# ANNUAL REPORT 

Statistical

Theory and Methodology for

Remote Sensing Data Analysis


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## ACKNOWLEDGMENTS


#### Abstract

The research work documented in this final report is carried out for NASA Johnson Space Center, Houston, Texas, under Contract NAS9-13512 to The University of Texas at Dallas, Richardson, Texas, for the period June 1, 1973 to May 31, 1974. This was accomplished in collaboration with Dr. J. P. Basu and Dr. R. S. Chhikara, research scientists, and Mr. Lynn Ziegler, graduate student, all of UTD; and Dr. T. L. Bullion of the Texas Tech University Statistics Faculty and his graduate student, Mr. E. R. Knezek, Jr.


## PREFACE

With the recent developments in statistical methodology of clustering and classification as well as in automatic remote sensing data processing techniques, the scope of extracting useful information on earth resources from multispectral scanner data has increased considerably. Though the performance of most of the previously developed remote sensing data handling techniques has yet to be ascertained through the process of testing and numerical evaluations, it now appears that the development of a technology for performing a large area earth resources inventory is in sight.

For the earth resources project of EOD, NASA-JSC, presently a top priority has been given to the development of a system for performing an inventory of some crop or crops of economic interest over a large area. An important aspect of this project would be to devise a suitable crop acreage estimation procedure. A major part of the research work reported here in our final report is concerned with this problem.

First of all, a model is developed for the evaluation of crop acreages (proportions) in an agricultural area using the classification approach. The model takes into account the classification errors likely to arise in labeling remotely sensed data points under the classification algorithm used and evaluates the actual crop proportions by correcting the expected labeled crop proportions for the possible bias due to these errors.

If the goal is to determine crop acreages for a large area, it will be necessary to use only a subset of the full data obtained using a sampling technique since it would not be practical to collect and process a complete set of scanner data covering the entire area. Depending upon whether or not
any information is available on the area crop layouts, agricultural practices, etc., a suitable sampling scheme needs to be devised for acquiring a well representative sample of the unlabeled remotely sensed data. The precision of any crop acreage estimates will also depend upon the performance of the classification algorithm used in labeling remotely sensed data points and whether or not the associated classification errors are known. It is very unlikely to have these errors known in advance. As such, a certain amount of ground truth, preferably a sample of ground truth for each crop type in the area of interest, needs to be ascertained so that the classifier is properly trained and the classification errors estimated.

Considering these aspects of the problem, we have discussed the estimation of crop acreages for different possible cases. If the classification errors for the classification procedure used are assumed known, the estimation method provides best estimates for the crop acreages. In case of unknown classification errors, the estimates are consistent. Next, both the error analysis and the problem of sample size are investigated in general as we11 as for certain specific cases. Our results are given in reports 1 , 4 and 5 listed in the table of contents.

A study of classification errors for the Gaussian maximum likelihood classifier is made when due to interest in identifying elements of only one class, the other class is made of the remainder of the classes and then the problem is treated as a two-class classification problem. Given in report 6, depending upon the geometry and location of the classes under this practice, it is shown that the classification performance for elements in the class of main interest may or may not improve under this practice.

The estimation of optimum errors of classification and the dependence of these estimates upon the distance between classes are examined for the two-class problem, assuming classes to be univariate normal, in report 2. Considering both the maximum likelihood estimate and the minimum variance unbiased estimate for the optimum probability of misclassification, we give the relative efficiencies of these two estimates; theoretically as well as numerically, and investigate the bias of the former.

A simulation study in done on the relationship between the probability of misclassification using linear as well as quadratic discriminant rules, the number of features and the training sample size. As shown in report 3, when the sample size is small, use of a fewer number of features for discrimination leads to smaller probability of misclassification.

For classifying an observation into one of two given normal populations whose parameters are unknown, the usual practice in the absence of any training sample is to cluster past data into two nearest neighbor clusters and to design a sample based Bayes' classifier, treating the two clusters as training samples from the two populations. The use of $\ell_{1}$ norm is often advocated for such clustering. In report 7 it has been shown that such advocation is not always reasonable.

A unified theory of adaptive pattern recognition has been presented in report 8. It has been shown that all adaptive pattern recognition algorithms are just different means of approximating a function that separates the sets of training samples, choosing different criteria for best approximation.

The data obtained by remote sensing devices in Earth Resources Survey come from a large area and often over a large period of time. Due to changes
in spatial and temporal conditions the statistical characteristics of the data have been found to undergo changes. In such a situation, the performance of sample-based Bayes classifier designed on the assumption of normality of the populations can be enhanced by periodic updating of the parameter estimates. It has been shown in report 9 that all updating algorithms that can be found in literature are related to one particular model of the random environment.
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# ESTIMATION OF A LARGE AREA 

CROP ACREAGE INVENTORY USING
REMOTE SENSING TECHNOLOGY*

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# ESTIMATION OF A LARGE AREA CROP ACREAGE <br> INVENTORY USING REMOTE SENSING TECHNOLOGY 

## 0. SUMMARY

Based upon the existing remote sensing capabilities, the useful informatin about the acreage of some crop of economic interest can be obtained from multispectral scanner measurements acquired over an agricultural area. If the goal is to determine the acreages covered by various crops over some large area such as the continental United States, then some sampling procedure will be necessary since it would not be practical to collect and process a set of scanner data covering the entire area.

In this report we develop a model for the evaluation of acreages (proportions) of different crop-types over a geographical area using a classification approach and give methods for estimating the crop acreages. If prior informaltion is available on the classification errors associated with the classification algorithm used, the estimation method provides the best estimate for the crop acreages. Otherwise, the method would first require a certain amount of ground truth in the area of interest to be obtained so that the classifier can be trained and the classification errors estimated.

If the main interest lies in estimating the acreages of a specific croptype such as wheat, it is suggested to treat the problem as a two-crop problem: wheat vs. non-wheat, since this simplifies the estimation problem considerably. The error analysis and the sample size problem is investigated for the twocrop approach. Certain numerical results for sample sizes are given for a JSC-ERTS-1 data example on wheat identification performance in Hill County, Montana and Burke County, North Dakota. Lastly, for a large area crop acreages inventory we suggest a sampling scheme for acquiring sample data and discuss the problem of crop acreage estimation and the error analysis.

## 1. INTRODUCTION

In recent years the development of several automatic data processing techniques for statistical pattern recognition has enhanced considerably the scope of remote sensing technology for the study of earth resources, particularly in the field of agriculture. It now appears that a system for performing a large area crop inventory can be developed on the basis of existing remote sensing capabilities.

The data handling and analysis for remotely sensed agricultural resources over a large area may not be feasible both from technical and economical viewpoints if each scanned data point is being processed for its recognition. For example, if a complete recognition is desired for an ERTS scene, it would require processing over half a million data points. As such, an important requirement for any system to be developed for a large area crop inventory should be to have a suitable crop acreage estimation technique that uses only a sample of the unlabeled remotely sensed data obtained for the area of interest for the purpose of recognition.

In this report we discuss a large area crop acreage estimation procedure that would meet this requirement for the system. We develop a model for the evaluation of crop proportions for an agricultural area and provide methods for crop acreage estimation, taking into consideration the classification errors likely to arise in labeling remotely sensed data. The error analysis for the model is studied and expressions for variances of different estimates are given, in general as well as in specific cases. For the two-crop situation, the problem of sample size is investigated and certain numerical results for the sample size are provided. Next, we extend the scope of our study to investigate a large area crop inventory.

## 2. CROP PROPORTIONS MODEL

Suppose there are m different crops $\pi_{1}, \pi_{2}, \ldots, \pi_{\mathbb{R}}$ in the agricultural area of interest and that every data point is identifiable with respect to one of these crops. Let $p_{i}$ denote the proportion of pixels in $\pi_{i}, i=1,2, \ldots, m$. Considering a random sample of $n$ unlabeled remotely sensed data points, let $n_{i}$ be the number of points classified into $\pi_{i}, i=1,2, \ldots, m$, using a classification algorithm. Suppose $n(i \mid j)$ is the number of data points to be in $\pi_{j}$ but classified into $\pi_{i}$, then

$$
n_{i}=n(i \mid 1)+n(i \mid 2)+\ldots \ldots+n(i \mid m)
$$

and

$$
\begin{equation*}
\frac{n_{i}}{n}=\sum_{j=1}^{m} \frac{n(i \mid j)}{n}, \quad i=1,2, \ldots, m \tag{2.1}
\end{equation*}
$$

are the observed crop proportions for the sample data under the classification algorithm used. The observed proportion $n_{i} / n$ is a biased estimate of $p_{i}$ since it estimates unbiasedly $E\left[n_{i} / n\right]$ given by

$$
\begin{align*}
e_{i} & =\sum_{j=1}^{m} E\left[\frac{n(i \mid j)}{n}\right] \\
& =\sum_{j=1}^{m} p_{j} P(i \mid j) \tag{2.2}
\end{align*}
$$

where $P(i \mid j)$ denotes the probability of classifying a data point from $\pi_{j}$ into $\pi_{i}$ under the classification algorithm. It may be pointed out that processing of remotely sensed data for total recognition would lead to an evaluation of the expected classified crop proportions $e_{i}$ 's instead of the
actual crop proportions $P_{i}$ 's. Of course, if the classification algorithm performs so well that the classifiction errors are sufficiently small, $\mathbf{e}_{\mathbf{i}}$ will be close enough to $p_{i}, i=1,2, \ldots, m$. But most statistical pattern recognition techniques for processing of remotely sensed data are expected to be fallible and thereby the two types of proportions are not going to be near equal. Henceforth in our discussion we will assume that $P(i \mid j)>0$ for at least one $\mathbf{j}$ different from $i$.

Denoting the observed proportion $n_{i} / n$ by $\hat{e}_{i}, i=1,2, \ldots, m$, it follows from (2.2) that

$$
e=E[\hat{e}]
$$

or

$$
\begin{equation*}
\mathrm{e}=\mathrm{Pp} \tag{2.3}
\end{equation*}
$$

where

$$
e=\left[\begin{array}{c}
e_{1} \\
e_{2} \\
\vdots \\
e_{m}
\end{array}\right], \quad \quad p=\left[\begin{array}{c}
p_{1} \\
p_{2} \\
\vdots \\
p_{m}
\end{array}\right]
$$

and

Accordingly, the vector of actual crop proportions

$$
\begin{equation*}
p=P^{-1} e \tag{2.4}
\end{equation*}
$$

are obtained subject to $\sum_{i=1}^{m} p_{i}=1$ provided e and $P$ are known.

The vector of classified crop proportions, e, can only be known if the complete set of remotely sensed data is processed for total recognition using a classification algorithm. Hence, in general, e will be unknown. Regarding the classification error matrix $P$, two cases arise:
(i) P is known
(ii) P is unknown .

If $P$ is known, an unbiased estimate of $p$ is

$$
\begin{equation*}
\hat{p}=\mathrm{P}^{-1} \hat{e} \tag{2.5}
\end{equation*}
$$

But $P$ will generally be unknown. As such it would be necessary to obtain a certain amount of suitably selected gound truth in the area of interest, probably independent of the sample data used in estimating $e$, so that the classifier is trained and the classification error matrix estimated. Let $\hat{P}$ be an estimate of $P$. Then an estimate of $p$ when $P$ is unknown is given by

$$
\begin{equation*}
\hat{\hat{\mathrm{p}}}=\hat{\mathrm{P}}^{-1} \hat{\mathrm{e}} . \tag{2.6}
\end{equation*}
$$

Clearly, $\hat{\hat{p}}$ will generaily be a biased estimate. Both bias and mean $\hat{\wedge}$ square error of p will depend upon the performance of the classification algorithm as well as the degree to which the sample represents the population. The classification performance can be achieved desiredly by training the classifier sufficiently on the basis of a well representative sample for the ground truth. By adopting certain sampling schemes that may be suitable for the area of interest, appropriate samples for both the ground truth and the unlabeled remotely sensed data can be acquired. For our later discussion we assume that these two types of samples are independently obtained.

In appendix 1 we have derived general expressions for the covariance matrix of $p$, and both the bias and the mean square error matrix of $\hat{p}$. In the
case of $\hat{\mathrm{p}}$ not only the estimate $\hat{\hat{p}}$ itself but bias as well as mean square error quantities will also depend upon how $\hat{\mathrm{P}}$ is obtained. One solution for $\hat{P}$ and the probability distribution of its components is suggested therein, as well.

## 3. TWO-CROP APPROACH

Sometimes the main interest is in estimating the acreage of a specific crop type in the area of interest. In that case one approach to the acreage estimation problem lies in considering $\pi_{1}$ to be the specific crop type and $\pi_{0}$ to be the "other crop" consisting of the remainder of the crops, and then treating it as a two-crop situation. However, lumping of different crops together for the "other crop" would require certain caution and should be judged in terms of the classification performance for the two-crop case as against that for the case of the original set of crops. For the Gaussian maximum likelihood classifier, Basu and Odell (1973) have investigated this problem and have shown that the classification performance for the class of main interest may or may not improve when the classification is performed using the two-class approach. But the problem under this approach is greatly simplified and, barring extreme cases, perhaps it will provide satisfactory solutions in the remote sensing situation when interest only lies in ascertaining the acreage cover of one specific crop.

Now considering two crops $\pi_{1}$ and $\pi_{0}$, let $P(1 \mid 0)=\Phi_{1}$ and $P(0 \mid 1)=\Phi_{2}$ for the probabilities of misclassification when a certain classifier is used. We will assume that $\Phi_{1}+\Phi_{2} \neq 1$. Then

$$
P=\left[\begin{array}{ccc}
1 & -\Phi_{2} & \\
\Phi_{1} \\
\Phi_{2} & 1 & -\Phi_{1}
\end{array}\right]
$$

If $p_{1}$ and $p_{2}$ are the actual crop proportions of $\pi_{1}$ and $\pi_{0}$, respectively, whereas $e_{1}$ and $e_{2}$ are their respective expected classified crop proportions under the classifier used, it follows from (2.3) that

$$
\begin{equation*}
e_{1}=\left(1-\Phi_{2}\right) p_{1}+\Phi_{1}\left(1-p_{1}\right) \tag{3.1}
\end{equation*}
$$

and

$$
e_{2}=1-e_{1}
$$

On the other hand,

$$
\begin{equation*}
p_{1}=\frac{e_{1}-\Phi_{1}}{1-\Phi_{1}-\Phi_{2}} \tag{3.2}
\end{equation*}
$$

and

$$
p_{2}=\frac{e_{2}-\Phi_{2}}{1-\Phi_{1}-\Phi_{2}} \quad \text { or } \quad p_{2}=1-p_{1}
$$

Suppose from a random sample of $n$ unlabeled remotely sensed data points, $n_{1}$ points were classified into $\pi_{1}$ and $n_{2}=n-n_{1}$ points were classified into $\pi_{0}$ by the classifier. Then

$$
\begin{align*}
& \hat{\mathrm{e}}_{1}=\frac{\mathrm{n}_{1}}{\mathrm{n}},  \tag{3.3}\\
& \hat{\mathrm{p}}_{1}=\frac{\hat{\mathrm{e}}_{1}-\Phi_{1}}{1-\Phi_{1}-\Phi_{2}} \tag{3.4}
\end{align*}
$$

if $\Phi_{1}$ and $\Phi_{2}$ are known, and

$$
\begin{equation*}
\hat{\hat{p}}_{1}=\frac{\hat{\mathrm{e}}_{1}-\hat{\Phi}_{1}}{1-\hat{\Phi}_{1}-\hat{\Phi}_{2}} \tag{3.5}
\end{equation*}
$$

when $\Phi_{1}$ and $\Phi_{2}$ are unknown and have estimates $\hat{\Phi}_{1}$ and $\hat{\Phi}_{2}$ respectively. Clearly

$$
\begin{equation*}
\operatorname{Var}\left(\hat{p}_{1}\right)=\frac{1}{\left(1-\Phi_{1}-\Phi_{2}\right)^{2}} \operatorname{Var}\left(\hat{e}_{1}\right) \tag{3.6}
\end{equation*}
$$

For the estimate $\hat{\hat{p}}_{1}$ in (3.5), it easily follows that

$$
\begin{equation*}
\text { Bias }\left(\hat{p}_{1}\right)=\left(e_{1}-\Phi_{1}\right) E[T-\theta]-E\left[T\left(\hat{\Phi}_{1}-\Phi_{1}\right)\right] \tag{3.7}
\end{equation*}
$$

where

$$
T=\left(1-\hat{\Phi}_{1}-\hat{\Phi}_{2}\right)^{-1}
$$

and

$$
\theta=\left(1-\Phi_{1}-\Phi_{2}\right)^{-1}
$$

However, evalution of expected values in (3.7) may be quite a difficult job and so an exact bias value may not be accessible. An evaluation of the $\operatorname{MSE}\left(\hat{p}_{1}\right)$ in its exact form is even more difficult. As such we instead consider having it in the following approximate form obtained in Appendix 1 using the $\delta$-method. For a discussion on the method, see Rao (1965).
$\underset{\hat{M S E}}{\left(\hat{p}_{1}\right)}=\frac{1}{\left[1-\Phi_{1}-\Phi_{2}\right]^{2}}\left(\operatorname{Var}\left(\hat{\mathrm{e}}_{1}\right)+\left[1-\frac{\mathrm{e}_{1}-\Phi_{1}}{1-\Phi_{1} \Phi_{2}}\right]^{2} \operatorname{Var}\left(\hat{\Phi}_{1}\right)+\left[\frac{\mathrm{e}_{1}-\Phi_{1}}{1-\Phi_{1} \Phi_{2}}\right]^{2} \operatorname{Var}\left(\hat{\Phi}_{2}\right)\right)$
or
$\operatorname{MSE}\left(\hat{\hat{p}}_{1}\right)=\frac{1}{\left[1-\Phi_{1}-\Phi_{2}\right]^{2}}\left[\operatorname{Var}\left(\hat{e}_{1}\right)+\left(1-p_{1}\right)^{2} \operatorname{Var}\left(\hat{\Phi}_{1}\right)+\mathrm{p}_{1}^{2} \operatorname{Var}\left(\hat{\Phi}_{2}\right)\right]$
where $P_{1}$ is given by (3.2).
Sample Size
Considering simple random sampling with pixel as the sampling unit, we discuss the problem of sample size necessary to minimize the sampling cost or to achieve a desired amount of precision for the proportion estimate, given that the other is specified. Suppose total sample consists of $N=n+N_{1}+N_{2}$ data points selected randomly, where $n$ is the size of sample of unlabeled remotely sensed data used for estimating $e_{1}$, and $N_{1}$ and $N_{2}$ are sample sizes for ground truth data from $\pi_{1}$ and $\pi_{0}$ and are used to estimate $\Phi_{1}$ and $\Phi_{2}$, respectively. The estimates $\hat{\mathrm{e}}_{1}, \hat{\Phi}_{1}$ and $\hat{\Phi}_{2}$ are all obtained as observed sample proportions and thus it follows from (3.6) and (3.9) that

$$
\operatorname{Var}\left(\hat{p}_{1}\right)=\frac{e_{1}\left(1-e_{1}\right)}{n\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}
$$

and

$$
\begin{equation*}
\operatorname{MSE}\left(\hat{\hat{p}}_{1}\right)=\frac{1}{\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}\left[\frac{\mathrm{e}_{1}\left(1-\mathrm{e}_{1}\right)}{\mathrm{n}}+\left(1-\mathrm{p}_{1}\right)^{2} \frac{\Phi_{1}\left(1-\Phi_{1}\right)}{\mathrm{N}_{1}}+\mathrm{p}_{1}^{2} \frac{\Phi_{2}\left(1+\Phi_{2}\right)}{\mathrm{N}_{2}}\right] \tag{3.10}
\end{equation*}
$$

Suppose we want to obtain sample sizes necessary to minimize the sampling cost when $\operatorname{Var}\left(\hat{p}_{1}\right)$ and $\operatorname{MSE}\left(\hat{\mathrm{p}}_{1}\right)$ are specified, say each equal to or smaller than $\sigma^{2}$. In the case of $\Phi_{1}, \Phi_{2}$ known, the only cost involved is that of pracessing the remotely sensed sample data. Clearly, it will be minimum when the sample size $n$ is the smallest finteger greater than or equal to

$$
\begin{equation*}
\frac{e_{1}\left(1-e_{1}\right)}{\left(1-\Phi_{2}-\Phi_{2}\right)^{2} \sigma^{2}} \tag{3.11}
\end{equation*}
$$

For when $\Phi_{1}$ and $\Phi_{2}$ are unknown, there are two types of cost involved: one is the cost of processing the total sample data, say at the rate of $c_{1}$ dollars per data point and the other is the cost of obtaining ground truth, say at the rate of $c_{2}$ dollars per data point. Then the cost associated with a sample of size $N=n+N_{1}+N_{2}$ is of the form:

$$
C(N)=c_{1} n+\left(c_{1}+c_{2}\right)\left(N_{1}+N_{2}\right)
$$

The purpose is to find $N$ (i.e., $n, N_{1}$ and $N_{2}$ ) which minimize $C(N)$ subject to $\operatorname{MSE}\left(\hat{p}_{1}\right) \leq \sigma^{2}$. This is done in Appendix 2 where we derive explicit expressions for $n, N_{1}$ and $N_{2}$ in (A.9).

## 4. AN EXAMPLE

Certain sites in Hill County, Montana and Burke County, North Dakota were selected to investigate wheat identification performance for the ERTS-1 satellite data during 1973. For the sites in Hill County, there were three acquisition periods, covering both winter and spring wheat seasons, for which ERTS-1 labeled data were evaluated against the ground truth to ascertain wheat identification performance. In the case of the site in Burke County, there were only two acquisition periods covering the spring wheat season. For the classification identification performance results and other details, refer to Appendix 3.

Considering $\Phi_{1}$ to be the omission percentage for the non-wheat data points and $\Phi_{2}$ for the wheat data points, we give sample size results in Figure 1-7 for the various cases of omission percentages listed in Appendix 3, assuming different wheat proportions in the area and $\sigma_{2}=.01$. Based on these results, the following conclusions are drawn:

1. Expected labeled wheat proportion, $e_{1}$, increases as the actual proportion of wheat, $\mathrm{p}_{1}$, increases for the area, though not strictly. Though to a certain extent it depends upon the magnitude of the omission percentages for both non-wheat and wheat data points, it tends to centralize away from too low or too high values for the percentage.
2. Sample size for the unlabeled remotely sensed data first increases as the actual wheat proportion increases and then decreases later on; the point of decrease depends upon the size of the two omission percentages.
3. All sample sizes increase as the total omission rate $\Phi_{1}+\Phi_{2}$ increases.
4. Sample size, for the unlabeled remotely sensed data is much larger when $\Phi_{1}, \Phi_{2}$ are unknown compared to when these are known.
5. In the case of $\Phi_{1}, \Phi_{2}$ unknown, the sample size for the unlabeled remotely sensed data is proportional to $c_{2} / c_{1}$; the ratio of two types of cost.
6. Sample sizes for ground truth of wheat and non-wheat are inversely proportioned to $c_{2} / c_{1}$.
7. Sample size for the ground truth of wheat is larger than that for non-wheat when the expected labeled wheat proportion is below. 5 . Reverse is the case when such proportion is above .5. A similar trend holds against the actual wheat proportion, though not strictly.
8. Sample size for the ground truth increases as either of the two omission percentages increases when the other is held fixed.

For making a comparison of sample sizes irrespective of the wheat proportion which, in fact, is unknown, a suitable criterian is to determine the sample sizes against values for the coefficient of variation, C.V. $=\sigma / \mathrm{p}$. Generally the wheat coverage in any area of interest is expected to be somewhere in between 1 percent and 20 percent. As such we here give sample sizes for the unlabeled remotely sensed data and the ground truth of wheat as well as nonwheat by specifying $\sigma=.01$ and considering certain $C . V$. values in a 5 to 50 percent range. Numerical results are presented in Table 1 for all different cases of $\Phi_{1}, \Phi_{2}$ values that arise from the wheat identification performance
results given in Appendix 3. Moreover, for certain cases the sample sizes are sketched in Figure 8-14. The following conclusions are drawn:

1. All samples sizes increase as the total omission percentage $\Phi_{1}+\Phi_{2}$ increases.
2. Except for the sample size for the ground truth of wheat, sample sizes decrease as the coefficient of variation increases. These are generally very high in numbers for the 5 percent co-efficient of variation but levels off when the co-efficient of variation is 50 percent.
3. Sample size for the unlabeled remotely sensed data increases considerably if $\Phi_{1}, \Phi_{2}$ are unknown compared to their known case.
4. Again, all sample sizes depend upon the ratio $c_{2} / c_{1}$ as regards the two types of cost.
5. Sample size for the ground truth of wheat is consistently larger than that of non-wheat. Also, it shows very small changes over the range of co-efficients of variations being considered here. In cases where there is a high overall omission percentage, and particularly for the non-wheat, it tends to increase as the coefficient of variation increases.

TABLE 1: Sample sizes: $n$ for the unlabeled remotely sensed data, $N_{1}$ for the ground truth of wheat, and $\mathrm{N}_{2}$ for the ground truth of non-wheat when $\sigma=.01$


## 5. A LARGE AREA CROP ACREAGE ESTIMATION

Our previous discussion, in essence, applies to crop acreage inventory for an agricultural area which is homogeneous in respect to agricultural practices and thus is not expected to be large enough. Since a major objective of the JSC-EOD project is to perform or estimate crop acreages for a large area using available remote sensing capabilities, we here suggest a sampling procedure to procure sample data for the purpose of estimating a large area crop acreage inventory and discuss the error analysis associated with it.

Once again, we assume that the frame is made of agricultural areas; the non-agricultural areas in the region of interest can be easily excluded by way of a monitoring system. As a first step in the sampling procedure, we suggest having a geographical-based stratification which effects a division of the region into reasonably homogeneous areas with respect to physical and climatological conditions. Considering additional factors of (i) the predominance of various crop-types and (ii) the latitude and longitude, a finer stratification must be achieved. This is to obtain better discrimina-. tion for the underlying crop-types and to control variability which may otherwise dominate over the distinction that exists between the resolution classes for these crop-types.

Note that as a result of stratification one may only need to consider a part of the region for frame if crops of interest do not cover the whole region. So depending upon whether the frame would require consideration of the complete region or only a part of it, one should make a list of strata making up the frame for the purpose of sampling.

Remoting sensing data gathered by an ERTS satellite is documented in terms of scenes, each covering approximately an area of $100 \times 100 \mathrm{miles}$ and
divided into four strips where each strip has approximately 6,400 scanlines in it. As such, we suggest a three stage sampling plan to be independently carried out in each stratum: select randomly ERTS scenes at the first stage, strips within scenes at the second stage and scanlines within strips at the third stage. Of course, one may consider one more stage in selecting pixels within scanlines. However, sampling at this stage is excluded from the plan because it is inconvenient and uneconomical.

## Notations

Let $R$ be the region (in the sense of frame) of interest for estimating crop acreages. Suppose it is stratified into strata $R_{t}, t=1,2, \ldots, L$, with weights $W_{t}$, the proportion of pixels in $t$ th stratum, $t=1,2, \ldots, L$ so that

$$
R=\bigcup_{t=1}^{\mathrm{L}} \mathrm{R}_{t} \quad \text { with } \quad \sum_{t=1}^{L} w_{t}=1
$$

In stratum $R_{t}$, let $I_{t}$ be the number of scenes whereas $J, H$ and $n$ denote the number of strips per scene, number of scanlines per strip and number of pixels per scanlines, respectively. From the previous paragraph it is obvious that there is no need to distinguish between strata in the categories of strips per scene, scanlines per strip and pixels per scanline. Next, let $e_{t i j h}\left(\pi_{k}\right)$ be the expected proportion of pixels to be classified in $\pi_{k}$ from the hth scanline in $j$ th strip of $i t h$ scene for stratum $R_{t}, t=1,2, \ldots, L$. Then for $R_{t}$,

$$
e_{t i j}\left(\pi_{k}\right)=\sum_{h=1}^{H} e_{t i j h}\left(\pi_{k}\right)
$$

the expected proportion of pixels to be classified in $\pi_{k}$ from the $j$ th strip in ith scene,

$$
e_{t i}\left(\pi_{k}\right)=\sum_{j=1}^{J} \sum_{h=1}^{H} e_{t i j h}\left(\pi_{k}\right)
$$

the expected proportions of pixels to be classified in $\pi_{k}$ from the ith scene,

$$
e_{t}\left(\pi_{k}\right)=\sum_{i=1}^{I} \sum_{j=1}^{J} \sum_{h=1}^{H} e_{t i j h}\left(\pi_{k}\right),
$$

the expected proportion of pixels to be classified in $\pi_{k}$. Accordingly,

$$
\begin{equation*}
e\left(\pi_{k}\right)=\sum_{t=1}^{L} w_{t} e_{t}\left(\pi_{k}\right) \tag{5.1}
\end{equation*}
$$

is the expected proportion of pixels to be classified in $\pi_{k}, k=1,2, \ldots, m$, for the region $R$.

## Estimates

Suppose $m_{t}, r$ and $s$ denote the corresponding number of scenes, number of strips per scenes and number of scanlines per strip that one selected for $R_{t}, t=1,2, \ldots, L$, using the stratified three stage random sampling described earlier. Let $n_{\text {tijh }}\left(\pi_{k}\right)$ be the number of pixels classified into $\pi_{k}$ from the $h$ th selected scanline in $j$ th selected strip of the ith selected scene in $R_{t}$. Then considering the observed proportions of classified data points into different crops for estimates, one has

$$
\begin{aligned}
& \hat{e}_{t i j h}\left(\pi_{k}\right)=\frac{n_{t i j h}\left(\pi_{k}\right)}{n} \\
& \hat{e}_{t i j}\left(\pi_{k}\right)=\frac{1}{n s} \sum_{h=1}^{s} n_{t i j h}\left(\pi_{k}\right)
\end{aligned}
$$

$$
\begin{aligned}
& \hat{e}_{t i}\left(\pi_{k}\right)=\frac{1}{n s r} \sum_{j=1}^{r} \sum_{h=1}^{s} n_{t i j h}\left(\pi_{k}\right), \\
& \hat{e}_{t}\left(\pi_{k}\right)=\frac{1}{n s r m_{t}} \sum_{i=1}^{m} \sum_{j=1}^{r} \sum_{h=1}^{s} n_{t i j h}\left(\pi_{k}\right),
\end{aligned}
$$

and

$$
\begin{equation*}
\hat{e}\left(\pi_{k}\right)=\sum_{t=1}^{L} w_{t} \hat{e}_{t}\left(\pi_{k}\right), \quad k=1,2, \ldots, m . \tag{5.2}
\end{equation*}
$$

Next, expressions for $\operatorname{Var}\left(e_{t}\left(\pi_{k}\right)\right)$ and $\operatorname{Cov}\left(e_{t}\left(\pi_{k}\right), e_{t}\left(\pi_{k}\right)\right)\left(k \neq k^{1}\right)$ can be obtained without much difficulty. For example, refer to Section 10.8 in Cochran (1963) for the general discussion on three stage sampling plan. Hence, the covariance matrix of $\hat{e}$ is given by

$$
\operatorname{Var}\left(\hat{e}\left(\pi_{k}\right)\right)=\sum_{t=1}^{L} w_{t}^{2} \operatorname{Var}\left(\hat{e}_{t}\left(\pi_{k}\right)\right)
$$

$\operatorname{Cov}\left(\hat{e}\left(\pi_{k}\right), \hat{e}\left(\pi_{k},\right)\right)=\sum_{t=1}^{L} w_{t}^{2} \operatorname{Cov}\left(\hat{e}_{t}\left(\pi_{k}\right), \hat{e}_{t}\left(\pi_{k^{\prime}}\right)\right), k \neq k^{\prime}, \quad k=1,2, \ldots, m$.
Similarly, an estimate of the covariance matrix is obtained by replacing the unknown quantities by their estimates in (5.3). In this context, see Chhikara and Odell (1974) who have discussed such results in greater details. Now to obtain the actual crop proportions, there is a need to consider whether or not the classification error matrix is the same for each stratum. When the area is wide and large and the stratification is performed considering factors mentioned in the beginning of this section, it is quite likely that these classification error matrices will not be the same for different strata.

In that case, find an estimate of $\mathrm{p}_{\mathrm{k}}, \mathrm{k}=1,2, \ldots, \mathrm{~m}$, using (2.5) if the classification error matrix is known and (2.6) if it is unknown for each stratum. Denoting $p_{k}$ by $p_{t}\left(\pi_{k}\right)$ for stratum $R_{t}$, it then follows that

$$
\begin{equation*}
\hat{p}_{k}=\sum_{t=1}^{L} w_{t} \hat{p}_{t}\left(\pi_{k}\right) \quad, \quad k=1,2, \ldots, m \tag{5.4}
\end{equation*}
$$

when the classification matrices, say $P_{t}, t=1,2, \ldots, L$, are known, and

$$
\begin{equation*}
\hat{\hat{p}}_{k}=\sum_{t=1}^{L} \hat{W}_{t} \hat{p}_{t}\left(\pi_{k}\right) \quad \quad \mathrm{k}=1,2, \ldots, m \tag{5.5}
\end{equation*}
$$

when these are unknown and are separately estimated using ground truth data from each stratum. $\operatorname{Next}, \operatorname{Var}\left(\hat{\mathrm{p}}_{\mathrm{k}}\right)$ and $\operatorname{MSE}\left(\hat{\mathrm{p}}_{\mathrm{k}}\right)$ are respectively obtained from (A.2) and (A.7) after making an appropriate substitution from (5.2).

On the other hand, either there is the same classification error matrix for all strata or can be made so by proper adjustment of signatures in the classification algorithm for each stratum. For then an estimate of crop proportions $p_{k}, k=1,2, \ldots, m$ is directly given by (2.5) if the common classification error matrix is known and by (2.6) if it is unknown, using $\hat{e}\left(\pi_{k}\right)$, $\mathrm{k}=1,2, \ldots, \mathrm{~m}$ of (5.1) for e. Hence, both estimates and their error analyses are obtained by following the general procedure given in Section 2.

In fact, our approach in Section 2 is quite general and can be applied to perform any large area acreage inventory by considering an appropriate sampling scheme for both the unlabeled remotely sensed data and the ground truthed data.

Once again if interest lies in estimating only the wheat acreage, the two-crop approach of Section 3 can be applied. Then an estimate of wheat proportion is obtained from (3.4) or (3.5) as the case may be, either first
obtaining it stratumwise and then combining as we did above in (5.4) and (5.5) or directly, depending upon whether or not the classification error matrix is the same for different strata. Subsequently, the precision of this estimate and the sample size necessary to achieve a desired precision with minimum cost can be easily obtained by applying our technique of Section 3.

Sample Size
Taking the cost factor into consideration, suppose we want to determine the sample size that either minimizes the sampling cost for a specified precision or maximizes the precision of the estimate for a fixed cost. Though a large initial cost is involved in acquiring remotely sensed data, presently we are mainly concerned with the cost of the processing and labeling of the sampled data. In general, any such cost can be considered as

$$
c_{t}=c_{1} m_{t}+c_{2} m_{t} r+c_{3} m_{t} r s
$$

for the sample in stratum $R_{t}$, and

$$
c=\left(c_{1}+c_{2} r+c_{3} r s\right) \sum_{t=1}^{L} m_{t}
$$

for the area of interest.
In case of unknown classification error matrix or matrices, there is an additional cost of sampling the ground truth, say $C^{\prime}$. As such the total cost involved is $C=C+C^{\prime}$. Now if the cost is fixed, say $C^{\prime \prime} \leq C_{0}$, a determination of sample sizes for both the unlabeled remotely sensed data in all three categories and the ground truth for various crops can be achieved by solving equations obtained by equating the partial derivatives of

$$
\operatorname{MSE}\left(\hat{\hat{p}}_{k}\right)+\lambda\left(C^{\prime \prime}-C_{0}\right), \quad k=1,2, \ldots, m
$$

where $\lambda$ is a Lagrange multiplier, with respect to $\bar{m}_{t}, r, s$ and the ground truth sample sizes to zero. Similarly, when the MSE ( $\hat{p}_{k}$ ) is fixed, say $\sigma_{k}^{2}$, $\mathrm{k}=1,2, \ldots, \mathrm{~m}$, again this can be achieved by considering the function

$$
\left.\mathrm{C}^{\prime \prime}+\lambda_{\mathrm{k}}\left[\hat{\operatorname{MSE}\left(\hat{\mathrm{p}}_{\mathrm{k}}\right.}\right)-\sigma_{\mathrm{k}}^{2}\right], \quad \mathrm{k}=1,2, \ldots, \mathrm{~m}
$$

for minimization. This, of course, would lead to $k$ different values for various sample sizes unless we consider the minimization from the point of a specific crop-type proportion estimate. On the other hand, a unique determination can be obtained by considering the largest value obtained in each case.

