) = 0\}) = 1$, so by the Borel-Cantelli Lemma (Theorem 8.23), X_n converges to 0 with probability 1.

- **Convergence in ℓ_1 norm:** We seek to show that $\lim_{n \rightarrow \infty} \mathbb{E}(|X_n|) = 0$. Since $X_n \in [0, CX_{n-1}]$, $X_n \geq 0$, so $|X_n| = X_n$. Therefore

$$\lim_{n \rightarrow \infty} \mathbb{E}(|X_n|) = \lim_{n \rightarrow \infty} \mathbb{E}(X_n) = \lim_{n \rightarrow \infty} \left(\frac{C}{2}\right)^{n-1} \cdot \frac{1}{2}$$

Since $\sqrt{3} < C < 2$, $\sqrt{3}/2 < C/2 < 1$, so $C/2 < 1$. Therefore we have

$$\lim_{n \rightarrow \infty} \mathbb{E}(|X_n|) = \lim_{n \rightarrow \infty} \left(\frac{C}{2}\right)^{n-1} \cdot \frac{1}{2} = 0$$

so X_n converges to 0 in 1st mean.

- **Convergence in ℓ_2 norm:** We seek to show that $\lim_{n \rightarrow \infty} \mathbb{E}(|X_n|^2) \neq 0$. We have

$$\lim_{n \rightarrow \infty} \mathbb{E}(|X_n|^2) = \lim_{n \rightarrow \infty} \mathbb{E}(X_n^2) = \lim_{n \rightarrow \infty} \left(\frac{C^2}{3}\right)^{n-1} \cdot \frac{1}{3}$$

Since $\sqrt{3} < C < 2$, $3/3 < C^2/3 < 4/3$, so $C^2/3 > 1$. Therefore we have

$$\lim_{n \rightarrow \infty} \mathbb{E}(|X_n|^2) = \lim_{n \rightarrow \infty} \left(\frac{C^2}{3} \right)^{n-1} \cdot \frac{1}{3} = \infty \neq 0$$

so X_n does not converge to 0 in 2nd mean.

- (iv) From the above, it is clear that for convergence with probability one or in 1st mean we require $0 < C/2 < 1$ and for convergence in second mean we require $0 < C^2/3 < 1$. For $0 < C < \sqrt{3}$, we see that X_n would converge to zero in 2nd mean since this would imply that $0 < C^2/3 < 1$. It would also still converge to 0 in 1st mean (and with probability 1) since we would have ($0 < C/2 < \sqrt{3}/2 < 1$). For $C = \sqrt{3}$, X_n would still converge to 0 with probability one and in 1st mean for the same reasons. However, it would not converge in 2nd mean because we would have

$$\lim_{n \rightarrow \infty} \mathbb{E}(|X_n|^2) = \lim_{n \rightarrow \infty} \left(\frac{\sqrt{3}^2}{3} \right)^{n-1} \cdot \frac{1}{3} = \frac{1}{3} \neq 0$$

For $C \geq 2$, it would diverge in all three cases, since in this case $C/2 \geq 2/2 = 1$ and $C^2/3 \geq 4/3 > 1$.

- (b) **Probably won't be on midterm.** Note that

$$\lim_{n \rightarrow \infty} Y_n = \lim_{n \rightarrow \infty} \prod_{k=1}^n X_k = 0 \iff \log(Y_n) = \log \left(\prod_{k=1}^n X_k \right) = \sum_{k=1}^n \log(X_k) \rightarrow -\infty$$

Note that

$$\begin{aligned} \mathbb{E}[\log(Y_n)] &= \mathbb{E} \left(\sum_{k=1}^n \log(X_k) \right) = \sum_{k=1}^n \mathbb{E}[\log(X_k)] = \sum_{k=1}^n \mathbb{E}[\log(X_1)] = \sum_{k=1}^n \int_0^a (\log(x)/a) dx \\ &= \sum_{k=1}^n \frac{1}{a} [x \log x - x]_0^a = \sum_{k=1}^n \frac{a \log a - a}{a} = \sum_{k=1}^n (\log(a) - 1) = n(\log(a) - 1) \end{aligned}$$

As $n \rightarrow \infty$ we have

$$\mathbb{E}[\log(Y_n)] = \begin{cases} -\infty & a < e \\ 0 & a = e \\ \infty & a > e \end{cases}$$

Since $\mathbb{E}[\log(Y_n)] \rightarrow \infty$ for $a < e$, we have $\lim_{n \rightarrow \infty} Y_n = 0$ for $a < 3$. Therefore

$$\lim_{n \rightarrow \infty} Y_n = \lim_{n \rightarrow \infty} \prod_{k=1}^n X_k = 0 \iff a < e.$$

9 Linear Regression

These notes are based on my notes from *Time Series and Panel Data Econometrics* (1st edition) by M. Hashem Pesaran and coursework for Economics 613: Economic and Financial Time Series I at USC. I also borrowed from some other sources which I mention when I use them.

9.1 Chapter 1: Linear Regression

9.1.1 Preliminaries

Suppose the true model is $y_i = \alpha + \beta x_i + \epsilon_i$. Classical assumptions:

- (i) $\mathbb{E}(\epsilon_i) = 0$
- (ii) $\text{Var}(\epsilon_i | x_i = \sigma^2)$ (constant)
- (iii) $\text{Cov}(\epsilon_i, \epsilon_j) = 0$ if $i \neq j$
- (iv) ϵ_i is uncorrelated to x_i , or $\mathbb{E}(\epsilon_i | x_j) = 0$ for all i, j .

9.1.2 Estimation

$$\hat{\beta} = \frac{n \sum_{i=1}^n x_i y_i - \sum_{i=1}^n x_i \sum_{i=1}^n y_i}{n \sum_{i=1}^n x_i^2 - (\sum_{i=1}^n x_i)^2} = \frac{\sum_{i=1}^n x_i y_i - n \bar{x} \bar{y}}{\sum_{i=1}^n x_i^2 - n \bar{x}^2}$$

$$\hat{\alpha} = \bar{y} - \hat{\beta} \bar{x}$$

or

$$\hat{\beta} = \frac{\sum_{i=1}^n (x_i - \bar{x})(y_i - \bar{y})}{\sum_{i=1}^n (x_i - \bar{x})^2} = \frac{S_{XY}}{S_{XX}}$$

or

$$\hat{\beta} = r \frac{S_{YY}}{S_{XX}}$$

where r is the correlation coefficient.

Let

$$w_i = \frac{x_i - \bar{x}}{\sum_{i=1}^n (x_i - \bar{x})^2}$$

so that

$$\hat{\beta} = \sum_{i=1}^n w_i(y_i - \bar{y}) = \sum_{i=1}^n w_i y_i - \bar{y} \frac{\sum_{i=1}^n x_i - \bar{x}}{\sum_{i=1}^n (x_i - \bar{x})^2} = \sum_{i=1}^n w_i y_i$$

since $\sum_{i=1}^n x_i - \bar{x} = 0$. Then a simple expression for $\text{Var}(\hat{\beta})$ is

$$\text{Var}(\hat{\beta}) = \sum_{i=1}^n w_i^2 \text{Var}(y_i | x_i) = \sum_{i=1}^n w_i^2 \text{Var}(\epsilon | x_i) = \sigma^2 \sum_{i=1}^n w_i^2 = \frac{\sigma^2}{\sum_{i=1}^n (x_i - \bar{x})^2} = \frac{\sigma^2}{S_{XX}}$$

We can estimate these quantities as follows:

$$\hat{\sigma}^2 = \frac{1}{n-2} \cdot \sum_{i=1}^n (y_i - \hat{\alpha} - \hat{\beta} x_i)^2$$

Note that

$$\begin{aligned} \hat{\sigma}^2 &= \frac{1}{n-2} \sum_{t=1}^T (y_t - \hat{\alpha} - \hat{\beta} x_t)^2 = \frac{1}{n-2} \sum_{t=1}^T [(y_t - (\bar{y} - \hat{\beta} \bar{x})) - \hat{\beta} x_t]^2 = \frac{1}{n-2} \sum_{t=1}^T (y_t - \bar{y} - \hat{\beta}(x_t - \bar{x}))^2 \\ &= \frac{1}{n-2} \sum_{t=1}^T (y_t - \bar{y})^2 - 2\hat{\beta}(x_t - \bar{x})(y_t - \bar{y}) + \hat{\beta}^2(x_t - \bar{x})^2 \end{aligned}$$

In the case where there is no intercept, we have

$$\hat{\sigma}^2 = \frac{1}{T-1} \sum_{t=1}^T (y_t - \hat{\beta} x_t)^2 = \frac{1}{T-1} \sum_{t=1}^T \left(y_t^2 - 2r \frac{S_{YY}}{S_{XX}} x_t y_t + r^2 \frac{S_{YY}^2}{S_{XX}^2} x_t^2 \right)$$

Also,

$$\widehat{\text{Var}}(\hat{\beta}) = \frac{\hat{\sigma}^2}{S_{XX}} = \frac{1}{n-2} \cdot \frac{\sum_{i=1}^n (y_i - \hat{\alpha} - \hat{\beta} x_i)^2}{\sum_{i=1}^n (x_i - \bar{x})^2}$$

Correlation coefficient:

$$r^2 = \frac{(\sum_{t=1}^T x_t y_t)^2}{\sum_{t=1}^T x_t^2 \sum_{t=1}^T y_t^2}$$

$$r = \frac{1}{T-1} \frac{S_{XY}}{\sqrt{S_{XX} S_{YY}}}$$

Remark. The formulas for the coefficients in univariate OLS can also be derived by considering (x, y) as a bivariate normal distribution and calculating the conditional expectation of y given x . (See Proposition (6.51).)

9.2 Chapter 2: Multiple Regression

General OLS:

$$\hat{\beta} = (\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'y = (\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'(\mathbf{X}\beta + u) = (\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'\mathbf{X}\beta + (\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'u = \beta + (\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'u$$

$$\text{Var}(\hat{\beta}) = \text{Var}(\beta + (\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'u) = \text{Var}(\beta) + \text{Var}((\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'u) = 0 + \mathbb{E}[(\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'uu'\mathbf{X}(\mathbf{X}'\mathbf{X})^{-1}]$$

$$= \mathbb{E}[(\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'\mathbb{E}(uu' | \mathbf{X})\mathbf{X}(\mathbf{X}'\mathbf{X})^{-1}] = \sigma^2 \mathbb{E}[(\mathbf{X}'\mathbf{X})^{-1}\mathbf{X}'I_T\mathbf{X}(\mathbf{X}'\mathbf{X})^{-1}] = \sigma^2 \mathbb{E}[(\mathbf{X}'\mathbf{X})^{-1}]$$

$$= \sigma^2(\mathbf{X}'\mathbf{X})^{-1}$$

$$\hat{\sigma}^2 = \frac{\hat{\mathbf{u}}'\hat{\mathbf{u}}}{T-k}$$

9.3 Chapter 3: Hypothesis testing in regression

In this section, I borrow from C. Flinn's notes "Asymptotic Results for the Linear Regression Model," available online at <http://www.econ.nyu.edu/user/flinnc/notes1.pdf>.

Lemma 9.1.

$$\frac{1}{n} \cdot \mathbf{X}'\epsilon \xrightarrow{p} 0$$

Proof. Note that $\mathbb{E}\frac{1}{n} \cdot \mathbf{X}'\epsilon = 0$ for any n . Then we have

$$\text{Var}\left(\frac{1}{n} \cdot \mathbf{X}'\epsilon\right) = \mathbb{E}\left(\frac{1}{n} \cdot \mathbf{X}'\epsilon\right)^2 = n^{-2}\mathbb{E}(\mathbf{X}'\epsilon\epsilon'\mathbf{X}) = n^{-2}\mathbb{E}(\epsilon\epsilon')\mathbf{X}'\mathbf{X} = \frac{\sigma^2}{n} \frac{\mathbf{X}'\mathbf{X}}{n}$$

implying that $\lim_{n \rightarrow \infty} \text{Var}\left(\frac{1}{n} \cdot \mathbf{X}'\epsilon\right) = 0$. Therefore the result follows from Chebyshev's Inequality (Theorem 8.4). \square

Lemma 9.2. If ϵ is i.i.d. with $E(\epsilon_i) = 0$ and $\mathbb{E}(\epsilon_i^2) = \sigma^2$ for all i , the elements of the matrix \mathbf{X} are uniformly bounded so that $|X_{ij}| < U$ for all i and j and for U finite, and $\lim_{n \rightarrow \infty} \mathbf{X}'\mathbf{X}/n = Q$ is finite and nonsingular, then

$$\frac{1}{\sqrt{n}}\mathbf{X}'\epsilon \xrightarrow{d} \mathcal{N}(0, \sigma^2 Q)$$

Proof. If we have one regressor, then $n^{-1/2} \sum_{i=1}^n X_i \epsilon_i$ is a scalar. Let G_i be the cdf of $X_i \epsilon_i$. Let

$$S_n^2 = \sum_{i=1}^n \text{Var}(X_i \epsilon_i) = \sigma^2 \sum_{i=1}^n X_i^2$$

In this scalar case, $Q = \lim_{n \rightarrow \infty} n^{-1} \sum_i X_i^2$. By the Lindeberg-Feller Theorem, a necessary and sufficient condition for $Z_n \rightarrow \mathcal{N}(0\sigma^2 Q)$ is

$$\lim_{n \rightarrow \infty} \frac{1}{S_n^2} \sum_{i=1}^n \int_{|\omega| > \nu S_n} \omega^2 dG_i(\omega) = 0$$

for all $\nu > 0$. Now $G_i(\omega) = F(\omega/|X_i|)$. Then rewrite the above equation as

$$\lim_{n \rightarrow \infty} \frac{n}{S_n^2} \sum_{i=1}^n \frac{X_i^2}{n} \int_{|\omega/X_i| > \nu S_n / |X_i|} \left(\frac{\omega}{X_i} \right)^2 dF(\omega/|X_i|) = 0$$

Since $\lim_{n \rightarrow \infty} S_n^2 = \lim_{n \rightarrow \infty} n\sigma^2 \sum_{i=1}^n X_i^2/n = n\sigma^2 Q$, we have $\lim_{n \rightarrow \infty} n/S_n^2 = (\sigma^2 Q)^{-1}$, which is a finite and nonzero scalar. Then we need to show

$$\lim_{n \rightarrow \infty} \frac{1}{n} \sum_{i=1}^n X_i^2 \delta_{i,n} = 0$$

where

$$\delta_{i,n} = \int_{|\omega/X_i| > \nu S_n / |X_i|} \left(\frac{\omega}{X_i} \right)^2 dF(\omega/|X_i|)$$

But $\lim_{n \rightarrow \infty} \delta_{i,n} = 0$ for all i and any fixed ν since $|X_i|$ is bounded while $\lim_{n \rightarrow \infty} X_n = \infty$, so the measure of the set $\{|\omega/X_i| > \nu S_n / |X_i|\}$ goes to 0 asymptotically. Since $\lim_{n \rightarrow \infty} n^{-1} \sum_i X_i^2$ is finite and $\lim_{n \rightarrow \infty} \delta_{i,n} = 0$ for all i , $\lim_{n \rightarrow \infty} n^{-1} \sum_i X_i^2 \delta_{i,n} = 0$, so $\frac{1}{n} \cdot X' \epsilon \xrightarrow{p} 0$.

□

Theorem 9.3. Under the conditions of Lemma 9.2 (ϵ is i.i.d. with $E(\epsilon_i) = 0$ and $\mathbb{E}(\epsilon_i^2) = \sigma^2$ for all i , the elements of the matrix X are uniformly bounded so that $|X_{ij}| < U$ for all i and j and for U finite, and $\lim_{n \rightarrow \infty} X' X/n = Q$ is finite and nonsingular),

$$\sqrt{n}(\hat{\beta} - \beta) \xrightarrow{d} \mathcal{N}(0, \sigma^2 Q^{-1})$$

Proof.

$$\sqrt{n}(\hat{\beta} - \beta) = \left(\frac{X' X}{n} \right)^{-1} \frac{1}{\sqrt{n}} X' \epsilon$$

Since $\lim_{n \rightarrow \infty} (X' X/n)^{-1} = Q^{-1}$ and by Lemma 9.2

$$\frac{1}{\sqrt{n}} X' \epsilon \xrightarrow{d} \mathcal{N}(0, \sigma^2 Q)$$

then

$$\sqrt{n}(\hat{\beta} - \beta) \xrightarrow{d} \mathcal{N}(0, \sigma^2 Q^{-1} Q Q^{-1}) = \mathcal{N}(0, \sigma^2 Q^{-1})$$

□

t-test statistic:

$$t = \frac{\hat{\beta} - 0}{s.e.(\hat{\beta})}$$

F-test statistic:

$$F = \left(\frac{T - k - 1}{r} \right) \left(\frac{SSR_R - SSR_U}{SSR_U} \right)$$

Since

$$R^2 = \frac{\sum_t (y_t - \bar{y})^2 - \sum_t (y_t - \hat{y}_t)^2}{\sum_t (y_t - \bar{y})^2} = \frac{\sum_t (y_t - \bar{y})^2 - SSR_U}{\sum_t (y_t - \bar{y})^2}$$

we have

$$SSR_U = \sum_t (y_t - \bar{y})^2 - R^2 \sum_t (y_t - \bar{y})^2 = (1 - R^2) \sum_t (y_t - \bar{y})^2$$

yielding

$$F = \left(\frac{T - k - 1}{r} \right) \left(\frac{\sum_t (y_t - \bar{y})^2 - (1 - R^2) \sum_t (y_t - \bar{y})^2}{(1 - R^2) \sum_t (y_t - \bar{y})^2} \right) = \left(\frac{T - k - 1}{r} \right) \left(\frac{R^2}{1 - R^2} \right)$$

Confidence interval for sums of coefficients. (Two coefficient case.) Suppose we want to test $H_0 : \beta_1 + \beta_2 = k$. Let $\delta = \beta_1 + \beta_2 - k$, $\hat{\delta} = \hat{\beta}_1 + \hat{\beta}_2 - k$. Note that under the null hypothesis $\delta = 0$. We can construct a *t*-statistic

$$t_{\hat{\delta}} = \frac{\hat{\delta} - 0}{\sqrt{\hat{\text{Var}}(\hat{\delta})}} = \frac{\hat{\beta}_1 + \hat{\beta}_2 - k}{\sqrt{\hat{\text{Var}}(\hat{\delta})}}$$

where

$$\hat{\text{Var}}(\hat{\delta}) = \hat{\text{Var}}(\hat{\beta}_1) + \hat{\text{Var}}(\hat{\beta}_2) + 2\hat{\text{Cov}}(\hat{\beta}_1, \hat{\beta}_2)$$

This means that a 95% confidence interval for δ can be constructed in the following way:

$$\hat{\delta} \pm t^* \sqrt{\text{Var}(\hat{\delta})}$$

where t^* is the 95% critical value for the t -distribution.

9.4 Chapter 4: Heteroskedasticity

Under heteroskedasticity, the OLS estimator $\hat{\beta} = (X'X)^{-1}X'y$ is unbiased, but the true covariance matrix of $\hat{\beta}$ no longer matches the OLS formula. For instance, suppose we have

$$y_t = \sum_{i=1}^K \beta_i x_{ti} + u_t$$

where $\text{Var}(u_t) = \sigma^2 z_t^2$.

$$\begin{aligned} \hat{\beta} &= (X'X)^{-1}X'y = (X'X)^{-1}X'X\beta + (X'X)^{-1}X'u = \beta + (X'X)^{-1}X'u \\ &\implies \mathbb{E}(\hat{\beta}) = \mathbb{E}[\beta] + (X'X)^{-1}X'\mathbb{E}[u] = \beta \end{aligned}$$

since $\mathbb{E}(u)$ is still 0. However,

$$\begin{aligned} \text{Var}(\hat{\beta}) &= \mathbb{E}[(\hat{\beta} - \mathbb{E}(\hat{\beta}))(\hat{\beta} - \mathbb{E}(\hat{\beta}))'] = \mathbb{E}[(\beta + (X'X)^{-1}X'u - \beta)(\beta + (X'X)^{-1}X'u - \beta)'] \\ &= \mathbb{E}[(X'X)^{-1}X'u((X'X)^{-1}X'u)'] = \mathbb{E}[(X'X)^{-1}X'u u' X ((X'X)^{-1})'] \\ &= (X'X)^{-1}X' \mathbb{E}[uu' | X] X (X'X)^{-1} \end{aligned}$$

$$\begin{aligned} &= (X'X)^{-1}X' \begin{bmatrix} \sigma^2 z_1^2 & 0 & 0 & \dots & 0 \\ 0 & \sigma^2 z_2^2 & 0 & \dots & 0 \\ 0 & 0 & \sigma^2 z_3^2 & \dots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & 0 & \dots & \sigma^2 z_T^2 \end{bmatrix} X (X'X)^{-1} \\ &= \sigma^2 (X'X)^{-1} X' \begin{bmatrix} z_1^2 & 0 & 0 & \dots & 0 \\ 0 & z_2^2 & 0 & \dots & 0 \\ 0 & 0 & z_3^2 & \dots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & 0 & \dots & z_T^2 \end{bmatrix} X (X'X)^{-1} \end{aligned}$$

which is different from the OLS estimator of the covariance matrix $\sigma^2(X'X)^{-1}$. Therefore the estimate of the variances of $\hat{\beta}$ will be biased if the OLS formulas are used, and the usual t and F tests for $\hat{\beta}$ will be invalid.

9.5 Chapter 5: Autocorrelated disturbances

Generalized least squares model:

$$\mathbf{y} = \mathbf{X}\boldsymbol{\beta} + \mathbf{u}$$

where

$$\mathbb{E}(\mathbf{u} | \mathbf{X}) = 0 \quad \forall t$$

$$\mathbb{E}(\mathbf{u}\mathbf{u}' | \mathbf{X}) = \boldsymbol{\Sigma}$$

where $\boldsymbol{\Sigma}$ is a positive definite matrix.

$$\hat{\beta}_{GLS} = (X'\boldsymbol{\Sigma}^{-1}X)^{-1}X'\boldsymbol{\Sigma}^{-1}\mathbf{y}$$

$$\text{Var}(\hat{\beta}_{GLS}) = (X'\boldsymbol{\Sigma}^{-1}X)^{-1}$$

10 Time Series

These notes are based on my notes from *Time Series and Panel Data Econometrics* (1st edition) by M. Hashem Pesaran as well as coursework for Economics 613: Economic and Financial Time Series I at USC.

10.1 Chapter 6: ARDL Models

In an ARDL model, if the error are serially correlated, then the coefficient estimates are biased (even as $T \rightarrow \infty$).

10.2 Chapters 12 and 13: Intro to Stochastic Processes and Spectral Analysis

Stationarity conditions: $\{X_t\}$ is **strictly stationary** if the joint distribution functions of $\{X_{t_1}, X_{t_2}, \dots, X_{t_k}\}$ and $\{X_{t_1+h}, X_{t_2+h}, \dots, X_{t_k+h}\}$ are identical for all values of t_1, t_2, \dots, t_k and h and all positive integers k .

Definition 10.1. X_t is **weakly (or covariance) stationary** if it has a constant mean and variance and its covariance function $\gamma(t_1, t_2)$ depends only on the absolute difference $|t_1 - t_2|$, namely $\gamma(t_1, t_2) = \gamma(|t_1 - t_2|)$.

Definition 10.2. X_t is said to be **trend stationary** if $y_t = X_t - d_t$ is covariance stationary, where d_t is the perfectly predictable component of X_t .

The process $\{\epsilon_t\}$ is said to be a **white noise process** if it has mean zero, a constant variance, and ϵ_t and ϵ_s are uncorrelated for all $s \neq t$.

