

Construction of a Vector Equivalent to a Given Vector from the Point of View of the Analysis of Principal Components

Yves Escoufier, P. Robert, and J. Cambon, Montpellier

I. INTRODUCTION - All random vectors considered in this paper are assumed to be centered, with elements belonging to the set $L_2(\Omega, \mathcal{A}, \mathcal{P})$ of random variables with finite variances on $(\Omega, \mathcal{A}, \mathcal{P})$.

Let X_1 and X_2 be two such vectors; let Z_1 and Z_2 be the random vectors of their principal components, respectively. In section III we study the problem of attaching weights to the elements of X_2 in such a way that the vector of principal components Z_2^* of this weighted vector be as closed as possible to Z_1 . In section IV we study ways of choosing a subvector X_2^* of X_2 , simultaneously attaching weights to the selected elements, so that the vector Z_2^* of principal components of the resulting weighted X_2^* will best approximate Z_1 , over all choices of X_2^* and all weighting rules.

When conducting a study based on principal components of a vector difficult or costly to observe, one would welcome a theoretically sound method of substituting to it a vector easier to measure and having approximately the same principal components. At this time not much is available. On the one hand, it is known that the simplest non-singular transformation of the original vector may perturb deeply the principal components; on the other hand, the results to be presented here indicate that the currently proposed techniques ([BEALE et al 1967], [JOLIFFE 1973]) may suffer from insufficient theoretical foundations.

In section II we recall essential results obtained by ESCOUFIER ([1970], [1973-a], [1973-b]) which are the basis of the solutions to the problems investigated in this paper. To comply with printing space regulations, proofs are omitted (see quoted references).

II. MATHEMATICAL BASES - Let X be a $p \times 1$ random vector and $k \leq p$ be the rank of $E(\underline{X}\underline{X}')$. It is known that there exists a $p \times p$ orthogonal matrix H^* such that

$$H^{*'} E(\underline{X}\underline{X}') H^* = \begin{pmatrix} \underline{D} & & 0 \\ & \dots & \\ 0 & & 0 \end{pmatrix}$$

where \underline{D} is a $k \times k$ diagonal matrix with positive diagonal elements. Let \underline{H} be the $p \times k$ submatrix of H^* such that $\underline{H}' E(\underline{X}\underline{X}') \underline{H} = \underline{D}$.

Definition 1. The $k \times 1$ random vector $\underline{Z} = \underline{H}'\underline{X}$ is called the vector of principal components of \underline{X} .

Definition 2. The $p \times 1$ random vector \underline{X}_1 , with the vector of principal components \underline{Z}_1 , and the $g \times 1$ random vector \underline{X}_2 , with the vector of principal components \underline{Z}_2 , are said to be equivalent if there exist a positive integer $k \leq \min(p, g)$, a constant α and a $k \times k$ matrix \underline{C} such that :

i) $\underline{C}\underline{C}' = \underline{C}'\underline{C} = \alpha \underline{I}_k$, ii) $\underline{Z}_1 = \underline{C}'\underline{Z}_2$.
(\underline{I}_k is the $k \times k$ identity matrix).

It is easily verified that the relation established in Definition 2 is an equivalence relation. Its value, in our context, stems from the following properties [ESCOUFIER, 1973-b] :

Proposition 1. If \underline{X}_1 and \underline{X}_2 are equivalent vectors, then

i) for each diagonal element $\delta_i^{(1)}$ of $E(\underline{Z}_1 \underline{Z}_1')$ there is a diagonal element $\delta_j^{(2)}$ of $E(\underline{Z}_2 \underline{Z}_2')$ such that $\delta_j^{(2)} = \alpha \delta_i^{(1)}$;

ii) there exist permutation matrices \underline{P} and \underline{Q} such that

$$\underline{Q} \underline{C} \underline{P} = \begin{vmatrix} A_1 & 0 & 0 & \dots & 0 \\ 0 & A_2 & 0 & \dots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \dots & \dots & A_r \end{vmatrix}$$

where r is the number of distinct diagonal elements of $E(\underline{Z}_2 \underline{Z}_2')$;

iii) if all the diagonal elements of $E(\underline{Z}_2 \underline{Z}_2')$ are distinct, there is a permutation σ of the indices $1, 2, \dots, k$ for which $Z_{1i} = \sqrt{\alpha} Z_{2, \sigma(i)}$.

