# Smoothing Based on Stretched Interpolated Moving Average Approach 

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#### Abstract

In this thesis, some smoothing techniques in multivariate and functional data analysis such as, kernel smoothing, local linear regression (LLR), spline smoothing and smoothing together with principal components analysis through conditional expectation (PACE) methods are considered. Their details are studied and a new smoothing method benefiting from moving average concept and applicable under certain conditions is proposed. Due to the steps involved in its logic, the proposed method is named Strecthed Interpolated Moving Average (SIMA). Its application to different data sets produced better results in terms of involved error, compared with LLR and similar results when compared with PACE.


Keywords: Karhunen-Loève Expansion, Stretched Interpolated Moving Average, Principal Component Scores, Lag Interval, Weight Function.

## ÖZ

Bu tezde, çok değişkenli ve fonksiyonel veri analizinin; çekirdek pürüzsüzleştirme, yerel lineer regresyon (LLR), spline pürüzsüzleştirme, ve koşullu beklenti ile temel bileşenler analizi (PACE) gibi bazı pürüzsüzleştirme tekniklerine yer verilmiştir. Bunların ayrıntıları incelenmiş ve belirli koşullar altında hareketli-ortalamadan yararlanılarak yeni bir pürüzsüzleştirme tekniği önerilmiştir. Kendi mantığı içinde yer alan adımları nedeniyle önerilen yöntem Gerilmiş İnterpolasyonlu Hareketli-Ortalama (SIMA) diye adlandırılır. SIMA'nın farklı verilerde yapılan uygulamasında LLR uygulamasına kıyasla daha iyi sonuçlar elde edilmiş, PACE ile kıyaslandığında ise benzer sonuçlar elde edilmiştir.

Anahtar Kelimeler: Karhunen-Loève Açılımı, Gerilmiş Interpolasyonlu HareketliOrtalama, Temel Bileşenler Skorları, Gecikme Aralığı, Ağırlık Fonksiyonu.

To My Father

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## LIST OF SYMBOLS

b Local linear regression coefficient vector.
$G\left(t, t^{+}\right)$Covariance function.
G Covariance matrix of the functional representation of the data set.
$h \quad$ Band width or the local neighbourhood of smoothing parameter.
$\ell_{k} \quad$ Eigenvalue of $k^{\text {th }}$ eigenvector $w_{k}$.
$n \quad$ Number of objects or trajectories.
$m \quad$ Number of data values used for moving averaging.
$M \quad$ Total number of the moving averaged data values.
$p \quad$ Number of variables.
$p_{i} \quad$ Number of observations on the $i$ th trajectory.
$p_{x_{i}} \quad$ Projected coordinate of data.
$S \quad$ Sample covariance operator.
S Sample covariance matrix.
$\mathbf{u}_{i} \quad i^{\text {th }}$ unit eigenvector of $\mathbf{X}^{T} \mathbf{X}$ / weight function.
$\mathbf{v}_{j} \quad j^{\text {th }}$ unit eigenvector of $\mathbf{X X} \mathbf{X}^{T}$.
$x_{i j} \quad i^{\text {th }}$ observation on the $j^{\text {th }}$ object.
$\mathbf{x}_{(j)} \quad$ The column vector.
$\mathbf{x}_{i} \quad$ The row vector.
$X \quad$ The random variable.
$X_{i} \quad i^{\text {th }}$ random variable/ $i^{\text {th }}$ trajectory.
$X(t) \quad$ Functional variable at time $t$.
$\mathbf{X} \quad$ The $n \times p$ data matrix.
$z_{k} \quad$ Rowwise $k^{\text {th }}$ principal component.
$w_{k} \quad$ Columnwise $k^{\text {th }}$ principal component.
$W_{j} \quad j^{\text {th }}$ weight function.
$Y_{i j} \quad$ Functional representation of $j^{\text {th }}$ object on the $i^{\text {th }}$ trajectory.
$Y_{i l}^{*} \quad l^{\text {th }}$ stretched interpolated moving average value on the $i^{\text {th }}$ trajectory.
$\lambda_{k} \quad k^{\text {th }}$ eigenvalue of eigenvector $z_{k} / k^{\text {th }}$ eigenvalue of the covariance matrix.
$\phi_{k} \quad$ Eigenfunction of $G\left(t, t^{+}\right)$.
$\xi_{k} \quad k^{\text {th }}$ principal component scores.
$\boldsymbol{\Phi} \quad$ Matrix formed by basis vectors in fitting of ordinary least squares.
$\boldsymbol{\mu} \quad$ The mean vector.
$\Sigma \quad$ Theoretical covariance matrix.

## Chapter 1

## INTRODUCTION

Studies in statistical data analysis gained momentum at the beginning of the $20^{\text {th }}$ century. Substantial foundation work laid during the first half, and with the advent of computers in the second half of the same century, wide applications into all disciplines became common ground. Today the development of statistical theory and application of developed ideas using the continuously advancing computer technology has enabled the testing of abstract statistics theory, previously not possible. This resulted in rapid development of nonparametric statistical data analysis.

For the analysis of multivariate data that is considered in this study, one main issue is to smooth the data before processing in order to eliminate the effect of extreme values. Alternately smoothing the mean and covariance functions are commonly used in many data analysis methods. Functional data analysis (FDA) forms the theoretical foundation for multivariate data analysis (MDA). Theory related with MDA and FDA is summarized in Chapter 2 and Chapter 3 respectively. One of the main concepts used in MDA is the Principle Component Analysis (PCA) that enables dimension reduction of multivariate data. Pioneering work on this topic was initially carried out by Hotelling (Hotelling, 1933), who built his theory on the foundations laid out by Karl Pearson. One main issue in FDA and MDA is the smoothing of raw data, mean and covariance
functions. There are many smoothing methods developed over the years. Amongst some widely used ones are Kernel group of smoothers, Spline smoothers, Regression smoothers and Moving Average smoothers. In this work, Epanechnikov kernel, and local linear regression smoothers are used and a specific method for the moving average smoothers is proposed.

The Epanechnikov kernel and spline smoothers are used in the PCA through conditional expectation method (PACE) for smoothing the mean and covariance functions, Müller (2005). Further details on kernel smoothing is given by Härdle (1992). Structure of a smoothing spline is explained in detail by Ramsay and Silverman (2006). Principles of smoothing PCA is given in (Ramsay and Silverman, 2002). Local linear regression smoothing (LLR) is explained by Loader (1999).

The main idea of estimating a trajectory from available data, through MDA is widely studied by many different researchers. One note worthy method in this respect is PACE, which is summarized in Section 2.3.

Smoothing a trajectory, the mean function or the covariance matrix (surface) is possible using any of the above mentioned methods. The proposed moving average smoothing method is named as Stretched Interpolated Moving Average (SIMA), mainly because of the steps involved in its computation. Details are given in Chapter 4.

The proposed SIMA smoothing method is applied to two distinct data sets together with kernel, and LLR smoothers in Chapter 5. Obtained results from SIMA are compared
with those from other methods. SIMA performed better than LLR under the condition of weak correlation between the variables involved. Since kernel and spline smoothers are used in the PACE method, these smoothers are sometimes referred to as PACE smoothers. In the application to the data sets for the smoothing of mean and covariance, SIMA performed equally well with PACE smoothers. The measure used in comparing different smoothers is Mean Square Error (MSE) between an observed trajectory and its smooth estimate.

## Chapter 2

## MULTIVARIATE DATA ANALYSIS

### 2.1 Introduction to the Multivariate Data Analysis

Analyzing data where more than one variable is involved requires the use of MDA techniques. Representing the data in matrix format is essential in the process. Let $x_{i j}: i=1, \ldots, n ; j=1, \ldots, p$ be the set of $n \times p$ observations or data set. Then each column of the data belongs to the $j^{\text {th }}$ random variable $X_{j}$ and denoted by the column vector $\mathbf{x}_{(j)}=\left[\begin{array}{c}x_{1 j} \\ \vdots \\ x_{n j}\end{array}\right] ; j=1, \ldots, p$. Similarly the rows represent the data values belonging to each trajectory, denoted by the row vector $\mathbf{x}_{i}=\left[\begin{array}{c}x_{i 1} \\ \vdots \\ x_{i p}\end{array}\right] ; i=1, \ldots, n$. In this setup the row
vectors $\mathbf{x}_{1}^{T}, \ldots, \mathbf{x}_{n}^{T}$ represents a random sample of the trajectories while the column vectors $\mathbf{x}_{(1)}, \ldots, \mathbf{x}_{(p)}$ represents the values of the random variables $X_{j}$.

The vector representation of data given above forms the $n \times p$ data matrix $\mathbf{X}$ as given below.

$$
\mathbf{X}=\text { objects }\left\{\begin{array}{|cccc}
\left(\begin{array}{ccccc}
x_{11} & \cdots & x_{1 j} & \cdots & x_{1 p} \\
\vdots & \ddots & \vdots & & \vdots \\
x_{i 1} & \cdots & x_{i j} & \cdots & x_{i p} \\
\vdots & & \vdots & \ddots & \vdots \\
x_{n 1} & \cdots & x_{n j} & \cdots & x_{n p}
\end{array}\right)
\end{array}\right.
$$

When the number of variables and the trajectories are large, numerical processing needed in analyzing such data becomes prohibitive. On the other hand it is a fact that not every variable will have the same impact on the process under consideration. Therefore, one main concern of MDA is to identify the variables having major influence on the process under study. This in turn will enable the exclusion of the variables with minor or marginal effect, hence alleviating load of data processing while maintaining high level of accuracy (Mardia, et. al., 1979).

### 2.1.1 The Sample Statistics

The mean vector and covariance matrix of the multivariate data can be written by extending the univariate case to the multivariate form. The sample mean and sample variance of the $j^{t h}$ variable are given as in equations (2.1.1) and (2.1.2).

$$
\begin{equation*}
\bar{x}_{j}=\frac{\sum_{i=1}^{n} x_{i j}}{n}, \tag{2.1.1}
\end{equation*}
$$

The vector of means for $p$ variables is $\overline{\mathbf{x}}=\left[\begin{array}{c}\bar{x}_{1} \\ \vdots \\ \bar{x}_{p}\end{array}\right]=\frac{1}{n} \sum_{i=1}^{n} \mathbf{x}_{i}=\frac{1}{n} \mathbf{X}^{T} \mathbf{1}_{n}$. Here $\mathbf{1}_{n}=\left[\begin{array}{c}1 \\ \vdots \\ 1\end{array}\right]$.

$$
\begin{equation*}
s_{j}^{2}=\frac{\sum_{i=1}^{n}\left(x_{i j}-\bar{x}_{j}\right)^{2}}{n-1},(j=1, \ldots, p) . \tag{2.1.2}
\end{equation*}
$$

The sample covariance between the $j^{t h}$ and the $j^{* t h}$ variables is

$$
\begin{equation*}
s_{i j j^{*}}=\frac{\sum_{i=1}^{n}\left(x_{i j}-\bar{x}_{j}\right)\left(x_{i j^{*}}-\bar{x}_{j^{*}}\right)}{n-1}, j, j^{*}=1, \ldots, p \tag{2.1.3}
\end{equation*}
$$

It is evident that, $s_{i j^{*}}=s_{j}^{2}$, when $j=j^{*}$.

The $p \times p$ covariance matrix can be written as

$$
\mathbf{S}=\frac{1}{n} \sum_{i=1}^{n}\left(\mathbf{x}_{i}-\overline{\mathbf{x}}\right)\left(\mathbf{x}_{i}-\overline{\mathbf{x}}\right)^{T} .
$$

Using the centring matrix $\mathbf{H}=\mathbf{I}-\frac{1}{n} \mathbf{1 1}^{T}$, the covariance can also be denoted as

$$
\begin{equation*}
\mathbf{S}=\frac{1}{n} \mathbf{X}^{T} \mathbf{H} \mathbf{X} \tag{2.1.4}
\end{equation*}
$$

Since $\mathbf{H}$ is symmetric and idempotent, using a $p$-vector $\mathbf{v}, \mathbf{v}^{T} \mathbf{S v}=\frac{1}{n} \mathbf{v}^{T} \mathbf{X}^{T} \mathbf{H}^{T} \mathbf{H} \mathbf{X} \mathbf{v} \geq 0$ can be written, meaning the covariance matrix $\mathbf{S}$ in equation (2.1.4) is positive semidefinite.

### 2.1.2 Linear Transformation

Often in MDA linear transformation of data becomes necessary before analysis, due to linearly transformed data results in dimension reduction, simplifying computations. Hence, computation of statistics for linearly transformed data has to be formulated. Letting $\mathbf{a}^{T}=\left(a_{1}, \ldots, a_{p}\right)$ be the vector of coefficients to be used in the transformation, transformed data will be $y_{i}=\mathbf{a}^{T} \mathbf{x}_{i j} ; \mathrm{i}=1, \ldots, n, \mathrm{j}=1, \ldots, p$. Transformed data will have a mean

$$
\bar{y}=\frac{1}{n} \mathbf{a}^{T} \sum_{i=1}^{n} \mathbf{x}_{i}=\mathbf{a}^{T} \overline{\mathbf{x}}
$$

and variance

$$
s_{y}^{2}=\frac{1}{n} \sum_{i=1}^{n}\left(y_{i}-\bar{y}\right)^{2}=\frac{1}{n} \sum_{i=1}^{n} \mathbf{a}^{T}\left(\mathbf{x}_{i}-\overline{\mathrm{x}}\right)\left(\mathbf{x}_{i}-\overline{\mathrm{x}}\right)^{T} \mathbf{a}=\mathbf{a}^{T} \mathbf{S} \mathbf{a} .
$$

In a $q$ dimensional linear transformation, $\mathbf{A}_{q \times p}$ being the matrix of coefficients and $\mathbf{b}_{q}$ vector of constants, then,

$$
\mathbf{y}_{i}=\mathbf{A x}+\mathbf{b}, i=1, \ldots, n \rightarrow \mathbf{Y}=\mathbf{X A}^{T}+\mathbf{1} \mathbf{b}^{T}
$$

can be written. Then, the mean vector and covariance matrix of transformation will be

$$
\overline{\mathbf{y}}=\mathbf{A} \overline{\mathbf{x}}+\mathbf{b} \text { and } \mathbf{S}_{y}=n^{-1} \sum_{i=1}^{n}\left(\mathbf{y}_{i}-\overline{\mathbf{y}}\right)\left(\mathbf{y}_{i}-\overline{\mathbf{y}}\right)^{T}=\mathbf{A S A} \mathbf{A}^{T} .
$$

It can be shown that linear combinations of a multinormal vector are univariate normal (Mardia, et.al., 1979).