Autocovariance generating function: The autocovariance generating function for the general linear stationary process $y_t = \sum_{i=0}^{\infty} a_i \epsilon_{t-i}$ is given by:

$$G(z) = \sigma^2 a(z) a(z^{-1})$$

where $a(z) = \sum_{i=0}^{\infty} a_i z^i$.

Wold's Decomposition (Theorem 42, p. 275, Section 12.5) Any trend-stationary process $\{y_t\}$ can be represented in the form of $y_t = d_t + \sum_{i=0}^{\infty} \alpha_i \epsilon_{t-i}$ where $\alpha_0 = 1$ and $\sum_{i=0}^{\infty} \alpha_i^2 < K < \infty$. The term d_t is a deterministic component, while $\{\epsilon_t\}$ is a serially uncorrelated process: $\epsilon_t = y_t - \mathbb{E}(y_t | y_{t-1}, y_{t-2}, \dots)$.

Stationarity conditions for an ARMA(p, q) process: Consider the ARMA(p, q) process

$$y_t = \sum_{i=1}^p \phi_i y_{t-i} + \sum_{i=0}^q \theta_i \epsilon_{t-i}, \quad \theta_0 = 1$$

The MA part is stationary for any finite q . The AR part is stationary if the roots of the characteristic equation

$$\lambda^t = \sum_{i=1}^p \phi_i \lambda^{t-i}$$

lie strictly inside the unit circle. Alternatively, in terms of $z = \lambda^{-1}$, the process is stationary if the roots of

$$1 - \sum_{i=1}^p \phi_i z^i = 0$$

lie outside the unit circle. The ARMA process is **invertible** (so that y_t can be solved uniquely in terms of its past values) if all the roots of

$$1 - \sum_{i=1}^p \theta_i z^i = 0$$

fall outside the unit circle.

Spectral Density Function: Definition (Equation 13.3):

$$f(\omega) = \frac{1}{2\pi} \sum_{h=-\infty}^{\infty} \gamma(h) e^{ih\omega}, \omega \in (-\pi, \pi)$$

Equation (13.5):

$$f(\omega) = \frac{1}{2\pi} \left[\gamma(0) + 2 \sum_{h=1}^{\infty} \gamma(h) \cos(h\omega) \right], \quad \omega \in [0, \pi]$$

Can also be found using the autocovariance generating function. We have (Equation 13.6, section 13.3.1)

$$f(\omega) = \frac{1}{2\pi} G(e^{i\omega}) = \frac{\sigma^2}{2\pi} a(e^{i\omega}) a(e^{-i\omega})$$

Properties of spectral density function:

- (1) $f(\omega)$ always exists and is bounded if $\gamma(h)$ is absolutely summable.
- (2) $f(\omega)$ is symmetric.
- (3) The spectrum of a stationary process is finite at zero frequency; that is, $f(0) < \infty$.

Linear (time-domain) processes don't have to be stationary, but to write something as a frequency-domain process, it must be stationary.

10.2.1 Worked Examples

Midterm Problem 2 part (1) (chapter 12 exercise 6)

Midterm Problem 2 part (2) (exercise 7 in chapter 12; similar to exercise 1 in chapter 14).
Suppose $\{y_t\}$ has the following general linear process

$$y_t = \mu + \alpha(L)\epsilon_t, \quad \epsilon_t \sim i.i.d. (0, \sigma^2)$$

where $\alpha(L) = \alpha_0 + \alpha_1 L + \alpha_2 L^2 + \dots$; $\alpha_0 = 1$. Let

$$\bar{y}_T = \frac{1}{T} \sum_{t=1}^T y_t$$

$$\gamma(h) = \mathbb{E}[(y_t - \mu)(y_{t-h} - \mu)]$$

$$\hat{\gamma}(h) = \frac{1}{T} \sum_{t=h+1}^T (y_t - \bar{y}_T)(y_{t-h} - \bar{y}_T)$$

Derive the conditions under which

- (a) \bar{y}_T is a consistent estimator of μ as $T \rightarrow \infty$
- (b) For fixed h , $\hat{\gamma}(h)$ is a consistent estimator of $\gamma(h)$ as $T \rightarrow \infty$.

Solution.

- (a) This is an MA(∞) process. By Chebyshev's Inequality (Theorem 8.4), \bar{y}_T is a consistent estimator of μ as $T \rightarrow \infty$ if $\lim_{T \rightarrow \infty} \mathbb{E}(\bar{y}_T) = \mathbb{E}(y_T) = \mu$ and $\lim_{T \rightarrow \infty} \text{Var}(\bar{y}_T) = 0$. In this case in particular (MA(∞) process), we can write

$$\bar{y}_T = \frac{1}{T} \sum_{t=1}^T (\mu + \alpha(L)\epsilon_t) = \frac{1}{T} \cdot T\mu + \frac{1}{T} \sum_{t=1}^T \alpha(L)\epsilon_t = \mu + \frac{1}{T} \sum_{t=1}^T \alpha(L)\epsilon_t$$

Then we have

$$\mathbb{E}(\bar{y}_T) = \mu + \frac{1}{T} \mathbb{E} \left(\sum_{t=1}^T \alpha(L)\epsilon_t \right) = \mu + \frac{1}{T} \sum_{t=1}^T \mathbb{E}(\alpha(L)\epsilon_t) = \mu$$

$$\text{Var}(\bar{y}_T) = 0 + \frac{1}{T^2} \text{Var} \left(\sum_{t=1}^T \alpha(L)\epsilon_t \right) = \frac{1}{T^2} \sum_{t=1}^T \text{Var}[\alpha(L)\epsilon_t] = \frac{1}{T^2} \sum_{t=1}^T \mathbb{E}[\alpha(L)\epsilon_t]^2 = \frac{1}{T} \alpha(1)^2 \mathbb{E}[\epsilon_t]^2$$

$$= \frac{\sigma^2}{T} \alpha(1)^2$$

Therefore a sufficient condition for consistency is

$$\lim_{T \rightarrow \infty} \frac{\sigma^2}{T} \alpha(1)^2 = 0 \iff \alpha(1)^2 < \infty \iff \boxed{\sum_{i=0}^{\infty} \alpha_i = 0}$$

- (b) See section 10.4.2.

10.3 Some time series and their properties

10.3.1 White noise process:

$$x_t = \epsilon_t, \epsilon_t \sim IID(0, \sigma^2)$$

- Autocovariances:

$$\gamma(0) = \sigma^2$$

$$\gamma(h) = 0, \forall h \neq 0$$

- Spectral density function:

$$f_x(\omega) = \frac{1}{2\pi} \cdot \sigma^2 = \frac{\sigma^2}{2\pi} \text{ (flat spectrum)}$$

10.3.2 MA(1) process:

$$x_t = \epsilon_t + \theta \epsilon_{t-1} \text{ with } \epsilon_t \sim iid(0, \sigma^2), |\rho| < 1.$$

- Autocovariances: By Equation (12.2), the autocovariance function is

$$\text{Cov}(u_t, u_{t-h}) = \gamma(h) = \sigma^2 \sum_{i=0}^{1-|h|} a_i a_{i+|h|} \text{ if } 0 \leq |h| \leq 1$$

$$\implies \mathbb{E}(x_t^2) = \gamma(0) = (1 + \theta^2)\sigma^2$$

$$\mathbb{E}(x_t x_{t-1}) = \gamma(1) = \theta \sigma^2$$

$$\gamma(h) = 0 \quad \forall |h| > 1$$

So the covariance matrix is

$$\begin{pmatrix} \sigma^2(1 + \theta^2) & \sigma^2\theta & 0 & 0 & \cdots & 0 \\ \sigma^2\theta & \sigma^2(1 + \theta^2) & \sigma^2\theta & 0 & \cdots & 0 \\ 0 & \sigma^2\theta & \sigma^2(1 + \theta^2) & \sigma^2\theta & \cdots & 0 \\ \vdots & \vdots & \vdots & \vdots & \cdots & \vdots \\ 0 & 0 & \cdots & \sigma^2\theta & \sigma^2(1 + \theta^2) & \sigma^2\theta \\ 0 & 0 & \cdots & 0 & \sigma^2\theta & \sigma^2(1 + \theta^2) \end{pmatrix}$$

$$= \sigma^2(1 + \theta^2)I_T + \sigma^2\theta A$$

where A is defined as in section 14.3.2 (p. 304).

- Spectral density function:

$$f(\omega) = \frac{\sigma^2}{2\pi} [1 + 2\theta \cos(\omega) + \rho^2], \quad \omega \in [0, \pi]$$

10.3.3 MA(∞) process:

This process is covariance stationary.

- Autocovariances:

10.3.4 AR(1) process:

$$x_t = \phi x_{t-1} + \epsilon_t, |\phi| < 1, \epsilon_t \sim IID(0, \sigma^2).$$

- Yule-Walker Equations:

$$\mathbb{E}[x_t x_{t-h}] = \mathbb{E}[\phi x_{t-1} x_{t-h}] + \mathbb{E}[\epsilon x_{t-h}]$$

$$\gamma_h = \phi \gamma_{h-1} + \mathbb{E}[\epsilon x_{t-h}]$$

$$\implies \gamma_0 = \phi \gamma_1 + \sigma^2, \quad \gamma_h = \phi \gamma_{h-1} \quad \forall h \geq 1$$

- Autocovariances:

$$\gamma(0) = \frac{\sigma^2}{1 - \phi^2}$$

$$\gamma_h = \frac{\sigma^2 \phi^h}{1 - \phi^2} = \phi^h \gamma(0) \quad \forall h \geq 1$$

$$\implies \text{Cov}(x) =$$

$$\begin{pmatrix} \sigma^2/(1 - \phi^2) & \sigma^2 \phi/(1 - \phi^2) & \sigma^2 \phi^2/(1 - \phi^2) & \sigma^2 \phi^3/(1 - \phi^2) & \dots & \sigma^2 \phi^{T-1}/(1 - \phi) \\ \sigma^2 \phi/(1 - \phi) & \sigma^2/(1 - \phi^2) & \sigma^2 \phi/(1 - \phi^2) & \sigma^2 \phi^2/(1 - \phi^2) & \dots & \sigma^2 \phi^{T-2}/(1 - \phi^2) \\ \sigma^2 \phi^2/(1 - \phi^2) & \sigma^2 \phi/(1 - \phi^2) & \sigma^2/(1 - \phi^2) & \sigma^2 \phi/(1 - \phi^2) & \dots & \sigma^2 \phi^{T-3}/(1 - \phi^2) \\ \vdots & \vdots & \vdots & \vdots & \dots & \vdots \\ \sigma^2 \phi^{T-2}/(1 - \phi^2) & \sigma^2 \phi^{T-3}/(1 - \phi^2) & \dots & \sigma^2 \phi/(1 - \phi^2) & \sigma^2/(1 - \phi^2) & \sigma^2 \phi/(1 - \phi^2) \\ \sigma^2 \phi^{T-1}/(1 - \phi^2) & \sigma^2 \phi^{T-2}/(1 - \phi^2) & \dots & \sigma^2 \phi^2/(1 - \phi^2) & \sigma^2 \phi/(1 - \phi^2) & \sigma^2/(1 - \phi^2) \end{pmatrix}$$

- If stationary, can be written as an infinite MA process with absolutely summable coefficients

$$x_t = \sum_{i=0}^{\infty} \phi^i \epsilon_{t-i} = \left(\frac{1}{1 - \phi L} \right) \epsilon_t$$

- Autocovariance generating function:

$$G(z) = \left(\frac{\sigma^2}{1 - \phi^2} \right) \left(1 + \sum_{h=1}^{\infty} \phi^h (z^h + z^{-h}) \right)$$

- Spectral density function:

$$f(\omega) = \frac{1}{2\pi} \sum_{h=-\infty}^{\infty} \frac{\sigma^2 \phi^{|h|}}{(1 - \phi^2)} (e^{i\omega})^h = \frac{1}{2\pi} \frac{\sigma^2}{(1 - \phi e^{i\omega})(1 - \phi e^{-i\omega})} = \frac{1}{2\pi} \frac{\sigma^2}{1 - 2\phi \cos(\omega) + \phi^2}$$

10.3.5 AR(2) process:

$$x_t = \phi_1 x_{t-1} + \phi_2 x_{t-2} + \epsilon, |\phi_1| < 1, |\phi_2| < 1, \epsilon_t \sim IID(0, \sigma^2).$$

Can be written as

$$x_t = \frac{1}{1 - \phi L} \epsilon_t = \epsilon_t + \phi \epsilon_{t-1} + \phi^2 \epsilon_{t-2} + \dots$$

- Yule-Walker equations:

$$\mathbb{E}[x_t x_{t-h}] = \mathbb{E}[\phi_1 x_{t-1} x_{t-h}] + \mathbb{E}[\phi_2 x_{t-2} x_{t-h}] + \mathbb{E}[\epsilon x_{t-h}]$$

$$\gamma_h = \phi_1 \gamma_{h-1} + \phi_2 \gamma_{h-2} + \mathbb{E}[\epsilon x_{t-h}]$$

$$\implies \boxed{\gamma_0 = \phi_1 \gamma_1 + \phi_2 \gamma_2 + \sigma^2, \quad \gamma_1 = \phi_1 \gamma_0 + \phi_2 \gamma_1, \quad \gamma_2 = \phi_1 \gamma_1 + \phi_2 \gamma_0}$$

- Autocovariances:

10.3.6 AR(p) process:

$$x_t = \phi_1 x_{t-1} + \phi_2 x_{t-2} + \dots + \phi_p x_{t-p} + \epsilon, |\phi_i| < 1, \epsilon_t \sim IID(0, \sigma^2).$$

- Stationary if the eigenvalues of Φ lie inside the unit circle, which is equivalent to all the roots of

$$\phi(z) = 1 - \phi_1 z - \phi_2 z^2 - \dots - \phi_p z^p = 0$$

being strictly larger than unity. Under this condition the AR process has the infinite-order MA representation'

$$x_t = \sum_{i=0}^{\infty} \alpha_i \epsilon_{t-i}$$

where $\alpha_i = \phi_1 \alpha_{i-1} + \dots + \phi_p \alpha_{i-p}$.

- Autocovariance generating function:

$$G(z) = \frac{\sigma^2}{\phi(z)\phi(z^{-1})}$$

10.3.7 ARMA(1, 1) process:

$x_t = \phi x_{t-1} + \epsilon_t + \theta \epsilon_{t-1}$, with $|\phi| < 1$ (implying stationarity), $\mathbb{E}(\epsilon_t^2) = \sigma^2$, $\mathbb{E}(\epsilon_t \epsilon_s) = 0$ for $t \neq s$.

- Yule-Walker Equations:

$$\gamma(0) = \phi\gamma(1) + \sigma^2(1 + \theta^2)$$

$$\gamma(1) = \phi\gamma(0) + \sigma^2\phi^2$$

$$\gamma(h) = \phi\gamma(h-1) \quad \forall h \geq 2$$

- Autocovariances:

$$\gamma(0) = \sigma^2 \left(1_{\frac{(\phi+\theta)^2}{1-\phi^2}} \right)$$

$$\gamma(1) = \sigma^2 \left(\phi + \theta + \frac{(\phi + \theta)^2 \phi}{1 - \phi^2} \right)$$

$$\gamma(2) = \phi^{h-1}\gamma(1) \quad \forall h \geq 2$$

- Autocorrelation function:

$$\rho(h) = \begin{cases} 1 & h = 0 \\ \frac{(\phi+\theta)(1+\phi\theta)}{1+2\phi\theta+\theta^2} & h = 1 \\ \phi^{h-1}\rho(1) & h \geq 2 \end{cases}$$

- Autocovariance generating function: the autocovariance function of an ARMA(p, q) process $\phi(L)y_t = \theta(L)\epsilon_t$ is given by

$$f(\omega) = \sigma^2 \frac{\theta(z)\theta(z^{-1})}{\phi(z)\phi(z^{-1})}$$

Plugging in for the ARMA(1,1) case yields (**double-check**)

$$f(\omega) = \sigma^2 \frac{(1 + \theta)^2}{(1 - \rho)^2}$$

- Spectral Density Function: the spectral density function of an ARMA(p, q) process $\phi(L)y_t = \theta(L)\epsilon_t$ is given by

$$f(\omega) = \frac{\sigma^2}{2\pi} \frac{\theta(e^{i\omega})\theta(e^{-i\omega})}{\phi(e^{i\omega})\phi(e^{-i\omega})}, \quad \omega \in [0, 2\pi]$$

Plugging in for the ARMA(1,1) case yields

$$f_x(\omega) = \frac{\sigma^2}{2\pi} \frac{(e^{i\omega} - \theta e^{i\omega})(e^{-i\omega} - \theta e^{-i\omega})}{(e^{i\omega} - \phi e^{i\omega})(e^{-i\omega} - \phi e^{-i\omega})} = \frac{\sigma^2}{2\pi} \frac{1 - 2\theta + \theta^2}{1 - 2\phi + \phi^2}$$

- If $\phi = \theta$, the ARMA(1,1) process becomes a white noise process. We can see this two ways. The ARMA(1, 1) process can be represented in the following way:

$$(1 - \phi L)y_t = (1 - \theta L)\epsilon_t$$

Therefore $\phi(L) = \theta(L)$ yields $y_t = \epsilon_t$.

We can also see that when $\phi = \theta$, an ARMA(1,1) process is equivalent to a white noise process as follows. Plugging in $\phi = \theta$ to the spectral density function, we have

$$f_x(\omega) = \frac{\sigma^2}{2\pi} \frac{1 - 2\theta + \theta^2}{1 - 2\theta + \theta^2} = \frac{\sigma^2}{2\pi}$$

showing that if $\theta = \phi$, the spectral density function is constant and independent of θ and ϕ . We can see that it in fact is a white noise process. Since a white noise process has the following covariances:

$$\gamma(0) = \sigma^2$$

$$\gamma(h) = 0, \quad \forall h \neq 0$$

for a white noise process we have

$$f_x(\omega) = \frac{1}{2\pi} \cdot \sigma^2 = \frac{\sigma^2}{2\pi}$$

10.4 Chapter 14: Estimation of Stationary Time Series Processes

10.4.1 Sufficient conditions for ergodicity of mean. (Book section 14.2.1)

By Chebyshev's Inequality (see section 8.2), \bar{y}_T is a consistent estimator of μ as $T \rightarrow \infty$ if $\lim_{T \rightarrow \infty} \mathbb{E}(\bar{y}_T) = \mathbb{E}(y_T) = \mu$ and $\lim_{T \rightarrow \infty} \text{Var}(\bar{y}_T) = 0$. We have

$$\mathbb{E}(\bar{y}_T) = \frac{1}{T} \mathbb{E}\left(\sum_{t=1}^T y_t\right) = \frac{1}{T} \sum_{t=1}^T \mathbb{E}(y_t) = \mu$$

$$\text{Var}(\bar{y}_T) = \frac{1}{T^2} \text{Var}\left(\sum_{t=1}^T y_t\right) = \frac{1}{T^2} \left(\sum_{t=1}^T \text{Var}(y_t) + 2 \sum_{0 \leq i < j \leq T} \text{Cov}(y_i, y_j) \right)$$

$$\begin{aligned}
&= \frac{1}{T^2} \left(\sum_{t=1}^T \gamma(0) + 2 \sum_{0 \leq i < j \leq T} \gamma(j-i) \right) = \frac{1}{T^2} \left(T\gamma(0) + 2 \sum_{h=1}^{T-1} (T-h)\gamma(h) \right) \\
&= \frac{1}{T} \left[\gamma(0) + 2 \sum_{h=1}^{T-1} \left(1 - \frac{h}{T} \right) \gamma(h) \right] = \frac{1}{T^2} \mathbf{1}' \text{Var}(\mathbf{y}) \mathbf{1}
\end{aligned}$$

where $\mathbf{1}$ is a vector of ones and

$$\text{Var}(\mathbf{y}) = \begin{pmatrix} \gamma(0) & \gamma(1) & \cdots & \gamma(T-2) & \gamma(T-1) \\ \gamma(1) & \gamma(0) & \cdots & \gamma(T-3) & \gamma(T-2) \\ \vdots & \vdots & \ddots & \vdots & \vdots \\ \gamma(T-2) & \gamma(T-3) & \cdots & \gamma(0) & \gamma(1) \\ \gamma(T-1) & \gamma(T-2) & \cdots & \gamma(1) & \gamma(0) \end{pmatrix}$$

Notice that

$$\left| \gamma(0) + 2 \sum_{h=1}^{T-1} \left(1 - \frac{h}{T} \right) \gamma(h) \right| < \left| 2 \sum_{h=0}^{T-1} \gamma(h) \right| \leq 2 \sum_{h=0}^{T-1} |\gamma(h)|$$

Therefore

$$\sum_{h=0}^{\infty} |\gamma(h)| < \infty$$

is a sufficient condition for

$$\lim_{T \rightarrow \infty} \text{Var}(\bar{y}_T) = \lim_{T \rightarrow \infty} \frac{1}{T} \left[\gamma(0) + 2 \sum_{h=1}^{T-1} \left(1 - \frac{h}{T} \right) \gamma(h) \right] = 0$$

10.4.2 Estimation of autocovariances (Book section 14.2.2).

A moment estimator of $\gamma(h) = \mathbb{E}[(y_t - \mu)(y_{t-h} - \mu)]$ is

$$\hat{\gamma}(h) = \frac{1}{T} \sum_{t=h+1}^T (y_t - \bar{y}_T)(y_{t-h} - \bar{y}_T)$$

By Chebyshev's Inequality (Theorem 8.4), $\hat{\gamma}(h)$ is a consistent estimator of $\gamma(h)$ as $T \rightarrow \infty$ if $\lim_{T \rightarrow \infty} \mathbb{E}(\hat{\gamma}(h)) = \gamma(h)$ and $\lim_{T \rightarrow \infty} \text{Var}(\hat{\gamma}(h)) = 0$.

$$\hat{\gamma}(h) = \frac{1}{T} \sum_{t=h+1}^T (y_t - \bar{y}_T)(y_{t-h} - \bar{y}_T) = \frac{1}{T} \sum_{t=h+1}^T (y_t - \mu + \mu - \bar{y}_T)(y_{t-h} - \mu + \mu - \bar{y}_T)$$

$$\begin{aligned}
&= \frac{1}{T} \sum_{t=h+1}^T (y_t - \mu)(y_{t-h} - \mu) + (y_t - \mu)(\mu - \bar{y}_T) + (\mu - \bar{y}_T)(y_{t-h} - \mu) + (\mu - \bar{y}_T)^2 \\
&= \frac{1}{T} \sum_{t=h+1}^T (y_t - \mu)(y_{t-h} - \mu) + (\mu - \bar{y}_T) \frac{1}{T} \sum_{t=h+1}^T (y_t - \mu) + (\mu - \bar{y}_T) \frac{1}{T} \sum_{t=h+1}^T (y_{t-h} - \mu) + \frac{1}{T}(T-h)(\mu - \bar{y}_T)^2 \\
&\quad \vdots
\end{aligned}$$

Because where does this line come from? on page 300 of book/331 of pdf.

$$\bar{y}_T = \mu + \mathcal{O}_p(T^{-1/2})$$

and for any fixed h

$$T^{-1/2} \sum_{t=h+1}^T (y_t - \mu) = \mathcal{O}_p(1)$$

it follows that

$$\begin{aligned}
(\mu - \bar{y}_T) \frac{1}{T} \sum_{t=h+1}^T (y_t - \mu) &= \frac{\mu}{T} \sum_{t=h+1}^T (y_t - \mu) - \frac{\bar{y}_T}{\sqrt{T}} \cdot \frac{1}{\sqrt{T}} \sum_{t=h+1}^T (y_t - \mu) = \mathcal{O}_p(T^{-1}) \\
(\mu - \bar{y}_T) \frac{1}{T} \sum_{t=h+1}^T (y_{t-h} - \mu) &= \mathcal{O}_p(T^{-1}) \\
\frac{1}{T}(T-h)(\mu - \bar{y}_T)^2 &= (\mu - \bar{y}_T)^2 - \frac{h}{T}(\mu - \bar{y}_T)^2 = \mathcal{O}_p(T^{-1}) \\
\implies \hat{\gamma}(h) &= \frac{1}{T} \sum_{t=h+1}^T (y_t - \mu)(y_{t-h} - \mu) + \mathcal{O}_p(T^{-1})
\end{aligned}$$

For $\hat{\gamma}(h)$ to be consistent, we need

$$\frac{1}{T} \sum_{t=h_1}^T (y_t - \mu)(y_{t-h} - \mu) \xrightarrow{p} \gamma(h)$$

First we show that $(y_t - \mu)(y_{t-h} - \mu)$ is a martingale difference process:

$$\mathbb{E}[(y_t - \mu)(y_{t-h} - \mu) | F_{t-h}] = (y_{t-h} - \mu)\mathbb{E}[y_t - \mu | F_{t-h}] = 0$$

We need to show that

$$\mathbb{E}[(y_t - \mu)^2(y_{t-h} - \mu)^2] = \mathbb{E}\left[\left(\sum_{i=0}^{\infty} \alpha_i \epsilon_{t-i}\right)^2 \left(\sum_{j=0}^{\infty} \alpha_j \epsilon_{t-h-j}\right)^2\right] < \infty$$

By the Cauchy-Schwarz Inequality (Theorem 8.6), we have

$$\begin{aligned} \mathbb{E}\left[\left(\sum_{i=0}^{\infty} \alpha_i \epsilon_{t-i}\right)^2 \left(\sum_{j=0}^{\infty} \alpha_j \epsilon_{t-h-j}\right)^2\right]^2 &\leq \mathbb{E}\left[\left(\sum_{i=0}^{\infty} \alpha_i \epsilon_{t-i}\right)^2\right]^2 \mathbb{E}\left[\left(\sum_{j=0}^{\infty} \alpha_j \epsilon_{t-h-j}\right)^2\right]^2 \\ &< \infty \iff \mathbb{E}\left[\sum_{i=0}^{\infty} \alpha_i \epsilon_{t-i}\right]^4 < \infty, \quad \mathbb{E}\left[\sum_{j=0}^{\infty} \alpha_j \epsilon_{t-h-j}\right]^4 < \infty \end{aligned}$$

These conditions hold if $\mathbb{E}(\epsilon_t^4) < \infty$ and $\sum_{i=0}^{\infty} |\alpha_i| < \infty$. Then $\mathbb{E}[(y_t - \mu)^2(y_{t-h} - \mu)^2] < \infty$ holds and

$$\hat{\gamma} \xrightarrow{P} \gamma(h)$$

10.4.3 Worked examples

Midterm Problem 3 parts (3) and (4) (similar to 14.7 and 14.8 material). Consider the following ARMA(1, 1) model

$$y_t = \phi y_{t-1} + u_t + \theta u_{t-1}, \text{ for } t = -\infty, \dots, -1, 0, 1, \dots$$

where $|\theta| < 1$, $|\phi| < 1$, and u_t is i.i.d. with mean zero and variance σ_u^2 , $\mathbb{E}(u_t^4) < \infty$.