It may be pointed out that under this procedure, it will be difficult to give any closed form expression for any sample size and its carrying out would involve some optimization technique.

If the classification error matrix (or matrices) is known, the sample sizes $m_{t}(t=1,2, \ldots, L), r$ and $s$ can be easily determined by minimizing $\operatorname{Var}\left(\hat{e}\left(\pi_{k}\right)\right)+\lambda\left(C-C_{0}\right)$ or $C+\lambda_{k}\left[\operatorname{Var}\left(\hat{e}\left(\pi_{k}\right)\right)-\sigma_{k}^{2}\right]$ as the case may be. Moreover, the sample size problem in the case of unknown classification error matrix or matrices can be treated either by assuming the classification errors known or by investigating the two types of sampling separately.

## 6. FURTHER REMARKS

In actual practice it may not be possible to have every data point identified with one of the crops in the area of interest, particularly if the area is large. This may be caused by not knowing all crop-types that exist in the area or some data points representing pixels falling on the field boundaries. As such the model developed in this report may be viewed somewhat restricted. Its use for performing a large area crop inventory may be considered subsequent to obtaining information about the agricultural practices in the area.

It is extremely difficult to model the problem of a large area crop inventory in its full generality unless certain constraints are imposed. The condition of identifiability is one such constraint that one must have in order to deal with the problem analytically.

## APPENDIX 1 .

A.1. Variances of Components of $\hat{p}$

For $\hat{p}$ given in (2.5), the covariance matrix,

$$
E\left[(\hat{p}-p)(\hat{p}-p)^{T}\right]=P^{-1} E\left[(\hat{e}-e)(\hat{e}-e)^{T}\right]\left(P^{-1}\right)^{T}
$$

or

$$
\begin{equation*}
\sum_{\hat{p}}=\left(P^{-1}\right) V\left(P^{-1}\right)^{T} \tag{A.1}
\end{equation*}
$$

where $V$ denotes the covariance matrix of $\hat{e}$. Denoting the $(i, j)^{\text {th }}$ element of $P^{-1}$ by $p^{i j}$, it follows that the variance of $\hat{p}_{i}$, the th element of $\hat{p}$, is given by

$$
\begin{equation*}
\operatorname{Var}\left(\hat{p}_{i}\right)=\sum_{j \neq 1}^{m}\left(P^{i j}\right)^{2} \operatorname{Var}\left(\hat{e}_{j}\right)+\sum_{\substack{j=1 \\ j \neq k}}^{m} \sum_{\substack{k=1}}^{m} P^{i j} P^{i k} \operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right) \tag{A.2}
\end{equation*}
$$

where $V\left(\hat{e}_{j}\right)$ and $\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right)$ would depend upon the sampling scheme used for obtaining samples of unlabeled remotely sensed data points.

In the case of random sampling with sampling unit as pixel (i.e. one data point),

$$
\begin{equation*}
\operatorname{Var}\left(\hat{e}_{j}\right)=\frac{e_{j}\left(1-e_{j}\right)}{n} \tag{A.3}
\end{equation*}
$$

and

$$
\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right)=-\frac{e_{j} e_{k}}{n}, j \neq k, j, k=1,2, \ldots, m
$$

ignoring the finite population correction due to large population size. Next
an unbiased estimate of these quantities is given by

$$
\begin{aligned}
& \widehat{\operatorname{Var}}\left(\hat{e}_{j}\right)=\frac{\hat{e}_{j}\left(1-\hat{e}_{j}\right)}{n-\hat{1}} \\
& \widehat{\operatorname{Cov}}\left(\hat{e}_{j}, \hat{e}_{k}\right)=-\frac{\hat{e}_{j} \hat{e}_{k}}{n-1}, j \neq k, j, k=1,2, \ldots, m .
\end{aligned}
$$

On the other hand if the sampling unit is a $5 \times 6$ mile segment consisting of $r$ pixels then considering a random sample of $m$ segments (here for the sample size one may consider $n=m r$ data points) from the total of $M$ segments in the area of interest and again ignoring the finite population correction, one gets (Cochran, 1963)

$$
\operatorname{Var}\left(\hat{e}_{j}\right)=\frac{1}{m(M-1)} \sum_{i=1}^{M}\left(e_{j i}-e_{j}\right)^{2}
$$

and

$$
\begin{gather*}
\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right)=\frac{1}{m(M-\hat{1})} \sum_{i=1}^{M}\left(e_{j i}-e_{j}\right)\left(e_{k i}-e_{k}\right), j \neq k  \tag{A.4}\\
\\
j, k=1,2, \ldots, m
\end{gather*}
$$

where $e_{j i}$ denotes the proportion of classified data points in $\boldsymbol{\pi}_{\mathbf{j}}$ for the ith segment. Once again, for their unbiased estimates

$$
\widehat{\operatorname{Var}}\left(\hat{e}_{j}\right)=\frac{1}{m(m-1)} \sum_{i=1}^{m}\left(\hat{e}_{j i}-\hat{e}_{j}\right)^{2}
$$

and

$$
\begin{gathered}
\widehat{\operatorname{Cov}}\left(\hat{e}_{j}, \hat{e}_{k}\right)=\frac{1}{m(m-1)} \sum_{i=1}^{m}\left(\hat{e}_{j i}-\hat{e}_{j}\right)\left(\hat{e}_{k i}-\hat{e}_{k}\right), i \neq k, \\
j, k=1,2, \ldots, m
\end{gathered}
$$

Similarly, components of V and their estimates can be obtained for other types of sampling plans. Making appropriate substitution in (A.1) or (A.2), variances for the components of $\hat{p}$ and their estimates are then obtained.
A.2. Mean Square Errors of Components of $\hat{\hat{p}}$.

$$
\begin{align*}
& \text { First we calculate the bias of } \hat{\hat{p}} \text { given by } \\
& \qquad \begin{aligned}
\text { Bias }(\hat{\hat{p}}) & =E[\hat{\hat{p}}-\mathrm{p}] \\
& =E\left[\hat{\mathrm{P}}^{-1} \hat{e}-\mathrm{P}^{-1} \mathrm{e}\right] \\
& =E\left[\hat{\mathrm{P}}^{-1}(\hat{e}-e)+\left(\hat{\mathrm{P}}^{-1}-\mathrm{p}^{-1}\right) \mathrm{e}\right] \\
& =E\left[\hat{\mathrm{P}}^{-1}-\mathrm{p}^{-1}\right] \mathrm{e}
\end{aligned}
\end{align*}
$$

because the first term is zero due $E(\hat{e}-e)=0$ for a given $\hat{\mathrm{P}}^{-1}$. Clearly, the bias depends upon how much bias there is in $\hat{\mathrm{P}}^{-1}$, and $\operatorname{Bias}(\hat{\hat{p}})=\left(\operatorname{Bias}\left(\hat{\mathrm{P}}^{-1}\right)\right)$ e.
In order to find the mean square error of any component of $\hat{p}$, let us first consider the evaluation of matrix,

$$
E\left[(\hat{\hat{p}}-p)(\hat{\hat{p}}-p)^{T}\right]=E\left[\left(\hat{p}^{-1} \hat{e}-p^{-1} e\right)\left(\hat{\mathrm{p}}^{-1} \hat{e}-P^{-1} e\right)^{T}\right]
$$

so

$$
\begin{align*}
& =E\left[\left(\hat{P}^{-1}\right)(\hat{e}-e)(\hat{e}-e)^{T}\left(\hat{P}^{-1}\right)^{T}+\left(\hat{P}^{-1}-P^{-1}\right) e e^{T}\left(\hat{P}^{-1}-P^{-1}\right)^{T}\right] \\
& =E\left[\left(\hat{P}^{-1}\right) V\left(\hat{P}^{-1}\right)^{T}\right]+E\left[\left(\hat{P}^{-1}-P^{-1}\right) e e^{T}\left(\hat{P}^{-1}-P^{-1}\right)^{T}\right] \tag{A.6}
\end{align*}
$$

where $E$ stands for expectation with respect to $\hat{P}$. Again, denoting the (i, j) th element of $P^{-1}$ by $P^{i j}$ and that of $\hat{P}^{-1}$ by $\hat{P}^{i j}$, it follows from (A.6) that the mean square error of $\hat{\hat{p}}_{i}$, the $i^{\text {th }}$ component of $\hat{\hat{p}}$, is given by

$$
\begin{align*}
& \operatorname{MSE}\left(\hat{\hat{p}}_{i}\right)=E\left[\sum_{j=1}^{m} \hat{p}^{i j}\right)^{2} \operatorname{Var}\left(\hat{e}_{j}\right)+\sum_{\substack{j=1 \\
j \neq k}}^{m} \sum_{\substack{m=1}}^{m} \hat{\mathrm{P}}^{i j} \hat{\mathrm{P}}^{i k} \operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right) \\
& \left.+\sum_{j=1}^{m} e_{j}^{2} E\left[\hat{p}^{i j}-p^{i j}\right]^{2}+\sum_{\substack{j=1 \\
j \neq k}}^{m} \sum_{k=1}^{m} e_{j} e_{k} E\left[\left(\hat{p}^{i j}-P^{i j}\right)\left(\hat{P}^{i k}-P^{i k}\right)\right]\right] \tag{A.7}
\end{align*}
$$

$\mathrm{i}=1,2, \ldots, \mathrm{~m}$.
Once again, v, i.e. $\operatorname{Var}\left(\hat{e}_{j}\right)$ and $\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right), j$ and $k=1,2, \ldots, m$, may be obtained as in (A.3) and (A.4). If some other sampling plan is used for selecting remotely sensed data to obtain the estimates $\hat{e}_{j}$ 's, expression for $V$ can accordingly be obtained. To evaluate expectation in (A.7), one needs to find the distribution of $\hat{P}$. This will, of course, depend upon how $\hat{\mathrm{P}}$ is obtained. In general, it will be difficult to obtain any exact distribution of $\hat{P}$. However, if the sampling of ground truth involves separate independent samples from each crop and $\hat{\mathrm{P}}$ is obtained as the matrix of observed proportions among randomly selected pixels classified
into different crops using a classifier, each column vector of $\hat{P}$ has a multinomial distribution and is stochaotically independent of the others in $\hat{P}$. Since expectation in (A.7) is for elements of $\hat{\mathrm{P}}^{-1}$, it may not be easy to derive the $\operatorname{MSE}\left(\hat{\mathrm{p}}_{i}\right)$ in a closed form, especially if the number of crops is large.

## APPENDIX 2

## Two-Crop Case

First we derive the $\operatorname{MSE}\left(\hat{p}_{1}\right)$ as in (3.8).
Proof of (3.8)
Considering the estimates $\hat{\Phi}_{1}, \hat{\Phi}_{2}$ and $\hat{e}$, being obtained from independent sets of samples, it follows by an application of the $\delta$-method that

$$
\begin{aligned}
& \operatorname{MSE}\left(\hat{\hat{p}}_{1}\right) \doteq\left(\frac{\partial \mathrm{p}_{1}}{\partial \mathrm{e}_{1}}\right)^{2} \quad \operatorname{\nabla ar}\left(\hat{\mathrm{e}}_{1}\right)+\left(\frac{\partial \mathrm{p}_{1}}{\partial \Phi_{1}}\right)^{2} \operatorname{Var}\left(\dot{\hat{\Phi}}_{1}\right)+\left(\frac{\partial \mathrm{p}_{1}}{\partial \Phi_{2}}\right)^{2} \operatorname{Var}\left(\hat{\Phi}_{2}\right) \\
& \doteq \frac{1}{\left(1-\Phi_{1}-\Phi_{2}\right)^{2}} \quad \operatorname{Var}\left(\hat{e}_{1}\right)+\left[\frac{e_{1}-1+\Phi_{2}}{\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}\right]^{2} \operatorname{Var}\left(\hat{\Phi}_{1}\right) \\
& +\left[\frac{e_{1}-\Phi_{1}}{\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}\right]^{2} \cdot \operatorname{Var}\left(\hat{\bar{\theta}}_{2}\right) \quad .
\end{aligned}
$$

Hence

$$
\begin{aligned}
\operatorname{MSE}\left(\hat{\hat{p}}_{1}\right) & \doteq \frac{1}{\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}\left(\operatorname{Var}\left(\hat{e}_{1}\right)+\left[1-\frac{e_{1}-\Phi_{1}}{1-\Phi_{1}-\Phi_{2}}\right]^{2} \cdot \operatorname{Var}\left(\hat{\Phi}_{1}\right)\right. \\
& \left.+\left[\frac{e_{1}-\Phi_{1}}{1-\Phi_{1}-\Phi_{2}}\right]^{2} \quad \operatorname{Var}(\hat{\Phi})\right)
\end{aligned}
$$

Here dot with equality sign means equality with approximation. This establishes (3.8).

For a determination of sample size necessary to minimize the cost subject to MSE $\left(\hat{\mathrm{P}}_{\mathrm{I}}\right) \leq \sigma^{2}$ as discussed in section 3 , it is achieved by minimizing the function

$$
F=C(N)+\lambda\left(\operatorname{MSE}\left(\hat{p}_{1}\right)-\sigma^{2}\right)
$$

with respect to $n, N_{1}$ and $N_{2}$, where $C$ (N) is given in (3.12) and $\operatorname{MSE}\left(\hat{p}_{1}\right)$ is given in (3:10). By rewriting, we have

$$
\begin{aligned}
& F=c_{1} n+\left(c_{1}+\dot{c}_{2}\right)\left(N_{1}+N_{2}\right)+\lambda\left(1-\Phi_{1}-\Phi_{2}\right)^{-2} \\
& \cdot\left[\frac{e_{1}\left(1-e_{1}\right)}{n}+\left(1-p_{1}\right)^{2} \frac{\Phi_{1}\left(1-\Phi_{1}\right)}{N_{1}}+p_{1}^{2} \frac{\Phi_{2}\left(1-\Phi_{2}\right)}{N_{2}}-\sigma^{2}\left(1-\Phi_{1}-\Phi_{2}\right)^{2}\right]
\end{aligned}
$$

Taking partial derivatives of $F$ with respect to $n, N_{1}$ and $N_{2}$ and equating each to zero, one obtains the following set of equations.

$$
\begin{aligned}
& c_{1}-\lambda\left(1-\Phi_{1}-\Phi_{2}\right)^{-2} \frac{e_{1}(1-e)}{n^{2}}=0 \\
& \left(c_{1}+\dot{c}_{2}\right)-\lambda\left(1-\Phi_{1}-\Phi_{2}\right)^{-2}\left(1-p_{1}\right)^{2} \frac{\Phi_{1}\left(1-\Phi_{1}\right)}{N_{1}^{2}}=0 \\
& \left(\dot{c}_{1}+c_{2}\right)-\lambda\left(1-\Phi_{1}-\Phi_{2}\right)^{-2} p_{1}^{2} \frac{\Phi_{2}\left(1-\Phi_{2}\right)}{N_{2}^{2}}=0
\end{aligned}
$$

Considering only the admissible solution of these equations, one has

$$
\begin{align*}
& n=\sqrt{\frac{\lambda e_{1}\left(1-e_{1}\right)}{c_{1}\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}} \\
& N_{1}=\left(1-p_{1}\right) \sqrt{\frac{\lambda \Phi_{1}\left(1-\Phi_{1}\right)}{\left(c_{1}+c_{2}\right)\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}}  \tag{A.8}\\
& N_{2}=p_{1} \sqrt{\frac{\lambda \Phi_{2}\left(1-\Phi_{2}\right)}{\left(c_{1}+c_{2}\right)\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}}
\end{align*}
$$

Considering that $\operatorname{MSE}\left(\hat{p}_{1}\right)=\sigma^{2}$ and making substitution in (3.10) for $n, N_{1}$ and $N_{2}$ obtained in (A.8), one gets

$$
\sqrt{\lambda}=\frac{1}{\sigma^{2}\left(1-\Phi_{1}-\Phi_{2}\right)}\left[\sqrt{c_{1} e_{1}\left(1-e_{1}\right)+}\left(1-p_{1}\right) \sqrt{\left(c_{1}+c_{2}\right) \Phi_{1}\left(1-\Phi_{1}\right)}+p_{1} \sqrt{\left(c_{1}+c_{2}\right) \Phi_{2}\left(1-\Phi_{2}\right)}\right] .
$$

After substituting for $\sqrt{\lambda}$ in (A.8), the sample sizes $n, N_{1}$ and $N_{2}$ are obtained as following:

$$
\begin{aligned}
& n=\sqrt{\frac{e_{1}\left(1-e_{1}\right) / c_{1}}{\sigma^{2}\left(1-\Phi_{1}-\Phi_{2}\right)^{2}}}\left[\sqrt{c_{1} e_{1}\left(1-e_{1}\right)}+\left(1-p_{1}\right) \sqrt{\left(c_{1}+c_{2}\right) \Phi_{1}\left(1-\Phi_{1}\right)}+p_{1} \sqrt{\left(c_{1}+c_{2}\right) \Phi_{2}\left(1-\Phi_{2}\right)}\right] \\
& \begin{array}{r}
N_{1}=\frac{\left(1-p_{1}\right)}{\sigma^{2}\left(1-\Phi_{1}-\Phi_{2}\right)^{2}} \sqrt{\frac{\Phi_{1}\left(1-\Phi_{1}\right)}{c_{1}+c_{2}}\left[\sqrt{\epsilon_{1} e_{1}\left(1-e_{1}\right)}+\left(1-p_{1}\right) \sqrt{\left(c_{1}+c_{2}\right) \Phi_{1}\left(1-\Phi_{1}\right)}\right.} \\
\left.+p_{1} \sqrt{\left(c_{1}+c_{2}\right) \Phi_{2}\left(1-\Phi_{2}\right)}\right]
\end{array} \\
& \begin{aligned}
N_{2}=\frac{p_{1}}{\sigma^{2}\left(1-\Phi_{1}-\Phi_{2}\right)^{2}} \sqrt{\frac{\Phi_{2}\left(1-\Phi_{2}\right)}{c_{1}+c_{2}}}\left[\sqrt{c_{1}{ }^{e\left(1-e_{1}\right)}+\left(1-p_{1}\right) \sqrt{\left(c_{1}+c_{2}\right) \Phi_{1}\left(1-\Phi_{1}\right)}}\right. \\
\left.+p_{1} \sqrt{\left(c_{1}+c_{2}\right) \Phi_{2}\left(1-\phi_{2}\right)}\right] \text { (A.9) }
\end{aligned}
\end{aligned}
$$

It can be easily seen that n is a monotone increasing and $\mathrm{N}_{1}, \mathrm{~N}_{2}$ are monotone decreasing functions in $c_{2} / c_{1}$, the ratio of two types of cost. For when $e_{1}, \Phi_{1}, \Phi_{2}$, are unknown, estimates of $n, N_{1}$ and $N_{2}$ can be obtained from (A.9) by replacing these unknown quantities by their estimates.

## APPENDIX 3

ERTS-1 DATA INVESTIGATION
FOR

WHEAT IDENTIFICATION

1. Hil1 County, Montana

- Complete ground for evaluation in $2 \times 6$ mile area in Hill County North
- Ground identifications of wheat, barley, oats in Hill County South
- ERTS-1 data evaluated at three acquisition periods covering spring and winter wheat seasons

| Date | Winter Wheat Stage | Spring Wheat Stage |
| :--- | :---: | :---: |
|  | Greening |  |
| May | Heading |  |
| June-emergence |  |  |
| July | Mature | $100 \%$ cover |
|  |  | Headed |

- Classification performance results:

$$
\begin{array}{r}
W \text { - Spring/Winter Wheat } \\
N W \text { - Oats/Barley/Pasture }
\end{array}
$$

Commission/Omission Percentages



## 2. Burke County, North Dakota

- Complete ground truth for evaluation in $2 \times 10$ mile area
: ERTS-1 data evaluated at two acquisition periods

Date Spring Wheat Stage
June $5 \quad 3 "-4^{\prime \prime}$ growth
June 23 Jointing

- Classification performance results:

W - Spring Wheat
NW - Barley/Oats/Pasture/Summer Fallow
Commission/Omission Percentages
$\left.\left.\begin{array}{r} \\ \mathrm{W} \\ \mathrm{NW}\end{array} \begin{array}{rr}\mathrm{W} & \mathrm{NW} \\ 75 & 25 \\ 10 & 90\end{array}\right] \quad \begin{array}{r}\mathrm{W}\end{array} \begin{array}{rr}\mathrm{W} & \mathrm{NW} \\ 85 & 15 \\ 10 & 90\end{array}\right]$ June $5\left(t_{1}\right) \quad$ June $23\left(t_{2}\right)$
$\left.\begin{array}{r} \\ \text { WW }\end{array} \begin{array}{rr}\text { W } & \text { NW } \\ 90 & 10 \\ 5 & 95\end{array}\right]$

June :5, June 23

$$
\left(t_{1}, t_{2}\right)
$$

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FIGURE I


Case: $\Phi_{1}, \Phi_{2}$ unknown and $c_{2} / c_{1}=5$
FIGURE 2




Case: $\Phi_{1}, \Phi_{2}$ unknown and $c_{2} / c_{1}=20$
Figure 5


Case: $\Phi_{1}, \Phi_{2}$ unknown and $c_{2} / c_{1}=20$
Figure 6


Case: $\Phi_{1}, \Phi_{2}$ unknown and $c_{2} / c_{1}=20 \quad$ Figure 7


Case: $\Phi_{1}, \Phi_{2}$ known
Figure 8


Figure 10




FIGURE I2
Figure 13

Case: $\Phi_{1}, \Phi_{2}$ unknown and $c_{2} / c_{1}=20$


IABLE : Sample sizes: $n$ for the unlabeled remotely sensed data, and $N_{1}$ and $N_{2}$ for the ground truth of wheat and non-wheat, respectively, when precision for the proportion estimate is specified by $\sigma=.1$.



# ESTIMATION OF OPTIMUM ERRORS OF CLASSIFICATION FOR UNIVARIATE NORMAL POPULATIONS 

by
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1. Introduction

For the multivariate normal populations, the problem of classification using the Bayes' discriminant procedure involves certain difficulties as to an exact evaluation of actual errors of classification, i.e. optimum probabilities of misclassification, and an optimal estimation of these errors, particularly in the case of small samples. If the populations are assumed to be univariate normal, these errors of classification are easily expressible in a close form. Yet even for this simplified case the problem of their estimation has not been fully examined. For example, to the authors' best knowledge no minimum variance unbiased estimates of these errors are given so far in the literature.

Recently Sorum [8] investigated the estimation of both expected and optimum errors of classification associated with the linear discriminant function for univariate normal populations with known common variance. For the optimum errors of classification, she considered the maximum likelihood estimates and their various slight variants in her study involving large samples. Hill [6] too considered the same problem but his investigation was primarily motivated towards examining expected errors of classification.

For a brief statement of the problem, let $\pi_{1}$ and $\pi_{2}$ be two normally distributed populations with means $\mu_{1}$ and $\mu_{2}$, respectively, and common variance $\sigma^{2}$. Having obtained an observation $x$ from one of the two populations, the problem is to identify the population to which it belongs. Assuming equal a priori probabilities for $\pi_{1}, \pi_{2}$ and a constant cost of misclassifying any observation, the Bayes' discriminant criterion amounts to: classify the observation $x$ into $\pi_{1}$ if

$$
x \geq \frac{\mu_{1}+\mu_{2}}{2} \quad \text { when } \mu_{1}-\mu_{2}>0
$$

and

$$
x \leq \frac{\mu_{1}+\mu_{2}}{2} \quad \text { when } \mu_{1}-\mu_{2}<0
$$

otherwise classify $x$ into $\pi_{2}$. Without loss of generality, we assume $\mu_{1}-\mu_{2}>0$.

Suppose $P(i \mid j)$ denotes the probability of classifying an observation $x$ from $\pi_{j}$ into $\pi_{j}, i$ and $j=1,2$. Then the actual errors and non-errors of classification under the above discriminant rule are

$$
P(i \mid j)= \begin{cases}\Phi\left(-\frac{\Delta}{2}\right) & \text { if } i \neq j  \tag{1.1}\\ 1-\Phi\left(-\frac{\Delta}{2}\right), & \text { otherwise }\end{cases}
$$

Sometimes it is desirable to have the knowledge of $P(i \mid j)$ 's. It provides an assessment of the confusion likely to arise between individuals of two groups and thereby it is helpful in correcting for bias, etc. For example, see Cochran [3] who deals with several aspects of statistical inference in presence of certain types of classification errors.

In this paper we consider the problem of estimating $P(i \mid j), i$ and $j=1,2$ when sample observations are available from the populations. In view of (1.1), one, however, only needs to consider the estimation of $\Phi(-\Delta / 2)$ when an estimation of $P(i \mid j)$ 's is desired.

Let $x_{1}, x_{2}, \ldots, x_{n}$ be a random sample from population $\pi_{1}$ and $y_{1}, y_{2}, \ldots, y_{n_{2}}$ be another random sample from population $\pi_{2}$. Considering the two parametric cases (i) $\mu_{1}, \mu_{2}$ unknown and $\sigma^{2}$ known and (ii) $\mu_{1}, \mu_{2}$ and $\sigma^{2}$ all unknown, the maximum likelihood estimate (MLE) of $\Phi(-\Delta / 2)$ is $\Phi(-\hat{\Delta} / 2)$ where $\hat{\Delta}=(\bar{x}-\bar{y}) / \sigma$ in case (i) and $\hat{\Delta}=(\bar{x}-\bar{y}) / s$ in case (i); here $\bar{x}$ and $\bar{y}$ denote the two sample means and $s$ is obtained by

$$
\left(n_{1}+n_{2}-2\right) s^{2}=\sum_{1}^{n_{1}}\left(x_{i}-\bar{x}\right)^{2}+\sum_{1}^{n_{2}}\left(y_{i}-\bar{y}\right)^{2}
$$

Below in section 2 we obtain the minimum variance unbiased estimate (MVUE) of $\Phi(-\Delta / 2)$ and then compare it with the MLE in section 3 by evaluating relative efficiencies of both MVUE and MLE with respect to the Rao-Cramer lower bound for variances of unbiased estimates of $\Phi(-\Delta / 2)$. It can be shown that the Rao-Cramer lower bound is

$$
\begin{equation*}
\frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right) \phi^{2}\left(-\frac{\Delta}{2}\right) \tag{1.2}
\end{equation*}
$$

in case (i) and

$$
\begin{equation*}
\frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}+\frac{\Delta^{2}}{2\left(n_{1}+n_{2}-1\right)}\right) \phi^{2}\left(-\frac{\Delta}{2}\right) \tag{1.3}
\end{equation*}
$$

in case (ii), where $\phi$ denotes the standard normal density function. The relative efficiency of an estimate is obtained by dividing the appropriate lower bound by the mean square error (MSE) of the estimate.
2. MVUE of $\Phi\left(-\frac{\Delta}{2}\right)$
(i) $\mu_{1}, \mu_{2}$ unknown and $\sigma^{2}$ known: The random variable

$$
\frac{1}{2}\left(z \sqrt{4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}+\frac{\bar{x}-\bar{y}}{\sigma}\right)
$$

where $Z$ is the standard normal variate, is distributed normally with mean $\frac{\Delta}{2}$ and variance one. As a result

$$
\begin{aligned}
\Phi\left(-\frac{\Delta}{2}\right) & =\operatorname{Prob}\left(z \sqrt{4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}+\frac{\bar{x}-\bar{y}}{\sigma} \leq 0\right) \\
& =E\left[\operatorname{Prob}\left(\left.z \leq-\frac{\bar{x}-\bar{y}}{\sigma \sqrt{4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}} \right\rvert\, \bar{x}, \bar{y}\right)\right] \\
& =E\left[\Phi\left(-\frac{\bar{x}-\bar{y}}{\left.\sigma \sqrt{4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}\right)}\right]\right.
\end{aligned}
$$

E denoting expectation with respect to $\bar{x}, \bar{y}$, a set of complete sufficient statistics. Thus by the RaO-Blackwell theorem we have the MVUE of $\Phi\left(-\frac{\Delta}{2}\right)$,

$$
\begin{equation*}
\hat{\Phi}\left(-\frac{\Delta}{2}\right)=\Phi\left(-\frac{\bar{x}-\bar{y}}{\sigma \sqrt{4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}}\right) . \tag{2.1}
\end{equation*}
$$

(ii) $\mu_{1}, \mu_{2}$ and $\sigma^{2}$, all unknown: Let $U$ be a beta variate, $\beta\left(\frac{v-1}{2}, \frac{v-1}{2}\right)$, stochastically independent of $s^{2}$ which is $\sigma^{2} x^{2} / v$ distributed with $v=\left(n_{1}+n_{2}-2\right)$ degrees of freedom (d.f.). Then by applying theorem 1 in Ellison [5], it follows that (2U-1) $\sqrt{v} s / \sigma$ has the standard normal distribution. Furthermore, the random variable

$$
\frac{1}{2 \sigma}\left[(2 u-1) s \sqrt{v\left(4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right.}+(\bar{x}-\bar{y})\right]
$$

has the normal distribution with mean $\Delta / 2$ and variance one. Accordingly,

$$
\begin{aligned}
\Phi\left(-\frac{\Delta}{2}\right) & =\operatorname{Prob}\left\{(2 U-1) s \sqrt{u\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right]\right.}+(\bar{x}-\bar{y}) \leq 0\right\} \\
& =E\left[\operatorname{Prob}\left\{\left.(2 U-1) s \sqrt{v\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right]}+(\bar{x}-\bar{y}) \leq 0 \right\rvert\, \bar{x}, \bar{y}, s\right\}\right]
\end{aligned}
$$

where $E$ denotes expectation with respect to the set of complete sufficient statistics $\vec{x}, \vec{y}$ and $s$. Thus the MVUE of $\Phi\left(-\frac{\Delta}{2}\right)$ is

$$
\begin{align*}
\hat{\Phi}\left(-\frac{\Delta}{2}\right) & =\operatorname{Prob}\left\{\left.(2 U-1) s \sqrt{\nu\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right]}+(\bar{x}-\bar{y}) \leq 0 \right\rvert\, \bar{x}, \bar{y}, s\right\} \\
& =\operatorname{Prob}\left\{\left.U \leq \frac{1}{2}-(\bar{x}-\bar{y}) / 2 s \sqrt{v\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right]} \right\rvert\, \bar{x}, \bar{y}, s\right\} \tag{2.2}
\end{align*}
$$

where $U$ has a $\beta\left(\frac{v-1}{2}, \frac{v-1}{2}\right)$ distribution. Since extensive incomplete-beta integral tables are available, (2.2) can be easily evaluated for any given values of $\bar{x}, \bar{y}$ and $s$ obtained from sample observations. Next, denoting

$$
w=\frac{1}{2}-\frac{\bar{x}-\bar{y}}{2 s \sqrt{v\left(4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right)}}
$$

(2.2) can be rewritten in the form

$$
\begin{equation*}
\hat{\Phi}\left(-\frac{\Delta}{2}\right)=\frac{1}{B\left(\frac{v-1}{2} \frac{v-1}{2}\right)} \int_{0}^{w}[u(1-u)]^{(v-3) / 2} d u \tag{2.3}
\end{equation*}
$$

3. Relative Efficiencies of MVUE and MLE of $\Phi\left(-\frac{\Delta}{2}\right)$

First we derive formulas for the variance of MVUE, $\hat{\Phi}(-\Delta / 2)$, and the MSE of MLE, $\Phi(-\hat{\Delta} / 2)$, in forms suitable for numerical computations, and then present their values as well as relative efficiencies of both types of estimates for certain parametric values of $\Delta$ in each of the two cases (i) and (ii).
3.1. R.E. of the MVUE

Case (i). Noting that $v=\frac{\bar{x}-\bar{y}}{\sigma}$ is distributed normally with mean $\Delta$ and variance $\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)$ and

$$
\hat{\Phi}\left(-\frac{\Delta}{2}\right)=\operatorname{Prob}\left\{\left.z \leq-\frac{v}{\left.\sqrt{4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right.}\right)} \right\rvert\, v\right\},
$$

we have

$$
\begin{align*}
E\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]^{2} & =\int_{-\infty}^{\infty} \Phi^{2}\left(-\frac{v}{\sqrt{4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right.}}\right) \phi\left(\frac{v-\Delta}{\sqrt{\frac{1}{n_{1}}+\frac{1}{n_{2}}}}\right) d v \\
& =\Phi\left(-\frac{\Delta}{2},-\frac{\Delta}{2}, \frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right) \tag{3.1}
\end{align*}
$$

where $\Phi(x, y ; \rho)$ is the bivariate standard normal c.d.f. The result in (3.1) easily follows by a simple probability argument, e.g. see Zacks and Evens [9]. Hence

$$
\begin{equation*}
\operatorname{Var}\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]=\Phi\left(-\frac{\Delta}{2},-\frac{\Delta}{2} ; \frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right)-\Phi^{2}\left(-\frac{\Delta}{2}\right) \tag{3.2}
\end{equation*}
$$

and

$$
\begin{equation*}
\text { R.E. }\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]=\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right) \phi^{2}\left(-\frac{\Delta}{2}\right) / 4 \operatorname{Var}\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right] . \tag{3.3}
\end{equation*}
$$

Case (ii). Letting $t=(\bar{x}-\bar{y}) / s\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)^{1 / 2}$, a non-central student $t$ variate with $v$ d.f. and non-centrality parameter $\Delta /\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)^{1 / 2}$, (2.2) can be written as

$$
\hat{\Phi}\left(-\frac{\Delta}{2}\right)=\operatorname{Prob}\left\{\left.(2 U-1) \leq-t \sqrt{\frac{1}{n_{1}}+\frac{1}{n_{2}}} / \sqrt{\nu\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right]} \right\rvert\, t\right\}
$$

Furthermore, letting $U_{1}$ and $U_{2}$ be two independent beta variables, each distributed as $B\left(\frac{v-1}{2}, \frac{v-1}{2}\right)$ and considering $W=\max \left(2 U_{1}-1,2 U_{2}-1\right)$, it follows that the density function of $W$ is

$$
f(w)=\frac{1}{2^{v-3_{B}\left(\frac{v-1}{2}, \frac{v-1}{2}\right)}\left(1-w^{2}\right)^{\frac{v-3}{2}} I_{(1+w) / 2}\left(\frac{v-1}{2}, \frac{v-1}{2}\right),-1 \leq w \leq 1}
$$

where

$$
I_{x}(a, b)=\frac{1}{B(a, b)} \int_{0}^{x} y^{a-1}(1-y)^{b-1} d y .
$$

Now we can write

$$
\begin{aligned}
{\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]^{2} } & =\operatorname{Prob}\left(\left.W \leq-\frac{t \sqrt{\frac{1}{n_{1}}+\frac{1}{n_{2}}}}{\sqrt{v\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right]}} \right\rvert\, t\right) \\
& =F\left(\left.-\frac{t \sqrt{\frac{1}{n_{1}}+\frac{1}{n_{2}}}}{\sqrt{\nu\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right]}} \right\rvert\, t\right),(\text { say }) .
\end{aligned}
$$

Then

$$
\begin{equation*}
E\left[\hat{\dot{\phi}}\left(-\frac{\Delta}{2}\right)\right]^{2}=\int_{-\infty}^{\infty} F\left(-\frac{t \sqrt{\frac{1}{n_{1}}+\frac{1}{n_{2}}}}{\sqrt{\nu\left[4-\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)\right]}}\right) \quad g(t ; \delta, \nu) d t \tag{3.4}
\end{equation*}
$$

where $\mathrm{g}(\mathrm{t} ; \delta, \mathrm{v})$ is a non-central student t density function with non-centrality parameter $\delta=\Delta /\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)^{1 / 2}$ and $v$ d.f. For numerical integration of the right side in (3.4), we have considered the following forms for the functions $F(w)$ and $g(t ; \delta, v)$ :

$$
\begin{aligned}
F(w)= & \frac{4}{(v-1) B^{2}\left(\frac{v-1}{2}, \frac{v-1}{2}\right)}\left[\int_{0}^{(1-w) / 2} y^{v-2}(1-y)^{v-2} d y\right. \\
& \left.+\sum_{k=0}^{\infty} \frac{B\left(\frac{v+1}{2}, k+1\right)}{B(v-1, k+1)} \int_{0}^{(1+w) / 2} y^{k+v-1}(1-y)^{v-2} d y\right]
\end{aligned}
$$

where exact integration was obtained using recursive scheme, and

$$
g(t, \delta, v)=\frac{1}{\sqrt{2 \pi v} 2^{v / 2} \Gamma\left(\frac{v}{2}\right)} \int_{0}^{\infty} y^{(v-1) / 2} \exp \left[-\frac{1}{2}\left\{\left(t \sqrt{\frac{y}{v}}-\delta\right)^{2}-y\right\}\right] d y
$$

where the 32 -point Gauss-Laguerre quadrature formula was used for numerical integrations. The final integration in (3.4) was done using the Rhomberg method of numerical integration. All of these evaluations were on double precision basis.