### 2.2 Multivariate Normal Theory

The univariate normal distribution is the most widely used distribution in many statistical application problems. Its multivariate version distribution similarly enables the solution of many multivariate estimation problems. Therefore, it plays a major role in MDA. It is wholly defined by its first and second moments and the p-variate normal distribution is given by

$$
f(\mathbf{x})=|2 \pi \boldsymbol{\Sigma}|^{-1 / 2} \exp (\mathbf{x}-\boldsymbol{\mu})^{T} \boldsymbol{\Sigma}^{-1}(\mathbf{x}-\boldsymbol{\mu})
$$

where $\boldsymbol{\Sigma}>0$ is the positive definite covariance matrix, $\mathbf{x}_{p}$ and $\boldsymbol{\mu}_{p}$ are the vectors of random variables and their means, respectively. Then, $\mathcal{N}(\boldsymbol{\mu}, \boldsymbol{\Sigma})$ denotes a multivariate normal distribution with parameters $\boldsymbol{\mu}$ and $\boldsymbol{\Sigma}$ (Park, 2008).

In multivariate normal distribution, for pairs $\left(X_{i}, X_{j}\right)$, correlation $\rho_{X_{i} X_{j}}=0, i \neq j$ implies independence and pairwise independence implies total independence.

Corollary 2.2: If $\mathbf{x}$ has a $p$-variate normal distribution, and if $\mathbf{y}=\mathbf{A x}+\mathbf{c}$ is the linear combination of the variables, with $q \times p$ dimensional matrix $\mathbf{A}$ and $q$-vector $\mathbf{c}$, then $\mathbf{y}$ has $q$-variate normal distribution. That is, if $\quad \mathbf{x} \square \mathcal{N}_{p}(\boldsymbol{\mu}, \boldsymbol{\Sigma})$, then $\mathbf{y} \square \mathcal{N}_{q}\left(\mathbf{A} \boldsymbol{\mu}+\mathbf{b}, \mathbf{A} \mathbf{\Sigma} \mathbf{A}^{T}\right)$, where $\boldsymbol{\mu}$ is the mean vector and $\boldsymbol{\Sigma}$ is the covariance matrix (Mardia et al., 1979).

Certain applications may require the use of partitioned matrices, due to the nature of the process. For example in a $p$ variate process, if it is required to show the independence of $k$ variables in a process from remaining $t=p-k$ elements of the process, then the covariance vector of variables $\mathbf{x}$ can be partitioned into two sub-vectors $\mathbf{x}_{1}$ and $\mathbf{x}_{2}$ with $k$ and $t$ elements, respectively as given in the following Theorem 2.2.1.

Theorem 2.2.1 (Mardia et al., 1979): Assume $\mathbf{x}=\binom{\mathbf{x}_{1}}{\mathbf{x}_{2}} \square \mathcal{N}_{p}(\boldsymbol{\mu}, \boldsymbol{\Sigma})$ where $\mathbf{x}_{1} \in \square^{k}$, $\mathbf{x}_{2} \in \square^{t}$, and $\mathbf{x}_{2.1}=\mathbf{x}_{2}-\boldsymbol{\Sigma}_{21} \Sigma_{11}^{-1} \mathbf{x}_{1}$ defined from the partitioned covariance matrix $\boldsymbol{\Sigma}=\left(\begin{array}{ll}\boldsymbol{\Sigma}_{11} & \boldsymbol{\Sigma}_{12} \\ \boldsymbol{\Sigma}_{21} & \boldsymbol{\Sigma}_{22}\end{array}\right), \quad$ then, $\quad \mathbf{x}_{1} \square \mathcal{N}_{k}\left(\boldsymbol{\mu}_{1}, \boldsymbol{\Sigma}_{11}\right)$ and $\mathbf{x}_{2.1} \square \mathcal{N}_{t}\left(\boldsymbol{\mu}_{2.1}, \boldsymbol{\Sigma}_{22.1}\right)$ are statistically independent, i.e. $\operatorname{Cov}\left(\mathbf{x}_{1}, \mathbf{x}_{2.1}\right)=0$, with $\boldsymbol{\mu}_{2.1}=\boldsymbol{\mu}_{2}-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \boldsymbol{\mu}_{1}$ and $\boldsymbol{\Sigma}_{22.1}=\boldsymbol{\Sigma}_{22}-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \boldsymbol{\Sigma}_{12}$.

Proof: Let $\mathbf{x}_{1}=\left(\begin{array}{ll}\mathbf{I}_{k} & 0\end{array}\right) \mathbf{x}$ and $\mathbf{x}_{2.1}=\left(\begin{array}{ll}-\Sigma_{21} \Sigma_{11}^{-1} & \mathbf{I}_{t}\end{array}\right) \mathbf{x}$. By Corollary 2.2, the multivariate normalities, $\mathbf{x}_{1} \square \mathcal{N}_{k}\left(\boldsymbol{\mu}_{1}, \boldsymbol{\Sigma}_{11}\right)$ and $\mathbf{x}_{2.1} \square \mathcal{N}_{t}\left(\boldsymbol{\mu}_{2}-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \boldsymbol{\mu}_{1}, \boldsymbol{\Sigma}_{22}-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \boldsymbol{\Sigma}_{12}\right)$ are clear. Consider the covariance for the independency.

$$
\begin{aligned}
\operatorname{cov}\left(\mathbf{x}_{1}, \mathbf{x}_{2.1}\right) & =\left(\begin{array}{ll}
\mathbf{I}_{k} & \mathbf{0}
\end{array}\right) \boldsymbol{\Sigma}\left(\begin{array}{ll}
-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} & \mathbf{I}_{t}
\end{array}\right) \\
& =\left(\begin{array}{ll}
\mathbf{I}_{k} & 0
\end{array}\right)\left(\begin{array}{ll}
\boldsymbol{\Sigma}_{11} & \boldsymbol{\Sigma}_{12} \\
\boldsymbol{\Sigma}_{21} & \boldsymbol{\Sigma}_{22}
\end{array}\right)\left(\begin{array}{ll}
-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} & \mathbf{I}_{t}
\end{array}\right) \\
& =\left(-\boldsymbol{\Sigma}_{11}\left(\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}\right)^{T}+\boldsymbol{\Sigma}_{12}\right)
\end{aligned}
$$

$$
\begin{aligned}
& =\left(-\boldsymbol{\Sigma}_{11} \boldsymbol{\Sigma}_{11}^{T} \boldsymbol{\Sigma}_{12}+\boldsymbol{\Sigma}_{12}\right) \\
& =0,
\end{aligned}
$$

i.e. $\mathbf{x}_{1}$ and $\mathbf{x}_{2.1}$ are statistically independent.

This theorem paves the way to the idea of conditional relationship between two subvectors of a vector of random variables. This is highlighted in the following theorem.

Theorem 2.2.2 (Mardia et al., 1979): From Corollary 2.2 and Theorem 2.2.1, conditional distribution of $\mathbf{x}_{2}$ for a certain value of $\mathbf{x}_{1}$ is approximately normally distributed, that is

$$
\begin{equation*}
\mathbf{x}_{2} \mid \mathbf{x}_{1} \square \mathcal{N}_{t}\left(\mu_{2}+\Sigma_{21} \Sigma_{11}^{-1}\left(\mathbf{x}_{1}-\mu_{1}\right), \Sigma_{22.1}\right) . \tag{2.2.1}
\end{equation*}
$$

with conditional mean

$$
\begin{equation*}
E\left[\mathbf{x}_{2} \mid \mathbf{x}_{1}\right]=\mu_{2.1}+\boldsymbol{\Sigma}_{21} \Sigma_{11}^{-1} \mathbf{x}_{1}=\mu_{2}+\Sigma_{21} \Sigma_{11}^{-1}\left(\mathbf{x}_{1}-\boldsymbol{\mu}_{1}\right) . \tag{2.2.2}
\end{equation*}
$$

Proof: Since $\mathbf{x}_{2.1}$ is independent of $\mathbf{x}_{1}$, its conditional distribution for a given value of $\mathbf{x}_{1}$ is same as its marginal distribution. Now $\mathbf{x}_{2}=\mathbf{x}_{2.1}+\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \mathbf{x}_{1}$ and this term is constant when $\mathbf{x}_{1}$ is given. Therefore the conditional distribution of $\mathbf{x}_{2} \mid \mathbf{x}_{1}$ is normal with conditional mean $E\left[\mathbf{x}_{2} \mid \mathbf{x}_{1}\right]=\boldsymbol{\mu}_{2.1}+\boldsymbol{\Sigma}_{\mathbf{2 1}} \boldsymbol{\Sigma}_{11}^{\mathbf{- 1}} \mathbf{x}_{1}$ (Mardia et al., 1979).

On the other hand, let $x_{1} \square \mathcal{N}\left(\mu_{1}, \sigma_{1}^{2}\right)$ and $x_{2} \square \mathcal{N}\left(\mu_{2}, \sigma_{2}^{2}\right)$. Since, by Normal Distribution Theory the marginal distribution of $x_{1}$ is,

$$
f_{1}\left(x_{1}\right)=\int_{-\infty}^{\infty} \frac{\exp \left(\frac{-1}{2-2 \rho^{2}}\left(\frac{\left(x_{1}-\mu_{1}\right)^{2}}{\sigma_{1}^{2}}-2 \rho \frac{x_{1}-\mu_{1} x_{2}-\mu_{2}}{\sigma_{1}}+\frac{\left(x_{2}-\mu_{2}\right)^{2}}{\sigma_{2}^{2}}\right)\right.}{2 \pi \sigma_{1} \sigma_{2} \sqrt{1-\rho^{2}}} d x_{2}=\frac{e^{\left.-\frac{1}{2} \frac{x_{1}-\mu_{1}}{\sigma_{1}}\right)^{2}}}{\sigma_{1} \sqrt{2 \pi}} .
$$

Moreover,

$$
E\left[x_{2} \mid x_{1}\right]=\int_{-\infty}^{\infty} x_{2} f\left(x_{2} \mid x_{1}\right) d x_{2}
$$

where the conditional distribution is as,

$$
f\left(x_{2} \mid x_{1}\right)=\frac{e^{-\frac{1}{2}\left(\frac{x_{2}-\left(\mu_{2}+\rho \frac{\sigma_{2}}{\sigma_{1}}\left(x_{1}-\mu_{1}\right)\right.}{\sigma_{2} \sqrt{1-\rho^{2}}}\right)^{2}}}{\sigma_{2} \sqrt{2 \pi\left(1-\rho^{2}\right)}} .
$$

Clearly by Normal theory, the conditional expectation can be written as,

$$
\begin{equation*}
E\left[x_{2} \mid x_{1}\right]=\mu_{2}+\rho \frac{\sigma_{2}}{\sigma_{1}}\left(x_{1}-\mu_{1}\right) \tag{2.2.3}
\end{equation*}
$$

and the general variance/covariance is as,

$$
\begin{equation*}
\operatorname{var}\left[x_{2} \mid x_{1}\right]=\sigma_{2}^{2}\left(1-\rho^{2}\right) \tag{2.2.4}
\end{equation*}
$$

Now, in terms of matrices; consider $\mathbf{x}_{1} \square N_{k}\left(\boldsymbol{\mu}_{1}, \boldsymbol{\Sigma}_{11}\right)$ and $\mathbf{x}_{2.1} \square N_{t}\left(\boldsymbol{\mu}_{2}-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \boldsymbol{\mu}_{1}, \boldsymbol{\Sigma}_{22}-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \boldsymbol{\Sigma}_{12}\right)$, then, substitute $\boldsymbol{\Sigma}_{i j}=\rho \sigma_{i} \sigma_{j}$ and $\boldsymbol{\Sigma}_{i i}=\sigma_{i}^{2}$ in (2.2.3) and (2.2.4) we have, (2.2.2) and $\operatorname{cov}\left[x_{2} \mid x_{1}\right]=\boldsymbol{\Sigma}_{22}-\boldsymbol{\Sigma}_{21} \boldsymbol{\Sigma}_{11}^{-1} \boldsymbol{\Sigma}_{12}=\boldsymbol{\Sigma}_{22.1}$.

This theorem will help in understanding the concept of principal component analysis under some conditional forms.

### 2.3 Eigenvalues and Eigenvectors

The symmetric covariance matrix is extensively used in MDA by means of eigenvalues and eigenvectors. Therefore two of the theorems pertaining to the eigenvalues and eigenvectors are presented for information.

The theorem that establishes links between the structure of a symmetric matrix and its eigenvalues and eigenvectors which can be found in most serious linear algebra books, is as follows.

Theorem 2.3.1: (Spectral or Jordan Decomposition) Any symmetric $p \times p$ matrix $\mathbf{A}$ can be written in terms of $\boldsymbol{\Lambda}$, the diagonal matrix of its eigenvalues and $\boldsymbol{\Gamma}$, an orthogonal matrix whose columns are standardized eigenvectors of A,

$$
\begin{equation*}
\mathbf{A}=\boldsymbol{\Gamma} \mathbf{\Lambda} \boldsymbol{\Gamma}^{T}=\sum_{j=1}^{p} \lambda_{j} \boldsymbol{\gamma}_{j} \boldsymbol{\gamma}_{j}^{T} \tag{2.3.1}
\end{equation*}
$$

where, $\boldsymbol{\Lambda}=\operatorname{diag}\left(\lambda_{1}, \ldots, \lambda_{p}\right)$ and $\boldsymbol{\Gamma}=\left(\boldsymbol{\gamma}_{1}, \boldsymbol{\gamma}_{2}, \ldots, \boldsymbol{\gamma}_{p}\right)$.

Therefore, (2.3.1) shows that a symmetric matrix is uniquely determined by its eigenvalues and eigenvectors. If $\lambda_{i}$ are distinct and written in decreasing order, then $\boldsymbol{\Gamma}$ is also uniquely determined.

Another important concept used in MDA under certain circumstances, i.e. when a matrix is not square, is the singular value decomposition theorem given below.

Theorem 2.3.2: (Singular Value Decomposition) Each $n \times p$ matrix A with rank $r$ can be decomposed by column orthonormal matrices, $\boldsymbol{\Gamma}(n \times r)$ and $\boldsymbol{\Delta}(p \times r)$, satisfying $\boldsymbol{\Gamma}^{T} \boldsymbol{\Gamma}=\boldsymbol{\Delta}^{T} \boldsymbol{\Delta}=\mathbf{I}_{r}$ and diagonal $\boldsymbol{\Theta}$ matrix of positive elements, so that $\mathbf{A}=\boldsymbol{\Gamma} \boldsymbol{\Theta} \boldsymbol{\Delta}^{T}$.

This is the generalization of the Jordan decomposition theorem.

### 2.4 Factoring the Data Matrices

In this section factoring data matrices is reviewed, since principal component analysis depends on the concepts developed here. The aim is to reduce the dimension of the data matrix by means of geometric approach with respect to a least-squares criterion. As a result, low dimensional graphical pictures of the data matrix area obtained. This is the process of decomposing the data matrix into factors, a concept used in many multivariate techniques. Dimension reduction facilitates the easy interpretation of the process estimated by the data.

### 2.4.1 Projecting Data from Higher to Lower Dimensional Space

Representing a multivariate data set in matrix format $\mathbf{X}_{n \times p}$ was introduced in Section 2.1. Projection of data values can be performed row or columnwise. In other words the column space $\mathbf{C}(\mathbf{X})$ or the row space $\mathbf{C}\left(\mathbf{X}^{T}\right)$ can be approximated by smaller subspaces. An important point in dimension reduction is not to lose much from the variation and structure of the data.