- (1) Suppose that we have the data $\{y_t : t = 0, 1, \dots, T\}$. Consider the following estimator of ϕ :

$$\hat{\phi}_T = \frac{\sum_{t=2}^T y_t y_{t-2}}{\sum_{t=2}^T y_{t-1} y_{t-2}}$$

Show that $\hat{\phi}$ is a consistent estimator of ϕ and derive the asymptotic distribution of $\sqrt{T}(\hat{\phi}_T - \phi)$. Comment on the case where $\theta = \phi$.

- (2) Suppose that $\sigma_u^2 = 1$ is known. Show that θ can be consistently estimated by

$$\hat{\theta}_T = \frac{1}{T} \sum_{t=1}^T y_t y_{t-1} - \frac{\hat{\phi}_T}{T} \sum_{t=1}^T y_{t-1}^2$$

Solution.

- (1) From the results in Question 2 part 2(b) (in section 10.2.1), since $\mathbb{E}(y_t) = \mathbb{E}(y_{t-1}) = \mathbb{E}(y_{t-2}) = 0$, we know that

$$\hat{\phi}_T = \frac{\sum_{t=2}^T y_t y_{t-2}}{\sum_{t=2}^T y_{t-1} y_{t-2}} = \frac{T^{-1} \sum_{t=2}^T y_t y_{t-2}}{T^{-1} \sum_{t=2}^T y_{t-1} y_{t-2}} \xrightarrow{p} \frac{\gamma(2)}{\gamma(1)}$$

By the result from Question 3 part (2), we have $\gamma(h) = \phi\gamma(h-1)$ for $h \geq 2$. Therefore $\gamma(2)/\gamma(1) = \phi$, so $\hat{\phi}_T$ is a consistent estimator for ϕ . To obtain the asymptotic distribution, note that

$$\begin{aligned} \sqrt{T}(\hat{\phi}_T - \phi) &= \sqrt{T} \left(\frac{T^{-1} \sum_{t=2}^T y_t y_{t-2}}{T^{-1} \sum_{t=2}^T y_{t-1} y_{t-2}} - \phi \right) \\ &= \frac{T^{-1/2} \sum_{t=2}^T (\phi y_{t-1} + u_t + \theta u_{t-1}) y_{t-2}}{T^{-1} \sum_{t=2}^T y_{t-1} y_{t-2}} - \frac{\phi T^{-1/2} \sum_{t=2}^T y_{t-1} y_{t-2}}{T^{-1} \sum_{t=2}^T y_{t-1} y_{t-2}} \\ &= \frac{T^{-1/2} \sum_{t=2}^T (u_t + \theta u_{t-1}) y_{t-2}}{T^{-1} \sum_{t=2}^T y_{t-1} y_{t-2}} \end{aligned}$$

In Question 2 part 2(b) (in section 10.2.1), we showed that

$$\frac{1}{T} \sum_{t=h_1}^T (y_t - \mu)(y_{t-h} - \mu) \xrightarrow{p} \gamma(h)$$

Therefore in the denominator, since $\mathbb{E}(y_{t-1}) = \mathbb{E}(y_{t-h}) = 0$, we have

$$T^{-1} \sum_{t=2}^T y_{t-1} y_{t-2} \xrightarrow{p} \gamma(1)$$

In the numerator,

$$\begin{aligned} T^{-1/2} \sum_{t=2}^T (u_t + \theta u_{t-1}) y_{t-2} &= \frac{1}{\sqrt{T}} \sum_{t=2}^T [u_t y_{t-2} + \theta u_{t-1} y_{t-2}] \\ &= \frac{1}{\sqrt{T}} \sum_{t=2}^T u_t y_{t-2} + \frac{1}{\sqrt{T}} \sum_{t=2}^T \theta u_{t-1} y_{t-2} = \frac{1}{\sqrt{T}} \left(\sum_{t=2}^{T-1} u_t y_{t-2} + u_T y_{T-2} \right) + \frac{1}{\sqrt{T}} \sum_{t'=1}^{T-1} \theta u_{t'} y_{t'-1} \\ &= \frac{1}{\sqrt{T}} \left(\sum_{t=2}^{T-1} u_t y_{t-2} + u_T y_{T-2} \right) + \frac{1}{\sqrt{T}} \left(\theta u_1 y_0 + \sum_{t=2}^{T-1} \theta u_t y_{t-1} \right) = \frac{1}{\sqrt{T}} \left(\sum_{t=2}^{T-1} u_t (y_{t-2} + \theta y_{t-1}) + \theta u_1 y_0 + u_T y_{T-2} \right) \end{aligned}$$

Since $\mathbb{E}(u_t(y_{t-2} + \theta y_{t-1}) | F_{t-1}) = 0$. Further, $T^{-1/2}(\theta u_1 y_0 + u_T y_{T-2}) = o_p(1)$. Then by the Central Limit Theorem in martingale difference processes (Theorem (8.39)):

Theorem 28 (Central limit theorem for martingale difference sequences). Let $\{x_t\}$ be a martingale difference sequence with respect to the information set Ω_t . Let $\bar{\sigma}_T^2 = \text{Var}(\sqrt{T}\bar{x}_T) = T^{-1} \sum_{t=1}^T \sigma_t^2$. If $\mathbb{E}(|x_t|^r) < K < \infty$, $r > 2$ and for all t , and

$$\frac{1}{T} \sum_{t=1}^T x_t^2 - \bar{\sigma}_T^2 \xrightarrow{p} 0$$

then

$$\sqrt{T} \cdot \frac{\bar{x}_T}{\bar{\sigma}_T} \xrightarrow{d} \mathcal{N}(0, 1)$$

we have

$$\sqrt{T} \cdot \frac{\bar{x}_T}{T^{-1/2} \sqrt{\sum_{t=1}^T \sigma_t^2}} \xrightarrow{d} \mathcal{N}(0, 1)$$

⋮

$$\frac{1}{\sigma^2} \frac{\gamma(1)^2}{(1+\theta)^2 \gamma(0) + 2\theta \gamma(1)} \sqrt{T} (\hat{\phi}_T - \phi) \xrightarrow{d} \mathcal{N}(0, 1)$$

$$\iff \sqrt{T} (\hat{\phi}_T - \phi) \xrightarrow{d} \mathcal{N}\left(0, \sigma^2 \frac{(1+\theta)^2 \gamma(0) + 2\theta \gamma(1)}{\gamma(1)^2}\right)$$

- (2) From the results of Question 2 part 2(b) (in section 10.2.1), where we showed that

$$\frac{1}{T} \sum_{t=h_1}^T (y_t - \mu)(y_{t-h} - \mu) \xrightarrow{p} \gamma(h)$$

(and since $\mathbb{E}(y_{t-1}) = \mathbb{E}(y_{t-h}) = 0$,

$$T^{-1} \sum_{t=2}^T y_t y_{t-1} \xrightarrow{p} \gamma(1), \quad T^{-1} \sum_{t=2}^T y_{t-1}^2 \xrightarrow{p} \gamma(0)$$

and by the Weak Law of Large Numbers (Theorem 8.30) we have

$$\hat{\theta}_T = \frac{1}{T} \sum_{t=1}^T y_t y_{t-1} - \frac{\hat{\phi}_T}{T} \sum_{t=1}^T y_{t-1}^2 \xrightarrow{p} \gamma(1) - \phi \gamma(0) = \phi \gamma(0) + \theta \sigma^2 - \phi \gamma(0) = \theta$$

Chapter 14 Question 3. The time series $\{y_t\}$ and $\{x_t\}$ are independently generated according to the following schemes:

$$y_t = \lambda y_{t-1} + \epsilon_{1t}, \quad |\lambda| < 1$$

$$x_t = \rho x_{t-1} + \epsilon_{2t}, \quad |\rho| < 1$$

for $t = 1, 2, \dots, T$, where ϵ_{1t} and ϵ_{2t} are non-autocorrelated and distributed independently of each other with zero means and variances equal to σ_1^2 and σ_2^2 respectively. An investigator estimates the simple regression

$$y_t = \beta x_t + u_t \quad t = 1, 2, \dots, T$$

by the OLS method. Show that

$$(a) \hat{\beta} \xrightarrow{p} 0 \text{ as } T \rightarrow \infty$$

(b)

$$t_{\hat{\beta}}^2 = \frac{\hat{\beta}^2}{\widehat{\text{Var}}(\hat{\beta})} = \frac{(T-1)r^2}{1-r^2}$$

(c)

$$Tr^2 \xrightarrow{p} \frac{1+\lambda\rho}{1-\lambda\rho} \text{ as } T \rightarrow \infty$$

where $\hat{\beta}$ is the OLS estimator of β , $\widehat{\text{Var}}(\hat{\beta})$ is the estimated variance of $\hat{\beta}$, and r is the sample correlation coefficient between x and y , i.e.

$$r^2 = \left(\sum_{t=1}^T x_t y_t \right)^2 \Bigg/ \left(\sum_{t=1}^T x_t^2 \sum_{t=1}^T y_t^2 \right)$$

What are the implications of these results for problems of spurious correlation in economic time series analysis?

Solution.

(a)

$$\hat{\beta} = \frac{\sum_{t=1}^T x_t y_t}{\sum_{t=1}^T x_t^2} = \frac{T^{-1} \sum_{t=1}^T x_t y_t}{T^{-1} \sum_{t=1}^T x_t^2}$$

By the Weak Law of Large Numbers (Theorem 8.30), we have

$$T^{-1} \sum_{t=1}^T x_t y_t = T^{-1} \sum_{t=1}^T x_t y_t \xrightarrow{p} \mathbb{E}(x_t y_t) = \text{Cov}(x_t, y_t) = 0$$

because of the i.i.d. distributions of ϵ_1 and ϵ_{2t} . By the Law of Large Numbers (Theorem 8.33),

$$T^{-1} \sum_{t=1}^T x_t^2 \xrightarrow{a.s.} \mathbb{E}(x_t^2) = \gamma_x(0) > 0$$

Therefore

$$\hat{\beta} \xrightarrow{p} \frac{0}{\gamma_x(0)} = 0$$

(b)

$$\widehat{\text{Var}}(\hat{\beta}) = \frac{\hat{\sigma}^2}{S_{XX}}$$

where

$$\begin{aligned}
\hat{\sigma}^2 &= \frac{1}{T-1} \sum_{t=1}^T (y_t - \hat{\beta}x_t)^2 = \frac{1}{T-1} \sum_{t=1}^T (y_t^2 - 2y_t\hat{\beta}x_t + \hat{\beta}^2x_t^2) \\
&= \frac{1}{T-1} \sum_{t=1}^T y_t^2 - \frac{1}{T-1} \sum_{t=1}^T \left(2y_t x_t \frac{\sum_{t=1}^T x_t y_t}{\sum_{t=1}^T x_t^2} \right) + \frac{1}{T-1} \sum_{t=1}^T \left(\left[\frac{\sum_{t=1}^T x_t y_t}{\sum_{t=1}^T x_t^2} \right]^2 x_t^2 \right) \\
&= \frac{1}{T-1} \sum_{t=1}^T y_t^2 - 2 \cdot \frac{1}{T-1} \frac{(\sum_{t=1}^T x_t y_t)^2}{\sum_{t=1}^T x_t^2} + \frac{1}{T-1} \frac{(\sum_{t=1}^T x_t y_t)^2}{\sum_{t=1}^T x_t^2} \\
&= \frac{1}{T-1} \left(\sum_{t=1}^T y_t^2 - \frac{(\sum_{t=1}^T x_t y_t)^2}{\sum_{t=1}^T x_t^2} \right) \\
\implies \hat{t}^2 &= \frac{(\sum_{t=1}^T x_t y_t)^2}{(\sum_{t=1}^T x_t^2)^2} \cdot \sum_{t=1}^T x_t^2 \cdot (T-1) \Bigg/ \left(\sum_{t=1}^T y_t^2 - \frac{(\sum_{t=1}^T x_t y_t)^2}{\sum_{t=1}^T x_t^2} \right) \\
&= \frac{(\sum_{t=1}^T x_t y_t)^2}{\sum_{t=1}^T x_t^2} \cdot (T-1) \Bigg/ \left(\sum_{t=1}^T y_t^2 - \frac{(\sum_{t=1}^T x_t y_t)^2}{\sum_{t=1}^T x_t^2} \right) \\
&= \left(\sum_{t=1}^T x_t y_t \right)^2 \cdot (T-1) \Bigg/ \left(\sum_{t=1}^T y_t^2 \sum_{t=1}^T x_t^2 - (\sum_{t=1}^T x_t y_t)^2 \right)
\end{aligned}$$

Note that

$$\begin{aligned}
r^2 &= \left(\sum_{t=1}^T x_t y_t \right)^2 \Bigg/ \left(\sum_{t=1}^T x_t^2 \sum_{t=1}^T y_t^2 \right), \quad 1 - r^2 = \left[\sum_{t=1}^T x_t^2 \sum_{t=1}^T y_t^2 - \left(\sum_{t=1}^T x_t y_t \right)^2 \right] \Bigg/ \left(\sum_{t=1}^T x_t^2 \sum_{t=1}^T y_t^2 \right) \\
\implies \frac{r^2}{1-r^2} &= \left(\sum_{t=1}^T x_t y_t \right)^2 \Bigg/ \left[\sum_{t=1}^T x_t^2 \sum_{t=1}^T y_t^2 - \left(\sum_{t=1}^T x_t y_t \right)^2 \right]
\end{aligned}$$

Therefore

(c)

$$\begin{aligned}
Tr^2 &= T \cdot \frac{(T-1) \sum_{t=1}^T x_t y_t)^2}{T^{-1} \sum_{t=1}^T x_t^2 \cdot T^{-1} \sum_{t=1}^T y_t^2} = \frac{T^{-1} (2 \sum_{1 \leq i < j \leq T} x_i x_j y_i y_j + \sum_{t=1}^T x_t^2 y_t^2)}{T^{-1} \sum_{t=1}^T x_t^2 \cdot T^{-1} \sum_{t=1}^T y_t^2} \\
&= \frac{2T^{-1} \sum_{1 \leq i < j \leq T} x_i x_j y_i y_j + T^{-1} \sum_{t=1}^T x_t^2 y_t^2}{T^{-1} \sum_{t=1}^T x_t^2 \cdot T^{-1} \sum_{t=1}^T y_t^2} = \boxed{\frac{(T-1)r^2}{1-r^2}}
\end{aligned}$$

Note that

$$2T^{-1} \sum_{1 \leq i < j \leq T} (x_i x_j)(y_i y_j) = \sum_{h=1}^{T-1} 2T^{-1}(T-h)(x_t x_{t-h})(y_t y_{t-h})$$

By the Weak Law of Large Numbers (Theorem 8.30) and the independence of x_t and y_t ,

$$\sum_{h=1}^{T-1} 2T^{-1}(T-h)(x_t x_{t-h})(y_t y_{t-h}) \approx 2T^{-1} \sum_{h=1}^{T-1} (x_t x_{t-h})(y_t y_{t-h}) \xrightarrow{p} 2 \sum_{h=1}^{\infty} \gamma_x(h) \gamma_y(h)$$

Recall that for an AR(1) process with coefficient ϕ , $\gamma(h) = \phi^h \gamma(0)$. (See section 10.3.4).

$$= 2 \sum_{h=1}^{\infty} \rho^h \gamma_x(0) \lambda^h \gamma_y(0) = 2\gamma_x(0)\gamma_y(0) \sum_{h=1}^{\infty} (\rho\lambda)^h$$

By the Weak Law of Large Numbers (Theorem 8.30) and the independence of x_t and y_t ,

$$T^{-1} \sum_{t=1}^T x_t^2 y_t^2 \xrightarrow{a.s.} \mathbb{E}(x_t^2 y_t^2) = \mathbb{E}(x_t^2) \mathbb{E}(y_t^2) = \text{Var}(x_t) \text{Var}(y_t) = \sigma_1^2 \sigma_2^2 = \gamma_x(0) \gamma_y(0)$$

Therefore in the numerator, we have

$$\begin{aligned} 2T^{-1} \sum_{1 \leq i < j \leq T} x_i x_j y_i y_j + T^{-1} \sum_{t=1}^T x_t^2 y_t^2 &\xrightarrow{p} 2\gamma_x(0) \gamma_y(0) \sum_{h=1}^{\infty} (\rho\lambda)^h + \gamma_x(0) \gamma_y(0) \\ &= \gamma_x(0) \gamma_y(0) \left(2 \frac{\rho\lambda}{1 - \rho\lambda} + 1 \right) = \gamma_x(0) \gamma_y(0) \left(\frac{1 + \rho\lambda}{1 - \rho\lambda} \right) \end{aligned}$$

Next we examine the denominator. By the Law of Large Numbers (Theorem 8.33),

$$T^{-1} \sum_{t=1}^T x_t^2 \xrightarrow{a.s.} \mathbb{E}(x_t^2) = \text{Var}(x_t) = \gamma_x(0) = \sigma_2^2$$

By the Law of Large Numbers (Theorem 8.33),

$$T^{-1} \sum_{t=1}^T y_t^2 \xrightarrow{a.s.} \mathbb{E}(y_t^2) = \gamma_y(0) = \sigma_1^2$$

Therefore in the denominator, we have

$$T^{-1} \sum_{t=1}^T x_t^2 \cdot T^{-1} \sum_{t=1}^T y_t^2 \xrightarrow{a.s.} \gamma_x(0) \gamma_y(0) = \sigma_1^2 \sigma_2^2$$

This yields

$$Tr^2 \xrightarrow{p} \frac{1 + \rho\lambda}{1 - \rho\lambda}$$

Lidan's explanation: Because as $T \rightarrow \infty$

$$t_{\hat{\beta}} \rightarrow \sqrt{Tr^2} = \sqrt{\frac{1+\lambda\rho}{1-\lambda\rho}}$$

so if $\lambda\rho \approx 1$, at very high probability we will reject the null $\hat{\beta} = 0$ when in fact $\hat{\beta} \xrightarrow{p} 0$.

My original explanation: Because $\hat{\beta}$ converges in probability to 0 and $t_{\hat{\beta}}^2$ is proportional to r^2 (which we would expect to be close to 0), this suggests that a regression of uncorrelated variables should result in an insignificant $\hat{\beta}$, but if r^2 is high due to a spurious correlation, the $\hat{\beta}$ could be found to be statistically significant even if there is no meaningful relationship between x and y .

10.5 Chapter 15: Unit Root Processes

Need to review concepts in this chapter.

10.5.1 Worked Problems

Problem 3. Suppose that a time series of interest can be decomposed into a deterministic trend, a random walk component, and stationary errors:

$$y_t = \alpha + \delta t + \gamma_t + v_t \quad (11)$$

$$\gamma_t = \gamma_{t-1} + u_t$$

with

$$v_t \sim iid \mathcal{N}(0, \sigma_v^2), \quad u_t \sim iid \mathcal{N}(0, \sigma_u^2), \quad u_t \perp\!\!\!\perp v_t$$

Let $\lambda = \sigma_u^2 / \sigma_v^2$.

- (a) Show that under $\lambda = 0$, y_t reduces to a trend stationary process.
- (b) Alternatively, suppose that y_t follows an ARIMA(0,1,1) process of the form

$$y_t = \delta + y_{t-1} + w_t \quad (12)$$

$$w_t = \epsilon_t + \theta \epsilon_{t-1}$$

where ϵ_t are iid $\mathcal{N}(0, \sigma_\epsilon^2)$. In this case show that under $\theta = -1$, y_t is a trend stationary process.

- (c) Derive a relation between λ and the MA(1) parameter θ , and hence or otherwise show that a test of $\theta = -1$ in (12) is equivalent to a test of $\lambda = 0$ in (11).
- (d) Show that (12) as a characterization of (11) implies $\theta < 0$.

Solution.

(a) $\lambda = 0 \implies \sigma_u^2 = 0 \implies u_t = 0$ (constant)

$$\implies \gamma_t = \gamma_{t-1} + 0 \iff \gamma_t = \gamma_0$$

$$\implies y_t = \alpha + \delta t + \gamma_0 + v_t = (\alpha + \gamma_0) + \delta t + v_t$$

which is trend stationary because $d_t = (\alpha + \gamma_0) + \delta t$ is perfectly predictable, and $y_t - d_t = v_t$ is covariance stationary.

Also note that $\text{Var}(y_t - \delta t) = \sigma_v^2$, $\text{Cov}([y_t - \delta t][y_{t-h} - \delta(t-h)]) = 0$, which implies trend stationarity of y_t . (Recall Definition 10.2: “ X_t is said to be **trend stationary** if $y_t = X_t - d_t$ is covariance stationary, where d_t is the perfectly predictable component of X_t .”)