After finding the variance of $\hat{\Phi}\left(-\frac{\Delta}{2}\right)$, we now obtain
R.E. $\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]=\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}+\frac{\Delta^{2}}{2\left(n_{1}+n_{2}-1\right)}\right) \phi^{2}\left(-\frac{\Delta}{2}\right) / 4 \operatorname{Var}\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]$.

### 3.2. R.E. of the MLE

For the MSE and Bias of the MLE, we need to find both $E[\Phi(-\hat{\Delta} / 2)]$ and $E\left[\Phi^{2}(-\hat{\Delta} / 2]\right.$. We will be using the following result of Ellison [5] in our discussion.

If $X$ is a normal variate with mean $\mu$ and variance $\sigma^{2}$, then

$$
\begin{equation*}
E[\Phi(X)]=\Phi\left(\frac{\mu}{\sqrt{1+\sigma^{2}}}\right) . \tag{3.6}
\end{equation*}
$$

Case (i): Since $-(\bar{x}-\bar{y}) / 2 \sigma$ has the normal distribution with mean $-\Delta / 2$ and variance $\frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)$, it follows from (3.6) that

$$
E\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]=\Phi\left(-\frac{\Delta}{2 \sqrt{1+\frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}}\right)
$$

Again, it can be easily shown that

$$
\begin{equation*}
E\left[\Phi^{2}\left(-\frac{\hat{\Delta}}{2}\right)\right]=\Phi\left(-\frac{\Delta}{2 \sqrt{1+\frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}},-\frac{\Delta}{2 \sqrt{1+\frac{1}{4}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}},\left(1+\frac{4 n_{1}^{n} n_{2}}{n_{1}^{+n_{2}}}\right)^{-1}\right) \tag{3.7}
\end{equation*}
$$

Hence

$$
\begin{equation*}
\operatorname{MSE}\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]=\Phi\left(\xi, \xi,\left(1+\frac{4 n_{1} n_{2}}{n_{1}+n_{2}}\right)^{-1}-2 \Phi(\xi) \Phi\left(-\frac{\Delta}{2}\right)+\Phi^{2}\left(-\frac{\Delta}{2}\right)\right. \tag{3.8}
\end{equation*}
$$

and

$$
\begin{equation*}
\operatorname{Bias}\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]=\Phi(\xi)-\Phi\left(-\frac{\Delta}{2}\right) \tag{3.9}
\end{equation*}
$$

where

$$
\xi=-\frac{\Delta}{2 \sqrt{1+\frac{1}{4}}\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}
$$

Next, the relative efficiency of $\Phi\left(-\frac{\hat{\Delta}}{2}\right)$,

$$
\begin{equation*}
\text { R.E. }\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]=\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right) \dot{\Phi}^{2}\left(-\frac{\Delta}{2}\right) / 4 \operatorname{MSE}\left[\Phi\left(-\frac{\dot{\hat{\Delta}}}{2}\right)\right] \text {. } \tag{3.10}
\end{equation*}
$$

Case (ii). Given $s^{2}$, conditionally $(\bar{x}-\bar{y}) / 2 s$ has a normal distribution with mean $\left(\mu_{1}-\mu_{2}\right) / 2 s$ and variance $\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right) \sigma^{2} / 4 s^{2}$. So by applying (3.6), we have

$$
\begin{aligned}
E\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right] & =E\left[E\left[\left.\Phi\left(-\frac{\hat{\Delta}}{2}\right) \right\rvert\, s^{2}\right]\right] \\
& =E\left[\Phi\left(-\Delta / \sqrt{\frac{4 s^{2}}{\sigma^{2}}+\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}\right)\right]
\end{aligned}
$$

Since $Q=v s^{2} / \sigma^{2}$ has a $\chi^{2}$ distribution with vd.f.,

$$
\begin{equation*}
E\left[\Phi\left(-\frac{\Delta}{2}\right)\right]=E\left[\Phi\left(-\Delta / \sqrt{\frac{4 Q}{v}+\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}\right)\right] \tag{3.11}
\end{equation*}
$$

where the expectation on the right side is with respect to the random variable Q.

Next by drawing analogy with the previous result in (3.7), it can easily be shown that

$$
\begin{align*}
E\left[\Phi^{2}\left(-\frac{\Delta}{2}\right)\right] & =E\left[E\left[\left.\Phi^{2}\left(-\frac{\Delta}{2}\right) \right\rvert\, s^{2}\right]\right] \\
& =E\left[\Phi\left(-\frac{\Delta}{\sqrt{\frac{4 Q}{v}+\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}},-\frac{\Delta}{\sqrt{\frac{4 Q}{v}+\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}}, \frac{1}{1+\frac{4 Q}{v}}\right)\right] . \tag{3.12}
\end{align*}
$$

Accordingly,

$$
\begin{equation*}
\operatorname{MSE}\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]=E\left[\Phi\left(n, n, \frac{1}{1+4 Q / v}\right)\right]-2 E[\Phi(n)] \Phi\left(-\frac{\Delta}{2}\right)+\Phi^{2}\left(-\frac{\Delta}{2}\right) \tag{3.13}
\end{equation*}
$$

and

$$
\begin{equation*}
\text { Bias }\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]=E[\Phi(\eta)]-\Phi\left(-\frac{\Delta}{2}\right) \tag{3.14}
\end{equation*}
$$

where

$$
n=-\frac{\Delta}{\sqrt{\frac{4 Q}{v}+\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}\right)}}
$$

Once again,

$$
\begin{equation*}
\text { R.E. }\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]=\left(\frac{1}{n_{1}}+\frac{1}{n_{2}}+\frac{\Delta^{2}}{2\left(n_{1}+n_{2}-1\right)}\right) \Phi^{2}\left(-\frac{\Delta}{2}\right) / 4 \operatorname{MSE}\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right] . \tag{3.15}
\end{equation*}
$$

Expressions in (3.1), (3.7) are in terms of the standard bivariate normal c.d.f. and can be easily evaluated using the method by Owen (1956), among others, for any given values of $n_{1}, n_{2}$ and $\Delta$. Next, for the final numerical integration in (3.11) and (3.12), we employed the Gauss-Laguarre quadrature formula.

## 4. Numerical Results

Certain numerical results are presented in tables 1 and 2 considering $n_{1}=n_{2}=n$ and specifying values for $n$ and $\Delta: n=5,10,15,30$ and $\Delta=.5,1.0,1.5$, $2.0,2.5,3.0$ in table 1 for $\sigma^{2}$ known case and $n=5,10$ and $\Delta=.5,1.0,1.5,2.0$, $2.5,3.0,3.5$ in table 2 for $\sigma^{2}$ unknown case. (Due to limited computational facilities, we considered only two values for $n$ in the latter case). The presented results are mainly designed to exemplify the comparison of the MVUE and the MLE in small sample case.

For the case of known $\sigma^{2}$ (figure 1), bias of the MLE is non-negative for all values of $\Delta$, maximizing at $\Delta=2.0$. But when $\sigma^{2}$ is unknown (figure 2), bias is negative for $\Delta$ equal or less than 2.0 and positive for $\Delta$ greater than 2.0. However, in both cases, as one would expect, it decreases uniformly as the sample size $n$ increases and is zero when $\Delta=0$.

Further, interestingly enough, we observe that whether or not $\sigma^{2}$ is known makes a great difference in relative efficiences of the two types of estimates. The relative efficiency of MLE is higher than that of MVUE for smaller values of $\Delta$ (i.e. when probability of misclassification is higher) when $\sigma^{2}$ is known (Table 1). But reverse is the case when $\sigma^{2}$ is unknown (Table 2).

TABLE 1
Relative Efficiencies of MVUE and MLE of $\phi\left(-\frac{\Delta}{2}\right)$ for $\sigma^{2}$ Known Case

| $\Delta$ | $n$ | $\operatorname{Var}\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]$ | $\operatorname{MSE}\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]$ | R.E. $\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]$ | R.E. $\left[\Phi\left(-\frac{\Delta}{2}\right)\right]$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| . 5 | 5 | . 01502 | . 01374 | . 9955 | 1.0881 |
|  | 10 | . 00749 | . 00716 | . 9982 | 1.0044 |
|  | 15 | . 00499 | . 00484 | . 9989 | 1.0297 |
|  | 30 | . 00249 | . 00245 | . 9995 | 1.0149 |
| 1.0 | 5 | . 01256 | . 01172 | . 9867 | 1.0573 |
|  | 10 | . 00624 | . 00603 | . 99361 | 1.0285 |
|  | 15 | . 00415 | . 00406 | . 9958 | 1.0189 |
|  | 30 | . 00207 | . 00205 | . 9980 | 1.0095 |
| 1.5 | 5 | . 00933 | . 00899 | . 9722 | 1.0082 |
|  | 10 | . 00460 | . 00452 | . 9861 | 1.0026 |
|  | 15 | . 00305 | . 00302 | . 9907 | 1.0014 |
|  | 30 | . 00152 | . 00151 | . 9954 | 1.0005 |
| 2.0 | 5 | . 00615 | . 00620 | . 9523 | . 9438 |
|  | 10 | . 00300 | . 00302 | . 9756 | . 9677 |
|  | 15 | . 00198 | . 00200 | . 9836 | . 9774 |
|  | 30 | . 00098 | . 00099 | . 9917 | . 9881 |
| 2.5 | 5 | . 00360 | . 00384 | . 9274 | . 8685 |
|  | 10 | . 00173 | . 00180 | . 9623 | . 9249 |
|  | 15 | . 00114 | .00117 | . 9745 | . 9475 |
|  | 30 | . 00056 | . 00057 | . 9871 | . 9725 |
| 3.0 | 5 | . 00186 | . 00212 | . 9014 | . 7907 |
|  | 10 | . 00088 | . 00095 | . 9500 | . 8798 |
|  | 15 | . 00058 | . 00061 | . 9653 | . 9144 |
|  | 30 | . 00028 | . 00029 | . 9816 | . 9538 |

TABLE 2
Relative Efficiencies of MVUE and MLE of $\Phi\left(-\frac{\Delta}{2}\right)$ for $\sigma^{2}$ Unknown Case

| $\triangle$ | $n$ | $\operatorname{Var}\left[\hat{\Phi}\left(-\frac{\Delta}{2}\right)\right]$ | $\operatorname{MSE}\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]$ | R.E. $\left[\hat{\phi}\left(-\frac{\Delta}{2}\right)\right]$ | $\text { R.E. }\left[\Phi\left(-\frac{\hat{\Delta}}{2}\right)\right]$ |
| :---: | :---: | :---: | :---: | :---: | :---: |
| . 5 | 5 | . 01639 | . 01726 | . 9438 | . 8965 |
|  | 10 | . 00780 | . 00814 | . 9900 | . 9485 |
| 1.0 | 5 | . 01491 | . 01494 | . 9467 | . 9448 |
|  | 10 | . 00712 | . 00722 | . 9843 | . 9714 |
| 1.5 | 5 | . 01249 | . 01178 | . 9526 | 1.0106 |
|  | 10 | . 00597 | . 00586 | . 9851 | 1.0027 |
| 2.0 | 5 | . 00951 | . 00849 | . 9576 | 1.0730 |
|  | 10 | . 00454 | . 00432 | . 9845 | 1.0340 |
| 2.5 | 5 | . 00653 | . 00564 | . 9549 | 1.1052 |
|  | 10 | . 00309 | . 00289 | . 9832 | 1.0514 |
| 3.0 | 5 | . 00403 | . 00346 | . 9374 | 1.0906 |
|  | 10 | . 00188 | . 00175 | . 9747 | 1.0441 |
| 3.5 | 5 | . 00224 | . 00189 | . 8988 | 1.0615 |
|  | 10 | . 00102 | . 00091 | . 9543 | 1.0717 |



Figure 1: Bias of the MLE of the true probability of misclassification when $\sigma^{2}$ is known.


Figure 2: Bias of the MLE of the true probability of misclassification when $\sigma^{2}$ is unknown.

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# -A SIMULATION STUDY OF POPULATION CLASSIFICATION USING SELECTED VARIATES 

## by

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## Summary

Considering the problem of estimating the optimum probability of misclassification under the maximum likelihood discriminant rule for univariate normal populations with common variance, we give its minimum variance unbiased estimate (MVUE) and derive the variance of the estimate for both cases of variance known and unknown. Next, we obtain the mean square error of the maximum likelihood estimate (MLE) and compare the two estimates by evaluating their relative efficiencies with respect to the RaoCramer lower bound for variances of any unbiased estimates. Also, bias of the MLE is investigated.

## CHAPTER I

## INTRODUCTION

## The Bayesian Solution of the Discrimination Problem

Consider $m$ populations $\Pi_{1}, \ldots, \Pi_{m}$ and suppose that each individual in the union of these populations possesses $p$ common observable characteristics $C_{1}, \ldots, C_{p}$. Let the observed values of an individual be denoted by $x=\left(x_{1}, \ldots, x_{p}\right)^{T}$, where $x_{j}$ denotes the observed values of $C_{j}$. The classical discriminant analysis problem consists of formulating a technique to divide the space of observations into m mutually exclusive and exhaustive regions $R_{1}, \ldots . R_{m}$ (hence classifying the individual in population $\Pi_{j}$ if $X$ falls in region $R_{j}$ ) in a manner such that the cost of misclassifying the individual is minimized.

There have been various techniques proposed for solving the problem, of which the Bayesian solution is optimal, in the sense that it minimizes the expected cost of misclassification. Anderson [1] states the Bayesian solution as follows:
"If $q_{i}$ is the a priori probability of drawing an observation from population $\Pi_{i}$ with density $p_{i}(X)(i=1, \ldots, m)$ and if the cost of misclassifying an observation from $\Pi_{i}$ as from $\Pi_{j}$ is
$C(j \mid i)$, then the regions of classification, $R_{1}, \ldots, R_{m}$, that minimize the expected cost are defined by assigning $X$ to $R_{k}$ if:

$$
\begin{equation*}
\sum_{\substack{i=1 \\ i \neq k}}^{m} q_{i} p_{i}(x) C(k \mid i)<\sum_{\substack{i=1 \\ i \neq j}}^{m} q_{i} p_{i}(x) C(j \mid i) \tag{1}
\end{equation*}
$$

for $j=1, \ldots, m, j \neq k . "$

## Estimation of Population Parameters

In actual situations the values $q_{1}, \ldots, q_{m}, C(i \mid j)$ for all $i, j=1, \ldots, m$, and the probability densities $p_{1}(X)$, $\mathrm{P}_{2}(\mathrm{X}), \ldots, \mathrm{p}_{\mathrm{m}}(\mathrm{X})$ may not be known. One solution is to assume the populations are equally likely, $C(i \mid j)=0$ for $i=j$ and $C(i \mid j)$ is constant for $i \neq j$, and the populations are normal, that is, when $X$ is from population $\Pi_{i}$,

$$
\begin{equation*}
X \sim N\left(\mu_{i}, \Sigma_{i}\right) \tag{2}
\end{equation*}
$$

where $\mu_{i}$ is the mean vector and $\Sigma_{i}$ is the covariance matrix of the ith population. In addition, if the parameters $\mu_{i}$ and $\Sigma_{i}, i=1, \ldots, m$, are unknown and a random sample of size $n_{i}$ (denoted $x_{1}^{(i)}, x_{2}^{(i)}, \ldots, x_{n_{i}}^{(i)}$ ) can be obtained from the ith population, then $\mu_{i}$ and $\Sigma_{i}$ can be estimated by the estimators $\bar{X}_{i}$ and $\hat{\Sigma}_{i}$ given by

$$
\begin{equation*}
\bar{x}_{i}=\frac{1}{n_{i}} \sum_{j=1}^{n_{i}} x_{j}^{(i)} \tag{3}
\end{equation*}
$$

and

$$
\begin{equation*}
\hat{\Sigma}_{i}=\frac{1}{\bar{n}_{i}-1} \sum_{j=1}^{n_{i}}\left(x_{j}^{(i)}-\bar{x}_{i}\right)\left(x_{j}^{(i)}-\bar{x}_{i}\right)^{T} \tag{4}
\end{equation*}
$$

respectively [1]. One can then estimate

$$
\begin{equation*}
P_{i}(X)=P_{i}\left(X, \mu_{i}, \Sigma_{i}\right) \tag{5}
\end{equation*}
$$

by

$$
\begin{equation*}
\hat{p}_{i}(X)=p_{i}\left(X ; \bar{X}_{i}, \hat{\Sigma}_{i}\right) \tag{6}
\end{equation*}
$$

or

$$
\begin{equation*}
\hat{p}_{i}(x)=(2 \pi)^{-\frac{p}{2}}\left|\hat{\varepsilon}_{i}\right|^{-\frac{1}{2}} \exp \left[-\frac{1}{2}\left(x-\bar{x}_{i}\right)^{T} \hat{\varepsilon}_{i}^{-1}\left(x-\bar{x}_{i}\right)\right] \tag{7}
\end{equation*}
$$

An individual with the observation vector $X=X_{o}$ will be classified as belonging to the jth population if

$$
\begin{equation*}
\max \hat{p}_{i}\left(X_{0}\right)=\hat{p}_{j}\left(X_{0}\right), i=1,2, \ldots, m \tag{8}
\end{equation*}
$$

## CHAPTER II

DIMENSION REDUCTION AND VARIATE SELECTION

## Misclassification as a Function of Separation

If $p$, the vector length, is large, the problem of classifying individuals becomes quite cumbersome and time consuming $[7,9]$. For example in the analysis of remote sensing data [10], when $p=12$ the amount of computational time is immense. Thus it is desirable to reduce $p$ while maintaining near optimal results. One technique is to pick an acceptable number of characteristics, say $s, s \leq p$, based on an examination of a measure of the separation between the populations, and proceed with Bayesian classification procedure using the s-variate marginal densities. Provided the separation is held constant or nearly so, it is feasible that fewer than $p$ variates may be used for classification purposes without significant degradation of accuracy.

In the case where $m=2$ and $\Pi_{1}$ and $\Pi_{2}$ are normal populations with unknown parameters $\mu_{1}, \mu_{2}$ and $\Sigma_{1}=\Sigma_{2}=\Sigma$, the classification procedure (8) can be replaced by an equivalent discriminant function [1]

$$
\begin{equation*}
v=x^{T} \hat{\Sigma}^{-1}\left(\bar{X}_{1}-\bar{x}_{2}\right)-\frac{1}{2}\left(\bar{X}_{1}+\bar{X}_{2}\right)^{T} \hat{\Sigma}^{-1}\left(\bar{x}_{1}-\bar{x}_{2}\right) \tag{9}
\end{equation*}
$$

where $\overrightarrow{\mathrm{X}}_{\mathrm{i}}$ is given by (3) and

$$
\begin{equation*}
\hat{\Sigma}=\frac{1}{n_{1}+n_{2}-2} \sum_{i=1}^{2} \sum_{j=1}^{n}\left(x_{j}^{(i)}-\bar{x}_{i}\right)\left(x_{j}^{(i)}-\bar{x}_{i}\right)^{T} . \tag{10}
\end{equation*}
$$

V is asymptotically distributed $\mathrm{N}(-\alpha / 2, \alpha)$ when $\mathrm{X} \sim \mathrm{N}\left(\mu_{2}, \Sigma\right)$ and $N(\alpha / 2, \alpha)$ when $X \sim N\left(\mu_{1}, \Sigma\right)$, where $\alpha$ is the Mahalanobis distance between the populations given by

$$
\begin{equation*}
\alpha=\left(\mu_{1}-\mu_{2}\right)^{T} \Sigma^{-I}\left(\mu_{1}-\mu_{2}\right) \tag{II}
\end{equation*}
$$

Thus the probability of misclassification is asymptotically dependent on $\alpha$ and not on $p$. Hence the procedure to select $s<p$ in a manner so as to hold the separation, $\alpha$, relatively constant, and then to classify on the basis of $s$ variates, is reasonable.

Development of Divergence (A Measure of Separation)
For the two population case, $\Sigma_{1} \neq \Sigma_{2}$, Kullback [5]
uses the divergence between the populations as the measure of separation or difficulty of discriminating between the populations. Defining the logarithm of the likelihood ratio, $\log \left[p_{1}\left(x_{0}\right) / p_{2}\left(x_{0}\right)\right]$, to be the information in $x=x_{0}$ for discimination in favor of $\Pi_{1}$ against $\Pi_{2}$, the mean information for discrimination in favor of $\Pi_{1}$ against $\Pi_{2}$ is

$$
\begin{equation*}
I_{12}=\int p_{1}(x) \log \frac{p_{1}(x)}{p_{2}(x)} d x \tag{12}
\end{equation*}
$$

The divergence between the populations, $J$, is defined by

$$
\begin{equation*}
J=I_{12}+I_{21} \tag{13}
\end{equation*}
$$

In the case where $\Pi_{1}$ and $\Pi_{2}$ are s-variate normal populations,

$$
\begin{equation*}
p_{i}(X)=(2 \pi)^{-\frac{S}{2}}\left|\Sigma_{i}\right|^{-\frac{1}{2}} \exp \left[-\frac{1}{2}\left(X-\mu_{i}\right)^{T} \Sigma_{i}^{-1}\left(X-\mu_{i}\right)\right] \tag{14}
\end{equation*}
$$

for $\mathrm{i}=1,2$. Then

$$
\begin{align*}
& \log \frac{p_{1}(x)}{p_{2}(x)}=\frac{1}{2} \log \frac{\left|\Sigma_{2}\right|}{\left|\Sigma_{1}\right|}-\frac{1}{2} \operatorname{tr} \Sigma_{1}^{-1}\left(x-\mu_{1}\right)\left(x-\mu_{1}\right)^{T} \\
&+\frac{1}{2} \operatorname{tr} \Sigma_{2}^{-1}\left(x-\mu_{2}\right)\left(x-\mu_{2}\right)^{T} \tag{15}
\end{align*}
$$

from which we get

$$
\begin{align*}
I_{12}=\frac{1}{2} \log & \frac{\left|\Sigma_{2}\right|}{\left|\Sigma_{1}\right|}+\frac{1}{2} \operatorname{tr} \Sigma_{1}\left(\Sigma_{2}^{-1}-\Sigma_{1}^{-1}\right)  \tag{16}\\
& +\frac{1}{2} \operatorname{tr} \Sigma_{2}^{-1}\left(\mu_{1}-\mu_{2}\right)\left(\mu_{1}-\mu_{2}\right)^{T}
\end{align*}
$$

Similarly

$$
\begin{align*}
I_{21}=\frac{1}{2} \log & \frac{\left|\Sigma_{1}\right|}{\left|\Sigma_{2}\right|}+\frac{1}{2} \operatorname{tr} \Sigma_{2}\left(\Sigma_{1}^{-1}-\Sigma_{2}^{-1}\right)  \tag{17}\\
& +\frac{1}{2} \operatorname{tr} \Sigma_{1}^{-1}\left(\mu_{2}-\mu_{1}\right)\left(\mu_{2}-\mu_{1}\right)^{T}
\end{align*}
$$

Thus

$$
J=I_{12}+I_{21}
$$

or

$$
\begin{align*}
J= & \frac{1}{2} \operatorname{tr}\left(\Sigma_{1}-\Sigma_{2}\right)\left(\Sigma_{2}^{-1}-\Sigma_{1}^{-1}\right) \\
& +\frac{1}{2} \operatorname{tr}\left(\Sigma_{1}^{-1}+\Sigma_{2}^{-1}\right)\left(\mu_{1}-\mu_{2}\right)\left(\mu_{1}-\mu_{2}\right)^{T} \tag{18}
\end{align*}
$$

Assuming equal covariance matrices, $\Sigma_{1}=\Sigma_{2}=\Sigma$,

$$
\begin{equation*}
J=\frac{1}{2} \operatorname{tr}\left[2 \Sigma^{-1}\left(\mu_{1}-\mu_{2}\right)\left(\mu_{1}-\mu_{2}\right)^{T}\right] \tag{19}
\end{equation*}
$$

or

$$
\begin{equation*}
J=\left(\mu_{1}-\mu_{2}\right)^{T} \Sigma^{-1}\left(\mu_{1}-\mu_{2}\right) \tag{20}
\end{equation*}
$$

That is, if $\Sigma_{1}=\Sigma_{2}=\Sigma$,

$$
\begin{equation*}
\boldsymbol{J}=\alpha \tag{21}
\end{equation*}
$$

where $\alpha$ is the Mahalanobis distance between the populations.
It may be noted in (18) that the first term of the expression for divergence is due to the difference of the covariance matrices, and the second term is due to the difference of the means. In actual situations $\Sigma_{1} \neq \Sigma_{2}$, however, one may choose to ignore the difference and compute $\alpha$ instead of J. Then a value to substitute for $\Sigma$ in (11) or (20) is

$$
\begin{equation*}
\bar{\Sigma}=\frac{\Sigma_{1}+\Sigma_{2}}{2} \tag{22}
\end{equation*}
$$

This assumption is made within this report, and the conclusions of Chapter IV indicate it is reasonable.

## Sample Size Considerations

If the parameters of $\Pi_{1}$ and $\Pi_{2}$ are unknown but estimated on the basis of training samples of sizes $n_{1}$ and $n_{2}$, respectively, then the choice of $s$ may be heavily dependent on the values $n_{1}$ and $n_{2}$. Consider 1) if $\Sigma_{1}=\Sigma_{2}=\Sigma$, $\left(n_{1}+n_{2}\right) p$ measurements are used to estimate the $2 p+\left(p^{2}+p\right) / 2$ distinct elements of $\bar{X}_{1}, \bar{x}_{2}$ and $\hat{\Sigma}$; and 2) if $\Sigma_{1} \neq \Sigma_{2},\left(n_{1}+n_{2}\right)$ p measurements are used to estimate the $3 p+p^{2}$ distinct elements of $\bar{x}_{1}, \overline{\mathrm{X}}_{2}, \hat{\Sigma}_{1}$ and $\hat{\Sigma}_{2}$. In either case, the number of elements to be estimated increases as a function of $p^{2}$. This suggests that for small sample sizes the probability of classification may actually be improved by considering fewer variates. The results contained herein support this hypothesis.

## CHAPTER III

## SIMULATED POPULATION CLASSIFICATION

## Classification Criteria

Anderson [l] discusses the classification when $m=2$ and $\mathrm{H}_{1}$ and $\mathrm{H}_{2}$ are normal populations with equal covariance matrices. The classification procedure (8) can be replaced by an equivalent discriminant function

$$
\begin{equation*}
\mathrm{v}=\mathrm{x}^{\mathrm{T}} \hat{\Sigma}^{-1}\left(\overline{\mathrm{x}}_{1}-\overline{\mathrm{x}}_{2}\right)-\frac{1}{2}\left(\overline{\mathrm{x}}_{1}+\overline{\mathrm{x}}_{2}\right)^{\mathrm{T}} \hat{\Sigma}^{-1}\left(\overline{\mathrm{x}}_{1}-\overline{\mathrm{x}}_{2}\right) \tag{23}
\end{equation*}
$$

where $\bar{X}_{i}$ and $\hat{\Sigma}$ are given in (3) and (10), respectively. If $n_{1}=n_{2}=n$ then the distribution of $V$ if $X$ is from $\mu_{1}$ is the same as the distribution of $-V$ if $X$ is from $\Pi_{2}$. Thus if $V \geq 0$ is the region of classification as $\Pi_{1}$, then the probability of misclassifying $X$ when it is from $\Pi_{1}$ is equal to the probability of misclassifying $X$ when it is from $\Pi_{2}$. Since V is asymptotically distributed $\mathrm{N}(-\alpha / 2, \alpha)$ when X is from $\pi_{2}, P(1 \mid 2)$, the probability of classifying an observation from $\Pi_{2}$ in $\Pi_{1}$, is approximately

$$
\begin{equation*}
\int_{0}^{\infty} \frac{1}{\sqrt{2 \pi \alpha}} e^{-(v+\alpha / 2)^{2} / 2 \alpha} d v \tag{24}
\end{equation*}
$$

Similarly

$$
\begin{equation*}
P(2 \mid 1) \simeq \int_{-\infty}^{0} \frac{1}{\sqrt{2 \pi \alpha}} e^{-(v-\alpha / 2)^{2} / 2 \alpha} d v \tag{25}
\end{equation*}
$$

and $P(2 \mid 1)=P(1 \mid 2)$.

In the event $\Sigma_{1} \neq \Sigma_{2}, \hat{p}_{1}(X)$ and $\hat{p}_{2}(X)$ are estimated on the basis of training samples of sizes $n_{1}$ and $n_{2}$, respectively. $X$ is classified in $\mathrm{H}_{2}$ if

$$
\begin{equation*}
\hat{p}_{2}(x)>\hat{p}_{1}(x) \tag{26}
\end{equation*}
$$

and in $\pi_{1}$ otherwise. Notice it is not necessary to compute $\hat{p}_{1}(X)$ and $\hat{p}_{2}(X)$ since (26) is equivalent to

$$
\begin{equation*}
\left|\hat{\Sigma}_{2}\right|^{\frac{1}{2}} \exp \left[-z_{1} / 2\right]<\left|\hat{\Sigma}_{1}\right|^{\frac{1}{2}} \exp \left[-z_{2} / 2\right] \tag{27}
\end{equation*}
$$

where

$$
\begin{equation*}
z_{i}=\left(X-\bar{X}_{i}\right)^{T} \hat{\Sigma}_{i}^{-1}\left(X-\bar{X}_{i}\right) \tag{28}
\end{equation*}
$$

$i=1,2$. Taking the natural logarithm of (27) and multiplying through by 2 gives the classification rule: classify $x$ in $\pi_{2}$ provided

$$
\begin{equation*}
\log \left(\operatorname{det} \hat{\varepsilon}_{1}\right)-z_{2}>\log \left(\operatorname{det} \hat{\varepsilon}_{2}\right)-z_{1} \tag{29}
\end{equation*}
$$

and in $\Pi_{1}$ otherwise. This rule is equivalent to (8) and (26). Monte Carlo Simulation Procedure

If the populations $\Pi_{1}$ and $\Pi_{2}$ are assumed to have equal covariances matrices, the probability $P(2 \mid 1)$ can be estimated from (25) provided the sample sizes $\mathrm{n}_{1}$ and $\mathrm{n}_{2}$ are large. For small (and moderately large) values of $n_{1}=n_{2}=n$, the probability $P(2 \mid 1)$ can be estimated by Monte Carlo methods. The process involves taking a sample of size $n$ from each of
the two normal populations. These samples are used to compute $\bar{X}_{1}, \bar{X}_{2}$ and $\hat{\Sigma}$ according to (3) and (10). A sample population of 50 is generated from $\Pi_{1} \sim N\left(\mu_{1}, \Sigma\right)$ and 50 values of $v$ are calculated. The probability of classifying an individual from the $\Pi_{1}$ population as belonging to $\pi_{2}$ is then estimated by

$$
\begin{equation*}
\hat{P}(2 \mid 1)=k / 50 \tag{30}
\end{equation*}
$$

where k is the number of values of $\mathrm{V}<0$. The sequence [training samples - estimation - population - classification $\hat{\mathbf{P}}(2 \mid 1)]$ is repeated 50 times, and the mean of the 50 values of $\hat{P}(2 \mid 1)$ is used as the estimate of $P(2 \mid 1)$. The results of Test Case 1 were obtained in this manner.

When the covariance matrices of $\Pi_{1}$ and $\Pi_{2}$ are not assumed to be equal, the procedure for estimating $\mathrm{P}(2 \mid 1)$ is basically the same as the preceding discussion. Differences arise in the estimation of $\hat{\Sigma}_{1}$ and $\hat{\Sigma}_{2}$ and in the classification procedure since $V$ is no longer valid. $\hat{\Sigma}_{1}$ and $\hat{\Sigma}_{2}$ are computed according to (4). The sample population of 50 is generated from $\mathbb{K}_{1} \sim N\left(\mu_{1}, \Sigma_{1}\right)$ and the 50 members are classified as belonging to $\Pi_{1}$ or $\Pi_{2}$ according to (29). Then

$$
\begin{equation*}
\hat{\mathrm{P}}(2 \mid 1)=k / 50 \tag{31}
\end{equation*}
$$

where $k$ is the number of times a member of the population is classified as belonging to $\Pi_{2}$. Again the sequence terminating in the computation of $\hat{P}(2 \mid 1)$ is repeated 50
times and the mean of the 50 values of $\hat{\mathrm{P}}(2 \mid 1)$ is used as the final estimate of $P(2 \mid 1)$.

When $\Sigma_{1} \neq \Sigma_{2}, \hat{P}(2 \mid 1) \neq \hat{P}(1 \mid 2)$. However, the estimation of $P(1 \mid 2)$ is accomplished by changing the order of the input of the parameters of the populatio:s. The results of Test Case 2 were obtained in this manner.

## CHAPTER IV

## RESULTS AND CONCLUSIONS

## Test Case 1: Feasibility

To demonstrate the feasibility of selecting a subset of the variates while holding the separation measure constant consider the populations $\Pi_{1}$ and $\Pi_{2}$ and the simulation results given in the following table:

| TABLE 1 | TEST CASE 1: FEASIBILITY |  |  |
| :---: | :---: | :---: | :---: |
| Number of | $\mu_{1}=0$ | Max |  |
| Variates | ${ }_{2}$ | $\alpha$ | $\underline{\mathrm{P}}(2 \mid 1)$ |
| 1 | (1) | 1 | . 318 |
| 2 | $(1,1)$ | 2 | . 256 |
| 3 | $(1,1,1)$ | 3 | . 216 |
| 4 | (1,1,1,1) | 4 | . 171 |
| 5 | (1,1,1,0,1) | 4 | . 192 |
| 6 | $(1,1,1,0,1,0)$ | 4 | . 195 |
| $\Pi_{1} \sim N(0, I)$ | $\mu_{2}=(1,1,1,0,1,0)$ |  |  |
| $\mathrm{H}_{1} \sim \mathrm{~N}\left(\mu_{2}, \mathrm{I}\right)$ | $\mathrm{n}_{1}=\mathrm{n}_{2}=15$ |  |  |

Table 1 illustrates a hypothetical situation where the maximum separation, $\alpha$, for the populations considered as 4-variate, 5-variate, and 6-variate populations is the same. Thus, if the training samples are large, an individual could be classified on the basis of four characteristics, namely $C_{1}, C_{2}, C_{3}$ and. $C_{5}$, with no degradation of accuracy.

For small sample sizes, in particular $n_{1}=n_{2}=15$, the simulation indicates that sample size considerations are significant, and better results are obtained using only 4 variates.

Test Case 2: Remote Sensing Data
To test the classification procedure using selected variates under meaningful conditions, actual data from a 12-channel sensor typical of those used in remote sensing of agricultural crops was obtained from NASA and used as population parameters for simulation purposes. $\pi_{1}$ was a composite of soybean fields and is represented by $\mu_{1}$ and $\Sigma_{1}$ as given in Table $2 . \mathrm{H}_{2}$ was a composite of corn fields and is. represented by $\mu_{2}$ and $\Sigma_{2}$ as given in Table 3. The divergence $J$ for all combinations of 12 channels taken $s$ at a time $(s=1,2, \ldots, 11)$ was computed by NASA. As many as 50 of the largest values of $J$ were printed for each value of $s$ and the values were included in the data package.

For purposes of comparing separation measures, $\alpha$ was computed [assuming $\Sigma=\left(\Sigma_{1}+\Sigma_{2}\right) / 2$ ] for $s=1,3,6$ and 12 . The maximum and minimum $\alpha$ and the maximum $J$ for the various values of $s$ are plotted in Figure 1 . The two distance measures are not readily comparable; however, the data does demonstrate the existence of combinations of fewer than 12 channels which yield almost the same separation as

```
Mean \mu
    169.56 174..76 193.24 192.64 169.11 166.80
Covariance Matrix }\mp@subsup{\Sigma}{1}{
    19.08
    14.47 14.28
    9.01 7.52 5.91
    9.40 8.49 5.12 6.50
    16.77 14.59 9.00 9.86 19.23
    13.73 11.98 7.24 8.01 14.22 12.84
\(7.97 \quad 7.01 \quad 4.68\)
        4.90 8.44 7.15
        5.31
    13.79 12.62 7.72.
        8.63 15.12
                            12.30 . 7.42 15.17
10.4
            10.0
            6.00
        7.00 12.08
            9.82
            6.14
            11.47
            10.88
                        5.97 10.32 8. 57
                            10.10
    9.46
            8.69
                            5.68
        6.28
            10.99
            9.26
-3.0
\(-3.48 \quad-1.33\)
                            -2.19
                    -2.81
                    -1.10
                                -. 20 -4.14 -4.02
                                -. }8
                            23.59
-4.73 -4.11 -2.73
                    -2.75
                    -4.24
                    -2.6
                    -1.67
                    -4. 11
                    -3.23
                            -1.43
                            6.55
                            9. }7
```

Table 2: MEAN AND COVARIANCE FOR SOYBEANS

```
Mean }\mp@subsup{\mu}{2}{
```



```
Covariance Matrix [ _2
23.92
18.21 17.21
\(11.57 \quad 9.33 \quad 7.25\)
\begin{tabular}{rrrrr}
10.95 & 9.28 & 6.01 & 6.74 & \\
23.40 & 19.28 & 12.65 & 12.59 & 28.43
\end{tabular}
\begin{tabular}{llllll}
24.52 & 20.50 & 13.18 & 13.47 & 28.94 & 32.71
\end{tabular}
\(13.29 \quad 11.06 \quad 7.52 \quad 7.61 \quad 15.95 \quad 17.52 \quad 10.84\)
```

| 15.21 | 13.26 | 8.88 | 9.19 | 19.14 | 20.45 | 11.46 |
| :--- | :--- | :--- | :--- | :--- | :--- | :--- |

16.43
$6.97 \quad 9.92$
$5.72 \quad 13.18$
15.85
5.17
8.79
7.52
5.
8.
8.12
5.41
5.76
11.49
12.1
9.31
12.40
18.61
9.01

7.76
$8.32 \quad 13.89$
$8.77 \quad 17.48 \quad 63.35$
4.94
9.57
18.00
$9.32 \quad 26.18$
18.21

Table 3: MEAN AND COVARIANCE FOR CORN

Separation


FIGURE 1. SEPARATION vs. NUMBER OF VARIATES
the full complement of channels. In particular, regardiess of the choice of separation measure, very little increase in maximum separation is obtained by considering more than 5 channels.