### 2.4.1.1 Projecting Data onto $\square^{q}$ from $\square^{p}$

Procedure for projecting the $p$-dimensional $n$ data points onto a subspace $\square^{q},(q \leq p)$ is explained. For simplicity details of projection from $\square^{p}$ onto $\square$ will be given. Projection onto $\square^{q}$ becomes an extension of the procedure undertaken for the one dimensional subspace.

Let $F_{1}$ be the line that passes through the origin, onto which data is to be projected. Direction of the line $F_{1}$ is determined by the unit vector $\mathbf{u}_{1}$. Projection is achieved by projecting the $i^{\text {th }}$ individual $x_{i} \in \square^{p}$ onto $\mathbf{u}_{1}$. Hence, the projection point $p_{x_{i}}$ will have the coordinate $p_{x_{i}}=\mathbf{x}_{i}^{T} \frac{\mathbf{u}_{1}}{\square \mathbf{u}_{1} \square}=\mathbf{x}_{i}^{T} \mathbf{u}_{1}$ on $F_{1}$. On the other hand $F_{1}$ has to be located such that $\mathbf{u}_{1} \in \square^{p}$ and $\sum_{i=1}^{n} \square x_{i}-p_{x_{i}} \square^{2}$ is minimum. This is equivalent to maximizing $\sum_{i=1}^{n} \square p_{x_{i}} \square^{2}$, reducing the problem to finding $\mathbf{u}_{1} \in \square^{p}$ such that $\sum_{i=1}^{n} \square p_{x_{i}} \square^{2}$ is maximum subject to the constraint $\square \mathbf{u}_{1} \sqcap=1$. Then projection is written as

$$
\left(\begin{array}{c}
p_{\mathbf{x}_{1}}  \tag{2.4.1}\\
p_{\mathbf{x}_{2}} \\
\vdots \\
p_{\mathbf{x}_{n}}
\end{array}\right)=\left(\begin{array}{c}
\mathbf{x}_{1}^{T} \mathbf{u}_{1} \\
\mathbf{x}_{2}^{T} \mathbf{u}_{1} \\
\vdots \\
\mathbf{x}_{n}^{T} \mathbf{u}_{1}
\end{array}\right)=\mathbf{X} \mathbf{u}_{1}
$$

and can also be expressed in quadratic form as

$$
\begin{equation*}
\max _{\mathbf{u}_{1}^{T} \mathbf{u}_{1}=1} \mathbf{u}_{1}^{T}\left(\mathbf{X}^{T} \mathbf{X}\right) \mathbf{u}_{1} \tag{2.4.2}
\end{equation*}
$$

(Härdle and Simar, 2003). For details see Appendix A.

The vector $\mathbf{u}_{1}$ is the eigenvector of $\mathbf{X}^{T} \mathbf{X}$ corresponding to the largest eigenvalue $\lambda_{1}$ of $\mathbf{X}^{T} \mathbf{X} . \mathbf{u}_{1}$ minimizes, $\sum_{i=1}^{n} \square x_{i}-p_{x_{i}} \square^{2}$. When the data is centered $(\bar{x}=0)$ and $\mathbf{S}=n^{-1} \mathbf{X}^{T} \mathbf{X}$ is the covariance matrix, then the quadratic form (2.4.2) is maximized with respect to $\mathbf{S}$.

Representation of $n$ trajectories on $F_{1}$ are given by $\mathbf{X} \mathbf{u}_{1}$ which is the first factorial variable $\mathbf{z}_{1}=\mathbf{X} \mathbf{u}_{1}$ and $\mathbf{u}_{1}$ is the first factorial axis. Then, $\mathbf{z}_{1}=u_{11} \mathbf{x}_{(1)}+\cdots+u_{p 1} \mathbf{x}_{(p)}$ represents a linear combination of trajectories with coefficients being the elements of $\mathbf{u}_{1}$ (Härdle and Simar, 2003).

Projection of data on $\square^{q}$ from $\square^{p}$ where $q \leq p$, is the extension of the above explained process from $\square^{p}$ to $\square$, except minimization of $\sum_{i=1}^{n} \square x_{i}-p_{x_{i}} \square^{2}$ will produce the best subspace $u_{1}, \ldots, u_{q}$ which are the orthonormal eigenvectors of $\mathbf{X}^{T} \mathbf{X}$. Corresponding eigenvalues of $\mathbf{X}^{T} \mathbf{X}$ are $\lambda_{1} \geq \lambda_{2} \geq \cdots \geq \lambda_{q}$. Then the coordinates of $n$ trajectories on the $k^{t h}$ factorial axis $u_{k}$ are given by $\mathbf{z}_{k}=\mathbf{X} \mathbf{u}_{k} ; k=1, \ldots, q$. The linear combination of the original variables $x_{(1)}, \ldots x_{(p)}$ whose coefficients are given by the $k^{\text {th }}$ vector
$\mathbf{u}_{k}: z_{i k}=\sum_{m=1}^{p} x_{i m} u_{m k}$ forms the factorial variables $\mathbf{z}_{k}=\left(z_{1 k}, \ldots, z_{n k}\right)^{T}$ (Härdle and Simar, 2003).

### 2.4.1.2 Projecting Data onto $\square^{q}$ from $\square^{n}$

The columns of $n \times p$ data matrix $\mathbf{X}$ represents the data as $p$ points in $\square^{n}$. That is each column or variable is a vector $\mathbf{x}_{(j)}=\left(x_{1 j}, \ldots, x_{n j}\right)^{T} \in \square^{n}$. Projecting the $n$-dimensional $p$ data points onto a subspace $\square^{q}(q \leq n)$ is sought. A similar approach used in Section 2.4.1.1 will be followed. Starting with projection onto a one dimensional space or a straight line $G_{1}$ defined by the unit vector $\mathbf{v}_{1} \in \square^{n}$ that will best fit for the $n$ dimensional $p$ points. It means finding $\mathbf{v}_{1}$ such that $\sum_{j=1}^{p} \square p_{x_{(j)}} \square^{2}$ is maximized. In other words the unit vector $\mathbf{v}_{1}$ should maximize, $\left(\mathbf{X}^{T} \mathbf{v}_{1}\right)^{T}\left(\mathbf{X} \mathbf{v}_{1}\right)=\mathbf{v}_{1}^{T}\left(\mathbf{X} \mathbf{X}^{T}\right) \mathbf{v}_{1}$. Then, the coordinates of $p$ variables on $G_{1}$ are given by $w_{1}=\mathbf{X}^{T} \mathbf{v}_{1}$. Hence, $w_{1 j}=v_{11} x_{1 j}+\cdots+v_{1 n} x_{n j} ; j=1, \ldots, p$, can be written (Härdle and Simar, 2003).

Extending to projection onto $\square^{q} ; q \leq n$ means generating a subspace through the orthonormal eigenvectors $\mathbf{v}_{1}, \ldots, \mathbf{v}_{q}$ corresponding to the eigenvalues $\ell_{1} \geq \cdots \geq \ell_{q}$ obtained from $\mathbf{X} \mathbf{X}^{T}$. Coordinates of the $p$ variables on the $k^{\text {th }}$ factorial axis are $\mathbf{w}_{k}=\mathbf{X}^{T} \mathbf{v}_{k} ; k=1, \ldots, p$ and $\mathbf{w}_{k}=\left(w_{k 1}, \ldots, w_{k p}\right)$. Then, $w_{k j}=\sum_{m=1}^{n} v_{k m} x_{m j}$.

### 2.4.1.3 Relationship Between the Projection of Data from $\square^{p}$ onto $\square^{q}$ and $\square^{n}$

onto $\square^{q}$

The $q(q \leq p)$ dimensional (column) subspace onto which data points are projected, is generated by the orthonormal eigenvectors $\mathbf{u}_{1}, \ldots, \mathbf{u}_{q}$ of $\mathbf{X}^{T} \mathbf{X}$. Respective eigenvalues are $\lambda_{1}, \ldots, \lambda_{q}$.

Projection of data from $\square^{n}$ onto $q(q \leq n)$, the row subspace is generated by the orthonormal eigenvectors $\mathbf{v}_{1}, \ldots, \mathbf{v}_{q}$ of $\mathbf{X X}^{T}, \ell_{1}, \ldots, \ell_{q}$ being the respective eigenvalues.

Taking into account the similar logic used in both projections, the eigenvector equations $\left(\mathbf{X}^{T} \mathbf{X}\right) \mathbf{u}_{k}=\lambda_{k} \mathbf{u}_{k}$ in $\square^{p}$ and $\left(\mathbf{X} \mathbf{X}^{T}\right) \mathbf{v}_{k}=\ell_{k} \mathbf{v}_{k}$ in $\square^{n}$ can be written. They satisfy the condition that, for $k \leq r, r$ being the rank of $\mathbf{X}$, the eigenvalues of $\mathbf{X}^{T} \mathbf{X}$ and $\mathbf{X X} \mathbf{X}^{T}$ are the same, and their eigenvectors are related by

$$
\begin{equation*}
\mathbf{u}_{k}=\frac{\mathbf{X}^{T} \mathbf{v}_{k}}{\sqrt{\lambda_{k}}}, \quad \mathbf{v}_{k}=\frac{\mathbf{X} \mathbf{u}_{k}}{\sqrt{\lambda_{k}}} \tag{2.4.3}
\end{equation*}
$$

(Härdle and Simar, 2003).

Let, $\mathbf{U}=\left[\mathbf{u}_{1}, \ldots, \mathbf{u}_{r}\right], \mathbf{V}=\left[\mathbf{v}_{1}, \ldots, \mathbf{v}_{r}\right]$, and $\boldsymbol{\Lambda}=\operatorname{diag}\left[\lambda_{1}, \ldots, \lambda_{r}\right]$, then singular value decomposition of the data matrix is $\mathbf{X}=\mathbf{V} \boldsymbol{\Lambda}^{1 / 2} \mathbf{U}^{T}$. From here, $x_{i j}=\sum_{k=1}^{r} \lambda_{k}^{1 / 2} v_{i k} u_{j k}$.

## Chapter 3

## FUNCTIONAL DATA ANALYSIS

### 3.1 Functional Modelling

In general many processes are continuous in nature, while available data is discrete. The question is how to express discrete observations in functional form for the assessment of the process in question. Processing of discrete data is dealt with in multivariate data analysis (MDA), which is explained in Section 2. A random variable $X$ is a functional variable if it takes values in infinite dimensional space also called functional space.

Observations over $X$ are denoted as $x$. If $T$ is a subset of $\square^{k} ; k=1,2, \ldots$ representing the range of time or space within which the process is taking place, then the random function $X=\{X(t) ; t \in T\}$ and its realizations are $x=\{x(t) ; t \in T\}$. Then the functional data set $x_{1}, \ldots, x_{n}$ is a particular realization of the $n$ functional variables $X_{1}, \ldots, X_{n}$ having the identical distribution as $X$.

In MDA a linear combination of variable values (the $k^{\text {th }}$ principal components) is taken by $z_{i k}=\sum_{m=1}^{p} x_{i m} u_{m k}, u_{m k}$ being the weights applied to the $x_{i m}$ observed value.

Obtaining the functional data from available discrete observations for each trajectory, would mean linear interpolation between successive observations and smoothing of trajectories. Hence each trajectory can be denoted by $X_{1}(t), \ldots, X_{n}(t), t$ representing the time or space coordinate of a trajectory. The discrete index used in MDA is replaced by the continuous index $t$ in functional PCA. Then, the principal component score in functional data becomes,

$$
\begin{equation*}
\xi_{i}=\int u x_{i}=\int u(t) x_{i}(t) d t \tag{3.1.1}
\end{equation*}
$$

Estimating the mean function from $n$ trajectories is possible. The obtained mean function $\bar{X}(t)$ should be smoothed to avoid undesirable fluctuations in $\bar{X}(t)$ stemming from noisy data.

The functional principal components weight function $u(t)$ is defined for each component over the range of $t$ such that $\int u(t)^{2} d t=1$. Then, the principal component scores $\xi_{i}$ in (3.1.1) for the sample data is given by $\xi_{i}=\int u(t) X_{i}(t) d t$.

In the first functional principal component the aim is to determine $u_{1}(t)$ such that it maximizes the variance $\operatorname{Var}\left(\xi_{i}\right)=n^{-1} \sum_{i} \xi_{i 1}^{2}=n^{-1} \sum_{i}\left(\int_{t} u_{1}(t) x_{i} d t\right)^{2}$ under the constraint $\int u(t)^{2} d t=1$. Second and higher order principal components can be defined in the same way except, they have to satisfy the additional mutual orthogonality property. That is,

$$
\int u_{j}(t) u_{1}(t) d t=\int u_{j}(t) u_{2}(t) d t=\cdots=\int u_{j}(t) u_{j-1}(t) d t=0
$$

(Ramsay and Silverman, 2002).

An optimal empirical orthonormal basis is needed for application purpose. That is finding $K$ orthonormal functions $u_{m}$, enabling the expansion of each curve or trajectory in terms of these basis functions, such that the trajectory is approximated as accurate as possible. As a result seeking an expansion of the form,

$$
\begin{equation*}
\hat{x}_{i}(t)=\sum_{k=1}^{K} \xi_{i k} u_{k}(t) \text { where } \xi_{i k}=\int x_{i}(t) u_{k}(t) d t \tag{3.1.2}
\end{equation*}
$$

is necessary.

The fitting criterion

$$
\begin{equation*}
\square x_{i}-\hat{x}_{i} \square^{2}=\int[x(t)-\hat{x}(t)]^{2} d t \tag{3.1.3}
\end{equation*}
$$

is used.

Then, from equation (3.1.3) the sum of the squares of errors for PCA, $S S E_{P C A}=\sum_{i=1}^{n} \square x_{i}-\hat{x}_{i} \square^{2}, \quad$ can be used as a measure of approximation. The basis that minimizes $S S E_{P C A}$ corresponds to the same set of PC weight functions that maximize variance components.

### 3.1.1. Brief Comparison of MDA and FDA

Given a data matrix $\mathbf{X}_{n \times p}$, its covariance matrix $\mathbf{S}_{p \times p}=n^{-1} \mathbf{X}^{T} \mathbf{X}$ or as in equation (2.1.4) with eigenvalues, $\lambda_{1} \geq \lambda_{2} \geq \cdots \geq \lambda_{q}$ and corresponding orthogonal eigenvectors $\mathbf{u}_{1}, \ldots, \mathbf{u}_{q}$, are computed. Projection of data onto column subspace is subject to the constraint, $\max _{\mathbf{u} \mathbf{u}=1}\left(\mathbf{u}^{T} \mathbf{S u}\right)$. Solution of the eigen-equation $\mathbf{S u}=\boldsymbol{\lambda} \mathbf{u}$ gives the eigenvalues in descending order, $\lambda_{1} \geq \lambda_{2} \geq \cdots \geq \lambda_{q}$. Due to centering of $\mathbf{X}$, its rank is at most, $n-1$, meaning that the $\mathbf{S}_{p \times p}$ covariance matrix will have $\min \{p, n-1\}$ nonzero eigenvalues. In functional PCA, the covariance function $s\left(t, t^{+}\right)$can be written for $n$ available data as, $s\left(t, t^{+}\right)=n^{-1} \sum_{i=1}^{n} x_{i}(t) x_{i}\left(t^{+}\right)$.