(b) $\theta = -1 \implies w_t = \epsilon_t - \epsilon_{t-1} = (1 - L)\epsilon_t$

$$\implies (1 - L)y_t = \delta + (1 - L)\epsilon_t \iff y_t = (1 - L)^{-1}\delta + \epsilon_t$$

which is trend stationary because $d_t = (1 - L)^{-1}\delta$ is perfectly predictable, and $y_t - d_t = \epsilon_t$ is covariance stationary.

(c) **Lidan's solution:** From (11), let

$$\begin{aligned} z_t &= y_t - y_{t-1} = \alpha + \delta t + \gamma_{t-1} + u_t + v_t - (\alpha + \delta(t-1) + \gamma_{t-1} + v_{t-1}) \\ &= \delta + u_t + v_t - v_{t-1} \end{aligned}$$

From (12), let

$$a_t = y_t - y_{t-1} = \delta + y_{t-1} + w_t - y_{t-1} = \delta + \epsilon_t + \theta\epsilon_{t-1}$$

Calculate the autocovariances for each and set them equal (using the serial independence of u_t , v_t , and ϵ_t as well as the independence of u_t , v_t , and ϵ_t for all t):

•

$$\gamma(0) : \text{Var}(z_t) = \text{Var}(a_t) \iff \text{Var}(u_t + v_t - v_{t-1}) = \text{Var}(\epsilon_t + \theta\epsilon_{t-1})$$

$$\iff \sigma_u^2 + 2\sigma_v^2 = \sigma_\epsilon^2(1 + \theta^2) \tag{13}$$

- $\gamma(1) : \text{Cov}(z_t, z_{t-1}) = \text{Cov}(a_t, a_{t-1}) \iff \text{Cov}(-v_{t-1}, v_{t-1}) = \text{Cov}(\theta\epsilon_{t-1}, \epsilon_{t-1})$
- $\iff -\sigma_v^2 = \theta\sigma_\epsilon^2 \quad (14)$
- $\gamma(h) (h \geq 2) : 0 = 0$

Plugging (14) into (13) and using $\lambda = \sigma_u^2/\sigma_v^2$ we have

$$\begin{aligned} \sigma_u^2 + 2\sigma_v^2 &= -\frac{\sigma_v^2}{\theta}(1 + \theta^2) \iff \sigma_u^2 = \sigma_v^2(-1/\theta - \theta - 2) \iff \lambda = -1/\theta - \theta - 2 \\ \iff \theta^2 + (2 + \lambda)\theta + 1 &= 0 \iff \theta = \frac{-(2 + \lambda) \pm \sqrt{(2 + \lambda)^2 - 4}}{2} \\ \iff \theta &= \boxed{\frac{-(2 + \lambda) \pm \sqrt{\lambda^2 + 4\lambda}}{2}} \end{aligned}$$

Clearly if $\lambda = 0$ then $\theta = -1$. The reverse is also true:

$$\theta = -1 \implies \lambda = -1/(-1) - (-1) - 2 = 1 + 1 - 2 = 0$$

Therefore a test of $\theta = -1$ in (12) is equivalent to a test of $\lambda = 0$ in (11).

original solution:

$$y_t = \alpha + \delta t + \gamma_{t-1} + u_t + v_t$$

$$y_{t-1} = \alpha + \delta(t-1) + \gamma_{t-1} + v_{t-1}$$

$$\implies y_t = y_{t-1} + \delta + u_t - v_{t-1} + v_t$$

Comparing this to the second expression, $y_t = \delta + y_{t-1} + \epsilon_t + \theta\epsilon_{t-1}$, they match if $u_t - v_{t-1} + v_t = \epsilon_t + \theta\epsilon_{t-1}$. Since the distributions of ϵ_t and v_t are both i.i.d. normal, this is the case if $\lambda = 0$ (so that $\sigma_u^2 = 0$ and $u_t = 0 \forall t$), $\theta = -1$, and $\sigma_\epsilon^2 = \sigma_v^2$.

(d) **Lidan's Solution:** By (14) (which follows from (12)),

$$-\sigma_v^2 = \theta\sigma_\epsilon^2 \implies \theta < 0$$

My solution: Again consider (2)

$$y_t = \delta + y_{t-1} + \epsilon_t + \theta\epsilon_{t-1}$$

which is clearly a unit root process with a drift, compared to a re-written version of (1)

$$y_t = \delta + y_{t-1} + u_t + v_t - v_{t-1}$$

which has a strong resemblance to a unit root process with a drift. These match up if $\epsilon_t = u_t + v_t$ and $\theta < 0$.

Problem 4

Homework 4 Problem 3. Let $\{u_t\}$ be an i.i.d. sequence with mean zero and variance σ^2 , and let

$$y_t = u_1 + u_2 + \dots + u_t$$

with $y_0 = 0$.

(a) Show that

$$T^{-3/2} \sum_{t=1}^T y_{t-1} = T^{-1/2} \sum_{t=1}^{T-1} u_t - T^{-3/2} \sum_{t=1}^{T-1} t u_t$$

(b) Show that

$$\begin{bmatrix} T^{-1/2} \sum_{t=1}^T u_t \\ T^{-3/2} \sum_{t=1}^T y_{t-1} \end{bmatrix} \xrightarrow{d} \mathcal{N}\left(0, \begin{bmatrix} 1 & 1/2 \\ 1/2 & 1/3 \end{bmatrix}\right)$$

(c) Use the functional central limit theorem (Donsker's Theorem, Theorem 7.3) to show that

$$T^{-3/2} \sum_{t=1}^T y_{t-1} \xrightarrow{d} \sigma \int_0^1 W(r) dr$$

where $W(\cdot)$ is a standard Brownian motion process.

(d) Use parts (a) - (c) to show that

$$T^{-3/2} \sum_{t=1}^T t u_t \xrightarrow{d} \sigma \cdot W(a) - \sigma \int_0^1 W(r) dr$$

Solution.

$$y_t = u_1 + u_2 + \dots + u_t = \sum_{i=1}^t u_i = \sum_{i=0}^t u_i$$

where the last equality follows because $y_0 = 0 \iff u_0 = 0$.

$$u_t \sim iid (0, \sigma^2), \quad y_0 = 0$$

(a)

$$y_T = \sum_{t=1}^T u_t$$

$$T^{-3/2} \sum_{t=1}^T y_{t-1} = T^{-3/2} \sum_{t=1}^T \left(\sum_{i=0}^{t-1} u_i \right) = T^{-3/2} \sum_{t=1}^T (T - (t-1)) u_{t-1} = T^{-3/2} \sum_{t'=0}^{T-1} (T - t') u_{t'}$$

$$\begin{aligned}
&= T^{-3/2} \cdot T \sum_{t'=0}^{T-1} u_{t'} - T^{-3/2} \sum_{t'=0}^{T-1} t' u_{t'} = T^{-1/2} \sum_{t'=1}^{T-1} u_{t'} - T^{-3/2} \sum_{t'=1}^{T-1} t' u_{t'} \\
&\implies \boxed{T^{-3/2} \sum_{t=1}^T y_{t-1} = T^{-1/2} \sum_{t'=1}^{T-1} u_{t'} - T^{-3/2} \sum_{t'=1}^{T-1} t' u_{t'}} \tag{15}
\end{aligned}$$

(b) By the Central Limit Theorem (Theorem 8.36), since $u_t \sim (0, \sigma^2)$, we have

$$\frac{1}{\sqrt{T}\sigma^2} \sum_{t=1}^T u_t \xrightarrow{d} \mathcal{N}(0, 1)$$

which implies

$$\boxed{\frac{1}{\sqrt{T}} \sum_{t=1}^T u_t \xrightarrow{d} \mathcal{N}(0, \sigma^2)} \tag{16}$$

The distribution of $T^{-3/2} \sum_{t=1}^T y_{t-1}$ is trickier. Note that from (15),

$$T^{-3/2} \sum_{t=1}^T y_{t-1} = T^{-1/2} \left(\sum_{t=0}^{T-1} u_t - \sum_{t=0}^{T-1} \frac{tu_t}{T} \right)$$

Note that $x_t = (tu_t)/T$ is a martingale difference process because

$$\mathbb{E}\left(\frac{tu_t}{T} \mid u_{t-1}, u_{t-2}, \dots\right) = 0$$

and $\text{Var}(x_t) = \frac{1}{T^2} \text{Var}(tu_t) = \frac{t^2}{T^2} \sigma^2 < \infty$. By the Central Limit Theorem in martingale difference processes (Theorem (8.39)):

Theorem 28 (Central limit theorem for martingale difference sequences). Let $\{x_t\}$ be a martingale difference sequence with respect to the information set Ω_t . Let $\bar{\sigma}_T^2 = \text{Var}(\sqrt{T}\bar{x}_T) = T^{-1} \sum_{t=1}^T \sigma_t^2$. If $\mathbb{E}(|x_t|^r) < K < \infty$ for any $r > 2$ and for all t , and

$$\frac{1}{T} \sum_{t=1}^T x_t^2 - \bar{\sigma}_T^2 \xrightarrow{p} 0$$

then

$$\sqrt{T} \cdot \frac{\bar{x}_T}{\bar{\sigma}_T} \xrightarrow{d} \mathcal{N}(0, 1)$$

since

$$T^{-1} \sum_{t=1}^T \sigma_t^2 = T^{-1} \sum_{t=1}^T \frac{t^2}{T^2} \sigma^2 = \frac{\sigma^2}{T^3} \cdot \frac{T(T+1)(2T+1)}{6} = \frac{\sigma^2}{6} \cdot \frac{2T^2 + 3T + 1}{T^2},$$

$$\mathbb{E}(|x_t|^r) = \mathbb{E}\left(\left|\frac{tu_t}{T}\right|^r\right) = \frac{t^r}{T^r} \mathbb{E}(|u_t|^r)$$

which is finite for $r = 4$ if u_t has finite fourth moment, and since by the Weak Law of Large Numbers (Theorem 8.30)

$$\begin{aligned}
T^{-1} \sum_{t=1}^T x_t^2 - T^{-1} \sum_{t=1}^T \frac{t^2}{T^2} \sigma^2 &= T^{-1} \sum_{t=1}^T \frac{t^2}{T^2} (u_t^2 - \sigma^2) \\
\left| T^{-1} \sum_{t=1}^T \frac{t^2}{T^2} (u_t^2 - \sigma^2) \right| &\leq \left| T^{-1} \sum_{t=1}^T (u_t^2 - \sigma^2) \right| \leq T^{-1} \sum_{t=1}^T |u_t^2 - \sigma^2| \xrightarrow{p} 0 \\
\implies T^{-1} \sum_{t=1}^T x_t^2 - T^{-1} \sum_{t=1}^T \frac{t^2}{T^2} \sigma^2 &\xrightarrow{p} 0
\end{aligned}$$

we have (if u_t has finite fourth moment)

$$\begin{aligned}
&\sqrt{T} \cdot \frac{1}{T} \left(\sum_{t=1}^T \frac{tu_t}{T} \right) \Bigg/ \sqrt{\frac{\sigma^2}{6} \cdot \frac{2T^2 + 3T + 1}{T^2}} \xrightarrow{d} \mathcal{N}(0, 1) \\
\iff T^{-3/2} \left(\sum_{t=1}^T tu_t \right) \cdot \sqrt{\frac{6}{\sigma^2} \cdot \frac{1}{2T^2 + 3T + 1}} &\xrightarrow{d} \mathcal{N}(0, 1) \implies \sqrt{T} \left(\sum_{t=1}^T tu_t \right) \cdot \sqrt{\frac{6}{\sigma^2} \cdot \frac{1}{2T^2}} \xrightarrow{d} \mathcal{N}(0, 1) \\
\iff T^{-3/2} \left(\sum_{t=1}^T \frac{tu_t}{T} \right) \cdot \sqrt{\frac{3}{\sigma^2}} &\xrightarrow{d} \mathcal{N}(0, 1) \iff T^{-1/2} \frac{\sqrt{3}}{\sigma} \sum_{t=0}^{T-1} \frac{tu_t}{T} \xrightarrow{d} \mathcal{N}\left(0, \frac{\sigma^2}{3}\right) \\
\iff T^{-1/2} \sum_{t=0}^{T-1} \frac{tu_t}{T} &\xrightarrow{d} \mathcal{N}\left(0, \frac{\sigma^2}{3}\right)
\end{aligned} \tag{17}$$

To get the covariance between these distributions, we have (using the serial independence of u_t)

$$\begin{aligned}
\text{Cov}\left(T^{-1/2} \sum_{t=0}^{T-1} u_t, T^{-3/2} \sum_{t=0}^{T-1} tu_t\right) &= \mathbb{E}\left[T^{-1/2} \sum_{t=0}^{T-1} u_t \cdot T^{-3/2} \sum_{t=0}^{T-1} tu_t\right] = T^{-2} \cdot \mathbb{E}\left[\sum_{t=0}^{T-1} u_t \cdot \sum_{t=0}^{T-1} tu_t\right] \\
&= T^{-2} \cdot \mathbb{E}\left[\sum_{t=0}^{T-1} tu_t^2\right] = T^{-2} \sum_{t=0}^{T-1} t \mathbb{E}(u_t^2) = \frac{\sigma^2}{T^2} \cdot \frac{T(T-1)}{2} \rightarrow \boxed{\frac{\sigma^2}{2}}
\end{aligned} \tag{18}$$

Putting (16), (17), and (18) together, we have $T^{-3/2} \sum_{t=1}^T y_{t-1} \xrightarrow{d} \mathcal{N}(0, \sigma^2 + \sigma^2/3 - 2 \cdot \sigma^2/2) = \boxed{\mathcal{N}(0, \sigma^2/3)}$.

Lastly, to get the covariance between these distributions, we have (again using the serial independence of u_t)

$$\begin{aligned}
\text{Cov}\left(T^{-1/2} \sum_{t=1}^{T-1} u_t, T^{-3/2} \sum_{t=1}^T y_{t-1}\right) &= T^{-2} \text{Cov}\left(\sum_{t=1}^{T-1} u_t, \sum_{t=1}^T (T-(t-1))u_{t-1}\right) \\
&= T^{-2} \text{Cov}\left(\sum_{t=1}^{T-1} u_t, \sum_{t'=0}^{T-1} (T-t')u_{t'}\right) = T^{-2} \mathbb{E}\left(\sum_{t=1}^{T-1} (T-t)u_t^2\right) = \frac{(T-1)T - T(T-1)/2}{T^2} \sigma^2 \\
&= \frac{(T-1)T}{2T^2} \sigma^2 \rightarrow \boxed{\sigma^2/2}
\end{aligned} \tag{19}$$

which yields the result.

(c) Let $r \in [0, 1)$, $t \in [0, T]$. Define

$$R_T(r) = \frac{1}{\sigma\sqrt{T}}y_{[rT]}$$

where $[rT]$ denotes the largest integer part of rT and $y_{[rT]} = 0$ if $[rT] = 0$. That is,

$$R_T(r) = \begin{cases} 0 & 0 \leq r < 1/T \\ \frac{y_1}{\sigma\sqrt{T}} & 1/T \leq r < 2/T \\ \frac{y_2}{\sigma\sqrt{T}} & 2/T \leq r < 3/T \\ \vdots & \vdots \\ \frac{y_{T-1}}{\sigma\sqrt{T}} & (T-1)/T \leq r < 1 \end{cases}$$

We have

$$\begin{aligned} T^{-3/2} \sum_{t=1}^T y_{t-1} &= T^{-3/2} \sum_{t'=0}^{T-1} y_{t'} = T^{-3/2} \sum_{t'=0}^{T-1} (y_{t'-1} + u_{t'}) = \sigma \sum_{t'=0}^{T-1} \int_{t'/T}^{(t'+1)/T} R_T(r) dr + o_p(1) \\ &= \sigma \int_0^1 R_T(r) dr + o_p(1) \implies \sigma \int_0^1 W(r) dr \text{ as } T \rightarrow \infty \end{aligned}$$

where the last step follows from Donsker's Theorem (Theorem 7.3, the functional central limit theorem) and the continuous mapping theorem (Theorem 7.4) and $W(r)$ is a standard Weiner process.

Donsker's Theorem, Theorem 43, p.335, Section 15.6.3. Let $a \in [0, 1)$, $t \in [0, T]$, and suppose $(J-1)/T \leq a < J/T$, $J = 1, 2, \dots, T$. Define

$$R_T(a) = \frac{1}{\sqrt{T}}s_{[Ta]}$$

where

$$s_{[Ta]} = \epsilon_1 + \epsilon_2 + \dots + \epsilon_{[Ta]}$$

$[Ta]$ denotes the largest integer part of Ta and $s_{[Ta]} = 0$ if $[Ta] = 0$. Then $R_T(a)$ weakly converges to $w(a)$, i.e.,

$$R_T(a) \rightarrow w(a)$$

where $w(a)$ is a Wiener process. Note that when $a = 1$, $R_T(1) = 1/\sqrt{T} \cdot S_{[T]} = 1/\sqrt{T} \cdot (\epsilon_1 + \epsilon_2 + \dots + \epsilon_T)$. Since ϵ_t 's are IID, by the central limit theorem, $R_T(1) \rightarrow \mathcal{N}(0, 1)$.

Continuous Mapping Theorem (Theorem 44 of Pesaran in 15.6.3). Let $a \in [0, 1)$, $i \in [0, n]$, and suppose $(J-1)/n \leq a < J/n$, $J = 1, 2, \dots, n$. Define $R_n(a) = n^{-1/2}S_{[n \cdot a]}$. If $f(\cdot)$ is continuous over $[0, 1]$, then

$$f[R_n(a)] \xrightarrow{d} f[w(a)]$$

(d) We have $T^{-1/2} \sum_{t=1}^T u_t = \sigma \cdot R_T(1) \implies \sigma \cdot W(1)$ by Donsker's Theorem. Using that and the result from (c), we have

$$T^{-3/2} \sum_{t=1}^{T-1} t u_t = T^{-1/2} \sum_{t=1}^{T-1} u_t - T^{-3/2} \sum_{t=1}^T y_{t-1} \xrightarrow{d} \sigma \cdot W(1) - \sigma \int_0^1 W(r) dr$$

10.6 Chapter 17: Introduction to Forecasting

Feel pretty good on concepts

10.6.1 17.7: Iterated and direct multi-step AR methods

Suppose y_t follows the AR(1) model:

$$y_t = a + \phi y_{t-1} + \epsilon_t, \quad |\phi| < 1, \epsilon_t \sim iid(0, \sigma_\epsilon^2) \quad (20)$$

$$\iff y_t = \frac{a}{1-\phi} + \sum_{i=0}^{\infty} \phi^i \epsilon_{t-i}$$

$$\iff y_t = a \left(\frac{1 - \phi^h}{1 - \phi} \right) + \phi^h y_{t-h} + \sum_{j=0}^{h-1} \phi^j \epsilon_{t-j} \quad (21)$$

We have two methods for forecasting y_{t+h} $h > 1$ steps ahead.

- (1) **Iterated method:** In this method, we first calculate the OLS estimates of \hat{a}_T and $\hat{\phi}_T$ in Equation (22) using all available data Ω_T . Then we use the form of Equation (21):

$$\hat{y}_{T+h|T}^* = \hat{a}_T \left(\frac{1 - \hat{\phi}_T^h}{1 - \hat{\phi}_T} \right) + \hat{\phi}_T^h y_T$$

- (2) **Direct method:** We directly calculate OLS estimates of the parameters in Equation (21) using all available data Ω_T :

$$\tilde{y}_{T+h|T}^* = \tilde{a}_{h,T} + \tilde{\phi}_{h,T} y_T$$

Proposition 10.1. (Pesaran Chapter 17 Proposition 45.) Suppose data is generated by Equation (22). If $u_t = \sum_{i=0}^{\infty} \phi^i \epsilon_{t-i}$ and $v_t = \sum_{j=0}^{h-1} \phi^j \epsilon_{t-j}$ are symmetrically distributed around zero and have finite second moments, and if $\mathbb{E}(\hat{\phi}_T)$ and $\mathbb{E}(\tilde{\phi}_{h,T})$ exist, then for any finite T and h we have

$$\mathbb{E}(\hat{y}_{T+h|T}^* - y_{T+h}) = \mathbb{E}(\tilde{y}_{T+h|T}^* - y_{T+h}) = 0$$

10.6.2 Worked Problems

Problem 4 (Homework 5 Question 3)—fine on all but part (c), which even Lidan is hazy on.

Consider the AR(1) model

$$y_t = \phi y_{t-1} + u_t, \quad u_t \sim iid(0, \sigma^2) \quad (22)$$

- (a) Derive iterated and direct forecasts of y_{T+2} condition on y_T , and show that they can be estimated as

$$\text{Iterated: } \hat{y}_{T+2|T}^{(it)} = \hat{\phi}^2 y_T$$

$$\text{Direct: } \hat{y}_{T+2|T}^{(d)} = \hat{\phi}_2 y_T$$

where $\hat{\phi}$ and $\hat{\phi}_2$ are OLS coefficients in the regressions of y_t on y_{t-1} and y_{t-2} , respectively, using the M observations $y_{T-M+1}, y_{T-M+2}, \dots, y_T$.

- (b) Show that conditional on y_t ,

$$\mathbb{E}(y_{T+2} - \hat{y}_{T+2|T}^{(it)})^2 = \mathbb{E}(\phi^2 - \hat{\phi}^2)^2 y_T^2 + (1 + \phi^2)\sigma^2$$

$$\mathbb{E}(y_{T+2} - \hat{y}_{T+2|T}^{(d)})^2 = \mathbb{E}(\phi^2 - \hat{\phi}_2)^2 y_T^2 + (1 + \phi^2)\sigma^2$$

- (c) Hence, or otherwise, show that

$$\lim_{M \rightarrow \infty} \mathbb{E}(d_{T+2}) = 0$$

where d_{T+2} is the loss differential of the two forecasting methods, defined by

$$d_{T+2} = (y_{T+2} - \hat{y}_{T+2|T}^{(it)})^2 + (y_{T+2} - \hat{y}_{T+2|T}^{(d)})^2$$

Solution.