To further stucy the equivalence of $J$ and $\alpha$ as the criteria for the selection of variates, Table 4 was constructed. Table 4 contains some of the values obtained for $J$ and a for various subsets of variates. For a single variate, the largest three values of $J$ arise from considering the same three variates or channels that yielded the largest values of $\alpha$, and in the same order.


For the three variate case, the best three subsets by $J$ measure were the best three subsets by a measure, though in different order. In the case of 6 variates, from the

924 possible subsets of channels, the best three by $J$ measure were not among the best three by $\alpha$ measure; however, the best three by $J$ did yield values of a significantly close to the maximum $\alpha$.

The conclusion from Figure 1 and Table 4 is that either of the separation measures between populations, divergence and Mahalanobis distance, is a suitable selection criteria for the desired subset of variates.

Figures 2, 3, and 4 were constructed from simulation data based on training sample sizes of 15 . They support the theory that $P(2 \mid 1)$ is a decreasing function of $\alpha$. In most instances $\hat{P}(2 \mid 1)$ decreases as the distance measure increases. The relation is consistent with what could be expected from the asymptotic theory. Figure 2 is not too coherent because there are exactly 12 data points to consider. It does, however, indicate the trend. Figure 3 is very consistent in support of the relation. Figure 4 shows considerable scattering. If one looks ahead to Figure 6, the reason for the scatter in Figure 4 is evident. The sample size was too small. A training sample size of 30 would probably have produced a more consistent figure.

Figures 5 and 6 are the significant figures of this report. They represent the same data presented from two different viewpoints. Figure 5 shows that for small fixed training sample sizes, there is a subset of fewer than 12
variates which yield the minimum probability of misclassification. Figure 6 holds $p$ constant and varies the training sample size. This figure shows that, for the populations considered, if one is constrained to $\mathrm{n} \leq 11$, one might just as well look at a single variate. For $11<n<30$, 3 channel data is adequate, and $n$ must be somewhere in the neighborhood of 100 before l2-variate data can be justified over 6-variate data.

The conclusion of this report is that for small sample sizes, a subset of the variates can be chosen so as to yield better classification results than the full complement of variates, and the selection of variates can be accomplished by considering the separation between the populations.


FIGURE 2. $\hat{P}(2 \mid 1)$ vs. $\alpha-$ SINGLE VARIATE


FIGURE 3. $\hat{\mathrm{P}}(2 \mid 1)$ vs. a - THREE VARIATES


FIGURE 4. $\hat{P}(2 \mid 1)$ vs. $a-\operatorname{SIX}$ VARIATES


FIGURE 5. $\hat{\mathrm{P}}(2 \mid 1)$ vs. NUMBER OF VARIATES FOR FIXED SAMPLE SIZES


FIGURE 6. $\hat{P}(2 \mid 1)$ vs. TRAINING SAMPLE SIZE FOR FIXED VARIATES

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# ACREAGE ESTIMATES FOR CROPS USING REMOTE SENSING TECHNIQUES: CLASSIFICATION ERROR MATRIX KNOWN 

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# ACREAGE ESTIMATES FOR CROPS USING REMOTE SENSING TECHNIQUES: <br> CLASSIFICATION ERROR MATRIX KNOWN 

## 1. Introduction

Remote sensing technology has shown a great potential for data collection for the earth resources in any given geographic region. As a result of this, it may be feasible to assess various earth resources and, thereby, answers to some of the important yet previously unsolvable problems may be available. At present we address ourselves to the problem of estimating crop sizes, i.e., amount of acreage under crops, in a large area using remote sensing technique. Even though it may not be difficult to obtain full data over an area by using remote sensors, crop acreage estimation on the basis of complete enumeration of data points in the area may not be feasible both from technical and economical viewpoints. As such it would be desirable to derive estimates based upon sampled data acquired by a suitable sampling process.

In this report we discuss a sampling scheme providing crop acreage estimates in a given geographic region, and investigate the precision of these estimates and the effect of misclassification on the estimates. Since observations are obtained on individual pixels, we consider the frame made of the collection of all pixels in the region of our interest. As to the actual layout for various crops, two cases arise. One is that no information is available and the other is that some a priori information on the physical situation of crops exists. For example, if the area of interest is small like a county or certain areas covering Western Oklahoma, it is
possible to have some prior information for the location of different crops. In our discussion we consider each of these cases and point out the difference that exists in precisions of two estimates.

## 2. Formulation of the Problem

Let $C_{1}, C_{2}, \ldots, C_{m}$ be $m$ different crops and $A_{1}, A_{2}, \ldots, A_{m}$ be their corresponding acreage areas in a region. Assume the whole region consists of $N$ total number of pixels. Since a pixel on the basis of observation taken by a remote sensor is subject to uncertainity in its correct identification, let $P(i \mid j)$ denote the probability of misclassifying a pixel from crop $C_{j}$ into crop $C_{i}$, and $P(i \mid i)$ denote the probability of correctly classifying a pixel into crop $C_{i}$. Accordingly, the expected acreages associated with $C_{i}$ on the basis of remote sensing data is

$$
\begin{equation*}
E_{i}=\sum_{j=1}^{m} A_{j} P(i \mid j) \tag{2.1}
\end{equation*}
$$

$\mathrm{i}=1,2, \ldots, \mathrm{~m}$.
Given the total number of pixels and a pixel size, it is sufficient to consider proportions of pixels associated with different crops in the region. Let $p_{1}, p_{2}, \ldots, p_{m}$ be the proportions of pixels for $C_{1}, C_{2}, \ldots, C_{m}$, respectively. Then the expected proportions of pixels likely to be identified in $C_{i}$, $\mathrm{i}=1,2, \ldots, \mathrm{~m}$, are

$$
\begin{equation*}
e_{i}=\sum_{j=1}^{m} p_{j} P(i \mid j) \tag{2.2}
\end{equation*}
$$

$e_{i}$ coincides with $p_{i}, i=1,2, \ldots, m$, when every pixel is correctly classified. Since it is almost impossible to expect so in remote sensing, in our discussion we first consider estimation of $e_{i}, i=1,2, \ldots, m$, and then seek estimates
for $p_{i}, i=1,2, \ldots, m$, assuming that $P(i \mid j)>0$ at least for one $j$ different from i. This will be done considering two cases: (a) $P(j \mid i)$ 's are known and (b) $P(i \mid j)^{\prime}$ 's are unknown.

## 3. The Sampling Procedure

As we mentioned earlier in Section 1 , we consider the following two cases:
(i) Crop boundaries (i.e. the physical layout for crops) are unknown.
(ii) Crop boundaries are enhanced and so known.
(i) In this case, a sampling procedure which appears to be useful is a twostage sampling scheme. At the first stage a specified number of flightlines, each of the same size, are randomly selected, and at the second stage an equal number of units (pixels) are randomly selected from each of the selected flightlines. However, in order for this scheme to be useful, it may require a fairly large number of flightlines to be selected. Otherwise, a simple random sampling with single-stage, though more difficult to execute, may produce estimates with smaller variance than the two-stage random sampling scheme.
(ii) When the physical layout for the crops is known to a certain degree, consider the region made of subregions, each consisting of flightlines covering area as homogeneous as possible. This could allow more than one crop in a subregion, and such subdivision should be restricted to a minimum possible number of subregions. For an illustration, see Figure 1 given at the end.

For a sampling procedure, we suggest a three-stage sampling scheme carried out in each subregion. At the first stage flightlines are randomly selected proportional to the size of a subregion; at the second stage blocks (or segments) are randomly selected within a flightline proportional to the size of a crop in the flightline; and at the third stage units within blocks are randomly selected in equal numbers. The number of blocks selected in each of the sampled flightlines is the same. Here by block we mean a sampling unit of specific size within a flightline. It may be pointed out that careful consideration should be given in specifying the size of a block. Neither should it be too small for any practical use, nor should it be too large to make any distinction for the crops when sampled. This way at the first two stages the sampling is proportional to the size of the underlying strata (i.e., subregions or crops, whichever may be the case) and at the last stage it is a simple random sampling. Though one may expect accumulation in sampling error due to three stages of sampling, it should lead to a smaller variance of the estimate for $e_{i}(i=1,2, \ldots, m)$ compared to a single-stage or double-stage sampling scheme.

## 4. Expected Proportions Estimates

(i) Let $N$ denote the total number of fightines for the region and let $n$ flightlines out of $N$ be randomly selected. Suppose each flightline consists of $R$ units from which $r$ units are randomly sampled for each of the sampled flightines. In the sample, let $m_{i t}$ denote the number of units from th selected flightline classified into $\mathcal{C}_{i}$. Now denoting

$$
\hat{e}_{i t}=\frac{m_{i t}}{r}, \quad t=1,2, \ldots, n
$$

an unbiased estimate of $e_{i}$ is given by

$$
\begin{equation*}
\hat{e}_{i}=\frac{1}{n r} \sum_{t=1}^{n} m_{i t}, \quad i=1,2, \ldots, m \tag{4.1}
\end{equation*}
$$

Since the total size of units, $N R$, is expected to be large in relation to the sample size, $n r$, the random vector $\hat{e}=\left(\hat{e}_{1}, \hat{e}_{2}, \ldots, \hat{e}_{m}\right)^{T}$ has a multinomial distribution, usually a satisfactory approximation in such a situation. So it follows that the variance of $\hat{e}_{i}$,

$$
\begin{equation*}
V\left(\hat{e}_{i}\right)=\left(1-\frac{n}{N}\right) \frac{1}{n(N-1)} \sum_{t=1}^{N}\left(e_{i t}-e_{i}\right)^{2}+\left(1-\frac{r}{R}\right) \frac{1}{n r} \frac{R}{N(R-1)} \sum_{t=1}^{N} e_{i t}\left(1-e_{i t}\right) \tag{4.2}
\end{equation*}
$$

and an unbiased estimate of $V\left(\hat{e}_{i}\right)$ is given by
$\hat{V}\left(\hat{e}_{i}\right)=\left(1-\frac{n}{N}\right) \frac{1}{n(n-1)} \sum_{t=1}^{n}\left(\hat{e}_{i t}-\hat{e}_{i}\right)^{2}+\left(1-\frac{r}{R}\right) \frac{1}{N n(r-1)} \sum_{t=1}^{n} \hat{e}_{i t}\left(1-\hat{e}_{i t}\right)$
where $e_{i t}$ denotes the expected proportion of $c_{i}$ in th flight-line and $e_{i}=\sum_{t=1}^{N} e_{i t} / N$, the same as defined in (2.2), (Cochran, 1963).
(ii) Again, let $N$ denote the total number of flightlines, each consisting of $R$ units. Suppose the region is divided into $k$ subregions, $R_{1}, R_{2}, \ldots, R_{k}$, and $N_{1}, N_{2}, \ldots, N_{k}$ are the corresponding number of flightlines for $R_{1}, R_{2}, \ldots, R_{k}$ such that

$$
N=\sum_{j=1}^{k} N_{j}
$$

Considering a sample of $n$ flightlines out of $N$, let $n_{j}$ flightlines be randomly selected from $N_{j}$ in $R_{j}$ such that $n_{j}=n\left(N_{j} / N\right), j=1,2, \ldots, k$, and

$$
n=\sum_{j=1}^{k} n_{j}
$$

For subregion $R_{j}$, let $B_{j i t}$ be the number of blocks covering ith crop in the th sampled flightlines; from this let $b_{j i t}$ blocks be randomly samples such that

$$
B=\sum_{i=1}^{m} B_{j i t}, \quad b=\sum_{i=1}^{m} b_{j i t}
$$

and

$$
\begin{aligned}
& b_{j i t}=\frac{b}{B} B_{j i t}, \\
& t=1,2, \ldots, n_{j}, \\
& i=1,2, \ldots, m .
\end{aligned}
$$

Considering block size L, suppose 1 units are randomly selected for each sampled block. Thus the sample size for subregion $R_{j} i s n_{j}$ bl and the complete sample consists of nbl units.

Let $u_{j i t}$ denote the number of sample units being classified in $C_{i}$ for $t$ th flightline selected from $R_{j}$. Then

$$
\begin{equation*}
\hat{e}_{j i}=\frac{1}{n_{j} b l} \sum_{t=1}^{n_{i}} u_{j i t} \tag{4.4}
\end{equation*}
$$

is an estimate of $e_{j i}$, the expected proportion of units in $C_{i}$ for subregion $R_{j}, i=1,2, \ldots, m$ and $j=1,2, \ldots, k$. Thus

$$
\begin{equation*}
\hat{e}_{i}=\frac{1}{N} \sum_{j=1}^{k} N_{j} \hat{e}_{j i}, \quad i=1,2, \ldots, m . \tag{4.5}
\end{equation*}
$$

Next,

$$
V\left(\hat{e}_{i}\right)=\frac{1}{N^{2}} \sum_{j=1}^{k} N_{j} v\left(\hat{e}_{j i}\right)
$$

where

$$
V\left(\hat{e}_{j i}\right)=\frac{N}{n}\left(1-\frac{n}{N}\right) \cdot \frac{e_{j i}\left(1-e_{j i}\right)}{N_{j}-1},
$$

assuming the same variance in each flightine of $R_{j}$ for a fixed $C_{i}$. Thus

$$
\begin{align*}
V\left(\hat{e}_{j}\right)=\frac{1}{n N}\left(1-\frac{n}{N}\right) & \sum_{j=1}^{k} \frac{N_{j}^{2}}{N_{j}-1} e_{j i}\left(1-e_{j i}\right)  \tag{4.6}\\
i & =1,2, \ldots, m
\end{align*}
$$

## 5. Actual Proportions Estimates

Denoting

$$
\begin{aligned}
& e=\left[\begin{array}{c}
e_{1} \\
e_{2} \\
\vdots \\
e_{m}
\end{array}\right], \quad p=\left[\begin{array}{c}
p_{1} \\
p_{2} \\
\vdots \\
p_{m}
\end{array}\right], \quad \text { and } \\
& Q=\left[\begin{array}{ccccc}
P(1 \mid 1) & P(1 \mid 2) & . & . & . \\
P(2 \mid 1) & P(2 \mid 2) & . & P(1 \mid m) \\
\cdot & . & . & P(2 \mid m) \\
\cdot & \cdot & . & . & . \\
P(m \mid 1) & P(m \mid 2) & . & . & . \\
P & P(m \mid m)
\end{array}\right],
\end{aligned}
$$

(2.2) can be written as

$$
\begin{equation*}
\mathrm{e}=\mathrm{Qp} \tag{5.1}
\end{equation*}
$$

Observe that $Q$ may be a singular matrix. In view of this the equations in (5.I) may or may not have a solution for $p$ for specified $e, Q$. In case any solution exists, it is given by

$$
\begin{equation*}
p=Q^{I} e+p_{0} \tag{5.2}
\end{equation*}
$$

where $Q$ is a generalized inverse of $Q$ matrix satisfying

$$
\begin{equation*}
Q Q_{Q}=Q \tag{5.3}
\end{equation*}
$$

and $P_{0}$ satisfying $Q p_{0}=0$.
In the second case when no solution exists, we look for the vector $p$ that minimizes the squared length of $e-Q p$ for a given $e, ~ Q$. Any such vector is given by

$$
\begin{equation*}
p=Q^{S} e+P_{0} \tag{5.4}
\end{equation*}
$$

where $Q^{S}$ is a right pseudo inverse of $Q$ matrix satisfying

$$
\begin{equation*}
\left(Q Q^{S}\right)^{T}=Q Q^{S} \tag{5.5}
\end{equation*}
$$

and $P_{0}$ satisfying $Q P_{0}=0$.
Since more than one $p_{0}$ may satisfy the condition $Q p_{0}=0$, in either case, a unique solution for $p$ may not exist. Accordingly, either one would have a complete set of solutions $p$ given by (5.2) or a complete set of vectors $p$ which are a "best-fit' when obtained from (5.4). The techniques for finding $Q^{I}, Q^{S}$ and $P_{0}$ are well known (Boullion and Odell, 1972). Henceforth, by a solution we mean in either sense and will denote it by

$$
\begin{equation*}
p=Q^{G} e, \tag{5.6}
\end{equation*}
$$

assuming $P_{0}=0$ without loss of generality.

### 5.1. Q known

(i) Taking the estimate $\hat{e}$ given by (4.1) and the known value of $Q$, it follows from the above discussion that an estimate of $p$ is given by

$$
\begin{equation*}
\hat{p}=Q^{G \hat{e}} . \tag{5.7}
\end{equation*}
$$

Denoting the $(i, j)$ the element of $Q^{G}$ by $q_{i j}^{G}$, we have

$$
\begin{equation*}
\hat{p}_{i}=\sum_{j=1}^{m} q_{i j}^{G} \hat{e}_{j}, \quad i=1,2, \ldots, m \tag{5.8}
\end{equation*}
$$

These are unbiased estimates of $p_{i}, i=1,2, \ldots, m$. For the variance,

$$
\begin{equation*}
v\left(\hat{p}_{i}\right)=\sum_{j=1}^{m}\left(q_{i j}^{G}\right)^{2} v\left(\hat{e}_{j}\right)+\sum_{j=1}^{m} \sum_{j=1}^{m} q_{i j}^{G} q_{i j}^{G}, \operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{j}\right) \tag{5.9}
\end{equation*}
$$

where $V\left(\hat{e}_{j}\right)$ is given in (4.2) and

$$
\begin{aligned}
\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{j}\right)= & \left(1-\frac{n}{N}\right) \frac{1}{n(N-1)} \sum_{t=1}^{N}\left(e_{j t}-e_{j}\right)\left(e_{j^{\prime}} t^{-e_{j}}\right) \\
& -\left(1-\frac{r}{R}\right) \frac{1}{N r} \frac{R}{N(R-1)} \sum_{t=1}^{N} e_{j t^{\prime}} e^{\prime} t, \quad\left(j \neq j^{\prime}\right)
\end{aligned}
$$

Next, an unbiased estimate of $V\left(\hat{e}_{j}\right)$ is given in (4.3). Similarly an unbiased estimate of $\operatorname{Cov}\left(e_{j}, e_{j}\right.$ ) can be shown as

$$
\begin{aligned}
\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{j \prime}\right)= & \left(1-\frac{n}{N}\right) \frac{1}{n(n-1)} \sum_{t=1}^{n}\left(\hat{e}_{j t}-\hat{e}_{j}\right)\left(\hat{e}_{j^{\prime} t}-\hat{e}_{j \prime}\right) \\
& -\left(1-\frac{r}{R}\right) \frac{1}{N r(r-1)} \sum_{t=1}^{n} \hat{e}_{j t} \hat{e}_{j^{\prime} t} \quad\left(j \neq j^{\prime}\right) .
\end{aligned}
$$

Replacing the unknown quantities in (5.9) by their estimates, one thus obtains an unbiased estimate of $V\left(\hat{p}_{i}\right), i=1,2, \ldots, m$.
(ii) In this case using the estimate $\hat{e}$ given in (4.5) we obtain

$$
\begin{equation*}
\hat{p}=Q^{G} \hat{e} \tag{5.10}
\end{equation*}
$$

or

$$
\begin{equation*}
\hat{p}_{i}=\sum_{j=1}^{m} q_{i j}^{G} \hat{e}_{j}, \quad i=1,2, \ldots, m \tag{5.11}
\end{equation*}
$$

Once again,

$$
\begin{equation*}
v\left(\hat{p}_{i}\right)=\sum_{j=1}^{m}\left(q_{i j}^{G}\right)^{2} v\left(\hat{e}_{j}\right)+\sum_{j=1}^{m} \sum_{j^{\prime}=1}^{m} q_{i j}^{G} q_{i j}^{G}, \operatorname{cov}\left(\hat{e}_{i}, \hat{e}_{j}\right) \tag{5.12}
\end{equation*}
$$

where $V\left(\hat{e}_{j}\right)$ is given in (4.6) and

$$
\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{j^{\prime}}\right)=-\frac{1}{n N}\left(1-\frac{n}{N}\right) \sum_{i=1}^{k} \frac{N_{j}}{N_{j}-1} e_{i j} e_{i j^{\prime}} .
$$

Now replacing $V\left(\hat{e}_{j}\right)$ and $\operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{j}\right)$ by their estimates, we can get an estimate of $V\left(\hat{e}_{j}\right), i=1,2, \ldots, m$.

### 5.2. Q unknown

In this case one also needs to estimate $P(i \mid j)$, $i$ and $j=1,2, \ldots, m$. An obvious way to do so is to ascertain the ground truth for a certain number of additional observations taken independently of those used for estimating $e_{i}{ }^{\prime}$ s and utilize these to estimate $P(i \mid j)$ 's. In view of this $\hat{e}$ and $\hat{Q}$ will be stochastically independent and an estimate of $p$ is given by

$$
\begin{equation*}
\hat{p}=\hat{Q}^{G} \hat{e} \tag{5.13}
\end{equation*}
$$

or

$$
\begin{equation*}
\hat{p}_{i}=\sum_{j=1}^{m} \hat{q}_{i j}^{G} \hat{e}_{j}, \quad i=1,2, \ldots, m \tag{5.14}
\end{equation*}
$$

where $\hat{q}_{i j}^{G}$ is the $(i, j)$ th element of $\hat{Q}^{G}$.

For a given $\hat{Q}^{G}$, the conditional variance of $\hat{p}_{i}$ is given by (5.9) in case (i) and (5.12) in case (ii) with $\dot{q}_{i j}^{G}$ 's replaced by $\hat{q}_{i j}^{G}$ 's. For the unconditional variance, the expression will involve the variances and covariances for the elements of $\hat{Q}^{G}$. In general it will be difficult to express the unconditional variance explicitly.

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Figure I: A subdivision of a region with five crops for a three-stage sampling plan.

# ESTIMATION OF CROP ACREAGE THROUGH <br> SAMPLING OF REMOTELY SENSED DATA: <br> CLASSIFICATION ERROR MATRIX UNKNOWN* 

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# Estimation of Crop Acreage Through <br> Sampling of Remotely Sensed Data 

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## 1. INTRODUCTION

In this report we consider the problem of estimating crop acreages in an area using samples from remotely sensed data for the area. Rationale for using samples is to avoid enormous cost and time that might be involved otherwise if the full data is processed. In particular, such would be the case for a ERTS scene which generally has over half a million data points and covers approximately an area of $100 \times 100$ square miles.

While considering crop acreage estimation, it is desirable to assume that the underlying region is an agricultural area and that every data point is identifiable with respect to certain known types of crops. In a larger context of an arbitrary area, it is sufficient for the condition of identification to hold with respect to the underlying known earth resources.

To formulate the problem, let $\Pi_{1}, \Pi_{2}, \ldots, \Pi_{m}$ denote the $m$ crops in the area and $p_{1}, p_{2}, \ldots, p_{m}$ be the proportions of their acreages. Next, let $P(i / j)$ be the probability of classifying a resolution element (pixel) from $\pi_{j}$ into $\pi_{i}$ using a classification algorithm. Then associated with such classification algorithm processing of full remotely sensed data would amount to expecting proportion of $\Pi_{i}$ acreage given by.

$$
\begin{equation*}
e_{i}=\sum_{j=1}^{m} p_{j} P(i / j), i, j, 2, \ldots, m \tag{1}
\end{equation*}
$$

Equivalently,

$$
\begin{equation*}
\mathrm{e}=\mathrm{Pp} \tag{2}
\end{equation*}
$$

where

$$
e=\left[\begin{array}{c}
e_{1} \\
e_{2} \\
\cdot \\
\cdot \\
\cdot \\
e_{m}
\end{array}\right] \quad, \quad p=\left[\begin{array}{c}
p_{1} \\
p_{2} \\
\cdot \\
\cdot \\
\cdot \\
p_{m}
\end{array}\right] \quad \text { and } p=\left[\begin{array}{cccc}
P(1 / 1) & P(1 / 2) & \ldots & P(1 / m) \\
P(2 / 1) & P(2 / 2) & \ldots & P(2 / m) \\
\cdot & \cdot & \cdot & \cdot \\
\dot{p_{m}} & \cdot & \cdot & \cdots \\
\dot{p}(\dot{m} / \dot{1}) & \dot{P}(\dot{m} / 2) & \ldots & \dot{P}(\dot{m} / m)
\end{array}\right]
$$

If the vector $e$ and matrix $P$ are known, one gets $p$ by solving (2) subject to $\Sigma p_{i}=1$. Otherwise one needs to consider estimation of e or $P$ or both of these as the case may be in order to ascertain about $p$.

In general, e will be unknown. An estimate of $e$ is given by the vector of observed proportions of resolution elements processed and classified into $I_{i}, i=1,2, \ldots, m$ using sample data obtained under a sampling plan. Regarding $P$, two cases arise:
(i) $P$ is known
(ii) $P$ is unknown

Again, $P$ will generally be unknown and for an estimate of $P$ one would require some suitably selected amount of the ground truth, probably independent of the previous sample data used in estimating e. Moreover, care should be exercised in handling of the sampled ground truth. Most often, it would be desirable to use the sample for both training the classifier and obtaining estimate of $P$.

Clearly how much complex the estimation problem is Gepends upon how much one knows about $P$. In an entirely unknown situation of classification with untrained classifier, estimation of acreage proportions can be very misleading. As such we would assume for the classifier to be properly trained and thus are able to achieve reliable estimate of $P$.

## 2. PROPORTION ESTIMATES

Let $\hat{e}$ be an estimate of $e$ and $\hat{p}$ of $P$. Then an estimate of $p$ is given by
(i) $\hat{p}=P^{-1} \hat{e} \quad$ when $P$ is known
and
(ii) $\hat{\hat{p}}=\hat{p}^{-1} \hat{e} \quad$ when $P$ is unknown

In case (i) $\hat{p}$ is an unbiased estimate of $p$ whenever $e$ is an unbiased estimate of e. However, if estimate for both $e$ and $P$ are involved in estimating $p$ as in case (ii); it may hardly be possible to obtain an unbiased estimate of $p$ in $\hat{\hat{p}}$. Henceforth, we will assume $\hat{e}$, and so also $\hat{p}$, to be unbiased estimate and it is the vector of observed proportions of resolu-. tion elements processed and classified into different $\pi_{i}$ 's (This will be the case due to the sampling plans being considered in this report.) Also, we would restrict our investigation to case (ii) and for case (i), refer to Chhikara and Odell [1]. Though final results in [1] pertain to specific sampling plans, the discussion there is of general nature and results can be adopted in any sampling situation.

## 3. MEAN SQUARE ERRORS OF ESTIMATES

First we calculate the bias of $\hat{\hat{p}}$ given by
Bias $\{\hat{\hat{p}}\}=E[\hat{\hat{p}}-p]$

$$
\begin{align*}
& =E\left[\hat{p}^{-1} \hat{e}-p^{-1} e\right] \\
& =E\left[\hat{p}^{-1}(\hat{e}-e)+\left(\hat{p}^{-1}-p^{-1}\right) e\right] \\
& =E\left[\hat{p}^{-1}-p^{-1}\right]_{e} \tag{5}
\end{align*}
$$

because the first term is zero due $E(\hat{e}-e)=0$ for a given $\hat{P}^{-1}$. Clearly, the bias depends upon how much bias there is in $\hat{p}^{-1}$, and

$$
\begin{equation*}
\operatorname{Bias}\{\hat{\hat{p}}\}=\left(\operatorname{Bias}\left\{\hat{\mathrm{P}}^{-1}\right\}\right) \text { e. } \tag{6}
\end{equation*}
$$

In order to find the mean square error of any component of $\hat{\hat{p}}$, we first consider the evaluation of matrix,

$$
\begin{align*}
& E\left[(\hat{\hat{p}}-p)(\hat{\hat{p}}-p)^{\top}\right]=E\left[\left(\hat{p}^{-1} \hat{e}-P^{-1} e\right)\left(\hat{P}^{-1} \hat{e}-P^{-1} e\right)^{\top}\right] \\
& \quad=E\left[\left(\hat{P}^{-1}\right)(\hat{e}-e)(\hat{e}-e)^{\top}\left(\hat{P}^{-1}\right)^{\top}+\left(\hat{p}^{-1}-p^{-1}\right) e^{\top}\left(\hat{p}^{-1}-p^{-1}\right)^{\top}\right] \\
& \quad=E\left[\left(\hat{p}^{-1}\right) M\left(\hat{P}^{-1}\right)^{T}\right]+E\left[\left(\hat{p}^{-1}-P^{-1}\right) e e^{\top}\left(\hat{P}^{-1}-p^{-1}\right)^{\top}\right], \tag{7}
\end{align*}
$$

where $M$ denotes the covariance matrix of $\hat{e}$.
Denoting the ( $i, j$ ) th element of $\mathrm{P}^{-1}$ by $\mathrm{P}^{\mathrm{ij}}$ and that of $\hat{\mathrm{P}}^{-1}$ by $\hat{p}^{i j}$, it follows from (7) that the mean square error of $\hat{\hat{p}}_{i}$ is given by

$$
\left.\begin{array}{rl}
E\left[\sum_{k=1}^{m} \hat{p}^{i k} \sum_{j=1}^{m} \hat{p}^{i j} \operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right)\right. \\
& +\sum_{k=1}^{m} e_{k}\left(\hat{p}^{i k} k_{-p}^{i k}\right) \sum_{j=1}^{m}\left(\hat{p}^{i j} p_{-} i j\right.
\end{array} e_{j}\right] .
$$

where $E$ stands for expectation with respect to $\hat{\mathrm{P}}$. Denoting it by MSE $\left\{\hat{\hat{p}}_{\mathrm{j}}\right\}$, we have

$$
\begin{align*}
& \operatorname{MSE}\left(\hat{p_{i}}\right)=E\left[\sum_{j=1}^{m} \hat{p}^{i j}\right)^{2} \operatorname{Var}\left(\hat{e}_{j}\right)+\sum_{\substack{j=1 \\
j \neq k}}^{m} \sum_{\substack{k=1}}^{m} \hat{p}^{i j} \hat{\mathrm{p}}^{i k} \operatorname{Cov}\left(\hat{e}_{j}, \hat{e}_{k}\right) \\
& \left.+\sum_{j=1}^{m} e_{j}^{2} E\left[\hat{p}^{i j}-p^{i j}\right]^{2}+\sum_{\substack{m=1 \\
j \neq k}}^{m} \sum_{k=1}^{m} e_{j} e_{k} E\left[\left(\hat{p}^{i j}-P^{i j}\right)\left(\hat{p}^{i k}-P^{i k}\right)\right]\right] \tag{8}
\end{align*}
$$

$1 \pm 1,2, \ldots, m$.
The procedure for obtaining $\hat{P}$ would involve sampling of ground truth independent of samples taken for estimating e. In the following section we give expressions for $v\left(\hat{e}_{i}\right)$ and $\operatorname{Cov}\left(\hat{e}_{i}, \hat{e}_{j}\right)$ considering different sampling plans. To evaluate expectation in (8), one needs to find the distribution of $\hat{P}$. This will, of course, depend upon how $\hat{p}$ is obtained. In general, it will be difficult to obtain any exact distribution of $\hat{P}$. However, if the sampling of ground truth involves separate independent samples from each crop and $\hat{P}$ is obtained as the matrix of observed proportions among randomly selected pixels classified into different crops using a classifier, each column vector of $\hat{p}$ has a multinomial distribution and is stochaotically independent of the others in $\hat{P}$. Since expectation in (8) is for elements of $\hat{P}^{-1}$, it may not be easy to derive the MSE $\left\{\hat{\mathrm{p}}_{\mathrm{i}}\right\}$ in a closed form, especially if the number of classes is large. As such we now consider the two-class case and show the procedure for obtaining $\operatorname{Bias}\left\{\hat{\hat{p}}_{i}\right\}$ and $\operatorname{MSE}\left\{\hat{\hat{p}}_{\mathrm{i}}\right\}, \hat{i}=1,2$.

## Two-class Problem

Often interest lies in ascertaining acreage of a specified crop in an area. In view of this one may consider $\Pi_{1}$ to be the crop of main interest and $\mathrm{II}_{2}$ to be its complementary part. Without loss of generality, let us assume density functions for $\Pi_{2}$ and $\Pi_{2}$ to be symmetric about certain
location points." Then $\mathrm{P}(1 / 2)$ is same as $\mathrm{P}(2 / 1)$. Denoting this common value by $\Phi$, one gets

$$
P=\left[\begin{array}{cc}
1-\Phi & \Phi \\
\Phi & 1-\Phi
\end{array}\right]
$$

It now follows from (2) that

$$
e_{1}=p_{1}+\left(p_{2}-p_{1}\right) \Phi
$$

and

$$
e_{2}=p_{2}+\left(p_{1}-p_{2}\right) \Phi,\left(e_{1}+e_{2}=1\right)
$$

Assuming $\Phi \neq 1 / 2$, we have

$$
p_{1}=\frac{e_{1}-\Phi}{1-2 \Phi}
$$

and

$$
p_{2}=\frac{e_{2}-\Phi}{1-2 \Phi}
$$

Considering $\hat{e}_{1}, \hat{e}_{2}$ and $\hat{\phi}(\hat{\Phi} \neq 1 / 2)$ estimates for $e_{1}, e_{2}$ and $\Phi$, respectively, obtained according to the procedure outlined in the previous paragraphs, one has

$$
\begin{aligned}
& \hat{\hat{p}}_{1}=\frac{\hat{e}_{1}-\hat{\phi}}{1-2 \hat{\phi}} \\
& \hat{\hat{p}}_{2}=\frac{\hat{e}_{2}-\hat{\phi}}{1-2 \hat{\phi}}
\end{aligned}
$$

Clearly, $\hat{\hat{p}}_{1}+\hat{\hat{p}}_{2}=1$ because of $\hat{e}_{1}+\hat{e}_{2}=1$. Next, it follows that for $k=1,2$,

$$
\operatorname{Bias}\left\{\hat{\hat{p}}_{k}\right\}=\left(e_{k}-1 / 2\right)(E[T]-\odot)
$$

and

$$
\operatorname{MSE}\left\{\hat{\hat{p}}_{\mathrm{k}}\right\}=\operatorname{Var}\left(\hat{e}_{\mathrm{k}}\right) E\left[T^{2}\right]+\left(e_{\mathrm{k}}-1 / 2\right)^{2} E[T-\Theta]^{2}
$$

where

$$
T=(1-2 \hat{\Phi})^{-1} \quad \text { and } \quad \theta=(1-2 \Phi)^{-1}
$$

To obtain $E[T]$ and $E\left[T^{2}\right]$, one way is to find the distribution of $\hat{\phi}$. Let $r_{1}$ out of $N_{1}$ pixels sampled from $\Pi_{1}$ be misclassified into $\Pi_{2}$ and $r_{2}$ out of $N_{2}$ pixels sampled from $\Pi_{2}$ be misclassified into $\Pi_{1}$. Then

$$
\hat{\Phi}=\frac{r_{1}+r_{2}}{N_{1}+N_{2}}
$$

provides an estimate of $\Phi$. The condition of $\hat{\phi} \neq 1 / 2$ is easily met if we restrict $N_{1}+N_{2}$ to an odd number, a mild restriction which should not undermine the generality of present discussion. Denoting $r=r_{1}+r_{2}$ and $N=N_{1}+N_{2}$, the random variable $r$ has a Binomial distribution with proportion $\Phi$ and sample size N. Accordingly

$$
E[T]=\sum_{r=0}^{N} \frac{N}{N-2 r}\binom{N}{r} \quad \phi_{\Phi^{r}}(1-\Phi)^{N-r}
$$

and

$$
E\left[T^{2}\right]=\sum_{r=0}^{N} \frac{N^{2}}{(N-2 r)^{2}}\binom{N}{r} \Phi^{r}(7-\Phi)^{N-r}
$$

However, it is difficult to give any closed-form expression for $E[\mp]$ and $E\left[T^{2}\right]$. Due to $-1<\left(1-\frac{2 r}{N}\right)<1$,

$$
E[T]=\sum_{s=0}^{\infty}(2 / N)^{s} \mu_{s}
$$

and

$$
E\left[T^{2}\right]=\sum_{s=0}^{\infty}(s+1)(2 / N)^{s} u_{s}
$$

where $\mu_{s}$ is the sth moment of Binomial distribution about the origin and can be easily obtained by evaluating the sth derivative of the moment generating function at the origin, i.e.

$$
\mu_{s}=\left[\frac{\partial^{s}}{\partial t^{S}}\left(1-\Phi+\Phi e^{t}\right)^{N}\right]_{t=0}
$$

It can be shown that

$$
\begin{aligned}
& E[T]=(1-2 \Phi)^{-1}+0(1 / N) \\
& E\left[T^{2}\right]=(1-2 \Phi)^{-2}+0(1 / N)
\end{aligned}
$$

Then asymptotically, i.e. as N becomes large,

$$
\text { Bias }\left\{\hat{p}_{k}\right\}=0
$$

and

$$
\operatorname{MSE}\left\{\hat{\hat{p}}_{\mathrm{k}}\right\}=\operatorname{Var}\left(\hat{\hat{p}}_{\mathrm{k}}\right)=(1-2 \Phi)^{-2} \operatorname{Var}\left(\hat{e}_{\mathrm{k}}\right)
$$

4. SAMPLING PLAN AND COVARIANCE MATRIX OF $\hat{e}$

In a remote sensing situation involving collection of data over a large region, we suggest a stratified random sampling scheme with three stages. First of all, stratification will be most effective if strata are formed on the basis of at least
(i) predominance of various crops,
(ii) latitude,
(iii) longitude.