Proposition 1 should make clear what is to be understood by the "equivalence" of \underline{X}_1 and \underline{X}_2 .

The next step is to find an easily computable criterion for the equivalence of two vectors. ESCOUFIER [1973-a] associates to a vector \underline{X} a linear transformation $U_{\underline{X}}$ defined on $L_2(\Omega, \mathcal{A}, \mathcal{P})$ by

$$U_{\underline{X}}(Y) = \sum_{i=1}^p [E(X_i Y) \cdot X_i] \text{ for all } Y \in L_2(\Omega, \mathcal{A}, \mathcal{P}).$$

The following properties are then established :

a)- the set of operators $U_{\underline{X}}$ is a subset of the class of Hilbert-Schmidt operators on L_2 and so is a Hilbert space ;

b)- if $\underline{\Sigma} = \begin{pmatrix} \underline{\Sigma}_{11} & \underline{\Sigma}_{12} \\ \underline{\Sigma}_{21} & \underline{\Sigma}_{22} \end{pmatrix}$ is the covariance

matrix of the vector $\underline{X} = \begin{pmatrix} X_1 \\ X_2 \end{pmatrix}$, the scalar product for the Hilbert space considered is given by

$$\langle U_{\underline{X}_1}, U_{\underline{X}_2} \rangle = \text{Tr} (\underline{\Sigma}_{12} \underline{\Sigma}_{21}) ;$$

c)- if Z_k is a principal component of \underline{X} associated with the eigenvalue λ_k , then

$$U_{\underline{X}} (Z_k) = \lambda_k Z_k .$$

From the above, the following practical criterion can be deduced [ESCOUFIER, 1973-b] :

Proposition 2. The vectors \underline{X}_1 and \underline{X}_2 are equivalent in the sense of Definition 2 if and only if the coefficient

$$RV(\underline{X}_1, \underline{X}_2) = \text{Tr} (\underline{\Sigma}_{12} \underline{\Sigma}_{21}) / [\text{Tr} (\underline{\Sigma}_{11}) \cdot \text{Tr} (\underline{\Sigma}_{22})]^{1/2}$$

is equal to 1.

III. THE FIRST PROBLEM - From the last proposition it can be seen that the first problem can be restated as :

- Given vectors \underline{Y} and \underline{X} find a diagonal matrix $\underline{\Delta}$ so as to maximize $RV(\underline{Y}, \underline{\Delta}\underline{X})$, where

$$RV(\underline{Y}, \underline{\Delta}\underline{X}) = \text{Tr} [\underline{\Delta} E(\underline{X}\underline{Y}') E(\underline{Y}\underline{X}') \underline{\Delta}] / \{ \text{Tr} [E(\underline{Y}\underline{Y}')]^2 \cdot \text{Tr} [\underline{\Delta} E(\underline{X}\underline{X}') \underline{\Delta}]^2 \}^{1/2} ;$$

Hence look for $\underline{\Delta}$ to maximize

$$\text{Tr} [\underline{\Delta} E(\underline{X}\underline{Y}') E(\underline{Y}\underline{X}') \underline{\Delta}] / \{ \text{Tr} [\underline{\Delta} E(\underline{X}\underline{X}') \underline{\Delta}]^2 \}^{1/2} .$$