- (a) Note that we can write

$$y_t = \phi[\phi y_{t-2} + u_{t-1}] + u_t = \phi^2 y_{t-2} + u_t + \phi u_{t-1}$$

$$\implies y_{T+2} = \phi^2 y_T + u_{T+2} + \phi u_{T+1}$$

$$\implies \mathbb{E}(y_{T+2} | y_{T-M+1}, y_{T-M+2}, \dots, y_T) = \phi^2 y_T + \mathbb{E}(u_{T+2} | y_{T-M+1}, \dots, y_T) + \phi \mathbb{E}(u_{T+1} | y_{T-M+1}, \dots, y_T)$$

$$= \phi^2 y_T$$

$$\iff \mathbb{E}(y_{T+2} | y_{T-M+1}, y_{T-M+2}, \dots, y_T) = \phi^2 y_T \quad (23)$$

If we calculate the OLS estimate $\hat{\phi}$ in Equation (22), we can substitute that into Equation (23) to obtain the iterated estimate of y_{T+2} :

$$\hat{y}_{T+2|T}^{(it)} = \hat{\phi}^2 y_T$$

Since we have no intercept term, the OLS estimate would be simply

$$\hat{\phi} = (x'x)^{-1}x'y = \frac{\sum_{t=T-M+2}^T y_t y_{t-1}}{\sum_{t=T-M+1}^{T-1} y_t^2}$$

Alternatively, we could directly calculate the OLS estimate $\hat{\phi}_2$ of ϕ^2 in Equation (23)

$$\hat{y}_{T+2|T}^{(d)} = \hat{\phi}_2 y_T$$

The OLS estimate would be simply

$$\hat{\phi}_2 = (x'x)^{-1}x'y = \frac{\sum_{t=T-M+3}^T y_t y_{t-2}}{\sum_{t=T-M+1}^{T-2} y_t^2}$$

(b) We have

$$\begin{aligned} \mathbb{E}(y_{T+2} - \hat{y}_{T+2|T}^{(it)})^2 &= \mathbb{E}(y_{T+2}^2) + \mathbb{E}((\hat{y}_{T+2|T}^{(it)})^2) - 2\mathbb{E}(y_{T+2} \cdot \hat{y}_{T+2|T}^{(it)}) \\ &= \mathbb{E}((\phi^2 y_T + u_{T+2} + \phi u_{T+1})^2) + \mathbb{E}((\hat{\phi}^2 y_T)^2) - 2\mathbb{E}((\phi^2 y_T + u_{T+2} + \phi u_{T+1}) \cdot \hat{\phi}^2 y_T) \\ &= \mathbb{E}(\phi^4 y_T^2 + (u_{T+2} + \phi u_{T+1})^2 + 2\phi^2 y_T(u_{T+2} + \phi u_{T+1})) + y_T^2 \mathbb{E}(\hat{\phi}^2) - 2\phi^2 y_T \mathbb{E}(\hat{\phi}^2 y_T) \\ &= \mathbb{E}(\phi^4) y_T^2 + \mathbb{E}(u_{T+2}^2 + \phi^2 u_{T+1}^2) + y_T^2 \mathbb{E}(\hat{\phi}^2) - 2\phi^2 y_T^2 \mathbb{E}(\hat{\phi}^2) = \mathbb{E}(\phi^4) y_T^2 + (1 + \phi^2) \sigma^2 + y_T^2 \mathbb{E}(\hat{\phi}^2) - 2\phi^2 y_T^2 \mathbb{E}(\hat{\phi}^2) \\ &= y_T^2 [\mathbb{E}(\phi^4 + \hat{\phi}^2 - 2\phi^2 \hat{\phi}^2)] + (1 + \phi^2) \sigma^2 = \boxed{\mathbb{E}(\phi^2 - \hat{\phi}^2)^2 y_T^2 + (1 + \phi^2) \sigma^2} \\ \mathbb{E}(y_{T+2} - \hat{y}_{T+2|T}^{(d)})^2 &= \mathbb{E}(y_{T+2}^2) + \mathbb{E}((\hat{y}_{T+2|T}^{(d)})^2) - 2\mathbb{E}(y_{T+2} \cdot \hat{y}_{T+2|T}^{(d)}) \\ &= \mathbb{E}((\phi^2 y_T + u_{T+2} + \phi u_{T+1})^2) + \mathbb{E}((\hat{\phi}_2 y_T)^2) - 2\mathbb{E}((\phi^2 y_T + u_{T+2} + \phi u_{T+1}) \cdot \hat{\phi}_2 y_T) \\ &= \mathbb{E}(\phi^4 y_T^2 + (u_{T+2} + \phi u_{T+1})^2 + 2\phi^2 y_T(u_{T+2} + \phi u_{T+1})) + y_T^2 \mathbb{E}(\hat{\phi}_2) - 2\phi^2 y_T \mathbb{E}(\hat{\phi}_2 y_T) \\ &= \mathbb{E}(\phi^4) y_T^2 + \mathbb{E}(u_{T+2}^2 + \phi^2 u_{T+1}^2) + y_T^2 \mathbb{E}(\hat{\phi}_2) - 2\phi^2 y_T^2 \mathbb{E}(\hat{\phi}_2) = \mathbb{E}(\phi^4) y_T^2 + (1 + \phi^2) \sigma^2 + y_T^2 \mathbb{E}(\hat{\phi}_2) - 2\phi^2 y_T^2 \mathbb{E}(\hat{\phi}_2) \\ &= y_T^2 [\mathbb{E}(\phi^4 + \hat{\phi}_2 - 2\phi^2 \hat{\phi}_2)] + (1 + \phi^2) \sigma^2 = \boxed{\mathbb{E}(\phi^2 - \hat{\phi}_2)^2 y_T^2 + (1 + \phi^2) \sigma^2} \end{aligned}$$

(c) Begin by expanding $\hat{\phi}^2$ in a first order Taylor series about ϕ :

$$\hat{\phi} = \phi + \mathcal{O}_p(M^{-1/2}) \implies \hat{\phi}^2 = \phi^2 + 2\phi(\hat{\phi} - \phi) + \mathcal{O}_p(M^{-1})$$

Then

$$\begin{aligned}\mathbb{E}[(\hat{\phi}^2 - \phi^2)^2 | y_T] &= \mathbb{E}[(2\phi(\hat{\phi} - \phi) + \mathcal{O}_p(M^{-1}))^2 | y_T] = 4\phi^2 \mathbb{E}[(\hat{\phi} - \phi + \mathcal{O}_p(M^{-1}))^2 | y_T] \\ &= 4\phi^2 \mathbb{E}[(\hat{\phi} - \phi)^2 + 2\phi(\hat{\phi} - \phi)\mathcal{O}_p(M^{-1}) + \mathcal{O}_p(M^{-1})^2 | y_T]\end{aligned}$$

10.7 Chapter 18: Measurement and Modeling of Volatility

Maybe review a little, but feel pretty okay on concepts

GARCH(1, 1) model (Pesaran Equation 18.5):

$$h_t^2 = \alpha_0 + \alpha_1 \epsilon_{t-1}^2 + \phi_1 h_{t-1}^2, \quad \alpha_0 > 0 \quad (24)$$

- This process is unconditionally stationary if $|\alpha_1 + \phi_1| < 1$.
- The unconditional variance exists and is fixed if $|\alpha_1 + \phi_1| < 1$.
- The case where $\alpha_1 + \phi_1 = 1$ is known as the Integrated GARCH(1,1), or IGARCH(1,1) for short. The RiskMetrics exponentially weighted formulation of h_t^2 for large H is a special case of the IGARCH(1,1) model where α_0 is set to 0. RiskMetrics formulation avoids the variance non-existence problem by focusing on H fixed.

10.7.1 Higher order GARCH models (Pesaran Section 18.4.2)

The various members of the GARCH and GARCH-M class of models can be written compactly as

$$y_t = \beta' \mathbf{x}_{t-1} + \gamma h_t^2 + \epsilon_t \quad (25)$$

where

$$h_t^2 = \text{Var}(\epsilon_t | \Omega_{t-1}) = \mathbb{E}(\epsilon_t^2 | \Omega_{t-1}) = \alpha_0 + \sum_{i=1}^q \alpha_i \epsilon_{t-i}^2 + \sum_{i=1}^p \phi_i h_{t-i}^2 \quad (26)$$

and Ω_{t-1} is the information set at time $t - 1$ containing at least $(\mathbf{x}_{t-1}, \mathbf{x}_{t-2}, \dots, y_{t-1}, y_{t-2}, \dots)$. The unconditional variance of ϵ_t is determined by

$$\sigma_t^2 = \alpha_0 + \sum_{i=1}^q \alpha_i \sigma_{t-i}^2 + \sum_{i=1}^p \phi_i \sigma_{t-i}^2$$

and yields a stationary outcome if all the roots of

$$1 - \sum_{i=1}^q \alpha_i \lambda^i + \sum_{i=1}^p \phi_i \lambda^i = 0$$

lie outside the unit circle. In that case

$$\text{Var}(\epsilon_t) = \sigma^2 = \frac{\alpha_0}{1 - \sum_{i=1}^q \alpha_i - \sum_{i=1}^p \phi_i} > 0 \quad (27)$$

Clearly the necessary condition for (26) to be covariance stationary is given by

$$\sum_{i=1}^q \alpha_i + \sum_{i=1}^p \phi_i < 1$$

10.7.2 Testing for GARCH effects (Pesaran Section 18.5.1)

If we consider y_t as periodic data defined by $y_t = r_t - \bar{r}$ with r_t representing, say, asset return and \bar{r} representing the unconditional mean, then we have the GARCH(1,1) representation of volatility:

$$\text{Var}(y_t | \Omega_{t-1}) = h_t^2 = \bar{\sigma}^2(1 - \alpha - \beta) + \alpha y_{t-1}^2 + \beta h_{t-1}^2$$

Then the test for GARCH effects would test

$$H_0 : \alpha = 0$$

against

$$H_1 : \alpha \neq 0$$

GARCH(1,1) can be approximated by ARCH(q):

$$\begin{aligned} \text{Var}(y_t | \Omega_{t-1}) &= \frac{\bar{\sigma}^2(1 - \alpha - \beta)}{1 - \beta} + \alpha y_{t-1}^2 + \alpha \beta y_{t-2}^2 + \dots + \alpha \beta^{q-1} y_{t-q}^2 \\ &= \tilde{\alpha}_0 + \tilde{\alpha}_1 y_{t-1}^2 + \tilde{\alpha}_2 z_{t-2}^2 + \dots + \tilde{\alpha}_q z_{t-q}^2 \end{aligned}$$

which means that we can approximate this hypothesis test by instead using the Lagrange multiplier test proposed by Engle:

$$H_0 : \tilde{\alpha}_1 = \tilde{\alpha}_2 = \dots = \tilde{\alpha}_q$$

against

$$H_1 : \tilde{\alpha}_1 \neq 0, \tilde{\alpha}_2 \neq 0, \dots, \tilde{\alpha}_q \neq 0$$

10.7.3 Worked Problems

Problem 1. Consider the generalized autoregressive heteroskedastic model

$$y_t = h_t z_t$$

where

$$z_t | \Omega_{t-1} \sim IID\mathcal{N}(0, 1) \quad (28)$$

$$h_t^2 = \text{Var}(y_t | \Omega_{t-1}) = \mathbb{E}(y_t^2 | \Omega_{t-1}) = \bar{\sigma}^2(1 - \alpha - \beta) + \alpha y_{t-1}^2 + \beta h_{t-1}^2 \quad (29)$$

and Ω_t is the information set that contains at least y_t and its lagged values.

- (a) Derive the conditions under which $\{y_t\}$ is a stationary process.
- (b) Are the observations $\{y_t\}$ serially independent and/or serially uncorrelated?
- (c) Develop a test of the GARCH effect and discuss the estimation of the above model by the maximum likelihood method.
- (d) Discuss the relevance of GARCH models for the analysis of financial time series data.

Solution.

- (a) Note that (using the fact that y_t and h_t are conditionally independent given Ω_{t-1})

$$\mathbb{E}(y_t) = \mathbb{E}(h_t z_t) = \mathbb{E}[\mathbb{E}(h_t z_t | \Omega_{t-1})] = \mathbb{E}[\mathbb{E}(h_t | \Omega_{t-1})\mathbb{E}(z_t | \Omega_{t-1})] = \mathbb{E}[\mathbb{E}(h_t | \Omega_{t-1}) \cdot 0] = 0$$

We have

$$\begin{aligned} \text{Var}(y_t) &= \mathbb{E}[(y_t - \mathbb{E}(y_t))^2] = \mathbb{E}[y_t^2] = \mathbb{E}[\mathbb{E}(y_t^2 | \Omega_{t-1})] = \mathbb{E}(h_t^2) = \mathbb{E}(\bar{\sigma}^2(1 - \alpha - \beta) + \alpha y_{t-1}^2 + \beta h_{t-1}^2) \\ &= \mathbb{E}(\bar{\sigma}^2(1 - \alpha - \beta)) + \alpha \mathbb{E}(y_{t-1}^2) + \beta \mathbb{E}(h_{t-1}^2) = \bar{\sigma}^2(1 - \alpha - \beta) + (\alpha + \beta)\text{Var}(y_{t-1}) \end{aligned}$$

Therefore in order for this to be a stationary process, we require $|\alpha + \beta| < 1$.

more stuff I did on original homework

By Definition 10.1 $\{y_t\}$ is a stationary process if it has constant mean and its covariance function depends only on the absolute difference $|t_1 - t_2|$; that is,

$$\text{Cov}(y_{t_1}, y_{t_2}) = \gamma(t_1, t_2) = \gamma(|t_1 - t_2|) \text{ for all } t_1, t_2$$

⋮

In order for this to be true, we must first show that h_t is finite for all t . It is sufficient to find the conditions that make h_t stationary. Using Equation (26)

$$h_t^2 = \alpha_0 + \sum_{i=1}^q \alpha_i \epsilon_{t-i}^2 + \sum_{i=1}^p \phi_i h_{t-i}^2$$

From equation (29) we have

$$h_t^2 = \bar{\sigma}^2(1 - \alpha - \beta) + \alpha y_{t-1}^2 + \beta h_{t-1}^2$$

Therefore we note that $p = q = 1$ and we have a GARCH(1,1) model. From Equation (24):

$$h_t^2 = \alpha_0 + \alpha_1 \epsilon_{t-1}^2 + \phi_1 h_{t-1}^2, \quad \alpha_0 > 0$$

is unconditionally stationary if $|\alpha_1 + \phi_1| < 1$. In this case, we require $|\alpha + \beta| < 1$ for stationarity of h_t .

⋮

Assume without loss of generality that $t_2 \geq t_1$. (Note that this implies $z_{t_2} \perp z_{t_1}, h_{t_2}, h_{t_1} \mid \Omega_{t_2-1}$.)

$$\begin{aligned} \text{Cov}(y_{t_1}, y_{t_2}) &= \mathbb{E}[(y_{t_1} - \mathbb{E}(y_{t_1}))(y_{t_2} - \mathbb{E}(y_{t_2}))] = \mathbb{E}[y_{t_1} y_{t_2}] = \mathbb{E}[h_{t_1} h_{t_2} z_{t_1} z_{t_2}] \\ &= \mathbb{E}[\mathbb{E}(h_{t_1} h_{t_2} z_{t_1} z_{t_2} \mid \Omega_{t_2-1})] = \mathbb{E}[\mathbb{E}(h_{t_1} h_{t_2} z_{t_1} \mid \Omega_{t_2-1}) \mathbb{E}(z_{t_2} \mid \Omega_{t_2-1})] \\ &= \mathbb{E}[\mathbb{E}(h_{t_1} h_{t_2} z_{t_1} \mid \Omega_{t_2-1}) \cdot 0] = [0] \end{aligned}$$

Therefore given that h_t is finite (that is, given $|\alpha + \beta| < 1$), we have that $\mathbb{E}(y_t) = 0$, $\text{Cov}(y_{t_1}, y_{t_2}) = 0 \quad \forall t_1, t_2$ which implies that under these conditions y_t is stationary.

- (b) y_t is serially uncorrelated because $\text{Cov}(y_{t_1}, y_{t_2}) = 0 \quad \forall t_1, t_2$. However, it is clear that y_t is not serially independent since past values of y_t affect $h_t^2 = \text{Var}(y_t)$. Observe that

$$\Pr(y_t \leq y \mid y_{t-1}) = \Pr(h_t z_t \leq y \mid y_{t-1}) = \Pr\left(z_t \sqrt{\bar{\sigma}^2(1 - \alpha - \beta) + \alpha y_{t-1}^2 + \beta h_{t-1}^2} \leq y \mid y_{t-1}\right)$$

In other words, even though the mean of y_t remains constant, because y_{t-1} affects the variance of y_t , it affects the heaviness of the tails of y_t , changing the probability distribution of y_t . Therefore the conditional cumulative distribution function of y_t given y_{t-1} is not equal to the unconditional cdf, so y_t and y_{t-1} are not independent.

Lidan's Explanation: From (a) we know that $\{y_t\}$ is serially uncorrelated, but they are not necessarily independent because from Equation (29) we know that h_t is not independent from h_{t-1} .

- (c) See section 10.7.2.
- (d) **Lidan:** Usually financial time series data are fat-tailed. So the series may not look stationary, and therefore the local variance would be clustered in some very low and very high values. To capture this serial correlation and heterogeneity in volatility, we need to use a GARCH model.

Book: In financial econometrics, ARCH and GARCH are fundamental tools for analyzing the time-variation of conditional variance. In many applications in finance, the assumption that the conditional variance of the disturbances is constant over time is not valid. GARCH models allow for time variation in volatility, relating (unobserved) volatility to squares of past innovations in price changes. However, this approach only partly overcomes the deficiency of the historical measure and continues to respond very slowly when volatility undergoes rapid changes.

10.8 Chapter 21: Vector Autoregressive Models

Feel ok on concepts.

10.8.1 Worked Problems

Problem 1. Consider the bivariate autoregressive model:

$$Y_t = \Phi Y_{t-1} + U_t, \quad U_t \sim IID\mathcal{N}(0, \Sigma) \quad (30)$$

where

$$Y_t = (y_{1t}, y_{2t})', \quad Y_{t-1} = (y_{1,t-1}, y_{2,t-1})', \quad U_t = (u_{1t}, u_{2t})'$$

and

$$\Phi = \begin{bmatrix} \phi_{11} & \phi_{12} \\ \phi_{21} & \phi_{22} \end{bmatrix}, \quad \Sigma = \begin{bmatrix} \sigma_{11} & \sigma_{12} \\ \sigma_{21} & \sigma_{22} \end{bmatrix} \succ 0$$

- (a) Derive the conditional mean and variance of y_{1t} with respect to y_{2t} and lagged values of y_{1t} and y_{2t} .
- (b) Show that the univariate representation of y_{1t} is an ARMA(2,1) process.

Solution

- (a) Note that

$$y_{1t} = \phi_{11}y_{1,t-1} + \phi_{12}y_{2,t-1} + u_{1t} \quad (31)$$

$$y_{2t} = \phi_{21}y_{1,t-1} + \phi_{22}y_{2,t-1} + u_{2t} \quad (32)$$

First,

$$\begin{aligned} \mathbb{E}(y_{1t} \mid y_{2t}, y_{2,t-1}, y_{2,t-2}, \dots, y_{1,t-1}, y_{1,t-2}, \dots) &= \mathbb{E}(y_{1t} \mid y_{2t}, \Omega_{t-1}) \\ &= \phi_{11}y_{1,t-1} + \phi_{12}y_{2,t-1} + \mathbb{E}(u_{1t} \mid y_{2t}) = \phi_{11}y_{1,t-1} + \phi_{12}y_{2,t-1} + \mathbb{E}(u_{1t} \mid \phi_{21}y_{1,t-1} + \phi_{22}y_{2,t-1} + u_{2t}) \\ &= \phi_{11}y_{1,t-1} + \phi_{12}y_{2,t-1} + \mathbb{E}(u_{1t} \mid u_{2t}) \end{aligned}$$

Recall Proposition 6.51: for a bivariate normal distribution with mean 0, the conditional distribution of u_{1t} given u_{2t} is

$$u_{1t} \mid u_{2t} \sim \mathcal{N}\left(\rho \frac{\sqrt{\sigma_{11}}}{\sqrt{\sigma_{22}}} u_{2t}, (1 - \rho^2)\sigma_{11}\right)$$

where $\rho = \sigma_{12}/\sqrt{\sigma_{11}}\sqrt{\sigma_{22}}$.

$$\begin{aligned} \iff u_{1t} \mid u_{2t} &\sim \mathcal{N}\left(\frac{\sigma_{12}}{\sqrt{\sigma_{11}}\sqrt{\sigma_{22}}} \cdot \frac{\sqrt{\sigma_{11}}}{\sqrt{\sigma_{22}}} u_{2t}, \left[1 - \left(\frac{\sigma_{12}}{\sqrt{\sigma_{11}}\sqrt{\sigma_{22}}}\right)^2\right]\sigma_{11}\right) \\ \iff u_{1t} \mid u_{2t} &\sim \mathcal{N}\left(\frac{\sigma_{12}}{\sigma_{22}}u_{2t}, \sigma_{11} - \frac{\sigma_{12}^2}{\sigma_{22}}\right) \\ \implies \mathbb{E}(u_{1t} \mid u_{2t}) &= \frac{\sigma_{12}}{\sigma_{22}}u_{2t}, \quad \text{Var}(u_{1t} \mid u_{2t}) = \sigma_{11} - \frac{\sigma_{12}^2}{\sigma_{22}} \\ \implies \mathbb{E}(y_{1t} \mid y_{2t}, \Omega_{t-1}) &= \phi_{11}y_{1,t-1} + \phi_{12}y_{2,t-1} + \frac{\sigma_{12}}{\sigma_{22}}u_{2t} \\ &= \phi_{11}y_{1,t-1} + \phi_{12}y_{2,t-1} + \frac{\sigma_{12}}{\sigma_{22}}(y_{2t} - \phi_{21}y_{1,t-1} - \phi_{22}y_{2,t-1}) \\ &= \boxed{\left(\phi_{11} - \frac{\sigma_{12}}{\sigma_{22}}\phi_{21}\right)y_{1,t-1} + \left(\phi_{12} - \frac{\sigma_{12}}{\sigma_{22}}\phi_{22}\right)y_{2,t-1} + \frac{\sigma_{12}}{\sigma_{22}}y_{2t}} \end{aligned}$$

Second,

$$\text{Var}(y_{1t} \mid y_{2t}, \Omega_{t-1}) = \phi_{11}^2 \text{Var}(y_{1,t-1} \mid y_{2t}, \Omega_{t-1}) + \phi_{12}^2 \text{Var}(y_{2,t-1} \mid y_{2t}, \Omega_{t-1}) + \text{Var}(u_{1t} \mid y_{2t}, \Omega_{t-1})$$

$$= \text{Var}(u_{1t} \mid u_{2t}) = \boxed{\sigma_{11} - \frac{\sigma_{12}^2}{\sigma_{22}}}$$

(b) Again, using equations (31) and (32),

$$\begin{aligned} y_{1t} &= \phi_{11}y_{1,t-1} + \phi_{12}y_{2,t-1} + u_{1t}, \quad y_{2t} = \phi_{21}y_{1,t-1} + \phi_{22}y_{2,t-1} + u_{2t} \\ \implies (1 - \phi_{11}L)y_{1t} &= \phi_{12}Ly_{2t} + u_{1t} \end{aligned} \tag{33}$$

$$(1 - \phi_{22}L)y_{2t} = \phi_{21}Ly_{1t} + u_{2t} \tag{34}$$

We can multiply $(1 - \phi_{22}L)$ on both sides of (33) to yield

$$(1 - \phi_{22}L)(1 - \phi_{11}L)y_{1t} = \phi_{12}L(1 - \phi_{22}L)y_{2t} + (1 - \phi_{22}L)u_{1t}$$

Using (34), we have

$$\begin{aligned} (1 - (\phi_{22} + \phi_{11})L + \phi_{11}\phi_{22}L^2)y_{1t} &= \phi_{12}L[\phi_{21}Ly_{1t} + u_{2t}] + (1 - \phi_{22}L)u_{1t} \\ \iff (1 - (\phi_{22} + \phi_{11})L + (\phi_{11}\phi_{22} - \phi_{12}\phi_{21})L^2)y_{1t} &= \phi_{12}Lu_{2t} + (1 - \phi_{22}L)u_{1t} \end{aligned} \tag{35}$$

If we can show that the left side of (35) is an AR(2) process and the right side is an MA(1) process, we are done. **Lidan:** “It is obvious that the left hand side of (35) is an AR(2) process.” The left side is a stationary AR(2) process if the absolute values of all the roots of

$$1 - (\phi_{22} + \phi_{11})z + (\phi_{11}\phi_{22} - \phi_{12}\phi_{21})z^2 = 0$$

are greater than 1. The roots are

$$\begin{aligned} z &= \frac{\phi_{22} + \phi_{11} \pm \sqrt{(\phi_{22} + \phi_{11})^2 - 4(\phi_{11}\phi_{22} - \phi_{12}\phi_{21})}}{2(\phi_{11}\phi_{22} - \phi_{12}\phi_{21})} = \frac{\phi_{22} + \phi_{11} \pm \sqrt{\phi_{22}^2 + \phi_{11}^2 - 2\phi_{22}\phi_{11} + 4\phi_{12}\phi_{21}}}{2(\phi_{11}\phi_{22} - \phi_{12}\phi_{21})} \\ &= \frac{\phi_{22} + \phi_{11} \pm \sqrt{(\phi_{22} - \phi_{11})^2 + 4\phi_{12}\phi_{21}}}{2(\phi_{11}\phi_{22} - \phi_{12}\phi_{21})} \\ &\vdots \end{aligned}$$