The first factor would lead to homogeneity in various strata whereas the other two factors are important from the point of assessing $P$ appropriately.

Let $R$ denote the region for which crop acreages are to be estimated. With the consideration of (i) - (iii), suppose $R$ is stratified into strata
$R_{s t}, s=1,2, \ldots, a$ and $t=1,2, \ldots, b$, with weights $w_{s t}$, the proportion of acreage (number of pixels in a stratum divided by the total number of pixels in the region), i.e.

$$
R=\underset{s, t}{U} R_{s t}
$$

with

$$
1=\sum_{s, t} w_{s t}
$$

Due to (i), one may expect elimination of many scenes or even many strata if interest lies in only a few specific types of crops. Nevertheless, for sampling purposes, select ERTS scenes at the first stage, strips within scenes at the second stage and scanlines within strips at the third stage. Of course, one can go one more stage in selecting pixels within scanline. However, this would not be convenient as far as processing is concerned. As such, this stage is not considered for the sampling.

For stratum $\mathrm{R}_{\text {st }}$, we denote the following:

$$
\begin{aligned}
e_{s t ~} i j h k= & \text { expected proportions of pixels in } \pi_{i} \text { for kth scanline } \\
& \text { in hth strip of } j \text { th scene, } \\
e_{s t ~} i j h= & \text { expected proportions of pixels in } \pi_{i} \text { for hth strip in } \\
& j \text { th scene, } \\
e_{s t ~}^{i j}= & \text { expected proportion of pixels in } \pi_{i} \text { for } j \text { th scene, } \\
e_{s t ~}^{i}= & \text { expected proportion of pixels in } \pi_{i}
\end{aligned}
$$

Then

$$
e_{i}=\sum_{s=1}^{a} \sum_{t=1}^{b} w_{s t} e_{s t i}, \quad i=1,2, \ldots, m .
$$

Next, let $G_{s t}, H_{s t}, R_{s t}$ and $n$ denote the number of scenes, number of strips per scene, number of scanlines per strip and number of pixels per
scanline, respectively, for stratum $R_{s t}$. Suppose $g_{s t}, h_{s t}$ and $r_{s t}$ are the corresponding number of scenes, number of strips in a scene, number of scanlines in a strip that are selected randomly at three stages. Let $n_{s t}$ ijhk be the number of pixels classified into $\pi_{i}$ for $k$ th selected scanline in hth selected strip of jth selected scene. Then considering the observed proportions for estimates, one has

$$
\begin{aligned}
& \hat{e}_{s t ~ i j h k}=\frac{n_{s t} i j h k}{n} \\
& \hat{e}_{s t} i j h=\frac{1}{n r_{s t}} \sum_{k=1}^{r_{s t}} n_{s t} i j h k \\
& \hat{e}_{s t ~ i j}=\frac{1}{n r_{s t} h_{s t}} \sum_{h=1}^{h_{s t}} \sum_{k=1}^{r_{s t}} n_{s t} i j h k \\
& \hat{e}_{s t} i
\end{aligned}
$$

and

$$
\hat{e}_{i}=\sum_{s=1}^{a} \sum_{t=1}^{b} w_{s t} \hat{e}_{s t i}, i=1,2, \ldots, m
$$

For the covariance matrix,

$$
\begin{aligned}
\operatorname{Var}\left(\hat{e}_{s t} i\right. & =\left(1-\frac{g_{s t}}{G_{s t}}\right) \frac{1}{g_{s t}\left(G_{s t}-1\right)} \sum_{j=1}^{G_{s t}}\left(e_{s t} i j-e_{s t ~}\right)^{2} \\
& +\left(1-\frac{h_{s t}}{H_{s t}}\right) \frac{1}{g_{s t} h_{s t} G_{s t}\left(H_{s t}-1\right)} \sum_{j=1}^{G_{s t}} \sum_{h=1}^{H}\left(e_{s t} i j h^{-e} s t i j\right)^{2}
\end{aligned}
$$

$$
\begin{aligned}
& +\left(1-\frac{r_{s t}}{R_{s t}}\right) \frac{1}{g_{s t} h_{s t} r_{s t} G_{s t} H_{s t}\left(R_{s t}-1\right)} \sum_{j=1}^{G} \sum_{h=1}^{G_{s t}} \sum_{k=1}^{H_{s t}}\left(e_{s t} \quad i j h k^{-e_{s t}} i j h\right)^{2} \\
& \operatorname{Cov}\left(\hat{e}_{s t i}, \hat{e}_{s t} i^{i}\right)=\left(1-\frac{g_{s t}}{G_{s t}}\right) \frac{1}{g_{s t}\left(G_{s t}-1\right)} \sum_{j=1}^{G}\left(e_{s t i j}-e_{s t i}\right)\left(e_{s t} i j-e_{s t i}\right) \\
& +\left(1-\frac{h_{s t}}{H_{s t}}\right) \frac{1}{g_{s t} h_{s t} G_{s t}\left(H_{s t}-1\right)} \sum_{j=1}^{G} \sum_{h=1}^{G_{s t}}\left(e_{s t} \quad i j h^{-e_{s t}} i j\right)\left(e_{s t} \quad i j h^{-e_{s t}} i^{\prime} j\right) \\
& +\left(1-\frac{r_{s t}}{R_{s t}}\right) \frac{1}{g_{s t} h_{s t} r_{s t}{ }^{G} s t^{H}{ }_{s t}\left(R_{s t}-1\right)} \sum_{j=1}^{G_{s t}} \sum_{h=1}^{H_{s t}} \sum_{k=1}^{R_{s t}}\left(e_{s t i j h k}-e_{s t i j h}\right)\left(e_{s t i j h k}-e_{s t i^{\prime} j h}\right) .
\end{aligned}
$$

Thus the covariance matrix for $\hat{e}$ can be obtained because

$$
\begin{align*}
& \operatorname{Var}\left(\hat{e}_{j}\right)=\sum_{s=1}^{a} \sum_{t=1}^{b} w_{s t}^{2} \operatorname{Var}\left(\hat{e}_{s t i}\right)  \tag{9}\\
& \operatorname{Cov}\left(\hat{e}_{i}, \hat{e}_{i,}\right)=\sum_{s-1}^{a} \sum_{t=1}^{b} w_{s t}^{2} \operatorname{Cov}\left(\hat{e}_{s t i}, \hat{e}_{s t} i^{\prime}\right) . \tag{10}
\end{align*}
$$

For an estimate of the covariance matrix,

$$
\begin{aligned}
& \hat{\operatorname{Var}}\left(\hat{e}_{i}\right)=\sum_{s=1}^{a} \sum_{t=1}^{b} w_{s t}^{2} \hat{\operatorname{Var}}\left(\hat{e}_{s t}{ }_{i}\right) \\
& \left.\hat{\operatorname{Cov}}\left(\hat{e}_{i}, \hat{e}_{i} \prime\right)=\sum_{s=1}^{a} \sum_{t=1}^{b} w_{s t}^{2} \hat{\operatorname{Cov}\left(e_{s t} i\right.}, \hat{e}_{s t} i^{\prime}\right)
\end{aligned}
$$

where
where

$$
\begin{align*}
& {\hat{\operatorname{Var}}\left(\hat{e}_{s t i}\right)}=\left(1-\frac{g_{s t}}{G_{s t}}\right) \frac{1}{g_{s t}\left(g_{s t}-1\right)} \sum_{j=1}^{g_{s t}}\left(\hat{e}_{s t i j}-\hat{e}_{s t i}\right)^{2} \\
& \quad+\left(1-\frac{h_{s t}}{H_{s t}}\right) \frac{1}{G_{s t} g_{s t} h_{s t}\left(h_{s t}-1\right)} \sum_{j=1}^{g_{s t}} \sum_{h=1}^{h_{s t}}\left(\hat{e}_{s t ~} i j h_{h} \hat{e}_{s t} i j\right)^{2} \\
& \quad+\left(1-\frac{r_{s t}}{R_{s t}}\right) \frac{1}{G_{s t} H_{s t} g_{s t} h_{s t} r_{s t}\left(h_{s t}-1\right)} \sum_{j=1}^{g_{s t}} \sum_{h=1}^{h_{s t}} \sum_{k=1}^{r_{s t}}\left(\hat{e}_{s t i j h k}-\hat{e}_{s t i j h}\right)^{2} . \tag{11}
\end{align*}
$$

Similarly one can write $\operatorname{Cov}\left(\hat{e}_{s t} \hat{i}, \hat{e}_{s t} i^{\prime}\right)$ by replacing the sum of squares terms in the variance by the corresponding sum of product terms.

## 5. OPTIMUM SAMPLE SIZE

Taking the cost factor into consideration one may want to know the sample size that either minimizes the cost for a specified mean square error or minimize the mean square error for a given cost. In remote sensing, the sampling cost would mainly involve a large initial cost plus the processing cost. Although it would depend upon a situation, the cost function may be considered as

$$
c_{s t}=c_{1} g_{s t}+c_{2} g_{s t} h_{s t}+C_{3} g_{s t} h_{s t} r_{s t}
$$

for when sampling in stratum $R_{s t}$, and

$$
c=c_{1} \Sigma g_{s t}+c_{2} \Sigma g_{s t} h_{s t}+c_{3} \Sigma g_{s t} h_{s t} r_{s t}
$$

for the whole region.

In the present context, there is an additional cost of taking samples for estimating $P$ and this can be expressed in the form of
$C^{\prime}=\sum_{i=1}^{m} c_{0 i} n_{i}$. Thus the overall cost involved is given by $i=1$

$$
C^{\prime \prime}=C^{\prime}+C \text {. }
$$

Next, the mean square errors, $\operatorname{MSE}\left\{\hat{\mathrm{p}}_{\mathrm{k}}\right\}, k=1,2, \ldots, m$, are obtained from (8) after making substitution from (9) and (10). Now, if the cost is fixed, say $C^{\prime \prime} \leq C_{0}$, a determination of sample sizes, $n_{i}{ }^{\prime} s, g_{s t}{ }^{\prime} s$, $g_{s t}{ }^{\prime} s, h_{s t}$ 's and $r_{s t}{ }^{\prime} s$, can be achieved by solving equations obtained by equating the partial derivatives of $\operatorname{MSE}\left\{\hat{p}_{k}\right\}+\lambda\left(C^{\prime \prime}-C_{0}\right), k=1,2, \ldots, m$ and $\lambda$ a Lagrange multiplier, with respect to $\eta_{j}{ }^{\prime} s, g_{s t}$ 's, $h_{s t}$ 's and $r_{s t}{ }^{\prime} s$ to zero. Similarly, for any fixed values, say $\sigma_{k}^{2}$, for $\operatorname{MSE}\left\{\hat{\hat{p}}_{k}\right\}, k=1,2, \ldots, m$, this can again be achieved by considering the function $C^{\prime \prime}+\lambda_{K}\left(\operatorname{MSE}\left\{\hat{p}_{k}\right\}-\sigma_{k}^{2}\right), k=1,2, \ldots, m$, for minimization. Of course, this procedure may lead to $k$ different values for various sample sizes. For a unique determination, take the largest value in each case.

It may be noted that under this procedure, it is not possible to give any closed form expression for any sample size and its carrying out would involve some optimization technique.

One direct way to simplify the problem is to treat the two types of cost separately. For example, if we consider only the $\cos t C$ and variances given in (9) for the purpose of determining $g_{s t}$ ' $s, h_{s t}$ ' $s$ and $r_{s t}$ 's, the problem is greatly simplified. Denoting

$$
\left(G_{s t}-1\right) s_{s t ~}^{2}=\sum_{j=1}^{G_{s t}}\left(e_{s t i j}-e_{s t i}\right)^{2}
$$

$$
G_{s t}\left(H_{s t}-1\right) s_{s t 2}^{2}=\sum_{j=1}^{G_{s t}} \sum_{h=1}^{H_{s t}}\left(e_{s t i j h^{-e}}^{s t} i j\right)^{2}
$$

and

$$
G_{s t} H_{s t}\left(R_{s t}-1\right) s_{s t ~ 3}^{2}=\sum_{k=1}^{R_{s t}} \sum_{h=1}^{H_{s t}} \sum_{j=1}^{G_{s t}}\left(e_{s t i j h k}-e_{s t i j h}\right)^{2} .
$$

(9) may be written as

$$
\begin{gathered}
\operatorname{var}\left(\hat{e}_{j}\right)=\sum_{s, t} w_{s t}^{2}\left[\frac{1}{g_{s t}}\left(s_{s t}^{2}-s_{s t 2}^{2} / H_{s t}\right)+\frac{1}{g_{s t} h_{s t}}\left(s_{s t 2}^{2}-s_{s t ~}^{2} / R_{s t}\right)\right. \\
\\
\left.+s_{s t 3}^{2} / g_{s t} h_{s t} r_{s t}-s_{s t 1}^{2} / G_{s t}\right]
\end{gathered}
$$

which is a function of $g_{s t}, g_{s t}{ }^{h} s t$ and $g_{s t}{ }^{h} s t r_{s t}$ as is the case with the above cost function. Thus, the quantity $\operatorname{Var}\left(\mathrm{e}_{\mathrm{i}}\right)+\lambda\left(\mathrm{C}-\mathrm{C}_{0}\right)$ or $C+\lambda_{i}\left(\operatorname{Var}\left(\hat{e}_{j}\right)-V_{i}^{2}\right)$, and equivalently $\operatorname{Var}\left(\hat{e}_{i}\right)$ for fixed $C$ or $C$ for fixed $\operatorname{Var}\left(\hat{e}_{i}\right)$, is minimized when

$$
\begin{aligned}
& g_{s t}=w_{s t} \sqrt{\left(s_{s t}^{2} 1^{\left.-s_{s t ~}^{2} / H_{s t}\right) / \lambda C_{1}}\right.} \\
& g_{s t} h_{s t}=w_{s t} \sqrt{\left(s_{s t ~ 2}^{2}-s_{s t ~}^{2} / R_{s t}\right) / \lambda C_{2}} \\
& g_{s t}{ }_{s t} l_{s t}=w_{s t} \sqrt{s_{s t ~}^{2} / \lambda C_{3}}
\end{aligned}
$$

Accordingly,

$$
\begin{aligned}
& r_{s t}=s_{s t} 3 \sqrt{C_{2}} / \sqrt{C_{3}\left(S_{s t}^{2}-S_{s t ~}^{2} / R_{s t}\right)} \\
& h_{s t}=\sqrt{C_{1}\left(s_{s t ~}^{2} 2^{-S_{s t ~}^{2}} 2 / R_{s t}\right) / C_{2}\left(s_{s t ~}^{2}-s_{s t ~}^{2} / H_{s t}\right)}
\end{aligned}
$$

and

$$
g_{s t} \alpha w_{s t} \sqrt{\left(s_{s t ~}^{2}-s_{s t 2}^{2} / H_{s t}\right) / c_{1}} .
$$

These values are, of course, for the case of $\operatorname{Var}\left(\hat{e}_{\mathbf{j}}\right)$. Similarly, one can obtain sample sizes corresponding to other variances. By choosing the maximum of these values in different cases, a unique solution can be obtained.

## 6. FURTHER COMMENTS

When stratification is based upon factors of latitude and longitude, one might expect different P's over different strata. However, if proper adjustment can be made in the classification algorithm with respect to spectral variation so that $P$ remains the same, there is no need to make any change in the above procedure. On the other hand, one should find both the actual proportion estimate and its mean square error separately for each stratum and then by combining these one is able to obtain so called $\hat{\hat{p}}$ and its mean square error. In this situation the formula in (8) is still valid but stratumwise.

Our discussion given in Section 1-3 is quite general and can be specialized and different sampling schemes can be adopted. For example, if our inference set is only one scene, then proportion estimates and their mean square errors are again obtained as discussed in Section 1-3, but as to sampling scheme there may not be any need of stratification and the sampling is done in two stages rather than in three stages.

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# ON COMBINING POPULATIONS IN STATISTICAL <br> CLASSIFICATION USING REMOTE SENSING DATA 

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## ABSTRACT


#### Abstract

Observations may be known to be coming from $m+1$ normal populations having same dispersion matrix; but one may be interested in identifying observations from $\pi_{0}$ only. Then the usual practice is to merge the other populations $\pi_{1}, \ldots, \pi_{m}$ into one single population $\pi$ and assume it to be normal. This is done in order to achieve the computational convenience of two population classification. This paper investigates the effectiveness of such practice. It has been found that in some situations this practicle may decrease the proper classification probability of observations from $\pi_{0}$ considerably.


Key Words

Dispersion matrix, Bayes' procedure, misclassification probability, proper classification probability, prior probability, misclassification cost.

## 1. Introduction

Let us suppose that all of our observations come from ( $m+1$ ) p-variate normal populations $\pi_{0}, \pi_{1}, \pi_{2}, \ldots, \pi_{m}$ with densities $N_{p}\left({ }_{i}, V\right)(i=0,1,2, \ldots, m)$, where $\mu_{i}^{\top}=\left(\mu_{i 1}, \mu_{i 2}, \ldots, \mu_{i p}\right)$ is a vector of $p$ means of the populations $\pi_{i}$ and $V$ is the dispersion matrix, same for all populations. In general the problem of classification into more than two classes ( $m+1>p>1$ ) is more complex and computationally inconvenient than the problem of classification into two classes. In some situations, instead of finding which population an observation has come from, one may be interested in finding only if the observation has come from some particular population, say $\pi_{0}$. There, for achieving the convenience and simplicity of two-population classification problem, one may think of merging the populations $\pi_{1}, \pi_{2}, \ldots, \pi_{m}$ into one single population $\pi$ and further assume the resulting single population $\pi$ to be normal. Then the question that needs to be answered is how badly the classification procedure based on such assumptions affects the misclassification probabilities. This paper endeavors to answer this question.

In Bayes' procedure the prior probabilities play a significant role. Improper choice of prior probabilities affects the misclassification probabilities. The usual practice in remote sensing data analysis is to assume the prior probabilities $\mathrm{q}_{0}$ and q of $\mu_{0}$ and $\mu$ to be equal, even though it has been assumed that the prior probabilities $q_{0}, q_{1}, \ldots, q_{m}$ of $\pi_{0}, \pi_{1}, \ldots, \pi_{m}$ are equal. Therefore in answering the question raised in the above paragraph we should take into consideration our assumption about the prior probabilities.

Thus our study will be concerned with combination of following sets of assumptions.

## Assumptions regarding population densities

(1) Before merger the populations $\pi_{0}, \pi_{1}, \ldots, \pi_{m}$ have densities $N_{p}\left({ }_{i}, v\right), i=0,1, \ldots, m ;$
(2) After merger, $\pi_{0}$ has density $N_{p}\left(\mu_{0}, V\right)$ and the density of $\pi$ is a linear combination of the densities $N_{p}\left(\mu_{i}, V\right), i=1, \ldots, m$;
(3) After merger $\pi_{0}$ has density $N_{p}\left(\mu_{0}, V\right)$ and $\pi$ has a density $N_{p}\left(\mu, V^{\prime}\right)$, where $\mu$ and $V^{\prime}$ will be specified later.

Assumptions regarding prior probabilities
(a) $q_{0}=1 /(m+1), q_{1}=q_{2}=\ldots=q_{m}=1 /(m+1)$;
(b) $q_{0}=1 / 2 \quad, q_{1}=q_{2}=\ldots=q_{m}=1 / 2 m$.
2. Dispersion Matrix of the Population $\pi$

The true density $p(x)$ of the population $\pi$ is given by

$$
\begin{equation*}
p(x)=\sum_{i=1}^{m}\left[q_{i} /\left(1-q_{0}\right)\right] N_{p}\left(\mu_{i}, v\right) \tag{1}
\end{equation*}
$$

Now. since in both assumptions (a) and (b) we have
and

$$
\begin{aligned}
& q_{1}=q_{2}=\ldots=q_{m} \\
& q_{i} /\left(1-q_{0}\right)=q_{1} /\left(1-q_{0}\right)=1 / m
\end{aligned}
$$

then the density $p(x)$ under (a) or (b) is given by

$$
\begin{equation*}
p(x)=(1 / m) \sum_{i=1}^{m} N_{p}\left(u_{i}, v\right) \tag{2}
\end{equation*}
$$

Therefore the mean $\mu$ of $\pi$ is given by

$$
\begin{align*}
\mu=E_{\pi}(X) & =\int_{R_{p}} x p(x) d x=\frac{1}{m} \sum_{i=1}^{m} \int_{R_{p}} x N_{p}\left(\mu_{i}, V\right) d x \\
& =\frac{1}{m}\left(\mu_{1}+\ldots+\mu_{m}\right)=\bar{\mu}, \tag{3}
\end{align*}
$$

and the dispersion matrix $V^{\prime}$ of $\pi$ is given by

$$
\begin{align*}
V^{\prime} & =E_{\pi}\left[(x-\bar{\mu})(x-\bar{\mu})^{T}\right] \\
& =\frac{1}{m} \sum_{i=1}^{m} \int_{p}(x-\bar{\mu})(x-\bar{\mu})^{T} N_{p}\left(\mu_{i}, v\right) d x \\
& =\frac{1}{m} \sum_{i=1}^{m} \int_{R_{p}} \quad\left\{\left(x-\mu_{i}\right)\left(x-\mu_{i}\right)^{T}+2\left(\mu_{i}-\bar{\mu}\right)^{T}\left(x-\mu_{i}\right)+\left(\mu_{i}-\bar{\mu}\right)\left(\mu_{i}-\bar{\mu}\right)^{T}\right\} N_{p}\left(\mu_{i}, v\right) d x \\
& =V+\frac{1}{m} \sum_{i=1}^{m}\left(\mu_{i}-\bar{\mu}\right)\left(\mu_{i}-\bar{\mu}\right)^{\top} . \tag{4}
\end{align*}
$$

Here $E_{\pi}(\cdot)$ denotes expectation of the population $\pi$.
Therefore, if we want to use a normal density function for $\pi$ instead of the density $p(x)$ given by (2), we have to use $N_{p}\left(\bar{\mu}, v^{\prime}\right)$, where $v^{\prime}$ is given by (4).
Expression for (V')

Let us write $V_{0}=V$ and

$$
\begin{equation*}
v_{i}=v_{i-1}+\frac{1}{m}\left(\mu_{i}-\bar{\mu}\right)\left(\mu_{i}-\bar{\mu}\right)^{\top}, \quad(i=1,2, \ldots, m) \tag{5}
\end{equation*}
$$

Then $V_{m}=V^{\prime}$. We know (Rao, 1965; p. 29) that if $A$ is a nonsingular $p \times p$ matrix and $\alpha$ and $\beta$ are two $\beta \times 1$ vectors, then

$$
\begin{equation*}
\left(A+\alpha \beta^{T}\right)^{-1}=A^{-1}-\frac{A^{-1} \alpha \beta^{T_{A}^{-1}}}{1+\beta^{\top} A^{-1} \alpha} \text {. } \tag{6}
\end{equation*}
$$

Thus, $\quad V_{i}^{-1}=V^{-1}-\frac{1}{m} V^{-1}\left(\mu_{1}-\bar{\mu}\right)\left(\mu_{1}-\vec{\mu}\right)^{T} V^{-1} /\left\{1+\left(\mu_{1}-\bar{\mu}\right)^{\top} V^{-1}\left(\mu_{1}-\mu_{2}\right)\right\}$,
and

$$
\begin{equation*}
V_{i}^{-1}=V_{i-1}^{-1}-\frac{1}{m} V_{i-1}^{-1}\left(\mu_{i}-\bar{\mu}\right)\left(\mu_{i}-\bar{\mu}\right)^{\top} V_{i-1}^{-1} /\left\{1+\left(\mu_{i}-\bar{\mu}\right)^{\top} V_{i-1}^{-1}\left(\mu_{i}-\bar{\mu}\right)\right\}, \tag{7}
\end{equation*}
$$

( $i=2, \ldots, m$ ) Therefore

$$
\begin{equation*}
\left(V^{\prime}\right)^{-1}=v^{-1}-M(\text { say }) . \tag{8}
\end{equation*}
$$

The exact expression for $M$ that can be found recursively from (7) is not a simple one.

## 3. Acceptance Region For $\pi_{0}$ : Assumptions (1) and (2)

We will denote the acceptance region of $\pi_{0}$ given by a Bayes' classification procedure by $B . B_{2 a}$, for example, will denote the region $B$ obtained under the assumptions (2) and (a). If the populations $\pi_{0}, \pi_{1}, \ldots, \pi_{m}$ have densities $p_{0}(x), p_{1}(x), \ldots, p_{m}(x)$ and prior probabilities $q_{0}, q_{1}, \ldots, q_{m}$ and $C\left(\pi_{j} \mid \pi_{i}\right),(j \neq i)$, denote the cost of misclassifying an observation from $\pi_{i}$ into $\pi_{j}$, then it is well known (Anderson, 1958; p. 143) that the acceptance region $B$ for $\pi_{0}$ is given according to Bayes' procedure by

$$
\begin{equation*}
B==_{j=1}^{m}\left\{x: \sum_{i=1}^{m} q_{i} p_{i}(x) C\left(\pi_{0} \mid \pi_{i}\right)<\sum_{i \neq j} q_{i} p_{i}(x) C\left(\pi_{j} \mid \pi_{i}\right)+q_{o} p_{o}(x) C\left(\pi_{j} \mid \pi_{0}\right)\right\} . \tag{9}
\end{equation*}
$$

In case of assumption (1) if we assume all costs of misclassification to be equal, then we have

$$
\begin{align*}
B_{1} & =n_{j=1}^{m}\left\{x: q_{0} p_{0}(x)>q_{j} p_{j}(x)\right\} \\
& =\left\{x: q_{0} p_{0}(x)>\max _{1 \leq j \leq m}\left[q_{j} p_{j}(x)\right]\right\} . \tag{10}
\end{align*}
$$

Therefore, $\quad B_{l a}=\left\{x: p_{0}(x)>\max _{l \leq j \leq m} p_{j}(x)\right\}$
and

$$
\begin{equation*}
B_{1 b}=\left\{x: p_{0}(x)>(1 / m) \max _{l \leq j \leq m} p_{j}(x)\right\} . \tag{12}
\end{equation*}
$$

In case of assumption (2) we assume that

$$
C\left(\pi_{0} \mid \pi_{i}\right)=C\left(\pi_{i} \mid \pi_{0}\right)=C\left(\pi_{j} \mid \pi_{0}\right), \text { for all } \mathbf{i} \neq \mathbf{j} \neq 0
$$

and

$$
c\left(\pi_{i} \mid \pi_{j}\right)=0, \quad \text { for all } i, j \neq 0 .
$$

Then we have

Therefore,

$$
\begin{align*}
B_{2} & =\left\{x: q_{0} p_{0}(x)>q_{1} p_{1}(x)+\ldots+q_{m} p_{m}(x)\right\} \\
& =\left\{x: q_{0} p_{0}(x)>\left(1-q_{0}\right) \sum_{i=1}^{m}\left[q_{i} /\left(1-q_{0}\right)\right] p_{i}(x)\right\} . \tag{13}
\end{align*}
$$

and

$$
\begin{equation*}
B_{2 a}=\left\{x: p_{0}(x)>p_{1}(x)+\ldots+p_{m}(x)\right\} \tag{14}
\end{equation*}
$$

$$
\begin{equation*}
B_{2 b}=\left\{x: p_{0}(x)>(1 / m)\left[p_{1}(x)+\ldots+p_{m}(x)\right]\right\} . \tag{15}
\end{equation*}
$$

Now since $\max \left\{p_{1}(x), \ldots, p_{m}(x)\right\} \geq(1 / m)\left[p_{1}(x)+\ldots+p_{m}(x)\right]$, the following set theoretic inequalities are evident.

$$
\begin{equation*}
\mathrm{B}_{1 \mathrm{~b}} \supset \mathrm{~B}_{2 \mathrm{~b}} \supset \mathrm{~B}_{1 \mathrm{a}} \supset \mathrm{~B}_{2 \mathrm{a}} . \tag{16}
\end{equation*}
$$

When $p_{i}(x)=N_{p}\left(\mu_{i}, V\right), \quad(i=0,1, \ldots, m)$, then we know (Anderson, 1958; 147)
that $B_{1}$ can be expressed as follows.

$$
\begin{align*}
B_{1} & =\bigcap_{j=1}^{m}\left\{x: W_{j}(x) \geq \log \left(q_{j} q_{0}\right)\right\} \\
& =\left\{x: W_{j}(x)>\log \left(q_{j} \mid q_{0}\right), \text { for } j=1, \ldots, m\right\}, \tag{18}
\end{align*}
$$

where $W_{j}(x)=\left(\mu_{0}-\mu_{j}\right)^{\top} V^{-1}\left[x-(1 / 2)\left(\mu_{0}+\mu_{j}\right)\right]$.
Therefore $\quad B_{l a}=\left\{x: W_{j}(x) \geq 0\right.$, for $\left.j=1, \ldots, m\right\}$
and $\quad B_{1 b}=\left\{x: W_{j}(x) \geq-\log m\right.$, for $\left.j=1, \ldots, m\right\}$.
When the means $\mu_{1}, \mu_{2}, \ldots, \mu_{m}$ are collinear, without loss of generality we can assume the populations to be two dimensional and assume the means to be as follows:

$$
\mu_{0}=\left[\begin{array}{l}
a_{0}  \tag{20}\\
b_{0}
\end{array}\right], \quad \mu_{i}=\left[\begin{array}{l}
a_{i} \\
0
\end{array}\right], \quad(i=1,2, \ldots, m) .
$$

In case of three populations, we can take

$$
\mu_{0}=\left[\begin{array}{l}
a_{0}  \tag{21}\\
b_{0}
\end{array}\right], \quad \mu_{1}=\left[\begin{array}{c}
-a \\
0
\end{array}\right] \quad \text { and } \mu_{2}=\left[\begin{array}{c}
a \\
0
\end{array}\right]
$$

Let the $2 \times 2$ disperion matrix $V$ be given by $V=\left[v_{i j}\right]$. Then we have

$$
\begin{align*}
w_{1}= & (1 /|v|)\left[x_{1}-(1 / 2)\left(a_{0}-a\right)\right]\left[\left(a_{0}+a\right) v_{22}-b_{0} v_{12}\right] \\
& +(1 /|v|)\left[x_{2}-(1 / 2) b_{0}\right]\left[b_{0} v_{11}-v_{12}\left(a_{0}+a\right)\right] \tag{22}
\end{align*}
$$

and

$$
\begin{align*}
w_{2}= & (1 /|v|)\left[x_{1}-(1 / 2)\left(a_{0}+a\right)\right]\left[\left(a_{0}-a\right) v_{22}-b_{0} v_{12}\right] \\
& +(1 /|v|)\left[x_{2}-(1 / 2) b_{0}\right]\left[b_{0} v_{11}-v_{12}\left(a_{0}-a\right)\right] . \tag{23}
\end{align*}
$$

The acceptance region $B_{1}$ for $\pi_{0}$ will be bounded by two straight lines, viz, $W_{1}=d$ and $W_{2}=d$, where $d=0$ in case (a) and $d=-\log 2$ in case (b). The point of intersection of these two lines will be given by
where

$$
\begin{gathered}
c=\left[\left\{\left(a_{0}^{2}-a^{2}\right) v_{22}-2 a_{0} b_{0} v_{12}+b_{0}^{2} v_{11}\right\}+2 d|v|\right] \\
D=2 b_{0}\left(v_{11} v_{22}-v_{12}^{2}\right)=2 b_{0}|v|
\end{gathered}
$$

$|V|$ denoting the determinant of $V$.

From (14) and (15) it follows that the acceptance region $B_{2 a}$ and $B_{2 b}$ can be given the following general description.

$$
B_{2}=\left\{x: p_{0}(x)>(1 / \alpha)\left[p_{1}(x)+\ldots+p_{m}(x)\right]\right\}
$$

where $\alpha$ should be taken as 1 for $B_{2 a}$ and as $m$ for $B_{2 b}$. In case of three populations with means given by (21) we may note that the condition

$$
\alpha p_{0}(x)>p_{1}(x)+p_{2}(x)
$$

is equivalent to

$$
\log \left[\alpha p_{0}(x) / p_{1}(x)\right]>\log \left[p_{2}(x) / p_{1}(x)+1\right]
$$

Noting that $\mu_{1}=-\mu_{2}$, we can write the above condition as

$$
\begin{equation*}
\log a+\left(\mu_{2}+\mu_{0}\right)^{\top} v^{-1}\left[x+\left(\mu_{2}-\mu_{0}\right) / 2\right]>\log \left[1+\exp \left(2 \mu_{2}^{\top} v^{-1} x\right)\right] \tag{25}
\end{equation*}
$$

Here $\alpha$ should be taken as 1 for $B_{2 a}$ and as 2 for $B_{2 b}$.
When all means are collinear, without loss of generality we can assume the populations to be univariate and $\mu_{1}<\mu_{2}<\ldots<\mu_{m} . \quad V$ in this case is the common variance of the univariate populations. Now from (18) we have

$$
W_{j}(x)=\left[\left(\mu_{0}-\mu_{j}\right) / V\right]\left[x-(1 / 2)\left(\mu_{0}+\mu_{j}\right)\right]
$$

It now follows from (19) that if $\mu_{0}<\mu_{1}$, then

$$
\begin{equation*}
B_{1 a}=\left\{x: x<\left(\mu_{0}+\mu_{1}\right) / 2\right\}, \tag{25}
\end{equation*}
$$

if $\mu_{i}<\mu_{0}<\mu_{i+1}$, then

$$
\begin{equation*}
B_{l a}=x:\left\{\left(\mu_{0}+\mu_{i}\right) / 2<x<\left(\mu_{0}+\mu_{i+1}\right) / 2\right\} \tag{27}
\end{equation*}
$$

and if $\mu_{0}>\mu_{m}$, then

$$
\begin{equation*}
B_{1 a}=\left\{x: x>\left(\mu_{0}+\mu_{m}\right) / 2\right\} . \tag{20}
\end{equation*}
$$

The positions of the end points of the interval $B_{2 a}$ depend on Vlogm and the values of $\mu_{0}, \mu_{1}, \ldots, \mu_{m} \cdot B_{2 a}$ can not be given a general description like $B_{\text {la }}$.
4. Acceptance Region for $\pi_{0}$ : Assumption (3)

When $m$ populations $\pi_{1}, \ldots, \pi_{m}$ have been merged to form the population $\pi$, according to assumption (a) the prior probabilities $q_{0}$ and $q$ of $\pi_{0}$ and $\pi$ are then given by $q_{0}=1 /(m+1)$ and $q=m /(m+1)$ and according to assumption (b) they are given by $q_{0}=1 / 2$ and $q=1 / 2$. When $\pi$ is assumed to have density $N_{p}\left(\bar{\mu}, V^{\prime}\right)$, the acceptance region $B_{3 a}$ of $\pi_{0}$ will be given according to Bayes' procedure by

$$
\begin{aligned}
& B_{3 a}=\left\{x: N_{p}\left(\mu_{0}, v\right)>\operatorname{mN}_{p}\left(\bar{\mu}, v^{\prime}\right)\right\} \\
&=\left\{x:-(x-\bar{\mu})^{\top}\left(V^{\prime}\right)^{-1}(x-\bar{\mu})+\left(x-\mu_{0}\right)^{\top} V^{-1}\left(x-\mu_{0}\right)\right. \\
&<\log (|v '| /|v|)-2 \operatorname{logm}\} .
\end{aligned}
$$

Using (8) we can write

$$
\begin{equation*}
B_{3 a}=\{x: Q(x)<\log (|v \cdot| /|v|)-2 \log m\}, \tag{29}
\end{equation*}
$$

where $Q(x)=(x-\bar{\mu})^{\top} M(x-\bar{\mu})-2\left(\mu_{0}-\bar{\mu}\right)^{\top} V^{-1}\left[x-\left(\mu_{0}+\bar{\mu}\right) / 2\right]$.
We can similarly have

$$
\begin{equation*}
B_{3 b}=\{x: Q(x)<\log (|v| /|v|)\} . \tag{31}
\end{equation*}
$$

Obviously $B_{3 b}=B_{3 a}$.