Denote by m_j the j -th diagonal element of $\underline{\Delta}$, by v_{ij} the elements of $E(\underline{X}\underline{X}')$, by u_{ij} the elements of $E(\underline{Y}\underline{X}')$ and define the vectors \underline{U} , \underline{M} and the matrix \underline{A} by

$$\underline{U} = (U_j) \text{ where } U_j = \sum_i u_{ij} ,$$

$$\underline{M} = (M_j) \text{ where } M_j = m_j^2 ,$$

$$\underline{A} = (A_{ij}) \text{ where } A_{ij} = v_{ij} .$$

By simply expanding the traces it can be verified that

$$\text{Tr} [\underline{\Delta} E(\underline{X}\underline{Y}') E(\underline{Y}\underline{X}') \underline{\Delta}] = \underline{U}' \underline{M} ,$$

$$\text{Tr} [\underline{\Delta} E(\underline{X}\underline{X}') \underline{\Delta}]^2 = \underline{M}' \underline{A} \underline{M} .$$

Thus we are lead to the quadratic programming problem for \underline{M} :

$$\text{Problem P1} \quad \begin{cases} \text{Minimize } \underline{M}'\underline{A}\underline{M} \\ \text{Subject to } \underline{U}'\underline{M} = 1 \text{ and } \underline{M} \geq 0. \end{cases}$$

Note that if $E(\underline{X}\underline{X}')$ is positive semi-definite, so is A [GOWER, 1971]. In such a case, the function $(\underline{M}'\underline{A}\underline{M})$ is convex on the convex set $(\underline{U}'\underline{M} = 1 \text{ and } \underline{M} \geq 0)$.

It was found convenient to substitute to Problem P1 on the unknown \underline{M} the following equivalent problem on the unknown vector \underline{N} :

$$\text{Problem P2} \quad \begin{cases} \text{Minimize } \underline{N}' \underline{C} \underline{N}, \\ \text{Subject to } \underline{e}'\underline{N} = 1 \text{ and } \underline{N} \geq 0, \end{cases}$$

where $\underline{e}' = (1 \ 1 \dots \ 1)$ and where \underline{N} and \underline{C} are defined by :

$$\underline{N} = (N_j) \text{ where } N_j = U_j M_j = U_j m_j^2,$$

$$\underline{C} = (C_{ij}) \text{ where } C_{ij} = \frac{A_{ij}}{U_i U_j} = \frac{v_{ij}^2}{U_i U_j}.$$

The authors have solved Problem P2 for many sets of statistical data. Numerical evidence on these test cases will not be given here but reported on elsewhere [CAMBON, 1974].

To be noted is the fact that the constraint $\underline{e}'\underline{N} = 1$ will force most of the currently proposed quadratic programming algorithms into the difficulty of a degenerated case. For this reason the authors have developed a specific algorithm which will now be justified and summarized.

\underline{C} being $(n \times n)$, denote by I a subset of $\{1, 2, \dots, n\}$ and by J the complement of I . Denote by \underline{C}_I the submatrix of \underline{C} formed by the C_{ij} 's with $i \in I, j \in I$; by \underline{C}_J the submatrix of \underline{C} formed by the C_{ij} 's with $i \in J, j \in I$. From the Kuhn-Tucker conditions for Problem P2, it can be proved that :

Proposition 3. Let $\underline{N}_I = \underline{C}_I^{-1} \underline{e}$. If

$$\underline{N}_I \geq 0 \text{ and } \underline{C}_J \underline{N}_I \geq \underline{e}, \quad (1)$$

then the vector $\underline{N} = (N_i)$:

$$N_i = \begin{cases} \frac{1}{\underline{e}'\underline{N}_I} N_{II} & \text{if } i \in I, \\ 0 & \text{if } i \in J, \end{cases} \quad (2)$$

is an optimal solution of Problem P2, and $\underline{N}'\underline{C}\underline{N} = (\underline{e}'\underline{N}_I)^{-1}$.