Finding the principal components weights $u_{j}(t)$ requires the following concepts (Ramsay and Silverman, 2006, Appendix A.5).

1. In a general inner product space, the symmetric matrix is replaced by a selfadjoint linear operator, $A$, which satisfies $\langle x, A y\rangle=\langle A x, y\rangle, \forall x, y$.
2. $A$ is compact (completely continuous) symmetric transformation on Hilbert space.

Maximizing the inner product space $\langle x, A y\rangle$ subject to the constraint $\square x \square=1$ is similar to maximizing $\mathbf{x}^{T} \mathbf{A x}$ subject to $\mathbf{x}^{T} \mathbf{x}=1$ in the finite dimensional space.

Here the sequence $u_{j}$ is defined as the solutions to the set of optimization problems

$$
\begin{equation*}
\max \langle x, A x\rangle \text {, subject to } \square x \square=1 \text { and }\left\langle x, u_{i}\right\rangle=0 \text { for } i<j . \tag{3.1.4}
\end{equation*}
$$

The solution is obtained under the given conditions by assessing the eigenfunction problem $A u=\lambda u$ and normalizing the eigenfunctions $u$ to satisfy, $\square u \square=1$. Then, the first eigenfunction $u_{1}$ solves the optimization problem given in (3.1.4) resulting in a maximum value equal to $\lambda_{1}$. Subsequent eigenfunctions $u_{j}$ solve the constrained problem given by (3.1.4). Then, maximum at the $j^{\text {th }}$ stage is, $\left\langle u_{j}, A u_{j}\right\rangle=\lambda_{j} \square u_{j} \square^{2}=\lambda_{j}$.

Each of the principal components weight functions $u_{j}(t)$ should satisfy

$$
\begin{equation*}
\int s\left(t, t^{+}\right) u\left(t^{+}\right) d t^{+}=\lambda u(t) \tag{3.1.5}
\end{equation*}
$$

for a certain eigenvalue $\lambda$. Left hand side of equation (3.1.5) is the integral transform of the weight function $u$ by the covariance function $S$ given by $S u=\int s\left(., t^{+}\right) u\left(t^{+}\right) d t^{+}$. This is called the covariance operator, $S$. Hence, $S u=\lambda u$, can be written.

In MDA there are $p$ eigenvalue-eigenvector pairs, whereas in FDA it becomes the number of function values which are infinitely many.

### 3.1.2. Smoothing in Functional PCA

Data matrix described in MDA, rows are the subjects or trajectories, representing one realization of the random process governed by $p$ random variables. In FDA a trajectory represents a random function in continuum. The continuous environment $T$ is a bounded time or space interval over which the domain of the random process $X($.$) lies.$

The weight functions $u(t)$ used in the computation of principal components needs to be smoothed by controlling their roughness. This results in the smoothing of the principal components. For the first PC the function $u_{1}(t)$ maximizes the variance of the principal component scores subject to

$$
\begin{equation*}
\int\{u(t)\}^{2} d t+\alpha \int\left\{u^{\prime \prime}(t)\right\}^{2} d t=1 ; \alpha \geq 0 \tag{3.1.6}
\end{equation*}
$$

$\alpha$ in (3.1.6) is a control factor over the amount of smoothing required. For smoothing the second and higher order principal component scores in addition to constraint (3.1.6), the constraint

$$
\int u_{i}(t) u_{j}(t) d t+\alpha \int u_{i}^{\prime \prime}(t) u_{j}^{\prime \prime}(t) d t=0 ; i \neq j
$$

is required.

Smoothing the sample mean function has a major importance in FDA. A functional data set $X_{1}(t), \ldots, X_{n}(t)$ can be expanded in terms of basis functions, $f_{1}(t), \ldots, f_{m}(t)$. A
coefficient matrix $\mathbf{A}_{n \times m}$ can be defined such that, $X_{i}(t)=\sum_{j=1}^{m} a_{i j} f_{j}(t)$. Then the smoothed sample mean becomes,

$$
\bar{X}(t)=\sum_{j=1}^{m} \bar{a}_{j} f_{j}(t), \text { if } \bar{a}_{j}=n^{-1} \sum_{i} a_{i j} .
$$

### 3.1.3 Storing Functional Data/Observations

Taking a basis to mean a standard set of functions $\left(f_{1}(t), \ldots, f_{m}(t)\right)$ such that, any function of interest can be expanded in terms of $f_{j}(t)$. Then, a functional datum $x(t)$ can be expressed in terms of $m$ basis functions and the principal component weight ( $u$ ) as $x(t)=\sum_{j=1}^{m} u_{j} f_{j}(t)$.

In FDA basis functions are designed to represent the nature of the process under study. This is achieved by fitting basis coefficients to the observed data. Given values $x_{1}, \ldots, x_{n}$ observed at locations $t_{1}, \ldots, t_{n}$, basis functions $f_{j}(t)$ can be represented by the matrix $\mathbf{B}=f_{j}\left(t_{i}\right), i=1, \ldots, n, j=1, \ldots, m$.

If $\mathbf{u}$ is the vector of coefficients, then the vector of values at the observed locations will be Bu. When the number of basis functions are at most the same as the observation locations ( $m \leq n$ ), then the basis functions can be fit by minimizing the sum of squares of deviations, $\left[x_{i}-\sum_{j} u_{j} f_{j}(t)\right]^{2}$.

For $m=n$ the expansion $x(t)=\sum_{j} u_{j} f_{j}(t)$ gives an exact interpolation of the $x_{i}$ values.

For $m<n$ the expansion is the smooth version of the initial data.

If $m>n$, more basis functions than observed locations, then, a choice of $u$ values gives exact interpolation of the $x_{i}$ values. That is, $x_{k}(t)=\sum_{j=1}^{m} u_{j} f_{j}\left(t_{k}\right), k=1, \ldots, n$.

Then, the interpolation that includes the parameters that minimizes the roughness of the curve is selected.

### 3.2 Expressing the Random Process in Terms of Global Mean and

## Covariance

Let $\left\{X_{t}\right\}_{t \in[a, b]}$ be a random process in the interval $[a, b]$. Then the random trajectories $X$ in $L^{2}(I)$ assumed to have mean function $\mu(t)=E(X(t))$ and covariance function $G\left(t, t^{+}\right)=\operatorname{cov}\left(X(t), X\left(t^{+}\right)\right)$, where $t, t^{+} \in I$. It is assumed that the operators on covariance have a sequence of orthonormal eigenfunctions $\phi_{k}$ with eigenvalues $\lambda_{1} \geq \lambda_{2} \geq \cdots$.

The Hilbert- Schmidt kernel, $G$ is very useful in the expression of a random process in terms of global mean and covariance and is given as

$$
\begin{equation*}
G\left(t, t^{+}\right)=\sum_{k=1}^{\infty} \lambda_{k} \phi_{k}(t) \phi_{k}\left(t^{+}\right), \mathrm{t}, \mathrm{t}^{+} \in T . \tag{3.1.7}
\end{equation*}
$$

This is the orthogonal expansion of the covariance in $L^{2}$ in terms of eigenfunctions $\phi_{k}$ and corresponding eigenvalues $\lambda_{k}, k=1,2, \ldots$ and $\lambda_{1} \geq \lambda_{2} \geq \ldots$. Karhunen-Loéve Expansion and, Mercer's theorems guarantee the expansion given in equation (3.1.7) and its spectral decomposition. Details of the Karhunen-Loéve theorems are in Apendix C.

In classical functional PCA, for the process, $\left\{X_{t}\right\}_{t \in[a, b]}$, the random curve (trajectory) $X(t)$ can be expressed as,

$$
\begin{equation*}
X(t)=\mu(t)+\sum_{k=1}^{\infty} \phi_{k}(t) \xi_{k}, \mathrm{t} \in \mathrm{~T} . \tag{3.1.8}
\end{equation*}
$$

Here, $\xi_{k}$ are uncorrelated random variables with, $E\left(\xi_{k}\right)=0$ and $E\left(\xi_{k}^{2}\right)=\lambda_{k}, \sum_{k} \lambda_{k}<\infty$. For any finite $K$, (3.1.8) can be written as,

$$
\begin{equation*}
X(t)=\mu(t)+\sum_{k=1}^{K} \phi_{k}(t) \xi_{k} . \tag{3.1.9}
\end{equation*}
$$

$K$, determines the fraction of variance, $F(K)=\sum_{i=1}^{K} \lambda_{i} / \sum_{k=1}^{\infty} \lambda_{k}$, in the process under consideration and is required to be as high as possible (preferably above 0.8 ).

### 3.3 The Principal Components Analysis Through Conditional Expectation

Functional principal component scores, $\xi_{i k}$, play a major role in the estimation of a trajectory. They are uncorrelated random variables with mean zero and variances being the eigenvalues of covariance matrix $\mathbf{G}$. The functional principal components scores given directly by equation (3.1.1), and it works well when the density of the grid of measurements for each subject is sufficiently large.

The functional representation of a trajectory, $X(\cdot)$, for the $j^{\text {th }}$ observation of the $i^{\text {th }}$ subject made at time $t_{i j}, i=1, \ldots, n, j=1, \ldots, p_{i}$ is represented by $Y_{i j}$. Number of observations, $p_{i}$, made on each of the $i^{\text {th }}$ subjects are assumed to be i.i.d. random variables. If no error is involved, then, $Y_{i j}=X\left(t_{i j}\right)$. However, observations will inherently include some measurement errors $\varepsilon_{i j}$ that are also assumed to be i.i.d. with $E\left(\varepsilon_{i j}\right)=0$ and constant variance $\sigma^{2}$. Then, the model representing the $i^{\text {th }}$ subject based on observations can be written as

$$
\begin{equation*}
Y_{i j}=\mu\left(t_{i j}\right)+\sum_{k=1}^{\infty} \xi_{i k} \phi_{k}\left(t_{i j}\right)+\varepsilon_{i j}, t_{i j} \in T . \tag{3.1.10}
\end{equation*}
$$

Thus, instead of the integral, the estimated principal components scores are given by $\xi_{i k}=\sum_{j=1}^{p_{i}}\left(Y_{i j}-\mu\left(t_{i j}\right)\right) \phi_{k}\left(t_{i j}\right)\left(t_{i j}-t_{i j-1}\right)$. However, this does not work well when the data are
sparse, since substituting $Y_{i j}$ for $X_{i}\left(t_{i j}\right)$ causes biased FPC scores. Therefore, an alternative method, the Principal Components Analysis through Conditional Expectation (PACE) is proposed by Yao, et. al. (2005). The PACE method, assumes that the principal component scores $\xi_{i k}$ and error $\varepsilon_{i k}$ are jointly Gaussian. Let $X_{i}=\left(X_{i}\left(t_{i 1}\right), \ldots, X_{i}\left(t_{i p_{i}}\right)\right)^{T}, \quad Y_{i}=\left(Y_{i}\left(t_{i 1}\right), \ldots, Y_{i}\left(t_{i p_{i}}\right)\right)^{T}, \quad \mu_{i}=\left(\mu\left(t_{i 1}\right), \ldots, \mu\left(t_{i p_{i}}\right)\right)^{T}, \quad$ and $\phi_{i k}^{T}=\left(\phi_{k}\left(t_{i 1}\right), \ldots \phi_{k}\left(t_{i p_{i}}\right)\right)$. Thus, under Gaussian assumptions, the best prediction of the PC scores can be found by conditional expectation.

$$
\begin{equation*}
\xi_{i k}=E\left[\xi_{i k} \mid Y_{i}\right]=\lambda_{k} \phi_{i k}^{T} \Sigma_{Y_{i}}^{-1}\left(Y_{i}-\mu_{i}\right), \tag{3.1.11}
\end{equation*}
$$

where $\Sigma_{Y_{i}}=\operatorname{cov}\left(Y_{i}, Y_{i}\right)=\operatorname{cov}\left(X_{i}, X_{i}\right)+\sigma^{2} \mathbf{I}_{p_{i}}=G\left(T_{i j}, T_{i l}\right)+\sigma^{2} \delta_{j l}$.

Substituting the estimates of $\lambda_{k}, \phi_{i k}, \Sigma_{Y_{i}}$ and $\mu_{i}$ in equation (3.1.11), leading to

$$
\begin{equation*}
\xi_{i k}=E\left[\xi_{i k} \mid Y_{i}\right]=\lambda_{k} \phi_{i k}{ }^{T} \Sigma_{Y_{i}}^{-1}\left(Y_{i}-\mu_{i}\right), \tag{3.1.12}
\end{equation*}
$$

where, $\Sigma_{Y_{i}}$ is obtained from the whole data. Then, the infinite-dimensional processes are approximated by the projection on the functional space spanned by the first $K$ eigenfunctions of estimated $\sum_{Y_{i}}$ covariance matrix. Thus, in practice the estimated $i^{\text {th }}$ trajectory is

$$
\begin{equation*}
X_{i}(t)=\mu(t)+\sum_{k=1}^{K} \xi_{i k} \phi_{k}(t) \tag{3.1.13}
\end{equation*}
$$

The conditional method (3.1.12) under the Gaussian assumptions works well in both case of sparse and dense data and yields the best estimates. It is also worth mentioning that (3.1.11) is the best linear prediction of principal components scores $\xi_{i k}$ and works well weather the Gaussian assumption holds or not (Yao, et. al. , 2005). See Yao, et.al., (2003) for another estimation for functional principal component scores.

## Chapter 4

## SMOOTHING

### 4.1 Smoothing The Mean

In a random process $X(\cdot)$ the underlying function $f$ is generally unknown. Available data collected or observations made at $p$ different time or space locations, $Y_{i}\left(t_{1}\right), Y_{i}\left(t_{2}\right), \cdots, Y_{i}\left(t_{p}\right)$ give some idea about the likely behaviour of the function. Using available observations to predict the underlying random function representing the random process is not an easy task, since data is mostly contaminated by errors due to various agents. If there is sufficient evidence to indicate that data is error free, then some simple linear interpolation may be adequate to represent $X(\cdot)$ with available observations. However, in the presence of error $\varepsilon$ smoothing will be required to represent $X(\cdot)$. Here, $\varepsilon_{i}$ is i.i.d. with $E\left(\varepsilon_{i}\right)=0$ and $\operatorname{var}\left(\varepsilon_{i}\right)=\sigma^{2}$.

Let $Y_{i}=X\left(t_{i}\right)+\varepsilon_{i}, i=1, \ldots, n$ be the noisy representation of $X(\cdot)$. If $\mathbf{Y}$ is the matrix of observations, then $\operatorname{var}(\mathbf{Y})=\boldsymbol{\Sigma}_{e}=\sigma^{2} \mathbf{I}$.

There are many different smoothing techniques used in various application fields. Some of the important ones are introduced in the following section.