To check if the right side of (35) is MA(1), we will write $x_t = \phi_{12}Lu_{2t} + (1 - \phi_{22}L)u_{1t}$ as an MA(1) process; that is,

$$x_t = \phi_{12}u_{2,t-1} + u_{1t} - \phi_{22}u_{1,t-1} = \xi_t + \theta\xi_{t-1}$$

with $\xi_t \sim iid(0, \sigma_\xi^2)$, $|\theta| < 1$. If x_t is an MA(1) process, it must satisfy

$$\gamma(0) = \mathbb{E}(x_t^2) = (1 + \theta^2)\sigma_\xi^2, \quad \gamma(1) = \mathbb{E}(x_t x_{t-1}) = \theta\sigma_\xi^2$$

We have (using the serial independence of the u_{1t}, u_{2t})

$$\begin{aligned}
\mathbb{E}(x_t^2) &= \mathbb{E}(\phi_{12}u_{2,t-1} + u_{1t} - \phi_{22}u_{1,t-1})^2 = \mathbb{E}(\phi_{12}^2u_{2,t-1}^2 - 2\phi_{12}\phi_{22}u_{2,t-1}u_{1,t-1} + \phi_{22}^2u_{1,t-1}^2 + u_{1t}^2) \\
&= \phi_{12}^2\mathbb{E}(u_{2,t-1}^2) - 2\phi_{12}\phi_{22}\mathbb{E}(u_{2,t-1}u_{1,t-1}) + \phi_{22}^2\mathbb{E}(u_{1,t-1}^2) + \mathbb{E}(u_{1t}^2) \\
&= \phi_{12}^2\sigma_{22} - 2\phi_{12}\phi_{22}\sigma_{12} + \phi_{22}^2\sigma_{11} + \sigma_{11} = (1 + \phi_{22}^2)\sigma_{11} - 2\phi_{12}\phi_{22}\sigma_{12} + \phi_{12}^2\sigma_{22} \\
\mathbb{E}(x_tx_{t-1}) &= \mathbb{E}([\phi_{12}u_{2,t-1} + u_{1t} - \phi_{22}u_{1,t-1}][\phi_{12}u_{2,t-2} + u_{1,t-1} - \phi_{22}u_{1,t-2}]) \\
&= \mathbb{E}(\phi_{12}u_{2,t-1}u_{1,t-1} - \phi_{22}u_{1,t-1}^2) = \phi_{12}\sigma_{12} - \phi_{22}\sigma_{11}
\end{aligned}$$

Therefore we require

$$(1 + \theta^2)\sigma_\xi^2 = (1 + \phi_{22}^2)\sigma_{11} - 2\phi_{12}\phi_{22}\sigma_{12} + \phi_{12}^2\sigma_{22} \quad (36)$$

$$\theta\sigma_\xi^2 = \phi_{12}\sigma_{12} - \phi_{22}\sigma_{11} \quad (37)$$

Dividing (36) by (37) we have

$$\frac{1 + \theta^2}{\theta} = \frac{(1 + \phi_{22}^2)\sigma_{11} - 2\phi_{12}\phi_{22}\sigma_{12} + \phi_{12}^2\sigma_{22}}{\phi_{12}\sigma_{12} - \phi_{22}\sigma_{11}} \quad (38)$$

For simplicity, let the right side of (38) be $A \in \mathbb{R}$. Then we have

$$\theta^2 - A\theta + 1 = 0 \iff \boxed{\theta_1 = \frac{1}{2}(A + \sqrt{A^2 - 4}), \theta_2 = \frac{1}{2}(A - \sqrt{A^2 - 4})}$$

Then the corresponding σ_ξ^2 are

$$\boxed{\sigma_{\xi,1}^2 = \sigma_{\xi,2}^2 = \frac{\phi_{12}\sigma_{12} - \phi_{22}\sigma_{11}}{\theta_1}}$$

As a double check,

$$\mathbb{E}(x_tx_{t-2}) = \mathbb{E}([\phi_{12}u_{2,t-1} + u_{1t} - \phi_{22}u_{1,t-1}][\phi_{12}u_{2,t-3} + u_{1,t-2} - \phi_{22}u_{1,t-3}]) = 0$$

as expected for an MA(1) process. Therefore the right side of (35) is an MA(1) process (provided that $0 < |\theta_1| < 1$ and/or $0 < |\theta_2| < 1$). Since the left side is an AR(2) process, this proves that the univariate representation (35) of y_{1t} is an ARMA(2,1) process.

Problem 2. Consider the VAR(2) model in the m -dimensional vector Y_t :

$$Y_t = \mu + \Phi_1 Y_{t-1} + \Phi_2 Y_{t-2} + U_t, \quad U_t \sim (0, \Sigma) \quad (39)$$

where μ is an m -vector of fixed constants.

- (a) Derive the conditions under which the VAR(2) model defined in (39) is stationary.
- (b) Derive the error correction form of (39) and discuss what is meant by the process Y_t being cointegrated.
- (c) Suppose now that one or more elements of Y_t is I(1). Derive suitable restrictions on the intercepts μ such that despite the I(1) nature of the variables in (39), Y_t has a fixed mean. Discuss the importance of such restrictions for the analysis of cointegration.

Solution.

- (a) Let

$$Y_t^* = Y_t - (I - \Phi_1 - \Phi_2)^{-1}\mu. \quad (40)$$

Then

$$\begin{aligned} Y_t^* &= \mu + \Phi_1 Y_{t-1} + \Phi_2 Y_{t-2} + U_t - (I - \Phi_1 - \Phi_2)^{-1}\mu \\ &= \mu + \Phi_1 Y_{t-1} + \Phi_2 Y_{t-2} + U_t - [(I - \Phi_1 - \Phi_2) + \Phi_1 + \Phi_2](I - \Phi_1 - \Phi_2)^{-1}\mu \\ &= \mu - (I - \Phi_1 - \Phi_2)(I - \Phi_1 - \Phi_2)^{-1}\mu + \Phi_1(Y_{t-1} - (I - \Phi_1 - \Phi_2)^{-1}\mu) + \Phi_2(Y_{t-2} - (I - \Phi_1 - \Phi_2)^{-1}\mu) + U_t \\ &= \Phi_1 Y_{t-1}^* + \Phi_2 Y_{t-2}^* + U_t \end{aligned} \quad (41)$$

Equation (41) can be rewritten in the companion form as follows:

$$\begin{bmatrix} Y_t^* \\ Y_{t-1}^* \end{bmatrix} = \begin{bmatrix} \Phi_1 & \Phi_2 \\ I & 0 \end{bmatrix} \begin{bmatrix} Y_{t-1}^* \\ Y_{t-2}^* \end{bmatrix} + \begin{bmatrix} U_t \\ 0 \end{bmatrix} \quad (42)$$

which is a VAR(1) model. If Y_{-M+1}, Y_{-M+2} are given, equation (42) can be solved iteratively from $t = -M + 2$ to obtain

$$\begin{bmatrix} Y_t^* \\ Y_{t-1}^* \end{bmatrix} = \begin{bmatrix} \Phi_1 & \Phi_2 \\ I & 0 \end{bmatrix}^{t+M-2} \begin{bmatrix} Y_{-M+2}^* \\ Y_{-M+1}^* \end{bmatrix} + \sum_{j=0}^{t+M-3} \begin{bmatrix} \Phi_1 & \Phi_2 \\ I & 0 \end{bmatrix}^j \begin{bmatrix} U_{t-j} \\ 0 \end{bmatrix} \quad (43)$$

Then the condition for (43) to be covariance stationary is for all the eigenvalues of

$$\Phi = \begin{bmatrix} \Phi_1 & \Phi_2 \\ I & 0 \end{bmatrix}$$

to lie inside the unit circle; that is, the solutions of

$$|\Phi - \lambda I_4| = 0$$

must satisfy $|\lambda| < 1$. Equivalently, the stability condition can be written in terms of the roots of the determinantal equation

$$|I_2 - \Phi_1 z - \Phi_2 z^2| = 0$$

in which case the process Y_t^* will be stationary if all the roots lie outside the unit circle ($|z| > 1$). Then if Y_t^* is stationary, so is Y_t , so either of these conditions (plus the invertibility of $(I - \Phi_1 - \Phi_2)$) ensure stationarity of Y_t .

(b) From (41) we have

$$Y_t^* = \Phi_1 Y_{t-1}^* + \Phi_2 Y_{t-2}^* + U_t$$

Note that (letting $\Delta Y_t^* = Y_t^* - Y_{t-1}^*$)

$$Y_t^* - Y_{t-1}^* + Y_{t-1}^* = \Phi_1 Y_{t-1}^* + \Phi_2 (Y_{t-2}^* - Y_{t-1}^* + Y_{t-1}^*) + U_t$$

$$\iff \Delta Y_t^* + Y_{t-1}^* = \Phi_1 Y_{t-1}^* + \Phi_2 (Y_{t-1}^* - \Delta Y_{t-1}^*) + U_t$$

$$\iff \Delta Y_t^* = -(I - \Phi_1 - \Phi_2) Y_{t-1}^* - \Phi_2 \Delta Y_{t-1}^* + U_t \iff \boxed{\Delta Y_t^* = -\Pi Y_{t-1}^* + \Gamma \Delta Y_{t-1}^* + U_t}$$

where $\Pi = (I - \Phi_1 - \Phi_2)$ and $\Gamma = -\sum_{i=2}^2 \Phi_i = -\Phi_2$. Because

$$\Delta Y_t^* = Y_t^* - Y_{t-1}^* = Y_t - (I - \Phi_1 - \Phi_2)^{-1} \mu - [Y_{t-1} - (I - \Phi_1 - \Phi_2)^{-1} \mu] = Y_t - Y_{t-1} = \Delta Y_t,$$

this can also be written as

$$\begin{aligned} \Delta Y_t &= -\Pi [Y_{t-1} - (I - \Phi_1 - \Phi_2)^{-1} \mu] + \Gamma \Delta Y_{t-1} + U_t \\ &= -\Pi [Y_{t-1} - \Pi^{-1} \mu] + \Gamma \Delta Y_{t-1} + U_t = \Pi \cdot \Pi^{-1} \mu - \Pi Y_{t-1} + \Gamma \Delta Y_{t-1} + U_t \\ &\implies \boxed{\Delta Y_t = \mu - \Pi Y_{t-1} + \Gamma \Delta Y_{t-1} + U_t} \end{aligned} \tag{44}$$

For the definition of cointegration, see Definition 10.3 and Section 10.9.1. In this particular case, if $Y_{t-1}^* \sim I(1)$ and the linear combinations ΠY_{t-1}^* of Y_{t-1}^* are covariance stationary (that is, $\Pi Y_{t-1}^* \sim I(0)$), we say Y_t^* is cointegrated (and therefore so is Y_t).

(c) Since (44) is $I(0)$, for Y_t to have fixed mean, take expectations on both sides of (44):

$$\mu - (I_m - \Phi_1 - \Phi_2) \mathbb{E}(Y_{t-1}) = 0$$

If this restriction is violated, then (44) becomes a stationary process with a drift, which implies (39) has a time trend.

10.9 Chapter 22: Cointegration Analysis

Feel pretty good except for long run effects, examples we went over in class, the restrictions, and the 5 cases. Don't need to understand SURE.

10.9.1 22.4 Cointegrating VAR: multiple cointegrating relations and 22.5: Identification of long-run effects

Definition 10.3. We say that the m variables in Y_t are *cointegrated* if they are individually integrated (or have a random walk component) but there exist linear combinations of them which are stationary. That is, $y_{it} \sim I(1)$ for $i = 1, 2, \dots, m$, but there exists an $m \times r$ matrix β such that $\beta' Y_t = \xi_t \sim I(0)$.

- In this case r denotes the number of cointegrating vectors, also known as the dimension of the cointegration space.
- The cointegrating relations summarized in the $r \times 1$ vector $\beta' Y_t$ are also known as long-run relations.
- $r = \text{rank}(\Pi)$ is the dimension of the cointegration space.
- Cointegration is present if Π is rank-deficient; that is, $r < m$.

When $\text{rank}(\Pi) = r < m$, we can write Π as

$$\Pi = \alpha\beta' \quad (45)$$

where α and β are $m \times r$ matrices of full column rank. Then

$$\Pi y_{t-1} = \alpha\beta' y_{t-1} \sim I(0)$$

and the VECM can be written as

$$\Delta y_t = -\alpha\beta' y_{t-1} + \sum_{j=1}^{p-1} \Gamma_j \Delta y_{t-j} + u_t \quad (46)$$

Since α is full rank, we have

$$\beta' y_{t-1} \sim I(0)$$

where $\beta' y_t$ is the r -vector of cointegrating relations, also known as the long-run relations.

However, β as defined above is not uniquely determined. Consider a linear transformation of β by a non-singular $r \times r$ matrix Q : $\tilde{\beta} = \beta Q$. Then since $\text{rank}(\Pi) = r < m$, Π can be expressed as $\Pi = \alpha\beta'$ where α is a rank r $m \times r$ matrix. But consider that

$$\Pi = \alpha\beta = (\alpha Q'^{-1})(Q'\beta') = \tilde{\alpha}\tilde{\beta}'$$

with $\tilde{\alpha} = \alpha Q'^{-1}$. Therefore β is not uniquely determined without r^2 exact- or just-identifying restrictions, r restrictions on each of the r cointegrating relations.

10.9.2 Worked Problems

Problem 2: See Chapter 21 Problem 2.

10.10 Chapter 23: VARX Modeling

Lidan says will not be on final.

10.10.1 Worked Problems

10.11 Chapter 24: Impulse Response Analysis

Feel ok on concepts, should review and do an example problem.

10.11.1 Worked Problems

Chapter 24 Problem 1. Consider the VAR(2) model

$$x_t = \Phi_1 x_{t-1} + \Phi_2 x_{t-2} + \epsilon_t, \quad \epsilon_t \sim IID(0, \Sigma)$$

in the $m \times 1$ vector of random variables x_t and Σ is the covariance matrix of the errors with typical element σ_{ij} .

- (a) Derive the conditions under which this process is stationary, and show that it has the following moving average representation:

$$x_t = \sum_{j=0}^{\infty} A_j \epsilon_{t-j} \tag{47}$$

- (b) Derive the coefficient matrices A_j in terms of Φ_1 and Φ_2 .
(c) Using the above result, write down the orthogonalized (OIR) and generalized impulse (GIR) response functions of one standard error shock (i.e. $\sqrt{\sigma_{ii}}$ to the error of the i th equation, $\epsilon_{it} = s_i' \epsilon_t$, where s_i is an $m \times 1$ selection vector).
(d) What are the main differences between OIR and GIR functions?

Chapter 24 Problem 1 Solution.

- (a) For stationarity conditions, see Ch. 21 Problem 2 in Section 10.8.1. To get the MA representation

$$x_t = \sum_{j=0}^{\infty} A_j \epsilon_{t-j} \tag{48}$$

of (47) note that

$$(I - \Phi_1 L - \Phi_2 L^2)x_t = \epsilon_t$$

$$\implies x_t = (I - \Phi_1 L - \Phi_2 L^2)^{-1}\epsilon_t = \sum_{j=0}^{\infty} A_j \epsilon_{t-j}$$

for some $\{A_j\}$.

- (b) Now we seek to evaluate $(I - \Phi_1 L - \Phi_2 L^2)^{-1}$. To do this, let $(I - \Phi_1 L - \Phi_2 L^2)^{-1} = A_0 + A_1 L + A_2 L^2 + A_3 L^3 + \dots$ and note that

$$\begin{aligned} & (I - \Phi_1 L - \Phi_2 L^2)(I - \Phi_1 L - \Phi_2 L)^{-1} = I \\ \iff & (I - \Phi_1 L - \Phi_2 L^2)(A_0 + A_1 L + A_2 L^2 + A_3 L^3 + \dots + A_j L^j + \dots) = I \\ \iff & A_0 + (A_1 - \Phi_1)L + (A_2 - \Phi_1 A_1 - \Phi_2)L^2 + (A_3 - \Phi_1 A_2 - \Phi_2 A_1)L^3 + \dots \\ & + (A_j - \Phi_1 A_{j-1} + \Phi_2 A_{j-2})L^j + \dots = I \end{aligned}$$

In order for this equation to hold, all the lag terms must equal zero and the constant matrix A_0 must equal I .

$$\implies [A_0 = I]$$

$$A_1 - \Phi_1 = 0 \iff [A_1 = \Phi_1]$$

$$A_2 - \Phi_1 A_1 - \Phi_2 = 0 \iff A_2 - \Phi_1^2 - \Phi_2 = 0 \iff [A_2 = \Phi_1^2 + \Phi_2]$$

$$A_3 - \Phi_1 A_2 - \Phi_2 A_1 = 0 \iff A_3 - \Phi_1^3 - \Phi_1 \Phi_2 - \Phi_2 \Phi_1 = 0 \iff [A_3 = \Phi_1^3 + \Phi_1 \Phi_2 + \Phi_2 \Phi_1]$$

and, in general,

$$[A_j = \Phi_1 A_{j-1} + \Phi_2 A_{j-2}]$$

- (c) • **OIR:** We employ the Cholesky decomposition of Σ :

$$\Sigma = PP' \tag{49}$$

where P is a lower-triangular matrix. Then the MA representation (48) can be written as

$$x_t = \sum_{j=0}^{\infty} (A_j P)(P^{-1} u_{t-j}) = \sum_{j=0}^{\infty} B_j \eta_{t-j} \tag{50}$$

where $B_j = A_j P$, $\eta_t = P^{-1} u_t$, so we have

$$\mathbb{E}(\eta_t \eta'_t) = P^{-1} \mathbb{E}(u_t u'_t)(P^{-1})' = P^{-1} \Sigma (P^{-1})' = P^{-1} P P' (P^{-1})' = I_m$$

so the new errors $\eta_{1t}, \eta_{2t}, \dots, \eta_{mt}$ are contemporaneously uncorrelated. Then the orthogonalized impact of a unit shock at time t to the i th equation on y at time $t + n$ is given by

$$B_n e_i, n = 0, 1, \dots \quad (51)$$

where e_i is an $m \times 1$ selection vector. Written more compactly, the orthogonalized impulse response function of a unit (one standard error) shock to the i th variable on the j th variable is given by

$$OI_{ij,n} = e'_j A_n P e_i, \quad i, j = 1, 2, \dots, m \quad (52)$$

These orthogonalized impulse responses are not unique and depend on the particular ordering of the variables in the VAR. The orthogonalized responses are invariant to the ordering of the variables only if Σ is diagonal.

- **GIR:** If the VAR model is perturbed by a shock of size $\delta_i = \sqrt{\sigma_{ii}}$ to its i th equation at time t , by the definition of the generalized IR function we have

$$GI_y(n, \delta_i, \Omega_{t-1}^0) = \mathbb{E}(y_t | u_{it} = \delta_i, \Omega_{t-1}^0) - \mathbb{E}(y_t | \Omega_{t-1}^0) \quad (53)$$

Once again using the MA(∞) representation (48) we obtain

$$GI_y(n, \delta_i, \Omega_{t-1}^0) = A_n \mathbb{E}(u_t | u_{it} = \delta_i) \quad (54)$$

which is history invariant (i.e. does not depend on Ω_{t-1}^0). The computation of the conditional expectations $\mathbb{E}(u_t | u_{it} = \delta_i)$ depends on the nature of the multivariate distribution assumed for the disturbances u_t . In the case where $u_t \sim IID\mathcal{N}(0, \Sigma)$, we have

$$\mathbb{E}(u_t | u_{it} = \delta_i) = \begin{bmatrix} \sigma_{1i}/\sigma_{ii} \\ \sigma_{2i}/\sigma_{ii} \\ \vdots \\ \sigma_{mi}/\sigma_{ii} \end{bmatrix} \delta_i \quad (55)$$

where as before $\Sigma = [\sigma_{ij}]$. Hence for a unit shock $\delta_i = \sqrt{\sigma_{ii}}$ we have

$$GI_y(n, \delta_i = \sqrt{\sigma_{ii}}, \Omega_{t-1}^0) = \frac{A_n \Sigma e_i}{\sqrt{\sigma_{ii}}}, \quad i, j = 1, 2, \dots, m \quad (56)$$

The GIRF of a unit shock to the i th equation in the VAR(p) model

$$y_t = \Phi_1 y_{t-1} + \Phi_2 y_{t-2} + \dots + \Phi_p y_{t-p} + u_t, \quad u_t \sim IID(0, \Sigma) \quad (57)$$

on the j th variable at horizon n is given by the j th element of (56), expressed more compactly by

$$GI_y(n, \delta_i = \sqrt{\sigma_{ii}}, \Omega_{t-1}^0) = \frac{e'_j A_n \Sigma e_i}{\sqrt{\sigma_{ii}}}, \quad i, j = 1, 2, \dots, m \quad (58)$$

- (d) The GIRF circumvents the problem of the dependence of the orthogonalized impulse responses to the ordering of the variables in the VAR. Unlike the OIR responses in (52), the GIR responses in (58) are invariant to the ordering of the variables in the VAR. The two responses coincide only for the first variable in the VAR or when Σ is diagonal.

Chapter 24 Problem 4.

Chapter 24 Problem 4 Solution.

10.12 Chapter 33: Theory and Practice of GVAR Modeling

Lidan says will not be on final.

11 Convex Optimization

These are my notes from taking EE 588 at USC and the textbook *Convex Optimization* (Boyd and Vandenberghe) 7th printing.

Need to cover:

- Update rules for optimization problems (e.g. gradient descent, be able to write down gradient, etc.)
- Know which algorithms are useful in which settings
- Homework-like problems from first part of class (no proofs though) (Boyd homework is good practice)
- Understand how to derive algorithms
- Understand how to calculate gradients, proximal functions, etc.
- Understand examples, how to run algorithms
- Only conceptual thing: duality question (write down dual)
- Formulate problems as convex optimization problems

Do not need to cover:

- ADMM
- Proofs from 2nd half of class (rates of convergence, etc.)
- Coding

11.1 Convex Functions

Theorem 11.1. Jensen's Inequality: f is convex if and only if

$$\frac{f(a) + f(b)}{2} \geq f\left(\frac{a+b}{2}\right)$$

for all $a, b \in \text{dom}(f)$.

11.2 Schur Complement Trick

11.2.1 Definition

For a matrix $X \in \mathbf{S}^n$ partitioned as

$$X = \begin{bmatrix} A & B \\ B^T & C \end{bmatrix}$$

the Schur complement is (if $\det(A) \neq 0$)

$$S = C - B^T A^{-1} B$$

The Schur complement has two useful properties in convex analysis.

Theorem 11.2. (a) $X \succ 0$ if and only if $A \succ 0$ and $S \succ 0$.

(b) If $A \succ 0$, then $X \succeq 0$ if and only if $S \succeq 0$.

11.2.2 The Trick

Suppose we are trying to express a problem as a semidefinite program (SDP); that is, in the form

$$\begin{aligned} & \text{minimize} && c^T x \\ & \text{subject to} && x_1 F_1 + \dots + x_n F_n + G \preceq 0 \\ & && Ax = b \end{aligned}$$

where $G, F_1, \dots, F_n \in \mathbf{S}^k$ and $A \in \mathbb{R}^{p \times n}$. If we have a constraint of the form $c^T F(x)^{-1} c \leq t$ where $F(x)$ is symmetric and positive definite and $t \in \mathbb{R}$, by Theorem 11.2(b) we can write

$$c^T F(x)^{-1} c \leq t \iff \begin{bmatrix} F(x) & c \\ c^T & t \end{bmatrix} \succeq 0$$

in order to get our constraint in the form required for an SDP.