## Special Case

Let us suppose that the populations are two dimensional, their means are given by (20) subject to the condition $\bar{\mu}=0$ and their common dispersion matrix is given by $V=$ diag $\left\{v_{1}, v_{2}\right\}$. Then from (4) we have

$$
\begin{equation*}
v^{\prime}=v+\sum_{i}^{m} \mu_{i} \mu_{i}^{T} / m=\operatorname{diag}\left\{v_{1}+\sum \mu_{i} \mu_{i}^{T} / m, v_{2}\right\} \tag{32}
\end{equation*}
$$

Therefore

$$
\begin{equation*}
V^{\prime}\left|/|V|=1+\Sigma \mu_{i} i_{i} / m v_{1}\right. \tag{33}
\end{equation*}
$$

Writing

$$
\omega^{\top}=\left[\left(\Sigma \mathrm{a}_{\mathrm{i}}^{2} / \mathrm{m}\right)^{\frac{1}{2}} 0\right]
$$

we can express $\mathrm{V}^{\prime}$ as

$$
v^{\prime}=V+\omega \omega^{\top} .
$$

From (6) and (8) we now have

$$
M=\left(m_{i j}\right)=V^{-1} \omega \omega^{\top} V^{-1} /\left[1+\omega^{\top} V^{-1} \omega\right\}
$$

Obviously $M$ is a matrix such that

$$
m_{11}=\Sigma a_{i}^{2} /\left\{v_{1}\left(m v_{1}+a_{i}^{2}\right)\right\}
$$

and

$$
m_{12}=m_{21}=m_{22}=0 .
$$

Therefore, from (20) and (30) we have

$$
\begin{aligned}
Q(x) & =m_{11} x_{1}^{2}-2 \mu_{0}^{\top} v^{-1} x+\mu_{0}^{\top} v^{-1} \mu_{0} \\
& =m_{11} x_{1}^{2}-2 a_{0} x_{1} / v_{1}-2 b_{0} x_{2} / v_{2}+\left(a_{0}^{2} / v_{1}+b_{0}^{2 / v_{2}}\right)
\end{aligned}
$$

From (29) and (31) it now follows that the regions
$B_{35}(s=a, b)$ are bounded by

$$
\begin{equation*}
\left\{x_{1}-\left(a_{0} / v_{1} m_{11}\right)\right\}^{2}=2\left(b_{0} / m_{11} v_{2}\right)\left\{x_{2}-A+\left(v_{2} / b_{0}\right) \log \alpha\right\} \tag{34}
\end{equation*}
$$

where $\alpha$ should be taken as $m$ for $B_{3 a}$ and as 1 for $B_{3 b}$ and

$$
\begin{equation*}
A=b_{0}^{\prime} 2+a_{0}^{2} v_{2}\left(m_{1} v_{1}-1\right) /\left\{2 b_{0} m_{11}\right)-\left(v_{2} / 2 b_{0}\right) \log \left(\mid v^{\prime} 1 / 1 v 1\right) \tag{35}
\end{equation*}
$$

A boundary described by (34) is a parabola with vertex at the point

$$
\begin{equation*}
\left(a_{0} /\left(v_{1} m_{11}\right), A-\left(v_{2} / b_{0}\right) \log \alpha\right) \tag{36}
\end{equation*}
$$

and symmetric about the point

$$
\begin{equation*}
x_{1}=a_{0} / v_{1}^{m}{ }_{11} \tag{37}
\end{equation*}
$$

When there are three populations with means given by (21) and $V=\operatorname{diag}\left\{v_{1}, v_{2}\right\}$, we then have the regions $B_{3 s}(s=1,2)$ bounded by a parabola $\left(x_{1}-f\right)^{2}=4 k\left(x_{2}-g\right)$, where

$$
\begin{align*}
& f=a_{0}\left(1+a^{2}\right) / a^{2} v_{1} \\
& k=b_{0}\left(1+a^{2}\right) / 2 v_{2} a^{2} \tag{38}
\end{align*}
$$

and

$$
\begin{aligned}
g=b_{0} / 2 & +a_{0}^{2} 2\left(a^{2} v_{1}-1-a^{2}\right) /\left(2 a^{2} b_{0}\right) \\
& -\left(v_{2} / 2 b_{0}\right)\left\{\log \left(1+a^{2} / v_{1}\right)-2 \log \alpha\right\}
\end{aligned}
$$

$\alpha$ being equal to 2 for $B_{3 a}$ and to 1 for $B_{3 b}$.

## 5. Classification Probabilities

The acceptance region of $\Pi_{0}$ being given by $B_{i s}(i=1,2,3$ and $s=a, b)$ under different assumptions, the true probability of misclassification of an element from the 'other' population $\pi$ into $\Pi_{0}$ is given by

$$
\begin{equation*}
P\left(B_{i s} \mid \Pi\right)=\sum_{j=1}^{m}\left[q_{j} /\left(i-q_{o}\right)\right] \int N_{p}\left(\mu_{j}, V\right) \tag{39}
\end{equation*}
$$

and that of an element of $\Pi_{0}$ into $\Pi$ by

$$
\begin{equation*}
P\left(B_{i s}^{c} \mid \pi_{0}\right)=\int_{B_{i s}^{c}}^{C} N_{p}\left(\mu_{0}, v\right) \tag{40}
\end{equation*}
$$

where $A^{c}$ denotes the complement of a set $A$. As the regions $B_{2 a}, B_{2 b}$, $B_{3 a}$ or $B_{3 b}$ are bounded by conicoids, the integrals in (39) and (40) can not be evaluated in closed form.

In special cases, when the populations are two dimensional and the common dispersion matrix is diagonal, the boundaries of $B_{3 a}$ and $B_{3 b}$ are parabolas. The integrals in (39) and (40) cannot be evaluated in closed form even in this case. Therefore, it is not possible to state precisely
in terms of misclassification probability how much we have to pay if we want to reduce our many population classification problem into a two population classification problem and which one of the alternative assumptions $1 \mathrm{~b}, 2 \mathrm{a}, 2 \mathrm{~b}, 3 \mathrm{a}$ and 3 b is preferable next to the true assumption la. In special cases, it may be possible to make some inference looking at the graphs.

## Examples

In the following examples we consider three populations with densities $N_{2}\left(\mu_{i}, v\right), i=0,1,2$, where

$$
v=\left[\begin{array}{ll}
1 & 0 \\
0 & 4
\end{array}\right], \mu_{0}=\left[\begin{array}{l}
a_{0} \\
b_{0}
\end{array}\right], \mu_{1}=\left[\begin{array}{c}
-a \\
0
\end{array}\right] \text { and } \mu_{2}=\left[\begin{array}{l}
a \\
0
\end{array}\right]
$$

$a$, $a_{0}$, $b_{o}$ being specified separately in each example. All the populations have the same prior probability.

Example 1. Let $a=1, a_{0}=3$ and $b_{0}=6$. Then we have

$$
\begin{aligned}
& B_{1 a}=\{(x, y): 8 x+3 y-17 \geq 0,4 x+3 y-17 \geq 0\}, \\
& B_{1 b}=\{(x, y): 8 x+3 y-15 . b 2 \geq 0,4 x+3 y-15 . b 2 \geq 0\}, \\
& B_{2 a}=\{(x, y): 3 y \geq 2 \log [1+\exp (2 x)]-8 x+17\}, \\
& B_{2 b}=\{(x, y): 3 y \geq 2 \log [1+\exp (2 x)-8 x+15 . b 2\}, \\
& B_{3 a}=\left\{(x, y):(x-b)^{2} \leq b(y-0.23)\right\}, \\
& B_{3 b}=\left\{(x, y):(x-b)^{2} \leq b(y+0.23)\right\} .
\end{aligned}
$$

The classification procedure defining the partition ( $B_{l a}, B_{l a}^{C}$ ) is optimal in the sense that it has the minimum expected cost of misclassification (which in this case is equal to the total misclassification probability). From the graph it is evident that the next best partition is ( $B_{2 a}, B_{2 a}^{C}$ ), then $\left(B_{2 b}, B_{2 b}^{c}\right)$, then $\left(B_{1 b}, B_{1 b}^{c}\right)$. The two classification procedures based on the assumption of normality of the mixture perform relatively poorly.



FIGURE 2

Of the two, the partition $\left(B_{3 a}, B_{3 b}^{C}\right)$ has lower misclassification probability than the partition $\left(B_{3 b}, B_{3 b}^{c}\right)$.

Example 2. Let $a=2, a_{0}=0$ and $b_{0}=8$. Then we have

$$
\begin{aligned}
& B_{1 a}=\{(x, y): x+y-3 \geq 0, y-x-3 \geq 0\} \\
& B_{1 b}=\{(x, y): x+y-2.66 \geq 0, y-x-2.66 \geq 0\} \\
& B_{2 a}=\{(x, y): 2 y \geq \log [1+\exp (4 x)]-2 x+6\}, \\
& B_{2 b}=\{(x, y): 2 y \geq \log [1+\exp (4 x)]-2 x+5.31\}, \\
& B_{3 a}=\left\{(x, y): x^{2}<5(y-3.94)\right\} \\
& B_{3 b}=\left\{(x, y): x^{2}<5(y-3.6)\right\} .
\end{aligned}
$$

The classification procedure defining the partition ( $B_{1 a}, B_{1 a}{ }^{C}$ ) is optimal and from the graph we may infer that the next best is the one with partition ( $B_{2 a}, B_{2 a}^{c}$ ). Unfortunately, in this case, no comparison of other procedures can be made even from the graph.

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## 

## On ciustermant cassimation rencedens

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It is known that the parent popilations have the seace. dispersion matrix, then in designing the

## ABSTPACT

When we bave a single sample of observations from two noreal poptintions with sare covarfance matrix, but no training sample from either population, then for classifyitg future observations, the usual practice is to.clueter. the sample into two nearest nelghbor clusters and design a bayes' classifier treating the two clusters as two training samples. The use of $\varepsilon_{1}$-distance is often advocated for such clustering. It has been shorm in this paper that auch advocation is not alwas reasonable.

## 2. Introduction

We suppose that our data have been obtained through remote sensing devices from two p-variate nomal populations $H_{1}$ and $H_{2}$ vith densitics $N_{p}\left(H_{1}, \Sigma\right)$ and $N_{p}\left(\mu_{2}, \Sigma\right)$ respectively, the vactors $\psi_{1}$ and $\psi_{2}$ of means and the commen dispersion matrix $c$ belne, unknom, and that we do not have any training sample, that is, we do not know the actual source of our niservations; but the data is essumed to be such that two and only two distinet wodes can be deternfued from then. For classiGyin; future obervations into these two populations on the basjs of such data, hire wand practice is to cluster tim datia futo two noarest. nefibibor clucters $C_{1}$ and $C_{2}$ wing metric or destance function defined by $\delta_{3^{-}}, s_{2}$, or $\ell_{\infty}{ }^{-n o z a n}$; and then design a fayes' cinosffler with the assurption that the population densitice are $H_{p}\left(i_{j}, \dot{r}_{j}\right)(j=1,2)$, wane $n_{j}$ and $\dot{r}_{j}$ are resace-



placed by $\Sigma$, where

$$
\begin{align*}
& \hat{\Sigma}=\left\{\left(N_{1}-1\right) \hat{\Sigma}_{2}+\left(N_{2}-1\right) \hat{\Sigma}_{2}\right\} /\left(\because_{1}+\because_{2}-2\right), \tag{1}
\end{align*}
$$

$N_{j}$ denotine, the number of elements of $C_{j}$. $A$ classification procedure defincd in this fashion will be referred to as cluster based classification procedure.

Let $P_{o}$ denote the Bayes' procedure based on complete knonledge of $\mathrm{N}_{\mathrm{P}}\left(\mathrm{H}_{\mathrm{j}}, \Sigma\right)(\mathrm{j}=1,2)$ and $\mathrm{P}_{\mathrm{i}}$ the cluster based procedure when $l_{1}$-nore ( $i=1,2, \infty$ ) has been used for clustering and the sample sizes are very larze. Our objective in this paper is. to compare the misclassification probabilities of these clascification procedures. he have made the comparison a little simpler by talina $p=$, $\mu_{1}=0$ and $\Sigma=I$, that is, by assuming our population densities to be $\mathrm{N}_{2}(0, \mathrm{I})$ and $\mathrm{N}_{2}\left(\mu_{2}, I\right)$. Ja designing the classificre we have also assumed that both populations have equal prior probabilitics (1/2 eacin) and equal misclassification costs. Now, without any further loss of generality, it can be assused that $H_{2}{ }^{T}=[a, h]$ and $a \geq 0, b \geq 0$.
2. Classification Procedare $P_{0}$

Under the set of assumptions rade above, the Baycs' clarsifier is givea by

$$
\begin{equation*}
\left.v_{o}(z)=a x+b y-\left(a^{2}+b^{2}\right)\right\rangle /, z^{q}=[x, y] \tag{2}
\end{equation*}
$$

and hence the acceptance reqion. $B_{o}$ for $F_{1}$ is given by

$$
\begin{equation*}
B_{0}^{n}=\left\{(x, y): a x+b y<\left(a^{2}+b^{2}\right) / 2\right\} . \tag{3}
\end{equation*}
$$

If $\mathrm{P}(\mathrm{J} \mid \mathrm{x})$ denotes the probability of wisclasaifying an element from $\Pi_{i}$ fiato $\Pi_{j}$, then for this procedure ve lieve

$$
\begin{align*}
& \text { edure ve lave } \\
& r(2 \mid 1)=r(1 \mid 2)=6\left(-\sqrt{a^{2}+b^{2}} / 2\right),
\end{align*}
$$

where

$$
q(1)=-\int_{-\infty}^{t} \exp \left(-x^{2} / 2\right) d x / \sqrt{2}:
$$

If $a=1$ and $b=2$, then we have fron (4)

$$
P(2 \mid 1)=P(1 \mid 2)=(-\sqrt{5} / 2)=0.13
$$

3. Cunster gased Clamifintion

It is well bown (Lerean, 1.970) that as the sample sizes increase, the clugters $C_{1 i}$ and $C_{2 i}$ based on $\mathcal{R}_{i}$-noma coiverges with probsbility ane to the partition ( $s_{1 \mathrm{i}}, \mathrm{S}_{21}$ ) of $\mathrm{R}_{2}$, the two dimensional real vector space, where

$$
\begin{equation*}
s_{2 i}^{c}=s_{1 i}=\left(x=(x, y): \quad\|z\|_{i}<\| \|_{2}-2 \|_{i}\right) \tag{5}
\end{equation*}
$$

$A^{c}$ denotes the complonent of the set $A$ and $\|u-v\|_{i}$ denotes the distance between the points u and $v$ of $K_{2}$ accordina to $s_{i}$ nomm. $\hat{H}_{j i}$ and $\dot{\Sigma}_{j i}$, the mean and dispersion matrix of the sathle $C_{j i}$, $(j=1,2)$, can be viened as the sample mean and sample dispersion natrix of a population with density function $p_{j i}(i)$ given by

$$
\begin{align*}
p_{j i}(z) & \left.=(1 / 2) \varkappa_{2}(0, I)+N_{2}\left(v_{2}, I\right)\right\}, z^{\in} S_{j i}(6)  \tag{6}\\
& =0 \text { othervise. }
\end{align*}
$$

Thus, by the law of large numbers, with probabilIty one $\hat{\beta}_{j i}$ converges to $\mu_{j i}$ and $\vec{E}_{j i}$ to $\Sigma_{j i}$, where

$$
\begin{align*}
& u_{j i}=S_{j i} 2 p_{j i}(z) d z / p\left(S_{j i}\right)  \tag{7}\\
& \Sigma_{j i}=\int_{j i}\left(z-i j_{j i}\right)(z-\mu j i)^{T} p_{j i}(z) d z / T\left(S_{j i}\right) \tag{S}
\end{align*}
$$

and $p^{p}\left(S_{j i}\right)={\underset{S}{j i}}^{r_{j i}}(z) d z$.
In desigiing cluster based classifiers, as we have montioned earlier, we shall consider iwo cases-when we do not knew that the dispersion riztrices of the population $\Pi_{1}$ and $\Pi_{2}$ are equal and when we know that they are equal. In the first case, we sinall denote the cluater based procedire by $P_{j}$ (based on $\ell_{i}$-noms) and in the second case ty ' $P_{1}$ '. Thus $P_{1}^{\prime}$ is the Fayes' procedure based an the denslifes $N_{\lambda}\left(\mu_{j 1}, \Gamma_{i}\right)(j=1,2)$, where

$$
\begin{equation*}
\Sigma_{i}=P\left(S_{j i}\right) \Sigma_{1 i}+P\left(S_{2 i}\right) \Sigma_{2 i} \tag{10}
\end{equation*}
$$

Misclaseification lrenabiluty
If a claselfiration procedure derines a set $B$ as, the acceptanct rigion for the poundation ill wlth true density $\because \because(O, I)$ and $S^{c}$ as the accentance

 are etven $b ;$

$$
\begin{equation*}
1(2 \mid 1)=P\left(11^{c} \mid i .1\right) \cdot{ }_{B^{c}}^{t} c_{2}(0,1) d: x \tag{11}
\end{equation*}
$$


fof to the rayes' classification procedure in
which lie population densities for $H_{1}$ and $H_{2}$ have been Inserertentiy assumed to be $N(m, V)$ and $N$ ( $x^{\prime}$, 'r' $^{\prime}$ Iesiectively. Then

$$
\begin{equation*}
B=\{Z=(x, y): W(x, y) \quad 0\} \tag{12}
\end{equation*}
$$

were

$$
H(x, y)=2\left(\varepsilon_{1} x+g_{2} y\right)+k+2 e x y-d_{1} x^{2}-d_{2} y^{2}
$$

$$
\mathrm{d}_{1}=\mathrm{v}_{22} /\left|\mathrm{v}-1-\mathrm{v}_{22} /|\mathrm{v}|\right.
$$

$$
c_{2}=v_{1} /\left|v^{\prime}\right|-v_{1 l} /|v|
$$

$$
e=v_{12} /\left|v^{*}\right|-v_{12} /|v|
$$

$$
E!=\left(v_{2}^{2} m_{1}-v_{1} m_{2}\right) /\left|v^{\prime}\right|-\left(v_{22^{m}}-v_{12} m_{2}\right) /|v|
$$

$$
E=\Sigma^{T} V^{-1}-m^{-T} V^{\prime-1}-10 g(|V| /|V|
$$

Here $V=\left[v_{i j}\right], \nabla^{*}=\left[v_{i j}\right], m^{T}=\left[m_{1}, m_{2}\right]$ and

$$
n^{-\bar{T}}=\left[\Psi_{1}^{\prime}, m_{2}^{\prime}\right]
$$

The risclassification probabilities $P(2 \mid 1)$ and $P(1!2)$ can now be expressed in the form

$$
\begin{equation*}
P(2 \mid 1)=P\left(h^{\prime}(X, Y) \geq 0 \mid Z \in \pi_{3}\right) \tag{14}
\end{equation*}
$$

and $P(1 \mid 2)=P\left(U(X, Y)<0 \mid Z \varepsilon \Pi_{2}\right)$
If $\mathrm{Vf} \mathrm{Y}^{\prime}$, then the exact distribution of $\mathrm{K}(\mathrm{X}, \mathrm{Y})$ when $Z \in \Pi_{1}$ or $Z \in \Pi_{2}$ is intractable. But. if $d_{j}, d_{z}$ and e are nogligible compared to $g l$ and E2, then in (14) and (15), the contribution frou the seconj degree tems in $X$ and $Y$ will be negligible compared to that from $K+2\left(\varepsilon_{1} X \not g_{2}{ }^{\prime}\right)$. Thus we can periaps approxisate the distribution of $W(X, i) b$ tha distribution of $\hat{H}(X, Y)=\mathrm{E}+$
$2\left(g_{1} X+\varepsilon_{2} \dot{\prime}\right)$. Fut ff $V=V^{\prime}$, then $N=\hat{W}$. Now $\hat{W}$ can be shon to be distributed normally with wean

$$
\hat{E i n}=\left\{, \text { when } 2 \sim_{2}(0, I)\right.
$$

ard $\hat{E n}=\mathrm{F}+2\left(81 a^{+} 8, b^{2}\right)$,

$$
\begin{equation*}
\text { when } 2 \operatorname{Hin}_{2}\left(11_{2}, T\right), H_{2}^{T}=[a, b] \tag{16}
\end{equation*}
$$

and viriance

$$
\begin{equation*}
\operatorname{Var} \hat{\mathrm{H}}=4\left(\varepsilon_{1}{ }^{2}+\varepsilon_{2}{ }^{2}\right) \tag{17}
\end{equation*}
$$

In eititer cise. The approximate values of the Efsciassification probatilities $P(2 \mid I)$ and $P(1 \mid 2)$

$$
\begin{align*}
& \text { can be given by } \\
& \quad P(2!1) \doteq P\left(h \circ 0 / 2 \in \pi_{1}\right)=1-\left(-K / 2 \sqrt{1_{1}^{2}+E 2^{2}}\right) \tag{18}
\end{align*}
$$


4. Cuacitiratinalrocusure $P_{2}$ and $P_{2}$

Thee $\mathcal{L}_{2}$-norm fnelued distance lictween two
 $\left.\|u-\|_{2}=\left(u_{1}-v_{1}\right)^{2}+\left(u_{2}-v_{2}\right)^{2}\right)^{\lambda / 2}$.


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(8)
$\ldots \ell_{1}$
$\ldots-\ell_{2}$
$\ldots-l_{-\infty}$.
Limpline clusters using $1_{1}-, 1_{2}-$ and $1_{m}$-norm
the liafting, clusters $S_{12}$ and $S_{22}$ aree given by $S_{22}{ }^{c}=S_{12}=\left\{(x, y) ; \quad\right.$ axtby $\left.<\left(a^{2}+h^{2}\right) / 2\right\}$. Evidently $\mathrm{S}_{12}=\mathrm{B}_{\mathrm{o}}$, as given by (3). Now it follows from (6)-(9) that

$$
\begin{align*}
& P\left(S_{12}\right)=P\left(S_{22}\right)=1 / 2,  \tag{20}\\
& \mu_{12}=k_{1}\left[{ }_{b}^{a}\right] \text { and } \\
& \Sigma_{12}=\left[\begin{array}{ll}
1+a^{2} k_{i}\left(1-k_{j}\right) & a b k_{i}\left(1-k_{1}\right) \\
a b k_{i}\left(1-k_{1}\right) & 1+b^{2} k_{i}\left(1-k_{i}\right)
\end{array}\right], \tag{21}
\end{align*}
$$

( $1=1,2$ ), where

$$
\begin{align*}
(1=1,2), & \text { where } \\
k_{1}= & 8\left(-a^{2}+b^{2} / 2\right)-\left(2 n\left(a^{2}+b^{2}\right) / 2\right)^{-1}  \tag{22}\\
& \exp \left\{-\left(a^{2}+b^{2}\right) / 8\right\} \\
\text { and } k_{2}= & \Leftrightarrow\left(a^{2}+b^{2} / 2\right)+\left(2 \pi\left(a^{2}+b^{2}\right) / 2\right)^{-1}
\end{align*}
$$

$$
\exp \left\{-\left(a^{2}+b^{2}\right) / 8\right\}
$$

The set $\mathrm{B}_{2}$, the acceptance region of $\Pi_{1}$ according to $P_{2}$, can be obtained from (12) by letting $n=\mu_{12}, m^{\prime}=H_{22}, v=\Sigma_{12}$ and $v^{\prime}=\Sigma_{22}$ and the corresponding misclassification probabilities from (13), (18) and (19).

An Example
Let $a=1$ and $b=2$. Then
$S_{22}{ }^{c}=S_{12}=\{(x, y): \quad x+2 y \leq 2.5\}$
[figure (3)]
and it follows from (21), (22) and (23) that

$$
\begin{aligned}
& k_{1}=-0.06, k_{2}=1.06, v_{i 2}^{T}=k_{i}[1,2], \\
& \Sigma_{12}=\left[\begin{array}{cc}
0.94 & -0.1] \\
-0.11 & 0.78
\end{array}\right] \text { and } \Sigma_{22}=\left[\begin{array}{cc}
0.93 & -0.12 \\
-0.12 & 0.75
\end{array}\right] .
\end{aligned}
$$

In $\mathrm{P}_{2}$, for discrimineting observations from $\mathrm{I}_{1}$ and $\dot{r}_{2}$, the bayes' clessificr based on deasities $\mathrm{N}_{2}\left(\mu_{12}, \Sigma_{12}\right)$ and $\mu_{2}\left(\mu_{22}, \Sigma_{22}\right)$ is used instead of that based on the true densities $\mathrm{N}_{2}(0, I)$ and $\mathrm{N}_{2}\left([1,2\}^{\mathrm{T}}, \mathrm{I}\right)$, with are unknown. Then the approximate values of the resulting njeclessification probabilicies are obtained from (1s) and (19) as

$$
\mathrm{P}(1 \mid 2)=0.22 \text { and } \mathrm{P}(2 \mid 1)=0.078
$$

In procedure $P_{2}^{\prime}$, then Dayes' classifict is based on densitior. $\mathrm{H}_{2}\left(H_{12}, 5\right)$ and $\mathrm{N}_{2}\left(\mu_{22}, 5\right)$, where

$$
\Sigma=P\left(s_{12}\right) \varepsilon_{12}+P\left(s_{22}\right) \varepsilon_{22}=\left[\begin{array}{cc}
0.935 & -0.315 \\
-0.115 & 0.765
\end{array}\right] .
$$

The ofsclassification prombilitles are now obtained from (15) and (19) as

$$
P(1 \mid 2)=0.18 \text { and } P(2 \mid 1)=0.09
$$


The $8_{1}$-nam fubled distance betweon the points $"$ and $v$ is given ly

$$
\left\|w_{1}-v\right\|_{1}=\left|w_{1}-v_{1}\right|+\left|w_{i}-v_{2}\right| .
$$


points ( $x, y$ ) waltefying the following, conditions in three different cases.

Case 1. b>aro lfigure (2) and (3))
(i) $x \geqslant a, y<(b-a) / 2$
(ii) $0<x<a, x+y<(a+b) / 2$
(iii) $x<0, y<(a+b) / 2$.

In this case $S_{11}$ and $S_{21}$ are distinct from $S_{12}$ and $S_{22}$ given by (3).

Case 2. $b>0, a=0 \quad\lceil$ figure (1)]
(1) $\mathrm{y}<\mathrm{b} / 2$.

In this case $S_{11}$ colncides with $S_{12}$ and $S_{21}$ coincides with $\mathrm{S}_{22}$.

Case 3. $b=a>0$ [figure (4)]
(i) $x+y<a$

The $l_{1}$-noma distances of the point ( $x, y$ ) from $(0,0)$ and $(a, b)$ are equal when
$x+y \geq a$ and $x<0$
or $x+y>a$ and $y<0$.
These points lave been assigned to $5_{21}$ in order that no point escape clessification. In this case also the cluster $S_{11}$ coincides with $S_{12}$ and $S_{21}$ coinciues with $S_{22}$.

Thus the procedure $P_{1}\left(P_{i}\right)$ differs from the procedure $\mathrm{P}_{2}\left(\mathrm{P}_{2}\right)$ only in case 1 . In this case, the expression for the components of the means $u_{11}$ and $y_{21}$ and the dispersion matrices $\Sigma_{21}$ and $\Sigma_{21}$ are long and complicated erpressions involving the probabilitics o( $a$ ), $\hat{y}[(a+b) / 2]$ and $\hat{\Delta}[(b-a) / 2]$. For that reason ve have not included it here. As In the exasple in section 4 , if we take $a=1$ and

$$
\begin{aligned}
b=2, & \text { then we sha11 have } \\
\mu_{11} & =\left[\begin{array}{c}
-0.005 \\
0.197
\end{array}\right], \quad \nu_{21}=\left[\begin{array}{l}
0.423 \\
2.11
\end{array}\right] \\
\Sigma_{11} & =\left[\begin{array}{cc}
0.98 & -0.06 \\
-0.05 & 0.66
\end{array}\right] \text { and } \Sigma_{21}=\left[\begin{array}{ll}
1.82 & -.2 \\
-.2 & 0.8
\end{array}\right]
\end{aligned}
$$

From (13) it noy follows that

$$
\begin{aligned}
U(X, Y)= & 0.46 X^{2}+0.24 Y^{2}+2(0.52 X+2.46 Y) \\
& +0.06 X Y-9.36
\end{aligned}
$$

The exact distribution of $W$ is intractable. The nomal approximation to $W$ eiven in section 3 is not valld, since the coefficients 0.46 of $x^{2}$ and 0.24 of $\dot{r}^{2}$ are not nefligible compared to coefficlence 0.52 of $X$ and 2.16 of X . Lishig a very rough approxination, it can be shom that in this case $P(2 \mid 1)$ and $P(J \mid 2)$ are larger then those for $\mathrm{P}_{2}$.

For procelure P'í ve have
$\Sigma_{1}=1 \cdot\left(S_{11}\right) \Sigma_{11}+\Gamma\left(S_{21}\right) \Sigma_{21}$
$-0.455 \Sigma_{11}+0.505 \Sigma_{21}$

$$
\cdots=\left[\begin{array}{ll}
1.4 & -0.13 \\
-0.13 & 0.73
\end{array}\right]
$$

Fron (13) we now have

$$
W(X, Y)=1.11 S X+5.44 Y-6.541
$$

and from (18) and (19) we have
$P(2 \mid 1)=0.18$ and $P(1 \mid 2)=0.12$
6. Classfication Proceducc $P_{\infty}$ and $P_{\infty}^{*}$

The $5_{\infty}$-norm indiecd distance betwen the points $u$ and $v$ is given by
$\|u-v\|_{\omega}=\max \left\{\left|u_{1}-v_{1}\right|,\left|u_{2}-v_{2}\right|\right\}$.
The points ( $x, y$ ) in the linitine cluster
$S_{l_{\infty}}\left(=S_{2_{\alpha}}\right)$ satisfy tle folloning conditions.
Case 1. b>a>0 [figure (2) and (3)]

> (i) $a-b / 2<x<\max (b / 2, a)$ and $y<b / 2$,
> (ii) $x<a-b / 2$ and $x+y<a$
> (iii) $x>\max (b / 2, a)$ and $x+y<b$.

Case 2. $b=a s 0$ [figure (4)]
(i) $\mathrm{sity}<\mathrm{a}$

The $\ell_{\infty}$ - distances of a point $(x, y)$ from ( 0,0 ) and ( $a, b$ ) are equal when
$x+y>a$ and $x<0$
or $x+y>a$ and $y<0$.
Such points have been assigned to $S_{2_{\infty}}$ arbitrarily. Thus in this case $S_{j_{\infty}}$ and $S_{2_{\omega}}$ coincide with $S_{12}$ and $\mathrm{S}_{22}$ respectively.

Case 3. $b>0, a=0$ [figure (1)]
In this case, the performince of $s_{\infty}$-nerm in clusterfing is very poor. The $s_{\infty}$-distances of every point ( $x, y$ ) fron ( 0,0 ) and ( $0, b$ ) are cgual wienever

$$
x-y+2<0 \text { and } x+y<0
$$

or, $x+y-2>0$ and $x-y>0$
Horicver, he can arbitrarily define $\mathrm{S}_{\mathrm{I}_{\mathrm{D}}}$ as the set
$\{(x, y): y<b / 2\}$.
But $S_{1_{\infty}}$ then becones identical to $S_{11}$ or $S_{12}$.
Thus the procedure $\mathrm{P}_{6}\left(\mathrm{I}_{\infty}{ }_{\infty}\right)$ differs from $\mathrm{P}_{2}\left(\begin{array}{l}\mathrm{O}\end{array}\right)$ only in case 1. Sut in this case, the ckact nisclassifjeation probibility is hard to fand. In the spectal case, wh:c:a $a=1$ and $b=2$, (figure (3)) we find that the bomtary line of $S_{1_{\omega}}$ and $S_{2_{\infty}}$ can be obtafned as the reflection of the butudary line of $s_{11}$ and $S_{21}$ about the lioutary dine of $S_{12}$ and $S_{22}$. Frea than we my guces that, as fa cate of the pooctare $l_{1}$ and yí, the fotal Miselanciffeat
 than the (win) mit:clatsiffeathon frotabilition of $I_{2}$ and $1 ; i$.

## 7. Conclusion

Our flisines about the misclassification prebabilitics $P(1 \mid 2)$ and $P(2 \mid 1)$ for the procedures $P_{0}, P_{1}, r_{1}, P_{2}$ and $P_{2}$ in the siccial care then $a=1$ and $b=2$ can be seen at a elance from the following table.

Table
True Baycs

|  | $P_{0}$ | $P_{1}$ | $P_{1}$ | $P_{2}$ | $P_{2}$ |
| :--- | :--- | :--- | :--- | :--- | :--- |
| $P(1 \mid 2)$ | 0.13 |  | 0.12 | 0.22 | 0.18 |
| $P(2 \mid 1)$ | 0.13 |  | 0.18 | 0.078 | 0.09 |
| Total | 0.26 | $*$ | 0.30 | 0.299 | 0.27 | *Using rough approximation it has been found that the total misclassification probability of $P_{1}$ is greater than 0.229.

Thus, taking misclassification probebility as the criterion for comparing the performance of the classiffers, we find that in the example considered above $P_{2}$ is better than $P_{i}$ and $P_{2}$ is better than $P_{1}$. But, no such conclusion can be made in generaI. It is not known whether $P_{i}$ always has 10:ier miselassification probability than its competitor $P_{1}$ or $P_{\infty}^{\prime}$ and $P_{2}$ has lover misclassification probability than $P_{1}$ or $P_{\infty}$.

Among all nom-induced distances, $\ell_{1}$-distances are tide most easy to compute. When population means are known to be such that the straight line joinfug thea is inclined at an angle of $45^{\circ}$ to r -axis (ior exariple, when they are $(0,0)$ and ( $a, 2$ ), the classification procedure $P_{1}\left(P_{1}\right)$ and $P_{2}\left(r_{2}^{\prime}\right)$ becone dentical. Only in that case, $P_{1}$ or $\dot{P}_{1}$ is the optemi procedure, because we then acticive the computational case along with low misclassification probablity.

The procedures $P_{\infty}$ and $P_{\omega}^{*}$ are not desirable. If in the prucess of clusterine, two cluster centers are such that they have the same $x$-coordinate, then scme points are to be clustered according to $\mathcal{L}_{2}$-distance. This arbitrardess in allucation of points to the clusters mates $P_{\text {o }}$ or $\Gamma_{\omega}^{\circ}$ undesirabic.

Feference

1. Lermen, I. C. (1970). Ies Bases de 1a Clarci-


# ADAPTIVE PATTERN RECOGNITION: A SURVEY <br> J. P. Basu and P. L. Odell 

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## 1. Introduction

A classifier or discriminant function $W(x j c)$ is a function or a set of functions which determine the membership of anservation vector $x$ into one of several given sub-populations, $c$ denoting the vector of parameters of the classifier. The classifier usually considered is one which has minimum expected cost of misclassification among all classifiers. This optimal classifier is known as Bayes' classifier [2]. This classifier is obtained on the assumption of complete knowledge of misclassification costs, prior probabilities (the proportion of the given subpopulations in the over all population) and the probability densities of the given subpopulations. When the parametric family of the population densities are known, but the parameter themselves are unknown, then the usual practice is to replace the true values of these parameters by their respective estimates based on sets of past observations of known classification in the expressions giving the elements of $c$. A classifier $W(x, \hat{c})$, thus obtained, lacks in the above optimality property. A set of past observations known to be coming from a particular population is often referred to as a training sample. When the parametric family of the population densities are also unknown and sometimes the prior probabilities are also unknown, the methods by which a classifier $W(x, c)$ is obtained from the available training samples of the populations are referred to as nonparametric methods of classifier design. An adaptive pattern recognition scheme is one such nonparametric method.