### 4.1.1. Commonly Used Smoothing Methods

A simple smoothing will be the fitting of ordinary least squares function to the data defined by the basis function expansion,

$$
X\left(t_{j}\right)=\sum_{k}^{K} b_{k} \phi_{k}\left(t_{j}\right)=\mathbf{b}^{T} \boldsymbol{\Phi} .
$$

$\left\{\phi\left(t_{j}\right)\right\}_{k}$ are the basis functions and the coefficients vector $\mathbf{b}_{1 \times K}$ is determined by minimizing the sum of square errors (SSE),

$$
\operatorname{SSE}(Y \mid b)=\sum_{j=1}^{n}\left[Y_{j}-\sum_{k}^{K} b_{k} \phi_{k}\left(t_{j}\right)\right]^{2}=(\mathbf{Y}-\boldsymbol{\Phi b})^{T}(\mathbf{Y}-\boldsymbol{\Phi} \mathbf{b})=\|\mathbf{Y}-\boldsymbol{\Phi b}\|^{2} .
$$

Then, the estimated (fitted) vector $\hat{\mathbf{y}}$ is found by

$$
\hat{\mathbf{y}}=\boldsymbol{\Phi}\left(\boldsymbol{\Phi}^{T} \boldsymbol{\Phi}\right)^{-1} \boldsymbol{\Phi}^{T} \mathbf{Y} .
$$

This assumes equal weight assignment to all observations, regardless of observation time/space $t$. In the case of regular grid this may be acceptable, but for irregularly observed data a weighing method is necessary. Then, the weighted least squares (WLS) fit offers a solution given by

$$
\operatorname{SSE}(Y \mid b)=(\mathbf{Y}-\mathbf{\Phi} \mathbf{b})^{T} \mathbf{W}(\mathbf{Y}-\mathbf{\Phi} \mathbf{b})
$$

Weights $W_{j}$ to be assigned can be computed using different methods. Kernel functions, splines, moving averages are just a few that can be used. In the case of kernel functions, some of widely used ones are

Uniform: $W_{j}(t)=K\left(\frac{t_{s}-t_{p}}{h}\right)= \begin{cases}0.5 & \text { for }\left|\frac{t_{s}-t_{p}}{h}\right| \leq 1 . \\ 0, & \text { otherwise }\end{cases}$
Quadratic: $W_{j}(t)=K\left(\frac{t_{s}-t_{p}}{h}\right)= \begin{cases}0.75\left(1-\left(\frac{t_{s}-t_{p}}{h}\right)^{2}\right) & \text { for }\left|\frac{t_{s}-t_{p}}{h}\right| \leq 1 . \\ 0, & \text { otherwise }\end{cases}$
Gaussian: $W_{j}(t)=K\left(\frac{t_{s}-t_{p}}{h}\right)=\frac{e^{-\frac{1}{2}\left(\frac{t_{s}-t_{p}}{h}\right)^{2}}}{\sqrt{2 \pi}}$.

Nadaraya-Watson: $W_{j}(t)==\frac{K\left(\frac{t_{s}-t_{p}}{h}\right)}{\sum_{s} K\left(\frac{t_{s}-t_{p}}{h}\right)}$.

Widely used Epanechnikov kernel function takes the form $K(x)=0.75\left(1-x^{2}\right) \mathbf{1}_{[-1,1]}(x)$ is the univariate case and $K(x, y)=0.5625\left(1-x^{2}\right)\left(1-y^{2}\right) \mathbf{1}_{[-1,1]}(x) \mathbf{1}_{[-1,1]}(y)$ is the bivariate case with $\mathbf{1}_{A}(x)=1$ if $x \in A$ and 0 otherwise for any set $A$.

Smoothing the mean $\mu(t)$ and covariance $\operatorname{Cov}\left(X(t), X\left(t^{+}\right)\right)$of the set of observed curves is necessary in many applications (Rice and Silverman, 1991).

Using the spline smoothing to smooth the mean is to use the penalized least squares. Given the $i^{\text {th }}$ data vector $\mathbf{x}_{i}=\left(x_{i 1}, \ldots, x_{i p}\right)^{T}$, the estimated mean curve $\hat{\mu}$ should minimize

$$
\begin{equation*}
n^{-1} \sum_{i} \square \mathbf{x}_{i}-\mu \square^{2}+\alpha \int \mu^{\prime \prime}(t)^{2} d t \tag{4.1.1}
\end{equation*}
$$

Here, $\alpha$ is a positive smoothing parameter and the integral represents the roughness of $\mu$.(4.1.1) is equivalent to $\sum_{j}\left[\bar{X}_{j}-\mu\left(t_{j}\right)\right]^{2}+\alpha \int \mu^{\prime \prime}(t)^{2} d t$, which is the spline smoothing applied to pointwise averages. Choice of the smoothing parameter $\alpha$ is subjective, but methods such as cross validation (CV) are available to automate the choice of optimum $\alpha$ value (Heckman, 1986).

Similarly various moving averages techniques can also be used as the smoother of the mean and covariance functions. The stretches interpolated moving average technique developed during the course of this study and given in Section 4.3 is successfully applied to the smoothing of the mean. It is shown to be more robust than the local linear smoothing when the correlation between trajectories are weak.

Local Linear Regression is also used as a smoother and some details of this method are given in Section 4.3.3.

### 4.1.2. Kernel Smoothing

One other method used for smoothing the mean is the scatter plot smoothing utilizing kernel smoothers. Let $Y_{i j}(t), i=1, \ldots, n$ and $j=1, \ldots, p_{i}$, be the $j^{\text {th }}$ observation on the $i^{\text {th }}$ subject made at time $t$. If the model has no additional error then shortly $Y_{i j}=X\left(t_{i j}\right)$ can be written.

The local linear scatter plot smoother at time $t$ for $\mu(t)$ is obtained from the scatter plot $\left(t_{i j}, Y_{i j}\right)$ by minimizing

$$
\begin{equation*}
\sum_{i=1}^{n} \sum_{j=1}^{p_{i}} \kappa\left(\frac{t_{i j}-t}{h_{\mu}}\right)\left\{Y_{i j}-\beta_{0}-\beta_{1}\left(t-t_{i j}\right)\right\}^{2} \tag{4.1.2}
\end{equation*}
$$

with respect to $\beta_{0}$ and $\beta_{1}$, where $h_{\mu}$ is the bandwidth and univariate density $\kappa$ is the kernel function. Then the estimate of $\mu(t)$ is $\hat{\mu}(t)=\hat{\beta}_{0}(t)$. The minimization can be done by taking the derivative of (4.1.2) with respect to $\beta_{0}$ and $\beta_{1}$. Solution of the obtained equations yields the local linear estimator (smoothed) $\hat{\mu}(t)$

$$
\begin{equation*}
\hat{\mu}(t)=\hat{\beta}_{0}(t)=\frac{\sum_{i} \frac{1}{E p} \sum_{j} w_{i j} Y_{i j}}{\sum_{i} \frac{1}{E p} \sum_{j} w_{i j}}-\frac{\sum_{i} \frac{1}{E p} \sum_{j} w_{i j}\left(t_{i j}-t\right)}{\sum_{i} \frac{1}{E p} \sum_{j} w_{i j}} \hat{\beta}_{1}(t), \tag{4.1.3}
\end{equation*}
$$

where,

$$
\hat{\beta}_{1}(t)=\frac{\sum_{i} \frac{1}{E_{p}} \sum_{j} w_{i j}\left(t_{i j}-t\right) Y_{i j}-\frac{\sum_{i} \frac{1}{E p} \sum_{j} w_{i j}\left(t_{i j}-t\right) \sum_{i} \frac{1}{E_{p}} \sum_{j} w_{i j} Y_{i j}}{\sum_{i} \frac{1}{E_{p}} \sum_{j} w_{i j}}}{\sum_{i} \frac{1}{E_{p}} \sum_{j} w_{i j}\left(t_{i j}-t\right)^{2}-\frac{\left(\sum_{i} \frac{1}{E p} \sum_{j} w_{i j}\left(t_{i j}-t\right)\right)^{2}}{\sum_{i} \frac{1}{E_{p}} \sum_{j} w_{i j}}} .
$$

Here, $E p=\frac{\sum p_{i}}{n}$ ( $p_{i}:$ Number of observations on the $i^{\text {th }}$ trajectory $)$ and $w_{i j}=\kappa\left(\frac{t_{i j}-t}{h_{\mu}}\right) / n h_{\mu}$ is the weight function, $\kappa(x)=0.75\left(1-x^{2}\right) \mathbf{1}_{[-1,1]}(x)$ is the uni-variate Epanechnikov kernel function with $\mathbf{1}_{A}(x)=1$ if $x \in A$ and 0 otherwise for any set $A$ where $t$ is the starting data value of bandwidth $h_{\mu}$ (Yao et. al. 2005). Alternative formulas for (4.1.3) can be found in Hall et.al. (2006), and in Müller (2005).

Choi and Hall (1998) give interesting ideas about bias reduction in local linear smoothing depending on the kernel function.

### 4.2 Moving Average Approach to Smoothing

Moving average is a well-known technique in a broad range of disciplines and initially used for trend generation in mainly time dependent variables. Various versions of moving averages are in implementation, depending on the nature of the process under consideration. Simple, cumulative, weighted, exponential moving average techniques are some of the commonly used ones for trend generation.

Cumulative moving average is used when the average starting from a given time point up to the present is required for decision making. Weighted moving average is mostly used in cases where higher weights are to be assigned for the latest or new data. Different weight determination techniques are employed. Widely used weighing techniques are linearly or exponentially decreasing, by assigning the high weights to
most recent data and lower weights as data becomes old (Durbin, 1959). In finance, economy or in some medical applications weighted moving average is used.

In principle moving average is taken to be a simple smoothing method, but its examination in detail indicates that sophisticated moving average methods can be developed. In part of the research devoted to this thesis, some new ideas are developed, and due to the process involved, it is named as "Stretched Interpolated Moving Average" (SIMA). Obtained smoothing results are compared with those of Local Linear Smoothing (LLR) and PACE results, to verify the validity of SIMA. First step is to locate the computed moving averages at equal distance over the range of the data. Hence, named as, stretched moving average, details of which are given in the following section.

### 4.3. Stretched Interpolated Moving Average (SIMA)

### 4.3.1. Stretched Moving Average

The method developed handles the moving average process when data are available on a regular grid basis in terms of time or space coordinates. That means, each datum is located at equal time or space interval from its immediate neighboring data values. Considering the one dimensional case, when $p$ data values are regularly spaced on a trajectory, $m$ data points within a fixed lag interval $h$ are used for averaging. The number of averaged values will be, $M=p-m+1$. Process starts with the first or oldest data averaging the first $m$ data values (Kenny and Durbin, 1982). Let obtained average values be denoted by the random variable $Y$. Given $p_{i}$ observations located at equal
distance on the $i^{\text {th }}$ trajectory, the moving average for each lag with $m$ data values will be

$$
\begin{equation*}
Y_{i l}=\frac{1}{m} \sum_{j=l}^{m+l-1} X_{i j}, 2 \leq m \leq p_{i}, i=1, \ldots, n, l=1, \ldots, M . \tag{4.3.1}
\end{equation*}
$$

Every averaged value has to be assigned to a coordinate within the lag interval it belongs to. Assuming every averaged value $Y$ is assigned to the first point's coordinate of each lag, the final average obtained from the last or $M^{\text {th }}$ lag will fall $h$ units distance behind the present or newest datum. Alternately, averaged values can be assigned to the midpoint or last point of each lag. Assignment of the upper lag boundary would mean ignoring a time or space interval equivalent to $h$ from the beginning point of the data values. The idea proposed to overcome this handicap, is to assign the coordinate of the first datum to the first average value, and the coordinate of the last datum to the last ( $M^{\text {th }}$ ) average value. Remaining $M-2$ moving average values will be equally spaced over the span of data.

Mean and variance of the moving average values for the $i^{\text {th }}$ trajectory is

$$
\begin{aligned}
& E\left(Y_{i}\right)=\frac{1}{M} \sum_{l=1}^{M} Y_{i l}=\frac{1}{m M} \sum_{l=1}^{M} \sum_{j=l}^{m+l-1} X_{i j} \\
& \sigma_{Y_{i}}^{2}=\frac{1}{M m^{2}} \sum_{l=1}^{M}\left(\sum_{j=l}^{m+l-1} X_{i j}\right)^{2}-\left(\frac{1}{M m} \sum_{l=1}^{M} \sum_{j=l}^{m+l-1} X_{i j}\right)^{2} .
\end{aligned}
$$

Relationship between the moving average values $Y_{i l}$, with data values $X_{i l}$ can be found as follows (Mentz, 1975).

Consider the case, $m=2$, then

$$
Y_{i l}=\frac{1}{m}\left(X_{i l}+X_{i l+1}\right) .
$$

Via mathematical induction, $X_{i l+1}=m Y_{i l}-X_{i l}$, follows with,

$$
\begin{aligned}
X_{i l+1} & =m Y_{i l}-m Y_{i l-1}+X_{i l-1} \\
& =m Y_{i l}-m Y_{i l-1}+m Y_{i l-2}-X_{i l-2} \\
& =m Y_{i l}-m Y_{i l-1}+m Y_{i l-2}-m Y_{i l-3}+X_{i l-3} \\
& =\vdots \\
& =m \sum_{j=1}^{l}(-1)^{j+l} Y_{i j}+(-1)^{l} X_{i 1} .
\end{aligned}
$$

For the formulation of the Stretched Moving Average, let random variable $S$ be the time or space interval covered by $p_{i}$ data values on the $i^{\text {th }}$ trajectory. It is required to locate the averaged values equally spaced on the same interval $S$. In other words, $p_{i}$ data values $X_{i j}, j=1, \ldots, p_{i}$ on the $i^{\text {th }}$ trajectory covers an interval $S$, the same interval will also be covered by $M_{i}$ moving average values. Without loss of generality the data values can be assumed to be uniformly distributed on a trajectory with equal interval between points.

Then, distance between data values is $d=S /(p-1)$ and the distance between the moving average points will be $s=S /(M-1)$. Note that, $d \leq s$. Alternately if $p$ data values are uniformly located on $S$ with unit interval in between $(d=1)$, then $s=(p-1) /(M-1)$.

Moving average values will be located at $s$ distance apart such that the first one $Y_{i 1}$ will correspond to the location of $X_{i 1}$, and last one $Y_{i M}$ corresponding to the location of $X_{i p}$. Other moving average values ranging from $Y_{i 2}$ to $Y_{i M-1}$ will occupy locations on $S$ accordingly at equal distances. Steps followed in the assignment of coordinates to compute moving average values, is named as the Stretched Moving Average.

Clearly as the lag interval becomes larger, in simple moving average the computed final $Y_{i l}$ value will lag farther behind the latest data values. Magnitude of this distance is $(m-1) d$. Stretched moving average eliminates the handicap by assigning the computed $Y_{i l}$ values uniformly over the full interval $S$ covered by the trajectory.

Let $t \in S, s=\frac{s_{M}-s_{1}}{M-1}=\frac{S}{M-1}$, where $s_{1}=\min (t)$ and $s_{M}=\max (t)$, then the stretched moving average can be written as

$$
Y_{i l}\left(s_{i l}\right)=\frac{1}{m} \sum_{j=l}^{m+l-1} X_{i j}(t), l=1, \ldots, M
$$

(Tandoğdu and İyikal, 2013).