11.2.3 Example 1: Last Year's Final, Question 2(b)

Suppose we have the constraints

$$\begin{aligned} & Ax + b \geq 0 \\ & \frac{(c^T x)^2}{d^T x} \leq t \end{aligned}$$

which we would like to express in an SDP. By Theorem 11.2(b) we can write

$$\frac{(c^T x)^2}{d^T x} \leq t \iff d^T x - (c^T x)^T t^{-1} c^T x \geq 0 \iff \begin{bmatrix} t & c^T x \\ c^T x & d^T x \end{bmatrix} \succeq 0$$

Since

$$Ax + b \geq 0 \iff \text{diag}(Ax + b) \succeq 0$$

we can finally write our constraints as

$$\begin{bmatrix} \text{diag}(Ax + b) & 0 & 0 \\ 0 & t & c^T x \\ 0 & c^T x & d^T x \end{bmatrix} \succeq 0$$

11.2.4 Example 2: Last Year's Final, Question 4(b)

Suppose we have the constraints

$$\begin{aligned} Ax + b &\geq 0 \\ \frac{(c^T x)^2}{d^T x} &\leq t \end{aligned}$$

which we would like to express in an SDP. By Theorem 11.2(b) we can write

$$\frac{(c^T x)^2}{d^T x} \leq t \iff d^T x - (c^T x)^T t^{-1} c^T x \geq 0 \iff \begin{bmatrix} t & c^T x \\ c^T x & d^T x \end{bmatrix} \succeq 0$$

Since

$$Ax + b \geq 0 \iff \text{diag}(Ax + b) \succeq 0$$

we can finally write our constraints as

$$\begin{bmatrix} \text{diag}(Ax + b) & 0 & 0 \\ 0 & t & c^T x \\ 0 & c^T x & d^T x \end{bmatrix} \succeq 0$$

11.3 Duality

Theorem 11.3. Slater's condition/constraint qualification: Strong duality holds for a convex problem

$$\begin{aligned} &\text{minimize} && f_0(x) \\ &\text{subject to} && f_i(x) \leq 0, i = 1, \dots, m \\ & && Ax = b \end{aligned}$$

if it is strictly feasible, i.e., there exists at least one x in the domain of f_0 such that $f_i(x) < 0$, $i = 1, 2, \dots, m$, $Ax = b$.

11.4 MLE estimates

For linear estimates with iid noise

$$y_i = a_i^T x + v_i, i = 1, \dots, m$$

where a is observed and $x \in \mathbb{R}^n$ are the parameters to be estimated, the likelihood function is

$$p_x(y) = \prod_{i=1}^m \Pr(v_i = y_i - a_i^T x \mid x)$$

Therefore the log likelihood function is:

$$\ell_x(y) = \sum_{i=1}^m \log[\Pr(v_i = y_i - a_i^T x \mid x)]$$

11.5 Practice Final (2017 Final)

- (1) (a) Strictly convex. Multiply by x/x (allowed in this case since $x > 0$) to get $\frac{x^2}{x+1}$ which is a quadratic over linear, which is convex in \mathbb{R}^{++} according to CVX rules.
- (b) Not convex, it is convex for $x \geq -1$, but there is a boundary problem at $x = -1$. Note that Jensen's inequality (Theorem 11.1)

$$\frac{f(a) + f(b)}{2} \geq f\left(\frac{a+b}{2}\right)$$

is violated because

$$\frac{f(-1.3) + f(-0.9)}{2} = \frac{2.3 + 0}{2} = 1.15 \leq 2.2 = f(-1.1) = f\left(\frac{-1.3 + -0.9}{2}\right)$$

(c)

(d)

$$f(x) = \sup \log \left(\frac{p(t)}{q(t)} \right) = \sup \{ \log p(t) - \log q(t) \} = \sup \{ \log \left(\sum_{i=1}^n \exp(x_i \sin(it)) \right) - \sum_{i=1}^n x_i \sin(it) \}$$

(e) The proximal mapping is

$$\begin{aligned} \text{prox}_{\mathcal{R}}(z) &= \arg \min_y \frac{1}{2} \|z - y\|_2^2 + \mathcal{R}(y) = \arg \min_y \frac{1}{2} \sum_{i=1}^n (z_i - y_i)^2 + \sum_{i=1}^n w_i |y_i| \\ &= \arg \min_y \frac{1}{2} \sum_{i=1}^n [(z_i - y_i)^2 + w_i |y_i|] \end{aligned}$$

Taking the gradient of the inside quantity with respect to y , we have

$$\nabla(y) = \begin{pmatrix} \frac{1}{2} \cdot 2(z_1 - y_1) + \text{sign}(y_1)w_1 \\ \frac{1}{2} \cdot 2(z_2 - y_2) + \text{sign}(y_2)w_2 \\ \vdots \\ \frac{1}{2} \cdot 2(z_n - y_n) + \text{sign}(y_n)w_n \end{pmatrix} = \begin{pmatrix} z_1 - y_1 + \text{sign}(y_1)w_1 \\ z_2 - y_2 + \text{sign}(y_2)w_2 \\ \vdots \\ z_n - y_n + \text{sign}(y_n)w_n \end{pmatrix}$$

Setting equal to 0, we have

$$y = \begin{pmatrix} z_1 \pm w_1 \\ z_2 \pm w_2 \\ \vdots \\ z_n \pm w_n \end{pmatrix}$$

(2) (a) The constraint is convex (affine). The denominator is affine. Since $c^T x = x^T c$, the numerator

$$(c^T x)^2 = (c^T x)(c^T x) = x^T c c^T x = x^T (cc^T)x$$

is convex since cc^T is positive semidefinite.

(b) We start by using the epigraph trick to transform the problem:

$$\begin{aligned} & \text{minimize} && t \\ & \text{subject to} && \frac{(c^T x)^2}{d^T x} \leq t \\ & && Ax + b \geq 0 \end{aligned}$$

We are trying to express this problem as a semidefinite program (SDP); that is, in the form

$$\begin{aligned} & \text{minimize} && c^T x \\ & \text{subject to} && x_1 F_1 + \dots + x_n F_n + G \preceq 0 \\ & && Ax = b \end{aligned}$$

where $G, F_1, \dots, F_n \in \mathbf{S}^k$ and $A \in \mathbb{R}^{p \times n}$. The first constraint

$$\frac{(c^T x)^2}{d^T x} \leq t$$

can be expressed in the form

$$(c^T x)^2 \leq t d^T x \iff (c^T x c^T - t d^T)x \leq 0$$

We have a constraint

$$Ax + b \geq 0$$

which can be expressed in the form

$$Ax \geq -b$$

$$c^T F(x)^{-1} c \leq t$$

where $F(x)$ is symmetric and positive definite and $t \in \mathbb{R}$, by Theorem 11.2(b) we can write

$$c^T F(x)^{-1} c \leq t \iff \begin{bmatrix} F(x) & c \\ c^T & t \end{bmatrix} \succeq 0$$

in order to get our constraint in the form required for an SDP.

- (3) (a) Yes, g is convex over \mathcal{X} since it is quadratic over linear.
- (b) The only points satisfying the constraint have $x_1 = 0$. Therefore the primal optimal value (the only feasible value) is $e^0 = \boxed{1}$.
- (c) Lagrangian:

$$L(x, \lambda) = e^{-x_1} + \lambda(x_1^2/x_2)$$

The Lagrangian obtains its minimum value of 0 when $x_2 = x_1^3$ and $x_1 \rightarrow \infty$. Thus, its dual function ($g(\lambda) = \min_x L(x, \lambda)$) is

$$g(\lambda) = 0$$

The dual problem is then

| | |
|------------|------------------|
| maximize | 0 |
| subject to | $\lambda \geq 0$ |

- (d) The optimal value of the dual problem is 0. Strong duality does not hold since the optimum of the dual problem is less than the optimum of the primal problem. We can also tell this because Slater's Condition (Theorem 11.3) is violated; that is, there is no (x_1, x_2) that is strictly feasible since x_1 must equal 0, which is on the boundary of the feasible region.
- (e) Now for the primal problem, instead of $x_1 = 0$, we have

$$\frac{x_1^2}{x_2} \leq u \iff x_1^2 \leq ux_2 \implies -\sqrt{ux_2} \leq x_1 \leq \sqrt{ux_2}$$

Since e^{-x_1} is minimized as $x_1 \rightarrow \infty$, our optimal solution is $x_2 \rightarrow \infty, x_1 = \sqrt{ux_2} \rightarrow \infty$ yielding a primal optimal value of $\boxed{0}$. For the dual problem, we have

$$L(x, \lambda) = e^{-x_1} + \lambda \left(\frac{x_1^2}{x_2} - u \right)$$

Dual function ($g(\lambda) = \min_x L(x, \lambda)$):

$$\frac{x_1^2}{x_2} - u = 0 \implies x_2 = \frac{x_1^2}{u}$$

and let $x_1 \rightarrow -\infty$ to yield

$$g(\lambda) = 0$$

The dual problem is then

| | |
|----------|---|
| maximize | 0 |
|----------|---|

with optimal value 0, so there is no longer a duality gap. We can also tell this because Slater's Condition (Theorem 11.3) is satisfied; that is, there exists an (x_1, x_2) which is strictly feasible (say $(x_1, x_2) = (\sqrt{u}, 10)$).

(4) (a) Yes, the set is convex. If $(u_i, v_i) = \mathbf{u}_i$, each

$$\sqrt{(x - u_i)^2 + (y - v_i)^2} = \|\mathbf{x} - \mathbf{u}_i\|_2$$

is convex in \mathbf{x} . Therefore the function

$$\sum_{i=1}^k \|\mathbf{x} - \mathbf{u}_i\|_2$$

is convex. For any fixed d , this set is a sublevel set of this function, which is convex since the function is convex.

(b) This is a feasibility problem:

$$\begin{aligned} & \text{find} && \mathbf{x} \\ & \text{subject to} && \sum_{i=1}^k \|\mathbf{x} - \mathbf{u}_i\| \leq d \\ & && \sum_{i=1}^j \|\mathbf{x} - \mathbf{v}_i\| \leq e \end{aligned}$$

or

$$\begin{aligned} & \text{minimize} && 0 \\ & \text{subject to} && \sum_{i=1}^k \|\mathbf{x} - \mathbf{u}_i\| \leq d \\ & && \sum_{i=1}^j \|\mathbf{x} - \mathbf{v}_i\| \leq e \end{aligned}$$

for two sets of points in \mathbb{R}^2 $\mathbf{u}_1, \dots, \mathbf{u}_k, \mathbf{v}_1, \dots, \mathbf{v}_j$. We would like to express these constraints as matrix inequalities in order to have an SDP. To do this, first rewrite the problem as

$$\begin{aligned} & \text{minimize} && 0 \\ & \text{subject to} && \|\mathbf{x} - \mathbf{u}_i\| \leq t_i, i = 1, \dots, k \\ & && \|\mathbf{x} - \mathbf{v}_i\| \leq s_i, s = 1, \dots, j \\ & && \mathbf{1}^T \mathbf{t} \leq d \\ & && \mathbf{1}^T \mathbf{s} \leq e \end{aligned}$$

Then note that we can use the Schur trick:

$$(\mathbf{x} - \mathbf{u}_i)^T I (\mathbf{x} - \mathbf{u}_i) \leq t_i \iff \begin{bmatrix} I & \mathbf{x} - \mathbf{u}_i \\ (\mathbf{x} - \mathbf{u}_i)^T & t_i \end{bmatrix} \succeq 0$$

and write the optimization problem as an SDP:

| | |
|------------|--|
| minimize | 0 |
| subject to | $\begin{bmatrix} I & \mathbf{x} - \mathbf{u}_i \\ (\mathbf{x} - \mathbf{u}_i)^T & t_i \end{bmatrix} \succeq 0, i = 1, \dots, k$ $\begin{bmatrix} I & \mathbf{x} - \mathbf{v}_i \\ (\mathbf{x} - \mathbf{v}_i)^T & s_i \end{bmatrix} \succeq 0, s = 1, \dots, j$ $\mathbf{1}^T t \leq d$ $\mathbf{1}^T s \leq e$ |

(5) (a) To minimize the MSE:

$$\mathcal{L}(z) = \sum_r (y_r - |a_r^T x|^2)^2$$

For MLE estimate:

$$p_x(y) = \prod_{r=1}^m \Pr(w_r = y_r - (a_r^T x)^2 \mid x) = \frac{1}{(y_r - (a_r^T x)^2)!} \cdot \exp(-(a_r^T x)^2) \cdot (a_r^T x)^{2[y_r - (a_r^T x)^2]}$$

Therefore the log likelihood function is:

$$\begin{aligned} \ell_x(y) &= \sum_{i=1}^m \log[\Pr(y_i - a_i^T x \mid x)] = \sum_{i=1}^m \log \left[\frac{1}{(y_r - (a_r^T x)^2)!} \cdot \exp(-(a_r^T x)^2) \cdot (a_r^T x)^{2[y_r - (a_r^T x)^2]} \right] \\ &= \sum_{i=1}^m \log \left[\frac{1}{(y_r - (a_r^T x)^2)!} \right] - (a_r^T x)^2 + 2[y_r - (a_r^T x)^2] \cdot \log[(a_r^T x)] \end{aligned}$$

- (b) b
- (c) c
- (d) d
- (e) e

12 Abstract Algebra

These are my notes from reading *Elementary Abstract Algebra* by W. Edwin Clark, available for free download on his website: http://shell.cas.usf.edu/~wclark/#ELEMENTARY_ABSTRACT_ALGEBRA

12.1 Chapter 1: Binary Operations

Definition 1.1 A **binary operation** $*$ on a set S is a function from $S \times S$ to S . If $(a, b) \in S \times S$ then we write $a * b$ to indicate the image of the element (a, b) under the function $*$.

The following lemma explains in more detail exactly what this definition means.

Lemma 1.1 A binary operation $*$ on a set S is a rule for combining two elements of S to produce a third element of S . This rule must satisfy the following conditions:

- (a) $a \in S$ and $b \in S \implies a * b \in S$. [S is closed under $*$.]
- (b) For all a, b, c, d in S
 $a = c$ and $b = d \implies a * b = c * d$. [Substitution is permissible.]
- (c) For all a, b, c, d in S
 $a = b \implies a * c = b * c$.
- (d) For all a, b, c, d in S
 $c = d \implies a * c = a * d$.

Definition: A **function** f from the set A to the set B is a rule which assigns to each element $a \in A$ an element $f(a) \in B$ in such a way that the following condition holds for all $x, y \in A$:

$$x = y \implies f(x) = f(y)$$

To indicate that f is a function from A to B we write $f : A \rightarrow B$. The set A is called the **domain** of f and the set B is called the **codomain** of f .

A function $f : A \rightarrow B$ is said to be **one-to-one** or **injective** if the following condition holds for all $x, y \in A$:

$$f(x) = f(y) \implies x = y$$

A function $f : A \rightarrow B$ is said to be **onto** or **surjective** if the following condition holds:

$$\forall b \in B \exists a \in A \mid f(a) = b$$

A function $f : A \rightarrow B$ is said to be **bijective** if it is both one-to-one and onto. Then f is sometimes said to be a **bijection** or a **one-to-one correspondence** between A and B .

15. Let S , T , and U be nonempty sets, and let $f : S \rightarrow T$ and $g : T \rightarrow U$ be functions such that the function $g \circ f : S \rightarrow U$ is one-to-one (injective). Which of the following must be true?
- f is one-to-one.
 - f is onto.
 - g is one-to-one.
 - g is onto.
 - $g \circ f$ is onto.

Solution 15. (A) For a composition of functions, if the first function isn't one-to-one, there's no way the composite is. It's worth mentioning here that the opposite is true for onto: the second function had better be onto.

Let S be a set. The **power set** $\mathcal{P}(S)$ of S is the set of all subsets of S (including S itself).

Definition 1.2 Assume that $*$ is a binary operation on the set S .

1. We say that $*$ is **associative** if

$$x * (y * z) = (x * y) * z \quad \text{for all } x, y, z \in S.$$

2. We say that an element e in S is an **identity** with respect to $*$ if

$$x * e = x \text{ and } e * x = x \quad \text{for all } x \text{ in } S.$$

3. Let $e \in S$ be an identity with respect to $*$. Given $x \in S$ we say that an element $y \in S$ is an **inverse** of x if both

$$x * y = e \text{ and } y * x = e.$$

4. We say that $*$ is **commutative** if

$$x * y = y * x \quad \text{for all } x, y \in S.$$

5. We say that an element a of S is **idempotent** with respect to $*$ if

$$a * a = a.$$

6. We say that an element z of S is a **zero** with respect to $*$ if

$$z * x = z \text{ and } x * z = z \quad \text{for all } x \in S.$$

For each integer $n \geq 2$ define the set

$$\mathbb{Z}_n = \{0, 1, 2, \dots, n - 1\}$$

For all $a, b \in \mathbb{Z}_n$ let

$$a + b = \text{remainder when the ordinary sum of } a \text{ and } b \text{ is divided by } n$$

and

$$a \cdot b = \text{remainder when the ordinary product of } a \text{ and } b \text{ is divided by } n.$$

These binary operations are referred to as **addition modulo n** and **multiplication modulo n** . The integer n in \mathbb{Z}_n is called the **modulus**. The plural of modulus is **moduli**.

Let K denote any one of the following: $\mathbb{Z}, \mathbb{Q}, \mathbb{R}, \mathbb{Z}_n$.

$$M_n(K)$$

is the set of all $n \times n$ matrices containing elements of K .

$$GL(n, K)$$

is the set of all matrices in $M_n(K)$ with non-zero determinant. $(GL(n, k), \cdot)$ is called the **general linear group of degree n over K** . It is non-abelian.

$$SL(n, K) = \{A \in GL(n, K) \mid \det(A) = 1\}$$

$SL(n, K)$ is called the **Special Linear Group of degree n over K** .

12.2 Chapter 2: Groups

Definition A **group** is an ordered pair $(G, *)$ where G is a set and $*$ is a binary operation on G satisfying the following properties:

1. The binary operation is associative on G : $\forall x, y, z \in G$,

$$x * (y * z) = (x * y) * z$$

2. The binary operation contains a (unique) identity in G : $\exists e \in G \mid \forall x \in G$

$$e * x = x, x * e = x$$

3. Every element in G has a (unique) inverse on $*$ in G : $\forall x \in G \exists y \in G \mid$

$$x * y = e, y * x = e$$

A group $(G, *)$ is said to be **abelian** if $\forall x, y \in G$, $x * y = y * x$. A group is said to be **non-abelian** if it is not abelian.

Theorem 2.2 Let $(G, *)$ be a group with identity e . Then the following hold for all elements a, b, c, d in G :

1. If $a * c = a * b$, then $c = b$. [Left cancellation law for groups.]
2. If $c * a = b * a$, then $c = b$. [Right cancellation law for groups.]
3. Given a and b in G there is a unique element x in G such that $a * x = b$.
4. Given a and b in G there is a unique element x in G such that $x * a = b$.
5. If $a * b = e$ then $a = b^{-1}$ and $b = a^{-1}$. [Characterization of the inverse of an element.]
6. If $a * b = a$ for just one a , then $b = e$.
7. If $b * a = a$ for just one a , then $b = e$.
8. If $a * a = a$, then $a = e$. [The only idempotent in a group is the identity.]
9. $(a^{-1})^{-1} = a$.
10. $(a * b)^{-1} = b^{-1} * a^{-1}$.

12.3 Chapter 3: The Symmetric Groups

If n is a positive integer,

$$[n] = \{1, 2, \dots, n\}$$

A **permutation** of $[n]$ is a one-to-one, onto function from $[n]$ to $[n]$, and

$$S_n$$

is the set of all permutations of $[n]$.

The identity of S_n is the so-called **identity function**

$$\iota : [n] \rightarrow [n]$$

which is defined by the rule

$$\iota(x) = x, \quad \forall x \in [n]$$

The inverse of an element $\sigma \in S_n$: Suppose $\sigma \in S_n$. Since σ is by definition one-to-one and onto, the rule

$$\sigma^{-1}(y) = x \iff \sigma(x) = y$$

defines a function $\sigma^{-1} : [n] \rightarrow [n]$. This function σ^{-1} is also one-to-one and onto and satisfies

$$\sigma\sigma^{-1} = \iota \text{ and } \sigma^{-1}\sigma = \iota$$

so it is the inverse of σ in the group sense also.

Since the binary operation of composition on S_n is associative $[(\gamma\beta)\alpha = \gamma(\beta\alpha)]$, S_n under the binary operation of composition is a group (it is associative, it has an inverse, and it has an identity).

Definition 3.2 Let i_1, i_2, \dots, i_k be a list of k distinct elements from $[n]$. Define a permutation σ in S_n as follows:

$$\begin{array}{rcl} \sigma(i_1) & = & i_2 \\ \sigma(i_2) & = & i_3 \\ \sigma(i_3) & = & i_4 \\ & \vdots & \vdots \\ \sigma(i_{k-1}) & = & i_k \\ \sigma(i_k) & = & i_1 \end{array}$$

and if $x \notin \{i_1, i_2, \dots, i_k\}$ then

$$\sigma(x) = x$$

Such a permutation is called a **cycle** or a **k -cycle** and is denoted by

$$(i_1 \ i_2 \ \cdots \ i_k).$$

If $k = 1$ then the cycle $\sigma = (i_1)$ is just the identity function, i.e., $\sigma = \iota$.

Two cycles $(i_1 \ i_2 \ \dots \ i_k)$ and $(j_1 \ j_2 \ \dots \ j_l)$ are said to be **disjoint** if the sets $\{i_1, i_2, \dots, i_k\}$ and $\{j_1, j_2, \dots, j_l\}$ are disjoint.

So for example, the cycles $(1 \ 2 \ 3)$ and $(4 \ 5 \ 8)$ are disjoint, but the cycles $(1 \ 2 \ 3)$ and $(4 \ 2 \ 8)$ are not disjoint.

If σ and τ are disjoint cycles, then $\sigma\tau = \tau\sigma$.

Theorem 3.4 Every element $\sigma \in S_n$, $n \geq 2$, can be written as a product

$$\sigma = \sigma_1 \sigma_2 \cdots \sigma_m \tag{3.1}$$

where $\sigma_1, \sigma_2, \dots, \sigma_m$ are pairwise disjoint cycles, that is, for $i \neq j$, σ_i and σ_j are disjoint. If all 1-cycles of σ are included, the factors are unique except for the order. ■

The factorization (3.1) is called the **disjoint cycle decomposition of σ** .

An element of S_n is called a **transposition** if and only if it is a 2-cycle.

Every element of S_n can be written as a product of transpositions. The factors of such a product are not unique. However, if $\sigma \in S_n$ can be written as a product of k transpositions and if the same σ can also be written as a product of l transpositions, then k and l have the same parity.

A permutation is **even** if it is a product of an even number of transpositions and **odd** if it is a product of an odd number of transpositions. We define the function $\text{sign} : S_n \rightarrow \{1, -1\}$ by

$$\text{sign}(\sigma) = \begin{cases} 1 & \text{if } \sigma \text{ is even} \\ -1 & \text{if } \sigma \text{ is odd} \end{cases}$$

If $n = 1$ then there are no transpositions. In this case, to be complete we define the identity permutation ι to be even.

If σ is a k -cycle, then $\text{sign}(\sigma) = 1$ if k is odd and $\text{sign}(\sigma) = -1$ if k is even.