An adaptive pattern recognition scheme is a nonparametric method in which a classifier is assumed to be of the form

$$
\begin{equation*}
W(x, c)=c_{1} Q_{1}(x)+\ldots+c_{m} Q_{m}(x)=c^{\top} \Phi(x), \tag{1}
\end{equation*}
$$

where the components $Q_{1}(x), \ldots, Q_{m}(x)$ of the vector $\Phi(x)$ are prechosen linearly independent continuous functions of the observation vector $x$ and the unknown parameters $\left[c, \ldots, c_{m}\right]=c^{\top}$ are directly estimated recursively as in [1, 7] or nonrecursively as in [6, 8] from the available training samples, instead of being obtained terms of the estimates of the probability densities. A classifier obtained in this fashion will be referred to as an adaptive classifier.

The functional forms of the functions $Q_{i}(x)$ 's and their number $m$ are often choosen arbitrarily. This arbitrariness in the choice of the functions $Q_{i}(x)$ 's may create problem. A good classifier should be optimal for discriminating among the given populations. For example, it is well known [2] that when there are two p-dimensional normal populations with different covariance matrices, a linear classifier of the form

$$
\begin{equation*}
W(x, c)=c_{1} x_{1}+\ldots+c_{p} x_{p}+c_{0}=c^{\top} z \tag{2}
\end{equation*}
$$

where $c^{\top}=\left[c_{1}, \ldots, c_{p}, c_{0}\right], x^{\top}=\left[x_{1}, \ldots, x_{p}\right]$ and $z^{\top}=\left[x^{\top}, 1\right]$, is not optimal for discriminating between these two populations. Therefore, if the functional forms of $Q_{i}(x)$ 's are not properly chosen, then the classifier $W(x, c)$ given by (1) may not be optimal for discriminating among the given populations. Also in designing an adaptive classifier rarely any attention is paid to the optimality criterion, minimum expected cost of misclassification. However, adaptive classifiers are easy to design and are very attractive in fast classification for their simplicity. Besides, there are empirical cases in which they do no worse, if not better, than many other sample based classifiers [2].

A comprehensive survey of adaptive pattern recognition schemes can be found in $H_{0}$ and Agrawala [5] and in Tsypkin [9]. Our objective here is to present a unified theory of adaptive pattern recognition. Most of the adaptive pattern recognition schemes have been devised for discriminating between two populations and few of these schemes have been generalized in a very straight forward manner for discriminating among more than two populations. For this reason, we shall also restrict ourselves mostly to the case of two population classification.

## 2. Basic Theory

Let all the observations come from two $p$-dimensional population $\pi_{1}$ and $\pi_{2}$. Then the observations can be represented by points and the two training samples by two sets $T_{1}$ and $T_{2}$ of points in the p-dimensional real Euclidean space $E_{p}$. It will be evident later from the way an adaptive classifier is designed that the sets $T_{1}$ and $T_{2}$ should be disjoint. A real function $f(x)$, $f: E_{p} \rightarrow E_{1}$, is said to separate two disjoint sets $T_{1}$ and $T_{2}$ in $E_{p}$, if there exists a constant $t$ such that $f(x)>t$ for all $x \varepsilon T_{1}$ and $f(x)<t$ for all $x \varepsilon T_{2}$. If there exists a function $f(x)$ separating $T_{1}$ and $T_{2}$, then $f(x)$ can be considered to be an ideal classifier which can be used to classify all of the points of $T_{1}$ and $T_{2}$ correctly. It may not often be possible to construct such $f(x)$, even if it exists. But, it may be possible to obtain a function $W(x, c)$ as an approximation to the function $f(x)$ and use it as a classifier instead of the ideal classifier $f(x)$. This is precisely the way an adaptive classifier is designed.

The existence of function $f(x)$ separating the training samples $T_{1}$ and $T_{2}$ will be evident soon. The following theorem is well known (Dunford and Schwartz, [4], p. 417).

Theorem. In a topological vector space, any two disjoint convex sets, one of which has an interior point, can be separated by a nonzero continuous linear functional.

If $T_{1}$ and $T_{2}$ have disjoint convex hulls $C_{1}$ and $C_{2}$, then we can use the above theorem to prove the existence of function $f(x)$ of the form

$$
f(x)=c_{1} x_{1}+\ldots+c_{p} x_{p}+c_{0}
$$

such that $f(x)>0$ on $C_{1}$ and $f(x)<0$ on $C_{2}$. In this case the training samples $T_{1}$ and $T_{2}$ are said to be linearly separable. Therefore, when the training samples are linearly separable, then a linear function $f(x)$ separating $T_{1}$ and $T_{2}$ can be used as an ideal classifier. An adaptive classifier is then easily obtained as a linear function approximating this ideal classifier.

Unfortunately, the sets $T_{1}$ and $T_{2}$ of training samples are rarely linearly separable. Therefore, often there may not exist any linear function separating $T_{1}$ and $T_{2}$. However, there exists a function $f(x)$, not necessarily linear, which will separate $T_{1}$ and $T_{2}$ whenever the sets $C_{1}$ and $C_{2}$, the smallest closed sets containing $T_{1}$ and $T_{2}$ respectively, satisfy the conditions of Uryshon's theorem, well known in Topology ([4], p. 24).

Uryshon's Theorem. If $C_{1}$ and $C_{2}$ are disjoint subsets of a normal topological space $V$, then there exists a continuous function $g(x), g: V \rightarrow[0,1]$, such that $g(x)=0$ for all $x \& C_{2}$ and $g(x)=1$ on $C_{1}$.

Using the function $g(x)$ of Uryshon's theorem, we can easily define a continuous function $f(x)$ on $V$ such that $(a) f(x)=t_{1}$ on $C_{1}$ and $f(x)=t_{2}$ on $C_{2}$, where $t_{1}$ and $t_{2}$ are any two distinct preassigned numbers, or, (b) $f(x)>\delta$ on $c_{1}$ and $f(x)<-\delta$ on $C_{2}$, where $\delta$ is a preassigned positive number. Now, since
it is well known that $E_{p}$ is a normal topological space under any nontrivial topology and $T_{1}$ and $T_{2}$ are disjoint sets containing finite number of points, then we can always obtain two disjoint sets $C_{1}$ and $C_{2}$ as the smallest closed sets containing $T_{1}$ and $T_{2}$. Therefore, as long as $T_{1}$ and $T_{2}$ are disjoint sets of training samples, there always exists a function $f(x)$, linear or nonlinear, which separates $T_{1}$ and $T_{2}$.

As we have mentioned earlier, in an adaptive pattern recognition procedure the ideal classifier $f(x)$ is approximated by a function $W(x, c)$ of the form (2), namely, a linear function of an observation $x$ to obtain a IInear adaptive classifier, or, by a function $W(x, c)$ of the form (1), a linear combination of finite number of known nonlinear continuous functions, to obtain a nonlinear adaptive classifier, the parameter vector $c$ being selected in order to minimize certain given pay-off function.

As the probability densities of the populations are unknown, expected misclassification cost cannot be evaluated and therefore cannot be used as a pay-off function. Distances $D(f, W)$ between the functions $f(x)$ and $W(x, c)$ or some functions of it defined by some meaningful metric can be used as geometrically and intuitively appealing pay-off functions. For example, using Euclidean metric we can define

$$
\begin{equation*}
D_{2}(f, W)=\left.\right|_{x \in T_{1} U T_{2}} f(x)-\left.W(x, c)\right|^{2} \tag{3}
\end{equation*}
$$

as a pay-off function and select the parameter vector $c$ in order to minimize $D_{2}(f, W)$. In this case $W(x, c)$ is a least square approximation to $f(x)$. In order to assure the existence of a unique minimum of the pay-off function, a pay-off function of the form

$$
\begin{equation*}
L(f, W)=\underset{x \in T_{1} U T_{2}}{F(f(x)-W(x, c)) / n} \tag{4}
\end{equation*}
$$

where $n$ is the total number of points in $T_{1}$ and $T_{2}$ and in the fashion of decision theory $F$ is a convex function interpreted as a loss function, can also be used [9]. We may note that the pay-off function $\bar{D}_{2}(f, W)=$ $D_{2}(f, W) / n$, often referred to as mean square error criterion $[6,8]$, is a particular case of (4).

Let the subpopulation $\Pi_{i}$ have prior probability $q_{i}$ and probability density function $p_{i}(x)$. Then the population which is a mixture of $\Pi_{1}$ and $I_{2}$ has probability density function $q_{1} p_{1}(x)+q_{2} p_{2}(x)$. Therefore, the expected value of $F(f(x)-W(X, c))$ will be given by

$$
\begin{align*}
R(f, W) & =E F(f(x)-W(X, c)) \\
& =\sum_{i=1} \quad q_{i} \int F(f(x)-W(x, c)) p_{i}(x) d x
\end{align*}
$$

Now, if there are $n_{i}$ observations in $T_{i}$ and $n_{1}+n_{2}=n$, then we can write 2
and observe that $L(f, W)$ is the sample mean and $R(f, W)$ is the population mean of $F(f(x)-W(X, c))$.

It is important to note that there is some arbitrariness in the choice of the number $m$ and the continuous functions $Q_{1}(x), \ldots, Q_{m}(x)$, used in the definition of $W(x, c)$.

## 3. Linear Classifier

The linear classifiers are used in adaptive classification mainly due to following reasons.

1. They are relatively simple to implement as electronic circuits.
2. The circuits can be made adaptive very easily.
3. The Bayes' classifier with minimum expected cost of misclassification has a linear structure when the two populations have normal distribution with equal covariance matrices.

### 3.1. Nilsson's Classifier

The algorithm proposed by Nilsson [7] iteratively determines the parameter vector $c$ of $W(x, c)$, as defined in (2), according to the following rule. Let $x(i)$ denote the $i t h$ observation of known classification, that is, an element selected from $T_{1} U T_{2}$ at the ith step of an iteration, $z(i)$ the corresponding value of $z, c(i, n)$ that of $c$ at the ith step of the nth iteration and $W_{i n}=c^{\top}(i, n) z(i)$. Then at the $(i+1)$ th step of the nth iteration we have

$$
\begin{array}{rlrl}
c(i+1, n) & =c(i, n) & & i f x(i) \varepsilon T_{1} \text { and } W_{i n}>0 \\
& =c(i, n) & & \text { if } x(i) \varepsilon T_{2} \text { and } W_{i n}<0 \\
& =c(i, n)+d z(i) & i f x(i) \varepsilon T_{1} \text { and } W_{i n} \leq 0 \\
& =c(i, n)-d z(i) & \text { if } x(i) \varepsilon T_{2} \text { and } W_{i n} \geq 0,
\end{array}
$$

where $d$ is a positive number, so chosen that $c^{\top}(i+1, n) z(i)$ can correctly classify $x(i)$ for all $i$. The vector $c(1, n+1)$ is taken as the value of $c$ at the end of nth iteration. The process is continued until we have, for some $i, c(i, n)=c(j, n)=c(\ell, n+1)=\hat{c}$ (say) for all $j>i$ and $\ell<i$. When such $\hat{c}$ exists, we shall say that $c(i, n)$ has converged to $\hat{c}$.

Unfortunately, if the sets $T_{1}$ and $T_{2}$ of training samples are not linearly separable, then the sequence $c(i, n)$ may not converge. The algorithm is effective, that is, $c(i, n)$ converges to $\hat{c} \hat{c}$, only when $T_{1}$ and $T_{2}$ are linearly separable. It should be noted that $\hat{c}$ is not unique; it depends on the sequence in which we select the observations from $\mathrm{T}_{1} \mathrm{UT}_{2}$. The number of iterations needed to reach such $\hat{c}$ also depends on this sequence. The classifier $W(x, \hat{c})$ constructed in this fashion is not in general a Bayes classifier.

Example. Let $T_{1}=\left\{[0,1]^{\top},[1,1]^{\top}\right\}, T_{2}=\left\{[0,0]^{\top},[1,0]^{\top}\right\}$. Then choosing $d=1$ and the sequence of observations as in the Table we can obtain Nilsson's classifier in the following way.

| $\square$ | i | c |  |  | z |  |  | $W_{i}$ | Change c? |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
|  |  | $c_{1}$ | $c_{2}$ |  | ${ }_{1}$ | ${ }^{2}$ | 1 |  |  |
| 1 | 1 | 0 | 0 | 0 | 1 | 1 | 1 | 0 | Yes |
|  | 2 | 1 | 1 | 1 | 1 | 0 | 1 | 2 | Yes |
|  | 3 | 0 | 1 | 0 | 0 | 1 | 1 | 1 | No |
|  |  | 0 | 1 | 0 | 0 | 0 | 1 | 0 | Yes |
| 2. | 1 | 0 | 1 | -1 | 1 | 1 | 1 | 0 | Yes |
|  | 2 |  | 2 | 0 | 1 | 0 | 1 | 1 | Yes |
|  | 3 | 0 | 2 | -1 | 0 | 1 | 1 | 1 | No |
|  | 4 | 0 | 2 | -1 | 0 | 0 | 1 | $-1$ | No |
|  | 1 | 0 | 2 | $-1$ | 1 | 1 | 1 | 1 | No |
| 3 | 2 | 0 | 2 | -1 | 1 | 0 | 1 | -1 | No |

Therefore $\hat{c}^{\top}=[0,2,-1]$ and the classifier is $2 x_{2}-1$. A decision rule based on this classifier is to assign $x$ to $\Pi_{1}$ if $2 x_{2}-1>0$ and to $\Pi_{2}$ if $2 x_{2}-1<0$.

$$
\begin{align*}
& \ell  \tag{9}\\
& \qquad=d \hat{\Sigma}^{-1}\left(m_{2}-m_{1}\right) \text { and } c_{0}=-\left\{\ell\left(N_{1} m_{1}+N_{2} m_{2}\right)-\left(N_{1}-N_{2}\right)\right\} /\left(N_{1}+N_{2}\right),  \tag{10}\\
& \text { where } \quad d=2 /\left[\left(N_{1}+N_{2}\right)^{3} /\left\{N_{1} N_{2}\left(N_{1}+N_{2}-2\right)\right\}+\left(m_{1}-m_{2}\right)^{T^{-}-1}\left(m_{1}-m_{2}\right)\right]  \tag{II}\\
& \text { and } \quad \hat{\Sigma}=\left(N_{1} S_{1}+N_{2} S_{2}\right) /\left(N_{1}+N_{2}-2\right) .
\end{align*}
$$

When

$$
N_{1}=N_{2} \text {, then we have }
$$

$$
\begin{equation*}
c_{0}=\left(m_{1}-m_{2}\right)^{T} \hat{\Sigma}^{-1}\left(m_{1}+m_{2}\right)[N d / 4(n-1)] \tag{12}
\end{equation*}
$$

The classifier $W(x, c)=c^{\top} z$ in this case resembles the Bayes' classifier with parameters of the population densities replaced by their respective estimates, that is, Anderson's [ 2 ] sample based discriminant function for two normal population with equal covariance matrices.

The classifier attributed to Agmon and Mays [5] is obtained when

$$
\begin{aligned}
L(f, W) & =\sum[|W(x, c)-f(x)|-\{W(x, c)-f(x)\}]^{2} \\
& =x \in T_{1} U T_{2}
\end{aligned}
$$

is used as the pay-off function.
4. Nonlinear Classifier

The nonlinear classifiers are used in adaptive classification mainly due to following reasons.

1. If the sets $T_{1}$ and $T_{2}$ of training samples are not linearly separable, then a function $f(x)$ that can be used to separate them are often nonlinear.
2. The Bayes classifier has nonlinear structure when the two populations have (a) multivariate normal distributions with unequal covariance matrices, or (b) nonnormal multivariate distributions.

### 3.2 Other Linear Classifiers

An adaptive linear classifier can be obtained as linear approximation to a function $f(x)$ separating $T_{1}$ and $T_{2}$ and taking the value 1 on $T_{1}$ and -1 on $T_{2}$, the parameter vector $c$ being selected as a vector $\hat{c}$ that minimize a given pay-off function. Different pay-off functions can be used to obtain different classifiers.

## Koford-Groner Classifier

An adaptive linear classifier $W(x, c)$ may be obtained by selecting a vector $\hat{c}$ as the parameter vector that minimize the mean square error criterion $\bar{D}_{2}(f, W)$, or equivalently the squared Euclidean distance $D_{2}(f, W)$ given by (3), using gradient method or any other suitable optimization technique. An iterative algorithm based on gradient method for obtaining $\hat{c}$ can be given by

$$
\begin{align*}
c(n+1)=c(n)+c z \Sigma & \left\{f(x)-c^{\top} z\right\}  \tag{7}\\
& x \varepsilon T_{1} U T_{2}
\end{align*}
$$

Koford and Groner [6] observed that the criterion $D_{2}(f, W)$ can be expre expressed in the form

$$
D_{2}(f, W)=\sum_{j=1}^{2} N_{j}\left[\ell^{\top} S_{j} \ell+\left\{\ell^{T} m_{j}+c_{o}+(-1)^{j}\right\}^{2}\right]
$$

where

$$
c^{\top}=\left[l^{\top}, c_{o}\right], m_{j}=\underset{x \in \top_{j}}{\Sigma} x / N_{j}, s_{j}=\sum_{x \in \top_{j}}\left(x-m_{j}\right)\left(x-m_{j}\right)^{\top} / N_{j}
$$

and $N_{j}$ is the number of observations in $T_{j}$. The values of $\&$ and $c_{o}$ that minimize $D_{2}(f, W)$, given by ( 8 ), are as follows.

As mentioned earlier, a nonlinear classifier $W(x, c)$ is obtained in the form of a linear combination

$$
\begin{equation*}
W(x, c)=c^{\top} \Phi(x)=c_{1} Q_{1}(x)+\ldots+c_{m} Q_{m}(x) \tag{1}
\end{equation*}
$$

of linearly indeperdent nonlinear continuous functions $Q_{1}(x), \ldots, Q_{m}(x)$, whose functional forms and number are prechosen. The performance of the classifier $W(x, c)$, how often the classifier will be able to correctly classify an observation of known classification, depends primarily on the choice of these functions. But, often they are chosen arbitrarily on the basis of economic consideration and one's intuition. However, sometimes, when one has some prior knowledge of the population densities, they can be satisfactorily chosen. For example, when it is known that the populations have multivariate normal distributions with unequal covariance matrices, the function $W(x, c)$ may be chosen as a second degree polynomial in $x_{1}, \ldots$, $x_{p}$, the $p$ components of the observation vector $x$, and the functions $Q_{i}(x)$ 's as the monomials $x_{1}{ }^{k_{1}} x_{2}{ }^{k_{2}} \ldots x_{p}{ }^{k_{p}}$, where $k_{i}{ }^{\prime}$ s are 0 , 1 or 2 and $k_{1}+k_{2}+\ldots+k_{p}=2$. It has been suggested [5] that when the densities are unknown, the functions $Q_{i}(x)$ 's should be taken as orthogonal or orthonormal functions such as generalized Hermite functions given by

$$
H_{k_{1}}, k_{2}, \ldots, k_{p}(x)=(-1)^{n}(2 \pi) \frac{-n / 2}{\partial x_{1}{ }_{1} \ldots \partial x_{p}{ }^{n}} \exp \left(-\frac{1}{2} \sum x_{i=1}^{p}{ }^{2}\right)
$$

Unfortunately, there is no specific rule to guide us in selecting these functions judiciously.
4.1. Patterson - Womack Classifier

The adaptive classifier of Patterson and Womack [8] is a nonlinear classifier $W(x, c)$ obtained as an approximation to a function $f(x)$ that separates the training sample sets $T_{1}$ and $T_{2}$ and takes the value $c(2 \mid 1)$ on $T_{1}$ and $-C(1 \mid 2)$ on $T_{2}$, where $C(j \mid i)$ denotes the cost of misclassifying an element from $\pi_{i}$ into $\Pi_{j}$. The parameter vector $c$ is selected as a vector $\hat{c}$ that minimize the pay-off function

$$
\begin{align*}
M\left(N_{1}, N_{2}\right)= & \left(q_{1} / N_{1}\right) \sum_{x \in T_{1}}|W(x, c)-c(2 \mid 1)|^{2}+ \\
& \left(q_{2} / N_{2}\right) \sum_{x \in T_{2}}|W(x, c)+c(1 \mid 2)|^{2} \tag{14}
\end{align*}
$$

where $N_{i}$ is the number of elements in $T_{i}$ and the prior probability $q_{i}$ of the population $\Pi_{i}$ is assumed to be known. As in (5) and (6), it is not difficult to see that $M\left(N_{1}, N_{2}\right)$ is the sample average of $|W(X, c)-f(X)|^{2}$ and therefore, by law of large number, $M\left(N_{1}, N_{2}\right)$ converges to

$$
R(f, W)=E|W(X, c)-f(X)|^{2}
$$

as $N_{1} \rightarrow \infty$ and $N_{2} \rightarrow \infty$. Patterson and Womack [8] have shown that the vector $\hat{c}$ that minimize $R(f, W)$, as given above, will also minimize

$$
E|W(x, c)-D(x)|^{2}
$$

where $D(X)$ is the Bayes' classifier given by

$$
\begin{equation*}
D(x)=\left\{c(2 \mid 1) q_{1} p_{1}(x)-c(1 \mid 2) q_{2} p_{2}(x)\right\} /\left\{q_{1} p_{1}(x)+q_{2} p_{2}(x)\right\} \tag{15}
\end{equation*}
$$

and $p_{i}(x)$ is the density of $\Pi_{i}$, provided that $f(x)$ has been chosen in the above fashion. This has been the motivation behind using $C(2 \mid 1)$ and $c(1 \mid 2)$ as the preassigned values of the separating function $f(x)$.

When $q_{i}$ 's are unknown and to be estimated, then $N_{i} /\left(N_{1}+N_{2}\right)$ can be used as an unbiased estimate of $q_{i}$, where $N_{i}$ is the number of elements belonging to $T_{i}$ out of a sample of size $\left(N_{1}+N_{2}\right)$ from the mixed population. Then the criterion $M\left(N_{1}, N_{2}\right)$ may be replaced by

$$
\begin{aligned}
\bar{D}_{2}(f, W)= & \Sigma \quad|W(x, c)-f(x)|^{2} /\left(N_{1}+N_{2}\right), \\
& x \in T \quad U T_{2}
\end{aligned}
$$

the mean square error criterion. Patterson and Womack have called $M\left(N_{1}, N_{2}\right)$ the mean square error criterion.

Example. Let $T_{1}=\left\{[0 ; 1]^{\top},[1,0]^{\top},[2,1]^{\top}\right\}$ and $T_{2}=\left\{[0,0]^{\top},[2,0]^{\top},[0,-1]^{\top}\right\}$.
Then, assuming $W(x, c)$ to be of the form

$$
\begin{equation*}
w(x, c)=c_{1} x_{1}^{2}+c_{2} x_{2}^{2}+c_{3} x_{1} x_{2}+c_{4} x_{1}+c_{5} x_{2}+c_{6} \tag{16}
\end{equation*}
$$

and $c(1 \mid 2)=c(2 \mid 1)=1$, we can obtain the parameters $c_{1}, c_{2}, \ldots, c_{6}$ by the least square method. Solving the normal equations, we obtain

$$
c_{1}=-11 / 8, c_{2}=1 / 4, c_{3}=5 / 8, c_{4}=17 / 8, c_{5}=1 \text { and } c_{6}=-1 / 4
$$

and

$$
w(x, c)=(1 / 8)\left[2 x_{2}^{2}-11 x_{1}^{2}+5 x_{1} x_{2}+17 x_{1}+8 x_{2}-2\right]
$$

A decision rule may be defined in order to assign $x \in I t$, whenever $W(x, c)>0$ and to $\Pi_{2}$ otherwise. We may note that such a decision rule correctly classifies all the observations in $T_{1}$ and $T_{2}$.

In this example, the sets $T_{1}$ and $T_{2}$ can be separated by a quadric such as parabola, ellipse, hyperbola or circle. That is why a classifier assumed to be of the form (16) of a second degree polynomial in $x$ could classify all the points of $T_{1}$ and $T_{2}$ correctly. But there are cases when the sets $T_{1}$ and $T_{2}$ cannot be separated by any such curve. Then a second degree polynomial in $x$ will not be so efficient, that is, the classifier will not be able to classify so many points of $T_{1}$ and $T_{2}$.

### 4.2. Potential Function Method

The potential function method was introduced for adaptive classifier design by Aizerman, Braverman and Rozonoer [1]. They assumed that a function $f(x), f(x)>0$ on $T_{1}$ and $f(x)<0$ on $T_{2}$, that is, sign $f(x)=1$ on $T_{1}$ and sign $f(x)=-1$ on $T_{2}$, that can be used to separate $T_{1}$ and $T_{2}$ can be expressed in the form

$$
\begin{equation*}
f(x)=c_{1} Q_{1}(x)+\ldots+c_{m} Q_{m}(x) \tag{17}
\end{equation*}
$$

where $\left\{Q_{1}(x), \ldots, Q_{m}(x)\right\}$ is a finite set of orthogonal or orthonormal functions. This function $f(x)$, which can be used as a classifier, can be obtained iteratively using the following algorithm.

$$
\begin{equation*}
f_{n+1}(x)=f_{n}(x)+s(n+1) K(x, x(n+1)), f_{0}(x)=0 \tag{18}
\end{equation*}
$$

where $s(n+1)=\operatorname{sign} f(x(n+1))-\operatorname{sign} f_{n}(x(n+1)), f_{n}(x)$ is the approximation to $f(x)$ obtained at the nth step of iteration, $x(n)$ is the nth observation in a sequence of observations of known classification from the
mixture of the populations $\Pi_{1}$ and $\Pi_{2}$ and $K(x, y)$ is a potential function of the form

$$
\begin{equation*}
K(x, y)=\sum_{i=1}^{m} Q_{i}(x) Q_{i}(y) \tag{19}
\end{equation*}
$$

In order to avoid the problem of choice of the set $\left\{Q_{1}(x), \ldots\right.$, $Q_{m}(x)$ \} Braverman [3] made the following suggestions. Let us assume that $f(x)$ can be represented by

$$
\begin{equation*}
f(x)=\sum_{i=1}^{\infty} c_{i} Q_{i}(x), \sum_{i=1}^{\infty} c_{i}^{2}<\infty, \tag{20}
\end{equation*}
$$

where $\left\{Q_{i}(x)\right\}$ is a complete set of square integrable functions. Then $K(x, y)$ in the above algorithm (18) should be replaced by a function $K(x, y)$ of the form

$$
\begin{equation*}
K(x, y)=\sum_{i=1}^{\infty} Q_{i}(x) Q_{i}(y) \tag{21}
\end{equation*}
$$

Braverman showed that if we take

$$
\begin{equation*}
K(x, y)=F(d(x, y)) \tag{22}
\end{equation*}
$$

where $F$ is a continuous real function, $d(x, y)$ is any function defining distance between $x$ and $y$ and $F$ has a positive Fourier transform everywhere, then $K(x, y)$ can be represented in the form (21). For example, $K(x, y)$ may be chosen as

$$
\begin{equation*}
K(x, y)=\exp \left\{-\alpha_{\sum}^{2} \sum_{i=1}^{p}\left|x_{i}-y_{i}\right|^{2}\right\} \tag{23}
\end{equation*}
$$

where $\alpha$ is any finite real number. Braverman has shown that the sequence $f_{n}(x)$, where $x$ is an observation from $\pi_{1}$ or $\bar{\pi}_{2}$, converges in mean to the function $f(x)$.

Example. In the expression (23) for $K(x, y)$ we choose $\alpha=1$ and illustrate the construction of $f_{n}(x)$.

| $n$ | $x_{1}$ | $x_{2}$ | Member | $s(n)$ | $f_{n}(x)$ |
| :--- | :---: | :---: | :---: | :---: | :---: |
| 1 | 0 | 0 | $T_{2}$ | -1 | $-\exp \left(-\Sigma x_{i}^{2}\right)$ |
| 2 | 0 | 1 | $T_{1}$ | 2 | $2 \exp \left\{-x_{1}^{2}-\left(x_{2}-1\right)^{2}\right\}+f_{1}(x)$ |
| 3 | 2 | 0 | $T_{2}$ | 0 | $f_{2}(x)$ |
| 4 | 2 | 1 | $T_{1}$ | 0 | $f_{2}(x)$ |
| 5 | 0 | -1 | $T_{2}$ | 0 | $f_{2}(x)$ |
| 6 | 1 | 0 | $T_{1}$ | 2 | $f_{2}(x)+2 \exp \left\{-\left(x_{1}-1\right)^{2}-x_{2}^{2}\right\}$ |

### 4.3. Stochastic Approximation Method

Yau and Schumpert [10] have used stochastic approximation method in obtaining an adaptive classifier $W(x, c)$, as in (l), as an approximation to a function $f(x)$ that takes the value 1 on $T_{1}$ and -1 on $T_{2}$ using $G(c)$ given by

$$
G(c)=E|f(x)-W(X, c)|^{2},
$$

a special case of (5), as a pay-off function. $G(c)$ has a unique minimum, or, the equation $G^{\prime}(c)=0$ has a unique solution. But the density $q_{1} P_{1}(x)+$ $q_{2} P_{2}(x)$ of $X$ from the mixture of $I_{1}$ and $\Pi_{2}$ being unknown, $G(c)$ is unknown. The solution $c$ of $G^{\prime}(c)=0$ can be obtained iteratively using the following algorithm.

$$
c(n+1)=c(n)-a_{n}\left\{c^{\top}(n) \phi(x(n)-f(x(n))\} \Phi(x(n))\right.
$$

where $x(n)$ is the nth observation from the mixture of $\Pi_{1}$ and $\Pi_{2} \quad \Phi^{\top}(x)=$ $\left[Q_{1}(x), \ldots, Q_{m}(x)\right]$ and $\left\{a_{n}\right\}$ is a sequence of real numbers satisfying the conditions

$$
\text { (a) } \sum_{n=1}^{\infty}=\infty \text { and } \sum_{n=1}^{\infty} a_{n}^{2}<\infty \text {. }
$$

$\{1 / n\}$ is an example of one such sequence. The convergence of $c(n)$ to $c$ follows from the fact that $c(n)$ forms a Robbins - Monroe process.

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ON RECOGNITION OF WANDERING PATTERNS
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## ABSTRACT

Two recursive estimators for the current 'mean' of two stochastic processes used in mod. Iing patterns varying over space and/or time have been obtained. A brief survey of models and estimators so far proposed in literature has also been made.

## 1. Introduction

In the analysis of remotely sensed data, such as multispectral scannee (mss) data, it is usual practice to assume that the data generated come from populations having multivariate normal distribution. When the parameters of the distributions, namely, the means and the covariance matrices, are known, then assuming the misclassification costs to be the same for all populations the Bayes' classifier, the classifier with minimum probability of misclassification, is given by

$$
\left.W_{i j}(x)=\log p_{i} x\right)-\log p_{j}(x)+\log \left(q_{i} / q_{j}\right), i \neq j, i, j=1, \ldots, m
$$

where $m$ is the number of populations and $P_{i}(x)$ is the density and $q_{i}$ is the prior probability of the fth population. When the parameters are unknown the usual and useful practice in designing classifiers is to replace the unknown parameters in (1) by their respective estimates obtained from a training sample, a sample of known classification. These sample based classifiers, which we shall refer to as estimate plug-in classifiers, do not in general minimize the probability of misclassification. But they are useful in large number of applications and theoretically desirable since we know (Anderson [2]) that as the training sample sizes increase the estimates converge with probability one to the value of the unknown parameter, provided that the observations in a training sample are dentically distributed. Thus for large training sample sizes and each training sample having identically distributed observations a sample based classifier may be expected to perform satisfactorily.

The data obtained by remote sensing devices in the earth resources survey come from large areas and over a long period of time. Often the
statistical characteristics of the data have been found to undergo changes over space and time due to variation in spatial and temporal conditions. Thus observations in a training sample collected from a large area and/or over a long period of time are not identically distributed. Therefore, the estimate plug-in classifiers in which the estimates are based on such samples can no longer be expected to perform satisfactorily in classifying patterns that vary over space or time. The performance of these estimate plug-in classifiers can be improved by updating the estimates whenever necessary.

The estimate updating methods are applicable to all populations in the same way. So, without loss of generality, we can discuss these methods in the context of any single population.

The efficiency of the updated estimates depends to a great degree on the accuracy of the statistical model chosen to represent the statistical nature of the varying patterns. But the complexity of the statistical model may on the other hand lead to estimates that are not useful for practical purposes because of computational inconvenience. In this paper we shall consider two general models for estimation and some "quick" methods of estimation based on exponential smoothing techniques (Box and Jenkins [3], Brown [4]).

Throughout the paper we shall denote a sequence of observation veltors from the p - dimensional population II by $\mathrm{Y}_{1}, \mathrm{Y}_{2}, \ldots, \mathrm{Y}_{\mathrm{n}}, \ldots$ and assume

$$
\begin{equation*}
Y_{k}=X_{k}+W_{k}, \tag{1}
\end{equation*}
$$

where $X_{k}$ 's are the signals, true patterns or means of $Y_{k}$ 's and $W_{k}{ }^{\prime} s$ are
p - variate normal random noise vectors distributed independently and identically as $N(0, C)$. Henceforth we slall write $L(\cdot)$ to designate the distribution law of the random vector or vectors in the parenthesis.

## 2. Autoregressive Mode1

The pattern $X_{k}$ can be assumed to be a member of a r-th order autoregressive process. For the sake of simplicity and mathematical tractability we assume that $X_{k}$ is a member of a first order autoregressive vector process given by

$$
\begin{equation*}
X_{k}-A X_{k-1}=Z_{k-1} \tag{2}
\end{equation*}
$$

where $A$ is a pxp matrix of real numbers, $Z_{k}$ 's are independently and identically distributed as $N(0, \Sigma), Z_{k}^{\prime} s$ are independent of $W_{k}^{\prime} s$ and $\mathrm{L}\left(\mathrm{X}_{1}\right)=\mathrm{N}(0, \mathrm{~V})$. The process is known (Box and Jenkins [3]) to be stationary if $|A|<1$, where $|A|$ denotes the determinant of $A$, and nonstationary otherwise.

We shall denote by $L(P), L(P, Q)$ and $L(P \mid Q)$ the probability distribution of $P, P$ and $Q$ jointly and conditional distribution of $P$ given $Q$ respectively. We know that if $L(P, Q)=N(\mu, D), L(P)=N\left(\mu_{1}, D_{11}\right), L(Q)=$ $\mathrm{N}\left(\mu_{2}, \mathrm{D}_{22}\right)$ and

$$
D=\left[\begin{array}{ll}
D_{11} & D_{12} \\
D_{21} & D_{22}
\end{array}\right], \mu=\left[\begin{array}{l}
\mu_{1} \\
\mu_{2}
\end{array}\right],
$$

then

$$
L(P \mid Q)=N(Q, \phi), \text { where }
$$

$$
\begin{equation*}
Q=\mu_{1}+D_{12} D_{22}^{-1}\left(Q-\mu_{2}\right) \tag{3}
\end{equation*}
$$

and $\quad \phi=\mathrm{D}_{11}-\mathrm{D}_{12} \mathrm{D}_{22}^{-1} \mathrm{D}_{21}$.