To highlight the stretched moving average concept, a hypothetical data set consisting of $p=10$ data values $X_{i j}, j=1, \ldots, 10$ given at unit interval apart over the $i^{\text {th }}$ trajectory is used. Cumulative distances from the first datum towards the last one will be, $t_{1}=0, t_{2}=1, \ldots, t_{10}=9$. This means, data is spread over $S=9$ units of time or space interval. Hence, data can be represented by $X_{i j}\left(t_{j}\right)$. If a lag interval of $h=3$ units is
used, it will include $m=4$ data values for the averaging process. Then, $s=(p-1) /(M-1)=9 / 6=1.5$ units. First moving average value $Y_{i 1}\left(s_{1}\right)$ will be assigned to the same location as, $t_{1}$, i.e. $s_{1}=t_{1}$. Subsequent moving average values will be $Y_{i l}\left(s_{l}\right)$. The graph showing the raw data, simple moving average and stretched moving average is as shown in Figure 4.1.


Figure 4.1: Raw, Simple Moving and Stretched Moving Averages.

The simple moving average graph lags behind the latest raw data points $X_{i j}$ by $m-1$ unit distance. Stretching the locations of moving average values to cover the whole space $S$, is the first step towards the computation of the error involved in smoothing committed by the moving average process. Accurate computation of this error would require a one to one correspondence between the data points and the moving average values. However, the number of moving average values is less than the number of data points on a trajectory by $m-1$. A linear interpolation is introduced to equalize the number of moving average values with the number of data values on a trajectory.

Further, the normality of a trajectory, leads to the conclusion that the stretched moving average trajectory is also normal as explained below.

It is known that the random process $X_{i}$ has a p-variate normal distribution if and only if $\mathbf{a}^{T} X_{i}$ is univariate normal for all fixed $p$ vectors $\mathbf{a}$. Then a linear combination $\mathbf{y}=\mathbf{A} X_{i}+\mathbf{c}$ of $X_{i}$ has a q-variate normal distribution, where $\mathbf{A}$ is a $q \times p$ matrix of coefficients and $\mathbf{c}$ is a $q \times 1$ vector of constants (Mardia, et.al., 1979). In place of $\mathbf{y}$ the stretched moving average $Y_{i}(t)$ can be written, leading to the fact expressed in the following theorem.

Theorem 4.3.1: If the $i^{\text {th }}$ trajectory $X_{i}(t)$ of the random process has p-variate normal distribution, then the corresponding stretched moving average trajectory $Y_{i}(t)$ has an (p$\mathrm{m}+1$ )-variate normal distribution.

Proof of Theorem 4.3.1 is given in the Appendix B.

### 4.3.2. Linear Interpolation of the Stretched Moving Averages

Since the idea is to use the moving average as a smoother, the error committed in the smoothing process should be measurable. This measure is usually expressed in some form of the difference between the observations and the corresponding moving average values. However, as number of observations on a trajectory is always greater than the number of computed moving average values $\left(p_{i}>M_{i}\right)$, it is deemed necessary to compute a moving average value corresponding to each observation. This is possible by
linear interpolation of the stretched moving average values. Method followed to obtain these averages is named as "Stretched Interpolated Moving Average" (SIMA). Obtained new averages for the $i^{t h}$ trajectory are denoted by $Y_{i}^{*}$ and given by

$$
\begin{equation*}
Y_{i j}^{*}(t)=\left[Y_{i j}\left(s_{j}\right)-Y_{i j-1}\left(s_{j-1}\right)\right] w_{j}+Y_{i j-1}\left(s_{j-1}\right) \tag{4.3.2}
\end{equation*}
$$

where, $w_{j}=\frac{t-s_{j-1}}{s_{j}-s_{j-1}}$. In (4.3.2), the first stretched interpolated moving average value $Y_{i 1}{ }^{*}$ is equal to the first moving average value $Y_{i 1}$, the last stretched interpolated moving average value $Y_{i p}{ }^{*}$ is equal to the last moving average value $Y_{i p}$ (Tandoğdu and İyikal, 2013).

Sample covariance function's limit distribution when the variance is finite in the moving average process, is derived by Davis and Resnick (1986). Obviously the difference between the observed and smoothed values or the error is required to be a minimum.

In SIMA, random variable $p_{i}$ represents the number of observations on the $i^{\text {th }}$ trajectory and $p_{i}$ 's are i.i.d. Assuming $t_{j}$ and $X_{i j}: j \in J_{i}$ are independent of $p_{i}, E\left(p_{i}\right)<\infty$, $P\left(p_{i}>1\right)>0$, and $E\left(X_{i j}\right)=\mu$, then it is a well known fact that as $p_{i} \rightarrow \infty$ in a fixed interval $T$, and $m \rightarrow 1$,

$$
\sup _{t \in T} E\left|Y_{i l}^{*}(t)-X_{i l}(t)\right| \rightarrow 0 .
$$

For more details see (Davis and Resnick, 1986) and (Furrer et.al, 2006). The following theorem shows the variance of the error approaches to $\sigma^{2}$ under similar assumptions.

Theorem 4.3.2: Let $p_{i}$ be the number of observations on the $i^{\text {th }}$ trajectory, and also under the assumption that $X_{i j}(t)$ 's are i.i.d., with $E\left(X_{i j}\right)=0$ and $E\left(X_{i j}\right)^{2}=\sigma^{2}$, when $t \rightarrow s_{j}$, i.e. $w_{j} \rightarrow 1$, then,

$$
\sup _{t \in T} E\left[Y_{i l}^{*}(t)-X_{i l}(t)\right]^{2} \rightarrow \frac{m-1}{m} \sigma^{2} .
$$

Further, when $p_{i} \rightarrow \infty$ and $t \rightarrow s_{j}$ with very large $m$,

$$
\sup _{t \in T} E\left[Y_{i l}^{*}(t)-X_{i l}(t)\right]^{2} \rightarrow \sigma^{2}
$$

Proof is given in Appendix B.

Application of SIMA concept to data trajectories has yielded a smooth trend. Details of the application are explained in Chapter 5. Given a data set consisting of several variables and multiple observations on each variable, smoothing enables the construction of a functional relationship among the variables. Degree of smoothing obtained via SIMA is compared with the smoothing results obtained from LLR. Hence, a brief summary of LLR is given in Section 4.3.3.

### 4.3.3. Local Linear Regression

The Local Linear Regression (LLR) is a smoothing method primarily based on the idea that a smooth function can be approximated by a low degree polynomial (Fan and Gijbels, 1996). This translates into minimizing

$$
\sum_{i=1}^{n} w\left(\frac{x_{i}-x}{h}\right)\left(Y_{i}-\left(a_{0}+a_{1}\left(x_{i}-x\right)\right)\right)^{2},
$$

to obtain estimates for the coefficients $a_{0}$ and $a_{1}$. Here $w\left(\left(x_{i}-x\right) / h\right)$ is the kernel weight function with bandwidth $h$. The LLR estimate at point $x$ is given by $\hat{\mu}(x)=\sum_{i=1}^{n} b_{i}(x) Y_{i}$. The coefficients $b_{i}$ are found by

$$
\mathbf{b}(x)^{T}=\left(b_{1}(x), \ldots, b_{n}(x)\right)=q^{T}\left(\mathbf{X}^{T} \mathbf{W} \mathbf{X}\right)^{-1} \mathbf{X}^{T} \mathbf{W}
$$

where, $q^{T}=\left(\begin{array}{ll}1 & 0\end{array}\right)$, and $\mathbf{X}=\left[\begin{array}{cc}1 & x_{1}-x \\ \vdots & \vdots \\ 1 & x_{n}-x\end{array}\right]$ is the design matrix. Further details are given in Loader (1999).

LLR smoothing follows the observed trend closely when the correlation $\left(r_{i j}\right)$ between the predictor and response variables is high (preferably, $\left|r_{i j}\right|>0.7$ ), and deviations from actual trend increase as this correlation becomes lower (Breiman and Friedman, 1985) (Hoover et. al.,1998, Härdle, 1992).

Let $r_{i j}$ be the correlation coefficient between different trajectories, where $i=1, \ldots, n$ is the trajectory counter and $j=1, \ldots, p$ is the column counter. Then obtained $p \times p$
correlation coefficient matrix $\mathbf{R}$ is symmetric with, $r_{i j}=r_{j i}$. Close inspection of $\mathbf{R}$ gives an idea in between which trajectories one can expect good or poor LLR smoothing results.

Figures 5.5 and Figure 5.7 are the graphical representation of matrix $\mathbf{R}$ for the data sets used in this study. In examples where natural phenomenon is involved, such as those in Earth Sciences, trajectories spatially closer will tend to have high $r$ values.

### 4.4. Smoothing the Covariance

Observations made over a trajectory are likely to include errors or noise. Smoothing by expanding into orthogonal eigenfunctions will help to reduce the noise involved during sampling. Therefore, estimation of eigenfunctions becomes a crucial step in smoothing. Covariance matrix from observed data is defined as, $G_{i j l}\left(Y_{i j}-\hat{\mu}\left(t_{i j}\right)\right)\left(Y_{i l}-\hat{\mu}\left(t_{i l}\right)\right)$. The obtained covariance matrix can be plotted as a scatter plot surface $\left(\left(t_{i j}, t_{i l}\right), G_{i j l}\right), l, j=1, \ldots, p_{i}$, and $i=1, \ldots, n$ in $\square^{3}$. Smoothing is applied to this surface under certain regularity assumptions as given in Fang et.al., (2005). Then the local linear surface smoother becomes

$$
\begin{equation*}
\sum_{i=1}^{n} \sum_{1 \leq j \neq 1 \leq p_{i}} \kappa\left(\frac{t_{i j}-t}{h_{G}}, \frac{t_{i l}-t}{h_{G}}\right)\left\{G_{i}\left(t_{i j}, t_{i l}\right)-f\left(\beta,\left(t, t^{+}\right),\left(t_{i j}, t_{i l}\right)\right\}^{2}\right. \tag{4.4.1}
\end{equation*}
$$

where, $\quad f\left(\beta,\left(t, t^{+}\right),\left(t_{i j}, t_{i l}\right)\right)=\beta_{0}+\beta_{11}\left(t-t_{i j}\right)+\beta_{12}\left(t-t_{i l}\right)$. Minimization of (4.4.1) with respect to $\beta_{0}, \beta_{11}$, and $\beta_{12}$ yields the smooth covariance matrix $\hat{G}\left(t, t^{+}\right)=\hat{\beta}_{0}\left(t, t^{+}\right)$.

Based on SIMA concept in Sections 4.3.1 and 4.3.2, smoothing the covariance matrix can be performed with a similar thought expressed above.

The covariance surface can be smoothed by applying the SIMA concept in various directions. Smooth surfaces obtained from each direction are then averaged to obtain the final smooth surface. That is equivalent to averaging smooth values obtained for a given point, smoothed by different directions.

Assume $k$ is the number of different directions used in directional smoothing and $\mathbf{G}_{1}^{*}, \ldots, \mathbf{G}_{k}^{*}$ be directional smoothed covariance matrices. Smoothing in each direction is carried out according to SIMA methodology. Final smoothed covariance matrix $\mathbf{G}^{*}$ is obtained by averaging the corresponding elements of directional smoothed matrices.

$$
\begin{equation*}
\mathbf{G}^{*}=k^{-1} \sum_{k} \mathbf{G}_{i}^{*}, i=1, \ldots, k \tag{4.4.2}
\end{equation*}
$$

The directional moving average smoothing concept in (4.4.2) is still under study. Thus, only one direction is used to smooth the covariance in Chapter 5.

## Chapter 5

## APPLICATIONS

Multivariate and functional data analysis concepts introduced in Chapter 2 and Chapter 3 together with the new method of smoothing (SIMA) developed during the course of this study given in Chapter 4, sets up the foundation of this thesis. Estimation of the unknown values or parameters in a certain process is of prime importance. Equation (3.1.13) is the key to the estimation of trajectories. Terms involved in this equation can all be estimated as explained in appropriate sections. The global mean function $\hat{\mu}(t)$ is estimated from all available data. $\hat{\phi}(t)$, the eigenfunction of the covariance function is also estimated using all available data. $\hat{\xi}$, represents the principal component scores computed from the data. During the estimation process, some types of smoothing are used. Hence, smoothing becomes an important aspect in the estimation process. There exists a range of different smoothing techniques developed over the years. Kernel, splines, LLR are some of the widely used smoothers. However, there are still certain situations that require special attention. One such point was the use of moving averages as a smoother. When multivariate data are observed on a regular grid and especially when the correlation between variables are low, use of SIMA as explained in detail under Chapter 4, is at least as efficient as the kernel or LLR smoothing methods. Therefore, in the application of the aforementioned methodologies were tried on data
sets from different fields to test their validity, and sometimes compare different methodologies.

### 5.1 Smoothing the Global Mean Function Using Different Smoothing

## Techniques

Part of this thesis work involved evaluation of the new smoothing technique developed. Smoothing the global mean was undertaken using SIMA, PACE, and LLR methods, and results compared. SIMA is also applied to smoothing the covariance surface and graphically compared with the kernel smoothed covariance surface. Details of the applications are presented in the following sections of this chapter.

### 5.1.1. Using SIMA and PACE for Smoothing the Mean

There are several smoothing techniques used in various stages of data processing, i.e. direct smoothing of data values, smoothing the mean and covariance functions. These are the cases considered and used in this thesis. As an example, a data set consisting of the daily percent change in the value of shares over 30 consecutive working-days, belonging to 20 companies (trajectories) from İstanbul Stock Exchange is used to estimate the smooth mean function. Raw data and the estimated/smoothed data are shown in Figure 5.1. Negative values refer to a loss and positive values refer to a gain in the value of a share. Actual data set obtained from the online records of İstanbul Stock Exchange web site is given in Appendix D, Table 1.




Figure 5.1: Raw Trajectories of the Daily Percent Change in the Value of Shares Over 30 Consecutive Working-Days, Trajectories Smoothed by SIMA, and PACE Methods.

Extreme values are visible in the first few days as well as days 18 and 24 . It is a fact that the observation of an extreme value in a process would mean that there is/are unexpected situation(s) temporarily affecting the process. They pose a danger if they are included in a study by shifting the trend (mean and variance) out of its expected path. Therefore, smoothing of data prior to the generation of the estimated mean function is beneficial.

The Epanechnikov kernel smoother used in the PACE method, and the SIMA method are implemented to smooth the trajectories of the data. For each case the average of trajectories are computed and compared with the average of the raw trajectories. MSE functions between average raw and smooth averages obtained through PACE and SIMA are given in Figure 5.2. In general the MSE function obtained from SIMA appears to be lower than that of PACE MSE function, indicating better performance in smoothing by the SIMA method. $\left(M S E_{S I M A}=0.80\right.$ and $\left.M S E_{P A C E}=1.09\right)$


Figure 5.2: MSE Functions between Average Raw and Smooth Averages obtained through PACE and SIMA.

A second data set from a different field of study is chosen for smoothing the mean using SIMA and PACE methodologies. This will give an idea about the performance of the two methods for comparison.