Remark. Let $A = [a_{ij}]$ be an $n \times n$ matrix. The determinant of A may be defined by the sum

$$\det(A) = \sum_{\sigma \in S_n} \text{sign}(\sigma) a_{1\sigma(1)} a_{2\sigma(2)} \cdots a_{n\sigma(n)}.$$

For example, if $n = 2$ we have only two permutations ι and $(1 \ 2)$. Since $\text{sign}(\iota) = 1$ and $\text{sign}((1 \ 2)) = -1$ we obtain

$$\det(A) = a_{11}a_{22} - a_{12}a_{21}.$$

Definition: If $(G, *)$ is a group, the number of elements in G is called the **order** of G . We use $|G|$ to denote the order of G . Note that $|G|$ may be finite or infinite.

Let

$$A_n$$

be the set of all even permutations in the group S_n . A_n is called the **alternating group of degree n** .

12.4 Chapter 4: Subgroups

Definition: Let G be a group. A **subgroup** of G is a subset H of G which satisfies the following three conditions:

1. $e \in H$
2. $a, b \in H \implies ab \in H$
3. $a \in H \implies a^{-1} \in H$

If H is a subgroup of G , we write $H \leq G$. The subgroups $\{e\}$ and G are said to be **trivial** subgroups of G .

Every finite subgroup may be thought of as a subgroup of one of the groups S_n .

Let A_n be the set of all even permutations in the group S_n . A_n is then a subgroup of S_n . A_n is called the **alternating group of degree n** .

Let a be an element of the group G . If $\exists n \in \mathbb{N} \mid a^n = e$ we say that a has **finite order** and we define

$$\text{o}(a) = \min\{n \in \mathbb{N} \mid a^n = e\}$$

If $a^n \neq e \forall n \in \mathbb{N}$ we say that a has **infinite order** and we define

$$\text{o}(a) = \infty$$

In either case we call $\text{o}(a)$ the **order** of a . Note carefully the difference between the order of a group and the order of an element of a group. Note also that $a = e \iff \text{o}(a) = 1$. So every element of a group other than e has order $n \geq 2$ or ∞ .

Let a be an element of group G . Define

$$\langle a \rangle = \{a^i : i \in \mathbb{Z}\}$$

We call $\langle a \rangle$ the **subgroup of G generated by a** . Note that $e = a^0$ and a^{-1} are in $\langle a \rangle$.

Theorem. For each $a \in G$, $\langle a \rangle$ is a subgroup of G . $\langle a \rangle$ contains a and is the smallest subgroup of G containing a .

Proof of second statement. If H is any subgroup of G containing a , $\langle a \rangle \subseteq H$ since H is closed under taking products and inverses. That is, every subgroup of G containing a also contains $\langle a \rangle$. This implies that $\langle a \rangle$ is the smallest subgroup of G containing a .

Theorem. Let G be a group and let $a \in G$. If $\text{o}(a) = 1$, then $\langle a \rangle = \{e\}$. If $\text{o}(a) = n$ where $n \geq 2$, then

$$\langle a \rangle = \{e, a, a^2, \dots, a^{n-1}\}$$

and the elements $e, a, a^2, \dots, a^{n-1}$ are distinct; that is,

$$\text{o}(a) = |\langle a \rangle|$$

Proof Assume that $\text{o}(a) = n$. The case $n = 1$ is left to the reader. Suppose $n \geq 2$. We must prove two things.

1. If $i \in \mathbb{Z}$ then $a^i \in \{e, a, a^2, \dots, a^{n-1}\}$.
2. The elements $e, a, a^2, \dots, a^{n-1}$ are distinct.

To establish 1 we note that if i is any integer we can write it in the form $i = nq + r$ where $r \in \{0, 1, \dots, n - 1\}$. Here q is the quotient and r is the remainder when i is divided by n . Now using Theorem 2.4 we have

$$a^i = a^{nq+r} = a^{nq}a^r = (a^n)^q a^r = e^q a^r = ea^r = a^r.$$

This proves 1. To prove 2, assume that $a^i = a^j$ where $0 \leq i < j \leq n - 1$. It follows that

$$a^{j-i} = a^{j+(-i)} = a^j a^{-i} = a^i a^{-i} = a^0 = e.$$

But $j - i$ is a positive integer less than n , so $a^{j-i} = e$ contradicts the fact that $\text{o}(a) = n$. So the assumption that $a^i = a^j$ where $0 \leq i < j \leq n - 1$ is false. This implies that 2 holds. It follows that $\langle a \rangle$ contains exactly n elements, that is, $\text{o}(a) = |\langle a \rangle|$.

Theorem. If G is a finite group, then every element of G has finite order.

49. What is the largest order of an element in the group of permutations of 5 objects?

- (A) 5 (B) 6 (C) 12 (D) 15 (E) 120

Solution 49. (B) The greatest order is given by the product of a 2-cycle and a 3-cycle acting on disjoint elements. That gives order 6.

12.5 Chapter 5: The Group of Units of \mathbb{Z}_n

Let $n \geq 2$. An element $a \in \mathbb{Z}_n$ is said to be a **unit** if $\exists b \in \mathbb{Z}_n \mid ab = 1$ (where the product is multiplication modulo n).

The set of all units in \mathbb{Z}_n is denoted by

$$U_n$$

and is a group under multiplication modulo n called the **group of units of \mathbb{Z}_n** .

Theorem. For $n \geq 2$, $U_n = \{a \in \mathbb{Z}_n : \gcd(a, n) = 1\}$

Theorem. p is a prime $\implies \exists a \in U_p \mid U_p = \langle a \rangle$

Theorem. If $n \geq 2$ then U_n contains an element a satisfying $U_n = \langle a \rangle$ if and only if a has one of the following forms: 2, 4, p^k , or $2p^k$ where p is an odd prime and $k \in \mathbb{N}$.

12.6 Chapter 6: Direct Products of Groups

If G_1, G_2, \dots, G_n is a list of n groups we make the Cartesian product $G_1 \times G_2 \times \dots \times G_n$ into a group by defining the binary operation

$$(a_1, a_2, \dots, a_n) \cdot (b_1, b_2, \dots, b_n) = (a_1 \cdot b_1, a_2 \cdot b_2, \dots, a_n \cdot b_n)$$

Here for each $i \in \{1, 2, \dots, n\}$ the product $a_i \cdot b_i$ is the product of a_i and b_i in the group G_i . We call this group the **direct product** of the groups G_1, G_2, \dots, G_n .

The direct product contains an identity and an inverse, and is associative (since it is composed of groups which must themselves be associative), so it is a group per below:

Theorem. If G_1, G_2, \dots, G_n is a list of n groups, the direct product $G = G_1 \times G_2 \times \dots \times G_n$ as defined above is a group. Moreover, if for each i , e_i is the identity of G_i , then e_1, e_2, \dots, e_n is the identity of G , and if

$$\mathbf{a} = (a_1, a_2, \dots, a_n) \in G$$

then the inverse of \mathbf{a} is given by

$$\mathbf{a}^{-1} = (a_1^{-1}, a_2^{-1}, \dots, a_n^{-1})$$

where a_i^{-1} is the inverse of a_i in the group G_i .

12.7 Chapter 7: Isomorphism of Groups

Let $G = \{g_1, g_2, \dots, g_n\}$. Let $\text{o}(g_i) = k_i$ for $i = 1, 2, \dots, n$. We say that the sequence (k_1, k_2, \dots, k_m) is the **order sequence** of the group G . To make the sequence unique we assume the elements are ordered so that $k_1 \leq k_2 \leq \dots \leq k_n$.

Let $(G, *)$ and (H, \bullet) be groups. A function $f : G \rightarrow H$ is said to be a **homomorphism** from G to H if

$$f(a * b) = f(a) \bullet f(b)$$

for all $a, b \in G$. If in addition f is one-to-one and onto, f is said to be an **isomorphism** from G to H .

We say that G and H are **isomorphic** if and only if there is an isomorphism from G to H . We write $G \cong H$ to indicate that G is isomorphic to H .

Isomorphism is an equivalence relation: If G, H , and K are groups then

1. $G \cong G$
2. If $G \cong H$ then $H \cong G$, and
3. If $G \cong H$ and $H \cong K$, then $G \cong K$.

Theorem. Let $(G, *)$ and (H, \bullet) be groups and let $f : G \rightarrow H$ be a homomorphism. Let e_G denote the identity of G , and let e_H denote the identity of H . Then

1. $f(e_G) = e_H$

Proof: Let $x_G \in G$ and let $f(x_G) = x_H \in H$. Then

$$x_H = f(x_G) = f(e_G * x_G) = f(e_G) \bullet f(x_G) = f(e_G) \bullet x_H = e_H \bullet x_H.$$

2. $f(a^{-1}) = f(a)^{-1}$

Proof: $f(a)^{-1} \bullet f(a) = e_H = f(e_G) = f(a^{-1} * a) = f(a^{-1}) \bullet f(a)$

3. $f(a^n) = f(a)^n \forall n \in \mathbb{Z}$

Proof by induction.

Theorem. Let $(G, *)$ and (H, \bullet) be groups and let $f : G \rightarrow H$ be an isomorphism. Then $\text{o}(a) = \text{o}(f(a)) \forall a \in G$. It follows that G and H have the same number of elements of each possible order.

Theorem. If G and H are isomorphic groups, and G is abelian, then so is H .

Proof: Let $a_G, b_G \in G$ and let $f(a_G) = a_H \in H, f(b_G) = b_H \in H$.

$$a_H \bullet b_H = f(a_G) \bullet f(b_G) = f(a_G * b_G) = f(b_G * a_G) = f(b_G) \bullet f(a_G) = b_H \bullet a_H.$$

A group G is **cyclic** if there is an element $a \in G$ | $\langle a \rangle = G$. If $\langle a \rangle = G$ then we say that a is a **generator** for G .

Theorem. If G and H are isomorphic groups and G is cyclic then H is cyclic.

Theorem. Let a be an element of group G .

1. $\text{o}(a) = \infty \implies \langle a \rangle \cong \mathbb{Z}$.
2. $\text{o}(a) = n \in \mathbb{N} \implies \langle a \rangle \cong \mathbb{Z}_n$

Cayley's Theorem. If G is a finite group of order n , then there is a subgroup H of S_n such that $G \cong H$.

66. Let \mathbb{Z}_{17} be the ring of integers modulo 17, and let \mathbb{Z}_{17}^\times be the group of units of \mathbb{Z}_{17} under multiplication.

Which of the following are generators of \mathbb{Z}_{17}^\times ?

- I. 5
- II. 8
- III. 16

- (A) None (B) I only (C) II only (D) III only (E) I, II, and III

Solution 66. (B) We need to pick elements of order 16 in $\mathbb{Z}/17^\times$. It is easy to rule out 16 $\equiv -1$, since -1 has order 2. We see that $5^2 = 25 \equiv 8$, so there's no way that 8 can be a generator. We just need to verify that the order of 5 is more than 8, so we can check 5^8 :

$$5^4 = 8^2 = 64 \equiv -4, \quad 5^8 = (-4)^2 = 16 \neq 1.$$

That makes 5 a generator.

12.8 Chapter 8: Cosets and Lagrange's Theorem

Let G be a group and let H be subgroup of G . For each element a of G we define

$$aH = \{ah \mid h \in H\}$$

We call aH the **coset of H in G generated by a** .

Let $a, b \in G$. Then

1. $a \in aH$ (since H must contain an identity; specifically, the identity of G)
2. $|aH| = |H|$ (since ah is unique)
3. $aH \cap bH \neq \emptyset \implies aH = bH$

Lagrange's Theorem. If G is a finite group and $H \leq G$ then $|H|$ divides $|G|$.

Any group of prime order is cyclic; therefore, there is only one such group up to isomorphism.

Exercise 3. Use Lagrange's theorem to prove that any group of prime order is cyclic.

Proof. Let G be a group whose order is a prime p . Since $p > 1$, there is an element $a \in G$ such that $a \neq e$. The group $\langle a \rangle$ generated by a is a subgroup of G . By Lagrange's theorem, the order of $\langle a \rangle$ divides $|G|$. But the only divisors of $|G| = p$ are 1 and p . Since $a \neq e$ we have $|\langle a \rangle| > 1$, so $|\langle a \rangle| = p$. Hence $\langle a \rangle = G$ and G is cyclic. \square

We say that there are k **isomorphism classes of groups of order n** if there are k groups G_1, G_2, \dots, G_k such that

1. if $i \neq j$ then G_i and G_j are not isomorphic, and
2. Every group of order n is isomorphic to G_i for some $i \in \{1, 2, \dots, k\}$.

This is sometimes expressed by saying that "there are k groups of order n up to isomorphism" or that "there are k non-isomorphic groups of order n ."

12. For which integers n such that $3 \leq n \leq 11$ is there only one group of order n (up to isomorphism) ?
- (A) For no such integer n
 - (B) For 3, 5, 7, and 11 only
 - (C) For 3, 5, 7, 9, and 11 only
 - (D) For 4, 6, 8, and 10 only
 - (E) For all such integers n

Solution 12. (B) Any group of prime order is necessarily cyclic, and hence there is only one up to isomorphism. This limits our choices to (B), (C), and (E). But there are two groups of order 9 (at least): $\mathbb{Z}/3 \times \mathbb{Z}/3$ and $\mathbb{Z}/9$. This makes (B) our only option.

In more advanced courses in algebra, it is shown that the number of isomorphism classes of groups of order n for $n \leq 17$ is given by the following table:

| | | | | | | | | | | | | | | | | | |
|-----------------|---|---|---|---|---|---|---|---|---|----|----|----|----|----|----|----|----|
| <i>Order :</i> | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 | 9 | 10 | 11 | 12 | 13 | 14 | 15 | 16 | 17 |
| <i>Number :</i> | 1 | 1 | 1 | 2 | 1 | 2 | 1 | 5 | 2 | 2 | 1 | 5 | 1 | 2 | 1 | 14 | 1 |

This table means, for example, that one may find 14 groups of order 16 such that every group of order 16 is isomorphic to one and only one of these 14 groups.

There is only one isomorphism class of groups of order n if n is prime. But there are some non-primes that have this property; for example, 15.

The Fundamental Theorem of Finite Abelian Groups. If G is a finite abelian group of order at least 2, then

$$G \cong \mathbb{Z}_{p_1^{n_1}} \times \mathbb{Z}_{p_2^{n_2}} \times \cdots \times \mathbb{Z}_{p_s^{n_s}}$$

where for each i , p_i is a prime and n_i is a positive integer. Moreover, the prime powers $p_i^{n_i}$ are unique except for the order of the factors.

If the group G in the above theorem has order n then

$$n = p_1^{n_1} p_2^{n_2} \cdots p_s^{n_s}$$

So the p_i may be obtained from the prime factorization of the order of the group G . These primes are not necessarily distinct, so we cannot say what the n_i are. However, we can find all possible choices for the n_i . For example, if G is an abelian group of order $72 = 3^2 \cdot 2^3$ then G is isomorphic to one and only one of the following groups. Note that each corresponds to a way of factoring 72 as a product of prime powers.

| | |
|--|--|
| $\mathbb{Z}_9 \times \mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_2$ | $72 = 9 \cdot 2 \cdot 2 \cdot 2$ |
| $\mathbb{Z}_9 \times \mathbb{Z}_4 \times \mathbb{Z}_2$ | $72 = 9 \cdot 4 \cdot 2$ |
| $\mathbb{Z}_9 \times \mathbb{Z}_8$ | $72 = 9 \cdot 8$ |
| $\mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_2$ | $72 = 3 \cdot 3 \cdot 2 \cdot 2 \cdot 2$ |
| $\mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_4 \times \mathbb{Z}_2$ | $72 = 3 \cdot 3 \cdot 4 \cdot 2$ |
| $\mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_8$ | $72 = 3 \cdot 3 \cdot 8$ |

Thus there are exactly 6 non-isomorphic abelian groups of order 72.

Corollary. For $n \geq 2$, the number of isomorphism classes of abelian groups of order n is equal to the number of ways to factor n as a product of prime powers (where the order of the factors does not count).

12.9 Chapter 9: Introduction to Ring Theory

Definition 9.1 A **ring** is an ordered triple $(R, +, \cdot)$ where R is a set and $+$ and \cdot are binary operations on R satisfying the following properties:

A1 $a + (b + c) = (a + b) + c$ for all a, b, c in R .

A2 $a + b = b + a$ for all a, b in R .

A3 There is an element $0 \in R$ satisfying $a + 0 = a$ for all a in R .

A4 For every $a \in R$ there is an element $b \in R$ such that $a + b = 0$.

M1 $a \cdot (b \cdot c) = (a \cdot b) \cdot c$ for all a, b, c in R .

D1 $a \cdot (b + c) = a \cdot b + a \cdot c$ for all a, b, c in R .

D2 $(b + c) \cdot a = b \cdot a + c \cdot a$ for all a, b, c in R .

Terminology If $(R, +, \cdot)$ is a ring, the binary operation $+$ is called *addition* and the binary operation \cdot is called *multiplication*. In the future we will usually write ab instead of $a \cdot b$. The element 0 mentioned in A3 is called the **zero** of the ring. Note that we have not assumed that 0 behaves like a *zero*, that is, we have not assumed that $0 \cdot a = a \cdot 0 = 0$ for all $a \in R$. What A3 says is that 0 is an identity with respect to addition. Note that *negative* (as the opposite of *positive*) has no meaning for most rings. We do not assume that multiplication is commutative and we have not assumed that there is an identity for multiplication, much less that elements have inverses with respect to multiplication.

23. Let $(\mathbb{Z}_{10}, +, \cdot)$ be the ring of integers modulo 10, and let S be the subset of \mathbb{Z}_{10} represented by $\{0, 2, 4, 6, 8\}$. Which of the following statements is FALSE?

- (A) $(S, +, \cdot)$ is closed under addition modulo 10.
- (B) $(S, +, \cdot)$ is closed under multiplication modulo 10.
- (C) $(S, +, \cdot)$ has an identity under addition modulo 10.
- (D) $(S, +, \cdot)$ has no identity under multiplication modulo 10.
- (E) $(S, +, \cdot)$ is commutative under addition modulo 10.

Solution 23. (D) Examining the choices, we see $S \subset \mathbb{Z}/10$ is a subgroup of an abelian group. Therefore it still have an additive identity and the operation is commutative. It is also closed under addition and multiplication. While S does not contain the multiplicative identity of $\mathbb{Z}/10$, it does have a multiplicative identity. $6 \in S$ is such an identity, as

$$6x = (5 + 1)x = 5x + x.$$

Since $x \in S$ are all even, $5x = 0$, so $6x = x$.

50. Let R be a ring and let U and V be (two-sided) ideals of R . Which of the following must also be ideals of R ?

I. $U + V = \{u + v : u \in U \text{ and } v \in V\}$

II. $U \cdot V = \{uv : u \in U \text{ and } v \in V\}$

III. $U \cap V$

- (A) II only (B) III only (C) I and II only (D) I and III only (E) I, II, and III

Solution 50. (D) The sum of the ideals is still an ideal: it is clearly closed under addition (using commutativity of addition), and still under left and right multiplication due to the distributive property. The intersection of ideals is still an ideal, which is not too hard to work out. The product of ideals, however, need not be closed under addition. Consider, for example, $R = \mathbb{Z}[X]$, $U = (2, X)$, and $V = (3, X)$ (the ideals generated by two elements). Then we know that $-2X \in U \cdot V$ and $3X \in U \cdot V$, and hence we should expect $3X - 2X = X \in U \cdot V$. However, there is no way to get X as the product of an element of U and an element of V .

18. Let V be the real vector space of all real 2×3 matrices, and let W be the real vector space of all real 4×1 column vectors. If T is a linear transformation from V onto W , what is the dimension of the subspace $\{\mathbf{v} \in V : T(\mathbf{v}) = \mathbf{0}\}$?

- (A) 2 (B) 3 (C) 4 (D) 5 (E) 6

Solution 18. (A) We see that $\dim V = 6$ and $\dim W = 4$. Since $\dim \text{im } T = \dim W = 4$, we must have $\dim \ker T = 6 - 4 = 2$.

13 Miscellaneous

6. Which of the following circles has the greatest number of points of intersection with the parabola $x^2 = y + 4$?

- (A) $x^2 + y^2 = 1$
- (B) $x^2 + y^2 = 2$
- (C) $x^2 + y^2 = 9$
- (D) $x^2 + y^2 = 16$
- (E) $x^2 + y^2 = 25$

Solution 6. (C) We can try to do this algebraically, but non-algebraically is simpler. Graphing $y = x^2 - 4$ shows that the graph crosses the x -axis at ± 2 . Therefore a circle of radius 1 or $\sqrt{2}$ will not intersect the parabola at all. A circle of radius 3 will intersect four times – twice above and twice below the x -axis. A circle of radius 4 will only intersect at one point below the x -axis (and twice above), and a circle of radius 5 will only intersect at the two points above.

19. If z is a complex variable and \bar{z} denotes the complex conjugate of z , what is $\lim_{z \rightarrow 0} \frac{(\bar{z})^2}{z^2}$?

- (A) 0
- (B) 1
- (C) i
- (D) ∞
- (E) The limit does not exist.

Solution 19. (E) Let us represent $z = a + bi$. Then our limit becomes

$$\lim_{(a,b) \rightarrow 0} \frac{(a - bi)^2}{(a + bi)^2} = \lim_{(a,b) \rightarrow 0} \frac{a^2 - b^2 - 2abi}{a^2 - b^2 + 2abi}.$$

If we let $a = 0$ (for instance), it is easy to see that the limit is equal to 1. However, if we let $a = b$, then our limit becomes

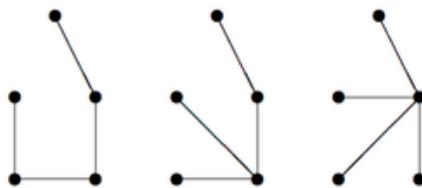
$$\lim_{a \rightarrow 0} \frac{-2a^2i}{2a^2i} = -1.$$

Therefore the limit does not exist.

29. A tree is a connected graph with no cycles. How many nonisomorphic trees with 5 vertices exist?

- (A) 1
- (B) 2
- (C) 3
- (D) 4
- (E) 5

Solution 29. (C) It's probably easiest to draw this out for yourself. The maximum degree of any vertex is 2, 3, or 4. If there is a vertex of degree 4, then our tree looks like a star. If the maximum degree of any vertex is 2, then we have a straight line. In the middle case, we obtain a 3-pointed star to which we attach one more vertex – the choice of branch yields isomorphic graphs. See Figure 1.



38. The maximum number of acute angles in a convex 10-gon in the Euclidean plane is

- (A) 1 (B) 2 (C) 3 (D) 4 (E) 5

Solution 38. (C) The total angle measure of a 10-gon is $180 \cdot 8 = 1440^\circ$. If the polygon is to be convex, all angles must be less than 180° . If we have 5 acute angles, then the remaining 5 angles would have to make up for $> 1440 - 5 \cdot 90 = 990$ degrees. This is impossible to do and remain convex. If we have 4 acute angles, the remaining 6 angles need to make up for $> 1440 - 4 \cdot 90 = 1080$ degrees. This is our edge case, so the answer must be 3 acute angles.

45. How many positive numbers x satisfy the equation $\cos(97x) = x$?

- (A) 1 (B) 15 (C) 31 (D) 49 (E) 96

Solution 45. (C) Certainly our solutions are concentrated in $[0, 1]$. We know that every $2\pi/97$ units in x , we get another period of $\cos(97x)$, and each period must meet $y = x$ twice. Therefore there are

$$\frac{1}{2\pi/97} = \frac{97}{2\pi} \approx \frac{97}{6.3} \approx 15$$

periods in $[0, 1]$ and about 30 meetings. There's only one answer in that range, so we'll stick with it.