Therefore, from (1) and (2) we have

$$
Y_{1}=X_{1}+W_{1}, L\left(Y_{1} \mid X_{1}\right)=N\left(X_{1}, C\right), L\left(Y_{1}\right)=N(0, V+C)
$$

and hence $L\left(Y_{1}, X_{1}\right)=N(0, Q)$, where

$$
\mathrm{Q}=\left[\begin{array}{cc}
\mathrm{v}+\mathrm{C} & \mathrm{v} \\
\mathrm{v} & \mathrm{v}
\end{array}\right]
$$

From (3) we now have $L\left(X_{1} \mid Y_{1}\right)=N\left(\mu_{1}, Q_{1}\right)$, where

$$
\begin{align*}
& \mathrm{u}_{1}=\mathrm{E}\left(\mathrm{X}_{1} \mid \mathrm{Y}_{1}\right)=\mathrm{V}(\mathrm{~V}+\mathrm{C})^{-1} \mathrm{Y}_{1}  \tag{4}\\
& \mathrm{Q}_{1}=\operatorname{Cov}\left(\mathrm{X}_{1}, X_{1} \mid \mathrm{Y}_{1}\right)=\mathrm{C}(\mathrm{~V}+\mathrm{C})^{-1} \mathrm{~V}
\end{align*}
$$

Similarly we have

$$
\mathrm{Y}_{2}=\mathrm{X}_{2}+\mathrm{W}_{2}=\mathrm{AX} \mathrm{X}_{1}+\mathrm{Z}_{1}+\mathrm{W}_{2}
$$

so, $L\left(Y_{2} \mid Y_{1}\right)=N\left(A \mu_{1}, A Q_{1} A^{T}+\Sigma+C\right)$ and $L\left(X_{2} \mid Y_{1}\right)=N\left(A H_{1}, A Q_{1} A^{T}+\Sigma\right)$

As in (4), we obtain from (3)

$$
\begin{aligned}
\mu_{2}=E\left(X_{2} \mid Y_{2}, Y_{1}\right)= & C\left(A Q_{1} A^{T}+\Sigma+C\right)^{-1} A \\
& +\left(A Q_{1} A^{T}+\Sigma\right)\left(A Q_{1} A^{T}+\Sigma+C\right)^{-1} Y_{2}
\end{aligned}
$$

and

$$
Q_{2}=\operatorname{Cov}\left(X_{2}, X_{2} \mid Y_{2}, Y_{1}\right)=C\left(A Q_{1} A^{T}+\Sigma+C\right)^{-1}\left(A Q_{1} A^{T}+\Sigma\right)
$$

thus

$$
\begin{align*}
& \mu_{k}=E\left(X_{k} \mid Y_{k}, \ldots, Y_{1}\right)= C\left(A Q_{k-1} A^{T}+\Sigma+C\right)^{-1} A \mu_{k-1}+\left(A Q_{k-1}\right. \\
&+\left(A Q_{k-1} A^{T}+\Sigma\right)\left(A Q_{k-1} A^{T}+\Sigma+C\right)^{-1} Y_{k-1} \\
& Q_{k}=\operatorname{Cov}\left(X_{k}, X_{k} \mid Y_{k}, \ldots, Y_{1}\right)=C\left(A Q_{k-1} A^{T}+\Sigma+C\right)^{-1}\left(A Q_{k-1} A^{T}+\Sigma\right) \tag{5}
\end{align*}
$$

and

From the above model we can obtain the model of Abramson and Braverman [1] by taking $A=a I_{p}$, where $I_{p}$ is the pep identity matrix and $a$ is a real number, and that of Scudder [12] by taking $A=I_{p}$. If we assume $|A|<1$, so that the process is stationary, then we obtain the model proposed by Tamura et al [14]. Following Abramson and Braverman, if we assume for slowly varying patterns that $Q_{k-1}=\lambda C_{:} \Sigma=\left\{\lambda^{2} /(1-\lambda)\right\} C$, $0<\lambda<1$, in order that $Q_{k-1}=Q_{k}$, then we have from (5)

$$
\mu_{k}=(1-\lambda) \mu_{k-1}+\lambda Y_{k}
$$

Writing $\quad \mu_{k}^{\prime}=E\left(X_{k+1} \mid Y_{k}, \ldots, Y_{1}\right)$, we obtain

$$
\begin{align*}
\mu_{k}^{\prime} & =E\left(X_{k+1} \mid Y_{k}, \ldots, Y_{1}\right) \\
& =E\left(X_{k}+Z_{k} \mid Y_{k}, \ldots, Y_{1}\right)=\mu_{k} \tag{6}
\end{align*}
$$

Therefore $\mu_{k}^{\prime}=(1-\lambda) \mu_{k-1}^{\prime}+\lambda Y_{k}$.

The above recursive formula is reminiscent of the exponentially weighted averages to be discussed in section 4.

In the model (2) it is implied that a new pattern is encountered at each observation, in other words, $\mathrm{Y}_{1}, \ldots, \mathrm{Y}_{\mathrm{n}}$ are observations on random vectors each having different means. The efficiency of the estimator (5)
will be poor if the changes in pattern are few and far between. In SecLion 6 we will come back to the question of how to improve the estimates (5).

## 3. Chernoff - Wacks Model

For analyzing time varying patterns Chernoff and Racks [6] proposed a model and obtained minimum variance linear unbiased (MVLU) estimate of the current mean. They also obtained Bayes' estimator and an "ad hoc" estimator, a simplified version of Bayes' estimator. But they are unisable for practical purposes, Bayes' estimator due to computational difficulties and the "ad hoc" estimator due to its being based on many restrictive assumptions. Chernoff and Racks dealt with univariate observations. There we have $Y_{k}, X_{i}, W_{i}$ as univariate random variables. The model assumes

$$
\begin{align*}
& Y_{i}=X_{i}+W_{i}, \quad i=1, \ldots, n  \tag{7}\\
& X_{i}=X_{i+1}+J_{i} Z_{i}, \quad i=1, \ldots, n-1 \tag{8}
\end{align*}
$$

where $W_{i}$ 's are independently and identically distributed as $N(0,1)$, $J_{i}$ 's are random variables having the value 1 if there is a change in the mean or pattern $X_{i}$ between $i$ th and $(i+1)$ th observation and the value 0 otherwise and $Z_{i}$ is a random variable representing the amount of change when a change takes place.

Let us write

Then noting that we can combine (7) and (8) to obtain the equation

$$
\begin{aligned}
Y_{i} & =X_{n}+W_{i}+\sum_{k=i}^{n-1} J_{k} Z_{k}, \underline{1 \leq i} \leq n-1, \\
& =X_{n}+W_{n},
\end{aligned}
$$

we can write

$$
\begin{equation*}
y=X_{n} e_{n}+w+M z, \tag{10}
\end{equation*}
$$

where

$$
\underset{\operatorname{nxn}}{M}=\left[\begin{array}{ccccc}
J_{1} & J_{2} & \ldots & J_{n-1} & 0 \\
& J_{2} & \ldots & J_{n-1} & 0 \\
& & \ldots & J_{n-1} & 0 \\
0 & & & & 0
\end{array}\right] \text { and } \begin{array}{r}
e_{n} \\
\\
\\
\end{array}
$$

If we assume

$$
\begin{equation*}
P\left(J_{k}=1\right)=p_{k}, k=1, \ldots, n-1 \tag{11}
\end{equation*}
$$

and

$$
L(Z)=N\left(0, \sigma^{2} S\right)
$$

where

$$
S=\left[\begin{array}{ll}
I_{n-1} & 0  \tag{12}\\
0 & 0
\end{array}\right]
$$

then we obtain from (10) that
where

$$
L\left(Y \mid X_{n}, J\right)=N\left(X_{n} e_{n}, Q(J)\right)
$$

$$
\begin{equation*}
Q(J)=I_{n}+\sigma^{2} M M^{T} \tag{13}
\end{equation*}
$$

From (11) we have

$$
\begin{align*}
\operatorname{EQ}(J) & =\sum_{k=1}^{n-1} E\left[p_{k}\left\{Q(J) \mid J_{k}=1\right\}+\left(1-p_{k}\right)\left\{Q(J) \mid J_{k}=0\right\}\right] \\
& =I_{n}+\sum_{k=1}^{n-1} p_{k} E\left(M^{T} \mid J_{k}=1\right) \\
& =I_{n}+\sum_{k=1}^{n-1} p_{k} R(n, k)
\end{align*}
$$

where $R(n, k)$ is a $n \times n$ matrix whose upper left submatrix is $E_{k}$, a kxk matrix all of whose elements are 1 , and the other elements axe zero.

When changes in the mean occur almost always or when there are almost no changes, the random vector $J$ can be assumed to be distributed independently of $Y$. In that case we obtain from (13) and (14)

$$
\begin{equation*}
L\left(Y \mid X_{n}\right)=N\left(X_{n} e_{n}, V_{n}\right) \tag{15}
\end{equation*}
$$

$$
\begin{equation*}
\text { where } \quad V_{n}=\operatorname{EQ}(J)=I_{n}+\sum_{k=1}^{n-1} \sigma_{k}^{2} p_{k} R(n, k) \text {. } \tag{16}
\end{equation*}
$$

If $\mu_{n}$ is the current mean, the mean of $Y_{n}$, then using standard arguements it can be shown from (15) that

$$
\begin{equation*}
\hat{\mu}_{n}=e_{n}^{T} V_{n}^{-1} Y / e_{n}^{T} V_{n}^{-1} e_{n} \tag{17}
\end{equation*}
$$

is a MVLU estimator of $\mu_{n}$. If $v(n, j)$ denotes the sum of the elements of the fth column of $V_{n}^{-1}$, then $\hat{\mu}_{n}$ can be written in the form

When we have $n+1$ observations, we can obtain in a way similar to (16)

$$
\begin{aligned}
V_{n+1} & =I_{n+1}+\sum_{k=1}^{n} \sigma^{2} p_{k} R(n+1, k) \\
& =\left[\begin{array}{ll}
V_{n} & 0 \\
0 & 1
\end{array}\right]+\sigma^{2} p_{n}\left[\begin{array}{l}
e_{n} \\
0
\end{array}\right]\left[\begin{array}{c}
e_{n}^{T}
\end{array} 0\right]
\end{aligned}
$$

We know (Rap [11] p. 29) that if

$$
\underset{\operatorname{pxp}}{\mathrm{B}}=\underset{\operatorname{pxp}}{A}+\underset{\mathrm{pxi}}{\mathrm{U} \mathrm{~V}^{\mathrm{T}}} \mathrm{lxp}^{\text {, }}
$$

then $\quad B^{-1}=A^{-1}-A^{-1} U V^{T} A^{-1} /\left(1+V^{T} A^{-1} U\right)$.

Therefore,

$$
v_{n+1}^{-1}=\left[\begin{array}{ll}
v_{n}^{-1} & 0 \\
0 & 1
\end{array}\right]-\frac{\sigma^{2} p_{n}}{1+\sigma^{2} p_{n} e_{n}^{T} v_{n}^{-1} e_{n}} K_{n+1}
$$

where $\quad K_{n+1}=\left[\begin{array}{c}\dot{v}(n, 1) \\ \vdots \\ v(n, n) \\ 0\end{array}\right]\left[\begin{array}{lllll}v(n, 1) & \ldots & v(n, n) & 0\end{array}\right]$.
So, $\quad v(n+1, n+1)=1$

$$
v(n+1, i)=v(n, i) /\left[1+\sigma^{2} p_{n} \sum_{k=1}^{n} v(n, k)\right], 1 \leq i \leq n,
$$

and. $1+\sum_{i=1}^{n} v(n+1, i)=\left[1+\left(1+\sigma^{2} p_{n}\right) \sum_{i=1}^{n} v(n, i)\right] /\left[1+\sigma^{2} p_{n} \sum_{k=1}^{n} v(n, k)\right]$
Therefore, $\quad \hat{u}_{n+1}=\left[\sum_{k=1}^{n} v(n+1, k) Y_{k}+Y_{n+1}\right] /\left[1+\sum_{k=1}^{n} v(n+1, k)\right]$

$$
=\frac{\left.\sum_{i=1}^{n} v(n, i)\right] \hat{\mu}_{n}+\left[1+\sigma^{2} p_{n} \sum_{i=1}^{n} v(n, i)\right] Y_{n+1}}{1+\left(1+\sigma^{2} p_{n}\right) \sum_{i=1}^{n} v(n, i)}
$$

;
where

$$
\begin{equation*}
=\left(1-\lambda_{\mathrm{n}}\right) \hat{\mu}_{\mathrm{n}}+\lambda_{\mathrm{n}} \mathrm{Y}_{\mathrm{n}+1} \tag{21}
\end{equation*}
$$

$$
\begin{equation*}
\lambda_{n}=\left\{1+\sigma^{2} p_{n} \sum_{i=1}^{n} v(n, i)\right\} /\left\{1+\left(1+\sigma^{2} p_{n}\right) \sum_{i=1}^{n} v(n, i)\right\} . \tag{22}
\end{equation*}
$$

Using (22) we can now recursively estimate the current mean.
4. Exponentially Weighted Moving Average

In analyzing observations on space and time varying patterns, especially time series data, it has been a popular practice (see Brown [4]) to use exponentially weighted moving average (EWMA) (Box and Jenkins [3]) as the current "mean", location or "level" of the process because of its success in a variety of applications. Let $\left\{\ldots, Y_{t-1}, Y_{t}\right\}$ be observations at .. $\mathrm{t}-1$, t (a stochastic process in discrete time). Then an exponenttially weighted moving average (EWMA) is a predictor or forecast of a future level at $t+h$ derived by weighting past observations "exponentially" (or geometrically) and expressed in the form

$$
\begin{array}{rlrl}
\bar{Y}(t, h ; \lambda) & =\lambda \sum_{r=0}^{\infty}(1-\lambda)^{r} Y_{t-r} & ,|\lambda|<1 \\
& =\lambda Y_{t}+(1-\lambda) \bar{Y}(t-1, h ; \lambda),|\lambda|<1 . \tag{24}
\end{array}
$$

For $h=1$, we shall waite

$$
\begin{equation*}
\bar{Y}(t, 1 ; \lambda)=\bar{Y}_{t}(\lambda) \tag{25}
\end{equation*}
$$

Thus from (24) and (25) we have

$$
\begin{align*}
\bar{Y}_{t}(\lambda) & =\lambda Y_{t}+(1-\lambda) \bar{Y}_{\mathrm{t}-1}(\lambda)  \tag{26}\\
& =\lambda \sum_{\mathrm{r}=0}^{\infty}(1-\lambda)^{r} Y_{\mathrm{t}-\mathrm{r}}
\end{align*}
$$

Mut [10] has shown that $\bar{Y}_{t}(\lambda)$, as given by (26), is optimal for some $\lambda$ for the following nonstationary moving average process. Let

$$
\begin{align*}
& Y_{t}=X_{t}+W_{t} \\
& X_{t}=X_{t-1}+Z_{t}, \tag{27}
\end{align*}
$$

where $W_{t}{ }^{\prime}$ 's are independently distributed as $N\left(0, \sigma^{2}\right)$ and $Z_{t}{ }^{\prime} s$ are independently distributed as $N\left(0, t^{2}\right)$. Then the predictor $\bar{Y}_{t}(\lambda)$ given by (26) minimizes the error variance

$$
\begin{equation*}
E\left(Y_{t+1}-\bar{Y}_{t}(\lambda)\right)^{2} \tag{28}
\end{equation*}
$$

when $\lambda$ is given by

$$
\begin{equation*}
\lambda=1+\frac{t^{2}}{2 \sigma^{2}}-\frac{t}{\sigma} \quad\left(1+\frac{t^{2}}{4 \sigma^{2}}\right)^{1 / 2} \tag{29}
\end{equation*}
$$

From (28) it also follows that

$$
\begin{equation*}
\bar{Y}_{t}(\lambda)=E\left(X_{t+1} \mid Y_{t}, Y_{t-1}, \ldots\right) \tag{30}
\end{equation*}
$$

When $Y_{t}$ is a pul vector, $L\left(W_{t}\right)=N(0, C), L\left(Z_{t}\right)=N(0, \Sigma)$, then following Math, it can be shown that $\bar{Y}_{t}(\lambda)$ minimizes the scatter of error given by

$$
E\left(Y_{t+1}-\bar{Y}_{t}(\lambda)\left(Y_{t+1}-\stackrel{\rightharpoonup}{Y}_{t}(\lambda)\right)^{T}\right.
$$

provided we take

$$
\begin{equation*}
\lambda=1+(|\Sigma| / 2|C|)-|\Sigma| /|c| \sqrt{1+|\Sigma| / 4|C|} \tag{31}
\end{equation*}
$$

In passing, it should be noted that the choice of the weight $\lambda$ is vital in the definition of EWMA. In many situations EWMA's may not be optimal, because the underlying process may not be of the form (27), or even if the process is of the form (27), the matrices $\Sigma$ and $C$ are unknown, so that it may be impossible to obtain the exact value of $\lambda$ given by (29) or (31). Even then they are attractive because they are easy to compute.

Motivated by the work of Abramson and Braverman [1], especially by the equation (6) which has been derived from a model of the form (27) and perhaps following the popular practice, Kriegler and Horwitz [8] have proposed that for simultaneously updating the current mean of several populations whose covariance matrices do not change any one of the following three algorithms should be adopted.
(1) Exponentially Weighted Running Estimates

If an observation $\lambda$ has been recognized as one coming from the class $\Pi_{i}$, then the present mean $M_{i}$ of this class is to be updated to a new value $\mathrm{M}_{\mathrm{i}}$, where

$$
\begin{equation*}
M_{i}^{\prime}=(I-\lambda) M_{i}+\lambda Y, \tag{32}
\end{equation*}
$$

and $\lambda$ is a constant.

Kriegler and Horwitz have recommended that $\lambda$ should be given a value between 0.001 and 0.003 . Evidently such small $\lambda$ will heavily weigh the past observations, especially the present mean $M_{i}$ and will attach very little weight to the current observation. Unless the patterns are extremely slowly varying, such small $\lambda$ will make the updated mean $M_{i}$ a very poor representative of the current pattern.
(2) Exponentially Weighted Running Estimates With Interaction

Assuming that at any time the mean $M_{i}$ of the class $\pi_{i}$ is related to the sum $M$ of means of all $M$ classes by the equation

$$
\begin{equation*}
M=m p_{i} M_{i}, \tag{33}
\end{equation*}
$$

where $p_{i}$ is constant for all time, the present mean $M_{j}$ of any class $\mathrm{II}_{\mathrm{j}}$ should be updated to the value $\mathrm{M}_{\mathrm{j}}^{-}$, where

$$
\begin{equation*}
M_{j}^{\prime}=p_{i} M_{i}^{\prime} / p_{j} \tag{3/4}
\end{equation*}
$$

and $M_{i}^{\prime}$ is the current updated mean of $H_{i}$, updated according to (32).
The assumption (33) is hard to justify.
(3) Posterior Probability Weighted Estimates

Let $k_{i}$ denote the fixed covariance matrix of $\Pi_{i}$. Then according to this algorithm an observation $Y_{t}$ is assigned to the class $\Pi_{i}$ if for some preassigned threshold value $t_{1}$,

$$
L_{i}\left(Y_{t}\right)>L_{j}\left(Y_{t}\right) \text { and } Q_{i}\left(Y_{t}\right)<t_{i} \text {, }
$$

where $\quad L_{i}\left(Y_{t}\right)=\left|K_{i}\right|^{-\frac{1}{2}} \exp \left\{-Q_{i}\left(Y_{t}\right) / 2\right\}$
and

$$
Q_{i}(Y)=\left(Y_{\bar{t}} M_{i}\right)^{T} K_{i}^{-I}\left(Y_{t}-Y_{i}\right)
$$

The mean $M_{i}$ should be updated to $M_{i}^{\prime}$ given by

$$
\begin{align*}
M_{i}^{\prime} & =1+\gamma R_{i}\left(Y_{t}\right)\left(Y_{t}-M_{i}\right)  \tag{35}\\
& =\{1-\lambda(t, i)\} M_{i}+\lambda(t, i) Y_{t}
\end{align*}
$$

where $R_{i}\left(Y_{t}\right)=L_{i}\left(Y_{t}\right) / \Sigma_{j=1}^{m} L_{j}\left(Y_{t}\right)$
and $\lambda(t, i)=\gamma R_{i}\left(Y_{t}\right), \quad \gamma$ being a constant $\quad(0<\gamma<1)$.
The updating formula (21) based on Cnemoff - Racks model and formula (35), both resemble EWMA, except that the weights are now variable and to be updated each time. The predictor (21) has been shown to be MVLU. But the predictor $M_{i}^{-}$given by (35) may not enjoy any such property. It is not known for what kind of process $M_{i}^{-}$will be optimal in some sense. Kriegler and Horwitz have suggested some values of $t_{1}, t_{2}$ and $\gamma$ on empirical grounds, but have failed to give a specific rule for the selection of these quantities. The performances of the predictors (35) in different situations are yet to be determined and compared.

The work of Kriegler and Horwitz [8] motivated Chang [5] to propose updating formulae for the mean and covariance matrix using variable weights. Let $Y_{j 1}, \ldots, Y_{j n(j)}$ denote the observation from the $j$ th training field, a set of identically distributed random vectors. Then the updated mean $M_{j}^{\prime}$ is given by

$$
\begin{equation*}
M_{j}^{\prime}=\gamma_{j} M_{j-1}^{\prime}+\left(1-\gamma_{j}\right) M_{j}, \tag{36}
\end{equation*}
$$

where

$$
\begin{aligned}
& M_{j}=\left(Y_{j 1}+\ldots+Y_{j n(j)}\right) / n(j) \\
& \gamma_{j}=a_{j-1} N_{j-1} / N_{j}, \\
& N_{k}=n(1)+\ldots+n(k)
\end{aligned}
$$

and $a_{j-1}$ is a number satisfying $0 \leq a_{j-1} \leq N_{j} / N_{j-1}$ and to be guessed from the nature of the data. The covariance matrix is to be updated to $K_{j}$ where

$$
\begin{aligned}
& K_{j}=\left\{N_{j} /\left(N_{j}-1\right)\right\}\left(Q_{j}^{-}-M_{j}^{-} M_{j}^{T}\right) \\
& Q_{j}^{\prime}=Y_{j-1} Q_{j-1}^{\prime}+\left(1-Y_{j-1}\right) Q_{j} \\
& Q_{j}=\sum_{i=1}^{n(j)} Y_{j i} Y_{j i}^{T} / n(j) .
\end{aligned}
$$

and

When the patterns are varying, the updated "mean" $M_{j}^{\prime}$ given by (36) is not an unbiased estimate of the current mean. It is not known in what way these updated estimates are optimal. The other disadvantage of this algorithm is that it leaves the choice of $Y_{i}$ to the grues:: of the user.

## 5. Updating By Kalian Filters

Consider the dynamic relationship

$$
\begin{align*}
& \mathrm{Y}_{\mathrm{k}}=\mathrm{H}_{\mathrm{k}} \mathrm{~S}_{\mathrm{k}}+\mathrm{W}_{\mathrm{k}}  \tag{38}\\
& \mathrm{~S}_{\mathrm{k}}=\mathrm{AS}_{\mathrm{k}-1}+\mathrm{Z}_{\mathrm{k}-1}
\end{align*}
$$

satisfied by the pol observation vectors $Y_{1}, \ldots, Y_{k}, \ldots$ and the qua state vectors $S_{1}, \ldots, S_{k}, \ldots$, where $W_{k}^{\prime} s$ are $p x 1$ random
noise vectors independently distributed as $N(0, C), Z_{k}{ }^{\prime} s$ are gui random vectors distributed independently of each other and of $W_{k}{ }^{\prime} s$ as $N(0, Q)$ and the pxq matrix $H_{k}$ and qua matrix $A$ may be time varying, but are known matrices of real numbers. It is well known (Lee [9]) that an estimate $\hat{S}_{k}$ of $S_{k}$ given $Y_{I}, \ldots, Y_{k}$ such that

$$
\begin{equation*}
E\left(\hat{S}_{k}\right)=E\left(S_{k} \mid Y_{k}, \ldots, Y_{1}\right) \tag{39}
\end{equation*}
$$

and

$$
E\left[\left(\hat{S}_{k}-S_{k}\right)^{T}\left(S_{k}-S_{k}\right)\right] \text { is minimum }
$$

can be obtained, using discrete Kalman filter, recursively as

$$
\begin{align*}
\hat{S}_{k} & =A \hat{S}_{k-1}+P_{k} H_{k}^{T} C^{-1}\left[Y_{k}-H_{k} A \hat{S}_{k-1}\right]  \tag{40}\\
P_{k} & =\left[\left(A P_{k-1} A^{T}+Q\right)^{-1}+H_{k}^{T} C^{-1} H_{k}\right]^{-1}  \tag{41}\\
& =\operatorname{Cov}\left(S_{k}, S_{k} \mid Y_{k}, \ldots, Y_{1}\right)
\end{align*}
$$

Assuming covariance matrices of all populations to be the same and not undergoing any change, Crane [7] used the estimation method by Kalman filters for simultaneously updating the means of the populations $\mathrm{I}_{1}, \ldots, \Pi_{m}$. Crane considered the following model:

$$
\begin{align*}
& \mathrm{Y}_{\mathrm{k}}=\mathrm{H}_{\mathrm{k}} \mathrm{~S}_{\mathrm{k}}+W_{k}  \tag{42}\\
& \mathrm{~S}_{\mathrm{k}}=\mathrm{S}_{\mathrm{k}-1}+\mathrm{Z}_{\mathrm{k}-1}  \tag{43}\\
& \mathrm{H}_{\mathrm{k}}=\mathrm{M}_{\mathrm{k}} \otimes \mathrm{I}_{\mathrm{p}} \tag{44}
\end{align*}
$$

where $M_{k}$ is a mol matrix with ith row ( $i=1, \ldots, m$ as 1 if $Y_{k}$ has been recognized to be coming from $I_{i}$ and zero otherwise and $A \otimes B$ denoting the Kronecker product of the matrices $A$ and $B$ (Anderson [2]). It has also been assumed that

$$
\begin{align*}
& Q=R Q \mathrm{C},  \tag{45}\\
& \text { Where } \quad \begin{aligned}
\mathrm{R} & =\mathrm{r}_{1}\left[\begin{array}{llll}
1 & r_{2} & \cdots & r_{2} \\
\mathrm{r}_{2} & 1 & \ldots & r_{2} \\
\vdots & \ddots & \vdots \\
r_{2} & & 1
\end{array}\right] .
\end{aligned} . \tag{46}
\end{align*}
$$

The state vector $S_{k}$ is a $m$ pul vector of $m$ stacked subvectors of pal vectors of present means (prior to updating). Let $X_{k}^{(i)}$ and $Z_{k}^{(i)}$ denote the th $p \times I$ vector component of $S_{k}$ and $Z_{k}$ respectively. Then assuming $Y_{k}$ to be an observation from $\Pi_{i}$, we obtain from (42), (43) and (44) that

$$
\begin{align*}
& Y_{k}=X_{k}^{(i)}+W_{k}  \tag{47}\\
& X_{k}^{(i)}+X_{k-1}^{(i)}+Z_{k-1}^{(i)} \tag{48}
\end{align*}
$$

This is the process for which EWMA s are optimal. Thus it is evident that Crane's model is a generalization of the above model so that in simultaneously updating the means of all populations their interaction has been taken into account.

Noting that

$$
\begin{aligned}
& {\left[\left(\mathrm{P}_{\mathrm{k}-1}+Q\right)^{-1}+H_{k}^{\mathrm{T}} \mathrm{C}^{-1} \mathrm{H}_{\mathrm{k}}\right]^{-1} H_{k}^{\mathrm{T}} \mathrm{C}^{-1}} \\
& \quad=\left(P_{k-1}+Q\right) H_{k}^{T}\left[H_{k}\left(P_{k-1}+Q\right) H_{k}^{T}+C\right]^{-1},
\end{aligned}
$$

we now obtain from (40) that

$$
\begin{equation*}
\hat{S}_{k}=\hat{S}_{k-1}+\left(\mathrm{P}_{k-1}+Q\right) H_{k}^{T}\left[H_{k}\left(P_{k-1}+Q\right) H_{k}^{T}+C\right]^{-1}\left(Y_{k}-H_{k} \hat{S}_{k-1}\right) \tag{49}
\end{equation*}
$$

Further assuming $L\left(S_{1}\right)=N\left(0, P_{0}\right)$ and

$$
\begin{equation*}
\underset{\operatorname{mpxmp}}{\mathrm{P}_{\mathrm{o}}}=\mathrm{T}_{\mathrm{o}} \otimes \underset{\mathrm{mxm}}{\mathrm{C}}, \tag{50}
\end{equation*}
$$

we can obtain from (45) that

$$
\begin{aligned}
& P_{1}+Q=\left(P_{o}^{-1}+H_{1}^{T} C^{-1} H_{1}\right)^{-1}+Q \\
& =\left(T_{0}^{-1}+M_{I} M_{I}^{T}\right)^{-1} \otimes C+R \otimes C
\end{aligned}
$$

$$
\begin{aligned}
& =T_{I} @ c
\end{aligned}
$$

and

$$
P_{2}+Q=\left[\left(P_{1}+Q\right)^{-1}+H_{2}^{T} c^{-1} H_{2}\right]^{-1}+R \otimes c
$$

$$
\begin{aligned}
& =\left(T_{1}^{-1}+M_{2} M_{2}^{T}\right)^{-1} \otimes c+R \otimes c \\
& =\left\{T_{1}-\frac{T_{1} M_{2} M_{2}^{T} T_{1}}{1+M_{2}^{T} T_{1} M_{2}}\right\} \otimes C+R \otimes C
\end{aligned}
$$

$$
=\mathrm{T}_{2} \otimes \mathrm{c}
$$

where $\quad T_{1}=T_{o}-\frac{T_{0} M_{1} M_{1}^{T} T_{o}}{1+M_{1}^{T_{1}} T_{0} M_{1}}+R$
and $\quad \mathrm{T}_{2}=\mathrm{T}_{1}-\frac{\mathrm{T}_{1} \mathrm{M}_{2} \mathrm{M}_{2}^{\mathrm{T}} \mathrm{T}_{1}}{1+\mathrm{M}_{2}^{\mathrm{T}} \mathrm{T}_{1} \mathrm{M}_{2}}+\mathrm{R}$.

By induction, we can obtain

$$
P_{k}+Q=T_{k} \otimes C
$$

mpxmp mpxnp mem exp
where $\quad T_{k}=T_{k-1}-\frac{T_{k-1} M_{k} M_{k}^{T} T_{k-1}}{1+M_{k}^{T} T_{k-1} M_{k}}+R$

Using (51) we can simplify (49) and obtain

$$
\begin{equation*}
\hat{S}_{k}=\hat{S}_{k-1}+\left[\left(1+M_{k}^{T} T_{k-1} M_{k}\right)^{-1} T_{k-1} M_{k} \otimes I_{p}\right]\left(Y_{k}-H_{k} \hat{S}_{k-1}\right) \tag{52}
\end{equation*}
$$

Crane has suggested some "ad hoc" value for $r_{1}$ and $r_{2}$ in (46) in order to define the interaction matrix $R$. But it seems more reasonable to estimate them from a part of the sample.

The above estimation procedure will be clear from the following numerical example.

$$
\begin{aligned}
& \text { Example. Let } m=2, p=2, T_{0}=I_{2}, C=I_{2}, r_{1}=0.01, r_{2}=0.02 \text {, } \\
& Y_{1}^{T}=\left[\begin{array}{ll}
1 & 2
\end{array}\right] \in \Pi_{1}, Y_{2}^{T}=\left[\begin{array}{ll}
-1 & 1
\end{array}\right] \in \Pi_{2} \text {. Then } \\
& M_{1}^{T}=\left[\begin{array}{ll}
1 & 0
\end{array}\right], \mathrm{m}_{2}^{\mathrm{T}}=\left[\begin{array}{ll}
0 & 1
\end{array}\right], \\
& M_{1}^{T} T_{o} M_{1}=1, \\
& T_{1}=I_{2}-\left\{I_{2}\left[\begin{array}{l}
1 \\
0
\end{array}\right]\left[\begin{array}{ll}
1 & 0
\end{array} I_{2}\right\} / 2+0.01\left[\begin{array}{ll}
1 & 0.02 \\
0.02 & 1
\end{array}\right]\right. \\
& =\left[\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right]-\left[\begin{array}{cc}
0.5 & 0 \\
0 & 0
\end{array}\right]+\left[\begin{array}{ll}
0.01 & 0.002 \\
0.002 & 0.01
\end{array}\right] \\
& =\left[\begin{array}{ll}
0.51 & 0.002 \\
0.002 & 1.01
\end{array}\right] ; \\
& \mathrm{M}_{2}^{\mathrm{T}} \mathrm{~T}_{1} \mathrm{M}_{2}=1.01, \\
& \mathrm{~T}_{2}=\mathrm{T}_{1}-\left\{\mathrm{T}_{1}\left[\begin{array}{l}
0 \\
1
\end{array}\right]\left[\begin{array}{ll}
0 & 1
\end{array}\right] \mathrm{T}_{1}\right\} / 2.01+\mathrm{R} \\
& =T_{1}-\frac{1}{2.01}\left[\begin{array}{ll}
0 & 0.002 \\
0 & 1.01
\end{array}\right]\left[\begin{array}{ll}
0.51 & 0.002 \\
0.002 & 1.01
\end{array}\right]+R .
\end{aligned}
$$

## 6. Detection of Change in Mean

The models on which we have based our estimation process so far implicitly assume that the means change almost at each observation. If this is the case, then the updating of means using the procedures so far discussed is just. But when we are considering spatial variation and
have observations from points spaced not too far from each other, then it is more likely that the sequence of observations $Y_{1}, Y_{2}, \ldots, Y_{n}$ can be grouped as $\left\{Y_{1}, Y_{2}, \ldots, Y_{k_{1}}\right\},\left\{Y_{k_{1}+1}, Y_{k_{1}+2}, \ldots, Y_{k_{2}}\right\}, \ldots$ such that members in the same group have the same mean while members in different groups have different means. In that case it will be more reasonable to use the arithmetic mean $\left(\mathrm{Y}_{\mathrm{k}_{\mathrm{i}}+1}+\ldots+\mathrm{Y}_{\mathrm{k}_{\mathrm{i}+1}}\right) /\left(\mathrm{k}_{\mathrm{i}+1}-\mathrm{k}_{\mathrm{i}}\right)$ as the mean of $Y_{j}\left(k_{i}+1 \leq j \leq k_{i+1}\right)$ than to use the mean obtained by the updating process. But often the points $K_{1}, k_{2}, \ldots$ at which the changes in mean take place is unknown. Therefore we have to use some statistical test in order to determine the points of change.

When $Y_{1}, Y_{2}, \ldots, Y_{n}$ are $p x l$ random vectors independently distribute normally with the same covariance matrix, say $I_{p}$, then for testing the hypothesis

$$
\begin{aligned}
& H_{0}: \mu_{1}=\mu_{2}=\ldots=\mu_{n} \quad \text { (equality of means) } \\
& \text { against } \quad H_{1}: \mu_{1}=\ldots=\mu_{r} \neq \mu_{r+1}=\ldots=\mu_{n},
\end{aligned}
$$

where the change point $r$ is unknown and $1 \leq r \leq N-1$, Sen and Srivastava [13] have used the test statistic U* given by

$$
U^{*}=N^{-2} \sum_{i=1}^{n-1}\left(\sum_{j=i}^{n-1}\left(Y_{j+1}-\bar{Y}\right)\right)^{T}\left(\sum_{j=i}^{n-1}\left(Y_{j+1}-\bar{Y}\right)\right),
$$

where $\quad \bar{Y}=\left(Y_{1}+\ldots+Y_{n}\right) / N$. They have also given the asymptotic percentile of $\mathrm{U*}$ for $\mathrm{p}=2$ through $\mathrm{p}=8$. Chernoff and Racks [6] have also given a test for $p=1$. For other references of work in this connedton we refer to [13].

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## Addendum to Papers 1, 4 and 5

Consider the matrix $P$ for the two crop situation. Say, $\hat{P}=\left[\begin{array}{ll}\hat{P}(1 \mid 1) & \hat{P}(1 \mid 2) \\ \hat{P}(2 \mid 1) & \hat{P}(2 \mid 2)\end{array}\right]$, where $\hat{P}(1 \mid 1)=x / N_{1} \quad \hat{P}(2 \mid 2)=y / N_{2}$ and $\hat{P}(2 \mid 1)=1-\hat{P}(1 \mid 1), \hat{P}(1 \mid 2)=1-\hat{P}(2 \mid 2) . \quad x$ is the number correctly classified into population one and $y$ is the number correctly classified into polalation two. Since $X$ and $Y$ are independently distributed as binomial variates, the probability that $\hat{\mathrm{p}}$ is singular is positive. This is illustrated for the case $N_{1}=N_{2}=3$ where $p_{i}$ represents the probability of correctly classifying an observation in population i.

Since $\hat{P}$ is singular iff $x y-(3-x)(3-y)=0$, iff $x+y=3$, there are 4 points yielding $\hat{\mathrm{P}}$ singular. They are $(1,2),(2,1),(0,3),(3,0)$ and the probability $\hat{P}$ is singular is $9 p_{1} q_{1}^{2} p_{2}^{2} q_{2}+9 p_{1}^{2} q_{1} p_{2} q_{2}^{2}+q_{1}^{3} p_{2}^{3}+p_{1}^{3} q_{2}^{3}$. In general, if $N_{1}=N_{2}$ there are $N_{2}+1$ points which yield $\hat{P}$ singular.

Hence, an alternate method must be used to estimate $\hat{P}$. One approach is to estimate $\mu_{i}, \Sigma_{i}$ for each population from training samples and then estimate $P(i \mid j)$ by $\hat{P}(i \mid j)=\int_{R_{i}} p_{j}\left(x ; \hat{\mu}_{j}, \hat{\Sigma}_{j}\right) d x$, where $R_{i}$ is the region corresponding to an observation being classified into the ith population, and $p_{j}$ is the (continuous) density function of the $j$ th population. The probability that $\hat{P}$ is non-singular with probability one can be established using advanced probabilistic arguments.

Another approach is to use the pseudoinverse of $\hat{P}$ rather than the inverse. However, this approach would affect all the analyses done and the results may be difficult to interpret. This approach is probably the best but needs to be studied to assure that subtleties have not been ignored.

Finally, one could make the analysis conditional on the fact that $\hat{P}$ is non-singular. Formulas for variances and mean squared error of P would necessarily be modified to incorporate this condition; yet the analysis would conceptually be straightforward and follow the arguments now presented.