The data set selected represents the coordinates of points in three dimensional space $(x, y, z) . x$ is the coordinate of a point in East-West, $y$ is the coordinate in the North - South direction, while $z$ is the elevation from sea level in meters.

Data is obtained from 1:2500 scale topographic maps. Selected area containing gentle slopes as well as locations where rapid changes in elevation are observed, and are selected for the purpose of this study.

The data set consists of 30 parallel trajectories at 50 meters apart and each trajectory containing 30 measurements taken on a regular grid bases at every 50 meters in the EastWest direction.

Figure 5.3 shows the raw trajectories of this data in the East-West direction, but there are no significant deviations in the values forming each trajectory. Smoothed trajectories by SIMA and PACE are also given in Figure 5.3 for comparison.

Data used in this study is given in Appendix D, Table 2. Hence, no extreme values.

Smoothing of the mean is undertaken using SIMA and PACE methods and associated MSE functions are generated. Figure 5.4 shows the MSE functions between raw trajectories and smoothed by SIMA and PACE methods. From Figure 5.4 it can be seen that the error functions are following a similar trend, indicating no significant difference
between the two methodologies. In fact the overall MSE of SIMA method and PACE method are $M S E_{S I M A}=8.47$ and $M S E_{P A C E}=7.97$ suporting the idea of close performance.

However, for the data values at locations 3, 4, 28, and 29 the increase in MSE for SIMA is due to rapid change in elevation in the Nort - South direction.

A similar increase in MSE at data locations 14, 15, 17, and 18 in PACE method is considered attributable to over smoothing of the mean, foloowing the detailed study of the raw and smooth mean values.

The shares data set with extreme values and the elevation data set without extreme values are both smoothed using SIMA and PACE methods. For each data set, associated MSE values indicates no significant difference between the two methodologies.


Figure 5.3: Raw and Smoothed Trajectories of Elevation Data.


Figure 5.4: MSE Functions for SIMA and PACE Exhibits a Similar Behaviour with a Few Exceptions.

### 5.1.2 Using SIMA and LLR for Smoothing the Mean.

In this section same data sets used in Section 5.1.1 are used to compare the performance of SIMA with the LLR smoothing method. It is known that the accuracy of smoothing with LLR gets better as the correlation coefficient between the variables increase. For the elevation data set the high correlation between trajectories is expected, since the nature forces shaping the topography are the same within close proximity.

In the shares data set, high correlation between trajectories should not be expected due to market conditions affecting shares differently.

### 5.1.2.1 High Correlation Between Variables Case

Area from where the elevation data is taken contains gentle slopes as well as locations where rapid changes in elevation are observed. Correlation between trajectories is
mostly high to very high $\mid r_{i j}>0.8$. Correlation matrix surface given in Figure 5.5 clearly shows this feature. On the diagonal correlation values are 1 and decline towards the ends as distances between trajectories increase.


Figure 5.5: Correlation Surface between Different Trajectories of Elevation Data.

Smoothing of trajectories through SIMA is compared with the smoothing results of LLR. For comparison the Root Mean Square Deviation (RMSD) or Root Mean Square Error (RMSE) between the observed and the smooth trajectories for SIMA and LLR are considered. Figure 5.6 shows the RMSD functions for the two smoothing methods.


Figure 5.6: RMSD Functions for SIMA and LLR Smoothing Methods.

As expected in the presence of high correlation between trajectories, smoothing carried out by LLR has lower trend, indicating smooth mean following closely the observed mean function. Higher RMSD values obtained in SIMA, meaning when the correlation is high between the variables LLR performs better than SIMA.

### 5.1.2.2 When Correlation is Low Between Variables

Based on the theory of linear regression it is known that as the correlation between two variables decrease, estimation using the obtained regression equation deteriorates, i.e. becomes less reliable. LLR is used as a smoother. Smoothing in the case when pairwise correlation between variables under study is low will result in high errors. Through trials using data sets where correlation between variables are low, agreed with this concept. The stock exchange shares data set introduced in Section 5.1.1 is used to compute the correlation between the daily performances of different shares. Results are shown in Figure 5.7. Apart from the diagonal elements where the correlation values are
$r_{i j}=1$, when $i=j$. Of the remaining 870 correlation values only $13 \%$ are above $50 \%$, leading to the conclusion of low correlation between 30 variables.


Figure 5.7: Correlation Values Between the Daily Performance of Different Shares.

Smooth mean functions are computed via SIMA and LLR. The RMSD between the average obtained from raw data and the smooth means are shown in Figure 5.8.

Clearly, SIMA resulted in lower errors compared with LLR, indicating better smoothing results can be obtained from a data set where the correlation between variables are low.


Figure 5.8: RMSD Functions Obtained for SIMA and LLR Smoothing.

In Figure 5.9, correlation coefficient $\left(r_{i j}\right)$ values between predictor and response trajectories, and RMSD values between observed and LLR smoothed trajectories are given for comparison. It can be seen that for correlation $r_{i j}$ values under 0.45, RMSD values tend to be not very sensitive to changes in $r_{i j}$, while RMSD starts decreasing for values of $r_{i j}$ above $45 \%$. This is an expected result due to the nature of the shares of different companies.


Figure 5.9: Relationship Between Correleation Coefficient and RMSD for the Shares Data Set.

### 5.2. Smoothing the Covariance Surface Using SIMA and Kernel

## Smoothers

The covariance matrix plays a major role in MDA and FDA. Smoothing is essential especially when high local variability is observed. Amount of smoothing in SIMA depends on the number of data points ( $m$ ) falling into each lag interval ( $h$ ), while in kernel smoothing used in PACE depends on the band width ( $h$ ).

In the studied elevation and shares data sets, for SIMA $m=3$ and for PACE $h$ is automatically selected by generalized cross validation (GCV) method (Müller and Prewitt, 1993). As a result both covariance surfaces are highly smoothed by the kernel smoother, while smoothing by SIMA remained relatively mild compared with kernel smoothing. Increasing the lag interval will increase the degree of smoothing, and similarly reducing the size of the bandwidth in kernel will reduce the degree of smoothing.

Figure 5.10 and Figure 5.11 show the raw, smoothed by SIMA and smoothed by kernel covariance surfaces for the elevation and shares data, respectively.


Figure 5.10: Raw Covariance Surface and Covariance Surfaces Smoothed by SIMA, and by Epanechnikov Kernel Methods for the Elevation Data.


Figure 5.11: Raw Covariance Surface and Covariance Surfaces Smoothed by SIMA, and by Epanechnikov Kernel Methods for the Shares Data.

### 5.3 Robustness of SIMA Method

An estimator is said to be robust if it is insensitive to changes in the underlying distribution and resistant against the presence of outliers.

A good robust estimator as given in Abu-Shawiesh (2008) should have,

- high efficiency, meaning minimum variance.
- high breakdown point, which is a measure of the maximum fraction of outliers.
- redescending influence function measuring the reaction of an estimator to a small fraction of outliers.
- low gross error sensitivity which measures the worst influence a small amount of contamination of fixed size can have on the value of the estimator.

The median absolute deviation from the sample median (MAD) is considered as one of the good robust estimators since it satisfies these requirements to notable level of significance. Other noteworthy references on robustnes are Lax (1985), and Parr and Schucany (1980).

MAD values are computed using the formula given in Abu-Shawiesh, (2008)

$$
M A D=1.4826 M D\left\{\left|X_{i}-M D\right|\right\}, \mathrm{i}=1,2, \ldots, \mathrm{n} .
$$

Here, $M D$ is the sample median. The factor 1.4826 is only used when samples come from Gaussian distribution, and should not be used when sample comes from non-

Gaussian distributions. The lower control limit ( $L_{c l}$ ) and upper control limit ( $U_{c l}$ ) of the S-control chart are given as,

$$
\begin{align*}
& L_{c l}=c_{4} S+3 S \sqrt{1-c_{4}^{2}}  \tag{5.3.1}\\
& U_{c l}=c_{4} S-3 S \sqrt{1-c_{4}^{2}}
\end{align*}
$$

where,

$$
c_{4}=\sqrt{\frac{2}{n-1}} \frac{\Gamma(n / 2)}{\Gamma(n-1 / 2)},
$$

(He and Grigoryan, 2002).

The robustness of the proposed SIMA method of smoothing is checked by using the Shewhart S-control chart. Column-wise standard deviations $S_{j}$ of SIMA values of the share and elevation data sets are calculated and plotted together with respective control limits from (5.3.1).

Figure 5.12 shows the standard deviation function together with the control limits for the elevation data, while Figure 5.13 shows the same for the shares data. In both cases the standard deviation function falls completely within the control limits indicating the robustness of the SIMA smoothing method.


Figure 5.12: Standard Deviation Function for the Elevation Data Totally within the Control Limits Indicating the Robustness of the SIMA Smoothing Method.


Figure 5.13: Standard Deviation Function for the Shares Data Set Completely within the Control Limits Indicating the Robustness of the SIMA Smoothing Method.

## Chapter 6

## CONCLUSION AND FURTHER STUDY

### 6.1 Conclusion

Application of the proposed SIMA smoothing method is applied to two distinct data sets together with kernel (PACE) and LLR smoothers. Obtained results from SIMA are compared with those from PACE and LLR.

SIMA performed better than LLR under the condition of weak correlation between variables involved. In the application to the data sets for the smoothing of mean and covariance, SIMA performed equally well with PACE method which uses Epanechnikov kernel smoother.

The measures used in comparing different smoothers are Mean Square Error (MSE) and Root Mean Square Deviation (RMSD) between an observed trajectory and its smoothed estimate.

SIMA smoothing method is checked for robustness using the standard deviation function control limits method (Shewhart S-control chart) and is found to be a robust smoother.

### 6.2 Further Research

Based on the research that led to the preparation of this thesis, the following topics are identified as possible areas of further research.

1. Minimizing the error for the proposed SIMA smoothing method by using areas between a trajectory and its estimated SIMA trajectory.
2. The directional moving average smoothing with SIMA concept is another area for further investigation. One immediate application area can be the smoothing of the covariance surface given in equation (4.4.2).

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## APPENDICES

## APPENDIX A: Projecting the Data

It is known that the angle $\theta$ between vectors $\mathbf{x}, \mathbf{y} \in \square^{p}$ is defined by

$$
\begin{equation*}
\cos \theta=\frac{\mathbf{x}^{T} \mathbf{y}}{\|\mathbf{x}\|\|\mathbf{y}\|} \tag{A.1}
\end{equation*}
$$

If $\mathbf{x}^{T} \mathbf{y}=0$, then $\theta=\pi / 2$, from (A.1). In a right angled triangle, the cosine of $\theta$ is equal to the length of the base $\left\|\mathbf{p}_{x}\right\|$, over, the length of the hypotenuse $\|\mathbf{x}\|$. Thus,

$$
|\cos \theta|=\frac{\left\|\mathbf{p}_{x}\right\|}{\|\mathbf{x}\|}=\frac{\mathbf{x}^{T} \mathbf{y}}{\|\mathbf{x}\|\|\mathbf{y}\|},
$$

can be written. Hence

$$
\begin{equation*}
\left\|\mathbf{p}_{x}\right\|=\|\mathbf{x}\||\cos \theta|=\frac{\mathbf{x}^{T} \mathbf{y}}{\|\mathbf{y}\|} . \tag{A.2}
\end{equation*}
$$

Here, $\mathbf{p}_{x}$ is the projection (coordinate) of $\mathbf{x}$ on $\mathbf{y}$. Therefore, the coordinate of $\mathbf{x}_{i}$ on $L_{1}$ is the projection point $p_{x_{i}}$,

$$
\begin{equation*}
p_{x_{i}}=\mathbf{x}_{i}^{T} \frac{\mathbf{u}_{1}}{\left\|\mathbf{u}_{1}\right\|}=\mathbf{x}_{i}^{T} \mathbf{u}_{1} . \tag{A.3}
\end{equation*}
$$

In "least-square" meaning; the best line $L_{1}$ is defined by finding a unit vector $\mathbf{u}_{1} \in \square^{p}$ which minimizes

$$
\begin{equation*}
\sum_{i=1}^{n}\left\|\mathbf{x}_{i}-p_{x_{i}}\right\|^{2} \tag{A.4}
\end{equation*}
$$

or, maximizes $\sum_{i=1}^{n}\left\|p_{x_{i}}\right\|^{2}$ under the constraint $\left\|\mathbf{u}_{1}\right\|=1$. From (A.3) the following can be written

$$
\left(\begin{array}{c}
p_{x_{1}} \\
p_{x_{2}} \\
\vdots \\
p_{x_{n}}
\end{array}\right)=\left(\begin{array}{c}
\mathbf{x}_{1}^{T} \mathbf{u}_{1} \\
\mathbf{x}_{2}^{T} \mathbf{u}_{1} \\
\vdots \\
\mathbf{x}_{n}^{T} \mathbf{u}_{1}
\end{array}\right)=\mathbf{X} \mathbf{u}_{1} \in \square^{n}
$$

The following theorem will be used to reformulate the maximization problem (Härdle and Simar, 2003).

Theorem A.1 : If $\mathbf{A}$ and $\mathbf{B}$ are symmetric and $\mathbf{B}>0$, then maximum of $\frac{\mathbf{x}^{T} \mathbf{A x}}{\mathbf{x}^{T} \mathbf{B} \mathbf{x}}$ is given by the largest eigenvalue of $\mathbf{B}^{-1} \mathbf{A}$. In general terms this can be expressed as,

$$
\max _{\mathbf{x}} \frac{\mathbf{x}^{T} \mathbf{A} \mathbf{x}}{\mathbf{x}^{T} \mathbf{B} \mathbf{x}}=\lambda_{1} \geq \lambda_{2} \geq \ldots \geq \lambda_{p}=\min _{\mathbf{x}} \frac{\mathbf{x}^{T} \mathbf{A} \mathbf{x}}{\mathbf{x}^{T} \mathbf{B} \mathbf{x}}
$$

where, $\lambda_{i}$ 's are the eigenvalues of $\mathbf{B}^{-1} \mathbf{A}$. If $\mathbf{x}^{T} \mathbf{B x}=1$ we obtain

$$
\max _{\left\{\mathbf{x}: \mathbf{x}^{T} \mathbf{B} \mathbf{x}=1\right\}} \mathbf{x}^{T} \mathbf{A} \mathbf{x}=\lambda_{1} \geq \lambda_{2} \geq \ldots \geq \lambda_{p}=\min _{\left\{\mathbf{x}: \mathbf{x}^{T} \mathbf{B} \mathbf{x}=1\right\}} \mathbf{x}^{T} \mathbf{A} \mathbf{x}
$$

Proof: By definition $\mathbf{B}^{1 / 2}=\boldsymbol{\Gamma}_{\mathbf{B}} \boldsymbol{\Lambda}_{\mathbf{B}}^{1 / 2} \boldsymbol{\Gamma}_{\mathbf{B}}^{T}$ is symmetric. Then $\mathbf{x}^{T} \mathbf{B x}=\square \mathbf{x}^{T} \mathbf{B}^{1 / 2} \square^{2}=\square \mathbf{B}^{1 / 2} \mathbf{x} \square^{2}$.

Let $\mathbf{y}=\frac{\mathbf{B}^{1 / 2} \mathbf{x}}{\square \mathbf{B}^{1 / 2} \mathbf{x} \square^{2}}$, then

$$
\begin{equation*}
\max _{\mathbf{x}} \frac{\mathbf{x}^{T} \mathbf{A x}}{\mathbf{x}^{T} \mathbf{B x}}=\max _{\left\{\mathbf{y} \mathbf{y}^{T} \mathbf{y}=1\right\}} \mathbf{y}^{T} \mathbf{B}^{-1 / 2} \mathbf{A} \mathbf{B}^{-1 / 2} \mathbf{y} \tag{A.5}
\end{equation*}
$$

From spectral decomposition theorem for symmetric square matrices
$\mathbf{B}^{-1 / 2} \mathbf{A} \mathbf{B}^{-1 / 2}=\boldsymbol{\Gamma} \boldsymbol{\Lambda} \boldsymbol{\Gamma}^{T}$. Letting $\mathbf{z}=\boldsymbol{\Gamma}^{T} \mathbf{y}$ we get $\mathbf{z}^{T} \mathbf{z}=\mathbf{y}^{T} \boldsymbol{\Gamma} \boldsymbol{\Gamma}^{T} \mathbf{y}=\mathbf{y}^{T} \mathbf{y}$.

Thus (A.5) can be written as

$$
\max _{\left\{\mathbf{z} \mathbf{z}^{T} \mathbf{z}=1\right\}} \mathbf{z}^{T} \boldsymbol{\Lambda} \mathbf{z}=\max _{\left\{\mathbf{z} \mathbf{z}^{T} \mathbf{z}=1\right\}} \sum_{i=1}^{p} \lambda_{i} \mathbf{z}_{i}^{2} \text { and } \max _{\mathbf{z}} \sum \lambda_{i} \mathbf{z}_{i}^{2} \leq \lambda_{1} \underbrace{\max _{\mathbf{z}} \sum_{i=1}^{p} \mathbf{z}_{i}^{2}}_{=1}=\lambda_{1}
$$

When $\mathbf{z}=(1,0, \ldots, 0)^{T}$ gives the maximum. That is $\mathbf{y}=\gamma_{1}$ the first column of $\boldsymbol{\Gamma}$, hence $\mathbf{x}=\mathbf{B}^{-1 / 2} \lambda_{1}$.

As $\mathbf{B}^{-1} \mathbf{A}$ and $\mathbf{B}^{-1 / 2} \mathbf{A B} \mathbf{B}^{-1 / 2}$ have the same eigenvalues, proof is complete.

In Theorem A.1, substituting $\mathbf{A}=\mathbf{X}^{T} \mathbf{X}$ and $\mathbf{B}=\mathbf{I}$, we have:

If $\mathbf{X}^{T} \mathbf{X}$ and $\mathbf{I}>0$ are symmetric, then the maximum of $\mathbf{u}_{1}^{T} \mathbf{X}^{T} \mathbf{X} \mathbf{u}_{1}$ under the constraint $\mathbf{u}_{1}^{T} \mathbf{u}_{1}=1$ is given by the largest eigenvalue of $\mathbf{X}^{T} \mathbf{X}$.

$$
\begin{equation*}
\left.\left.\max _{\left\{\mathbf{u}_{i} ; \mathbf{u}_{1}^{T} \mathbf{u}_{1}=1\right\}} \mathbf{u}_{1}^{T} \mathbf{X}^{T} \mathbf{X} \mathbf{u}_{1}=\lambda_{1} \geq \lambda_{2} \geq \ldots \geq \lambda_{p}=\min _{\left\{\mathbf{u}_{i}: \mathbf{u}_{1} T\right.} \mathbf{u}_{1}=1\right\}\right\} \tag{A.6}
\end{equation*}
$$

Thus, by (A.6) the problem reformulated as: Find $\mathbf{u}_{1} \in \square^{p}$, with $\left\|\mathbf{u}_{1}\right\|=1$ that maximizes the quadratic form $\left(\mathbf{X} \mathbf{u}_{1}\right)^{T}\left(\mathbf{X} \mathbf{u}_{1}\right)$ or $\max _{\left\{\mathbf{u}_{1}: \mathbf{u}_{1} \mathbf{u}_{1}=1\right\}} \mathbf{u}_{1}^{T} \mathbf{X}^{T} \mathbf{X} \mathbf{u}_{1}$. The unit vector $\mathbf{u}_{1}$ can be found by the following theorem.

Theorem A.2: The vector $\mathbf{u}_{1}$ is the eigenvector of $\mathbf{X}^{T} \mathbf{X}$ associated with the largest eigenvalue $\lambda_{1}$ of $\mathbf{X}^{T} \mathbf{X}$.

Proof: If eigenvector equation is used for $\mathbf{X}^{T} \mathbf{X}$, we have $\left(\mathbf{X}^{T} \mathbf{X}\right) \mathbf{k}=\left(\mathbf{u}_{1}^{T} \mathbf{X}^{T} \mathbf{X} \mathbf{u}_{1}\right) \mathbf{k}$, where $\mathbf{k}$ is the eigenvector of $\mathbf{X}^{T} \mathbf{X}$. The eigenvector equation can be rewritten as follows

$$
\begin{equation*}
\left(\mathbf{X}^{T} \mathbf{X}-\mathbf{u}_{1}^{T} \mathbf{X}^{T} \mathbf{X} \mathbf{u}_{1}\right) \mathbf{k}=\mathbf{X}^{T} \mathbf{X}\left(k-\mathbf{u}_{1}^{T} \mathbf{k} \mathbf{u}_{1}\right)=0 . \tag{A.7}
\end{equation*}
$$

From (A.7) it is clear that $\mathbf{k}=\mathbf{u}_{1}$.

## APPENDIX B: Some Proofs

Proof of Theorem 4.3.1: Let $\mathbf{J}$ be any $(p-m+1)$-vector and $Y_{i}(t)=\frac{1}{m} \mathbf{A}_{m} X_{i}(t)$. Then,
$\mathbf{J}^{T} Y_{i}(t)=\boldsymbol{\theta}^{T} X_{i}(t)$, where $\boldsymbol{\theta}=\frac{1}{m} \mathbf{A}_{m}{ }^{T} \mathbf{J}$ and the $(p-m+1) \times p \quad$ matrix $\mathbf{A}_{m}$ has the general form.

$$
\mathbf{A}_{m}=\left[\begin{array}{cccccccc}
1 & 1 & \cdots & 1 & 0 & 0 & \cdots & 0 \\
0 & 1 & 1 & \cdots & 1 & 0 & \cdots & 0 \\
\vdots & & \ddots & & & \ddots & & \vdots \\
\vdots & & & \ddots & & & \ddots & \vdots \\
0 & \cdots & 0 & 1 & 1 & \cdots & 1 & 0 \\
0 & 0 & \cdots & 0 & 1 & 1 & \cdots & 1
\end{array}\right] .
$$

Let $a_{i j}$ be the elements of $\mathbf{A}_{m}$. Then

$$
a_{i j}=\left\{\begin{array}{c}
1, \text { for }, i \leq j \leq i+m-1 \\
0, \text { for }, \text { elsewhere }
\end{array} .\right.
$$

Since $X_{i}(t)$ is p-variate normal, then $\boldsymbol{\theta}^{T} X_{i}(t)$ is univariate normal. Thus $\mathbf{J}^{T} Y_{i}(t)$ is also univariate normal for all fixed vectors $\mathbf{J}$, and $Y_{i}(t)$ is multivariate normal.

Proof of Theorem 4.3.2: Under the assumptions on $p_{i}$ and $t$,

$$
E\left[Y_{i l}^{*}(t)-X_{i l}(t)\right]^{2}=E\left[\frac{1}{m}\left(\sum_{j=l}^{m+l-1} X_{i j}(t)-\sum_{j=l-1}^{m+l-2} X_{i j}(t)\right) w_{j}+\frac{1}{m} \sum_{j=l-1}^{m+l-2} X_{i j}(t)-X_{i l}(t)\right]^{2}
$$

can be rewritten as,

$$
\begin{align*}
& E\left[Y_{i l}^{*}(t)-X_{i l}(t)\right]^{2}=\frac{w_{j}^{2}}{m^{2}} E\left[\sum_{j=l}^{m+l-1} X_{i j}(t)-\sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]^{2}+\frac{2 w_{j}}{m^{2}} E\left[\sum_{j=l}^{m+l-1} X_{i j}(t) \sum_{j=l-1}^{m+l-2} X_{i j}(t)-\left(\sum_{j=l-1}^{m+l-2} X_{i j}(t)\right)^{2}\right] \\
& -\frac{2 w_{j}}{m} E\left[X_{i l}(t) \sum_{j=l}^{m+l-1} X_{i j}(t)-X_{i l} \sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]-\frac{2}{m} E\left[X_{i l} \sum_{j=l-1}^{m+-2} X_{i j}(t)\right]+\frac{1}{m^{2}} E\left[\sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]^{2}+E\left[X_{i l}\right]^{2} \\
& =\frac{w_{j}^{2}}{m^{2}} E\left[\sum_{j=l}^{m+l-1} X_{i j}(t)\right]^{2}+\frac{2 w_{j}\left(1-w_{j}\right)}{m^{2}} E\left[\sum_{j=l}^{m+l-1} X_{i j}(t) \sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]+\frac{w_{j}\left(w_{j}-2\right)}{m^{2}} E\left[\sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]_{(\mathrm{B} .1)}^{2}  \tag{B.1}\\
& -\frac{2}{m} E\left[X_{i l} \sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]+\frac{1}{m^{2}} E\left[\sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]^{2}+E\left[X_{i l}\right]^{2} .
\end{align*}
$$

Under the i.i.d assumption on $X_{i j}$, and $E\left(X_{i j}\right)=0$ and $E\left(X_{i j}\right)^{2}=\sigma^{2}$, then

$$
E\left[\sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]^{2}=E\left[\sum_{j=l}^{m+l-1} X_{i j}(t)\right]^{2}=E\left[\sum_{j=l}^{m+l-1} X_{i j}(t) \sum_{j=l-1}^{m+l-2} X_{i j}(t)\right]=m \sigma^{2} .
$$

Therefore, (B.1) can be written as,

$$
\frac{1}{m^{2}} E\left[\sum_{j=l-1}^{m+l-1} X_{i j}(t)\right]^{2}+E\left[X_{i l}(t)\right]^{2}-\frac{2}{m} E\left[X_{i l}(t) \sum_{j=l-1}^{m+l-2} X_{i j}(t)\right] .
$$

As $t \rightarrow s_{j}$,

$$
E\left[Y_{i l}^{*}(t)-X_{i l}(t)\right]^{2} \rightarrow \frac{m-1}{m} \sigma^{2} .
$$

Moreover, for $p_{i} \rightarrow \infty$ and $t \rightarrow s_{j}$, with very large $m$,

$$
E\left[Y_{i l}^{*}(t)-X_{i l}(t)\right]^{2} \rightarrow \sigma^{2} .
$$

## APPENDIX C: Karhunen-Loève Theorem

The Karhunen-Loève theorem is an infinite linear combination of orthogonal functions representation of a stochastic process, as a Fourier series expansion of a function on a bounded interval. The theorem yields the best basis that minimizes the total mean squared error.

The coefficients in the Karhunen-Loève theorem are random variables and the expansion basis depends on the process. Hence, the orthogonal basis functions are determined by the covariance function of the process.

In the case of a centered stochastic process $\left\{X_{t}\right\}_{t \in I}$, i.e. $E\left(X_{t}\right)=0$ for all $t \in I$, where $I$ is a closed and bounded time or space interval, can be decomposed as

$$
\begin{equation*}
X_{t}=\sum_{k=1}^{\infty} \xi_{k} \phi_{k}(t) \tag{C.1}
\end{equation*}
$$

where, $\xi_{k}$ are pairwise uncorrelated random variables and the functions $\phi_{k}$ are continuous real-valued functions on $I$ that are pairwise orthogonal in $L^{2}(I)$. (C.1) is called Karhunen-Loève expansion or Karhunen-Loève decomposition. If the process is not centered, then in place of $X_{t}, X_{t}-E\left(X_{t}\right)$ is used.

The empirical version of the Karhunen-Loève theorem is called as the Karhunen-Loève transform, principal component analysis, proper orthogonal decomposition, Empirical orthogonal functions or the Hotelling transform.

A linear operator $T_{G_{X}}: L^{2}(I) \rightarrow L^{2}(I)$ defined for covariance function $G_{k}$, as follows,

$$
f(t) \rightarrow \int_{I} G_{X}(s, t) f(s) d s
$$

Since $T_{G_{X}}$ is a linear operator, its eigenvalues $\lambda_{k}$ and eigenfunctions $\phi_{k}$ can be found by solving the following equation.

$$
\begin{equation*}
\int_{I} G_{X}(s, t) \phi(s) d s=\lambda_{k} \phi_{k}(t) \tag{C.2}
\end{equation*}
$$

Mercer's Theorem: Let $X_{t}$ be a zero-mean square integrable stochastic process over $t \in I$, with continuous covariance function $G_{X}(s, t)$. Then, $G_{X}(s, t)$ is a Mercer kernel. Let $\phi_{k}$ be an orthonormal basis of $L^{2}(I)$ formed by the eigenfunctions of $T_{G_{X}}$ with respect to eigenvalues $\lambda_{k}, X_{t}$ has the representation $X_{t}=\sum_{k=1}^{\infty} \xi_{k} \phi_{k}(t)$ where the convergence is in $L^{2}$, uniform in $t$, and $\xi_{k}=\int_{I} X_{t} \phi_{k}(t) d t$. Further, $\xi_{k}$ are uncorrelated with $E\left(\xi_{k}\right)=0$ and $\operatorname{var}\left(\xi_{k}\right)=\lambda_{k}$.

Hence, Mercer's theorem says that, there exists a set of eigenvalues and eigenfunctions of $T_{G_{X}}$ from (C.2), forming an orthonormal basis of $L^{2}(I)$, such as,

$$
G_{X}(s, t)=\sum_{k=1}^{\infty} \lambda_{k} \phi_{k}(s) \phi_{k}(t) .
$$

More specifically, given any orthonormal basis $\left\{\varphi_{k}\right\}$ of $L^{2}(I)$, the process $X_{t}$ may decomposed $\quad$ as $\quad X_{t}(s)=\sum_{k=1}^{\infty} \zeta_{k}(s) \varphi_{k}(t) \quad$ where $\zeta_{k}(s)=\int_{I} X_{t}(s) \varphi_{k}(t) d t \quad$ and may approximate by the finite sum

$$
\begin{equation*}
\hat{X}_{t}(s)=\sum_{k=1}^{K} \zeta_{k}(s) \varphi_{k}(t) . \tag{C.3}
\end{equation*}
$$

The integer $K$ can be found by using the function $F(K)=\sum_{k=1}^{K} \lambda_{k} / \sum_{k=1}^{\infty} \lambda_{k}$. Claim of all such approximations (C.3) is that; the Karhunen-Loève approximation is the one that minimizes the total mean square error provided that eigenvalues are arranged in decreasing order.

## APPENDIX D: Data Sets.






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