A technique to find an optimal solution would be to start computing $\underline{N}_I = \underline{C}_I^{-1} \underline{e}$ for all subsets of indices I and to retain the first found to satisfy (1). The number of sets I to be considered can be greatly restricted in accordance with the following propositions.

Proposition 4. Let $\underline{P} = (P_i) = \frac{\underline{C}^{-1} \underline{e}}{\underline{e}' \underline{C}^{-1} \underline{e}}$, $I^* = \{i: P_i \geq 0\}$, $J^* = \{i: P_i < 0\}$. There is an optimal solution $\underline{N} = (N_i)$ of Problem P2 such that $N_j = 0$ for at least one index $j \in J^*$.

Proof. It is known that \underline{P} is an optimal solution of P2 without the positivity constraint. Consider any feasible vector $\underline{Q} = (Q_i)$ for P2 such that $Q_j > 0$ for all $j \in J^*$. Let $r \in J^*$ be such that $Q_r/P_r \geq Q_j/P_j$ for all $j \in J^*$, $k = (Q_r/P_r)/(1-Q_r/P_r)^{-1}$, and $\underline{Q}^k = (1-k)\underline{P} + k\underline{Q}$. Since the function F to be minimized is convex and $0 \leq k \leq 1$, $F(\underline{Q}^k) \leq F(\underline{Q})$. But \underline{Q}^k is feasible for P2 and $Q_r^k = 0$. The conclusion follows. ■

Proposition 5. An optimal solution of Problem P2 can be found as the best of the optimal solutions of the $|J^*|$ (n-1)-dimensional problems on \underline{N} :

$$\begin{cases} \text{minimize } \underline{N}' \underline{C} (\underline{n} - j) \underline{N}, \\ \text{subject to } \underline{e}' \underline{N} = 1, \underline{N} \geq 0, \end{cases} \quad (3)$$

where $\underline{n} = \{1, 2, \dots, n\}$ and where j spans J^* .

Proof. From Proposition 4, it is seen that one can find an optimal solution of P2 by looking at the solutions of each of the $|J^*|$ problems:

$$\begin{cases} \text{minimize } \underline{N}' \underline{C} \underline{N} \\ \text{subject to } \underline{e}' \underline{N} = 1, \underline{N} \geq 0 \\ \text{and the additional constraint } N_j = 0, \end{cases}$$

for $j \in J^*$. For each j, this last problem is equivalent to (3). ■

Clearly (3) is a problem of type P2 with dimensions reduced by 1. We may therefore repeatedly invoke the above proposition to justify the following algorithm. A queue Q is built which elements are sets of indices I for which $\underline{N}_I = \underline{C}_I^{-1} \underline{e}$ has to be calculated in accordance with Proposition 3. (We recall that in a "queue", elements are added "at the bottom" and removed "from the to

the second element moving up at the top).

- 1 - Set the queue \mathcal{Q} to the single element $\{1, 2, \dots, n\}$ and $\alpha = 0$;
- 2 - If \mathcal{Q} is empty, print error diagnostic and stop ; otherwise, select and remove the first element I . Compute $N_I = \underline{C}_I^{-1} \underline{e}$;
- 3 - Let $K = \{k : N_{Ik} < 0\}$. If K is empty, go to Step 5 ; otherwise :
- 4 - For each $k \in K$, if $(I - \{k\})$ is not a subset of an element of \mathcal{Q} , then add $(I - \{k\})$ to \mathcal{Q} . Go to Step 2 ;
- 5 - If $\alpha \geq \underline{e}' N_I$, go to Step 2 ; otherwise, set $\alpha = \underline{e}' N_I$;
- 6 - If $\underline{C}_{j \sim I} N_I > \underline{e}$ (see (1)), set N_i as per (2) and stop ; otherwise go to Step 2.

The numerical efficiency of the algorithm will depend greatly on the particular code used to compute $\underline{C}_I^{-1} \underline{e}$. It will be seen in Section IV that, by using the Choleski decomposition of \underline{C}_I^{-1} , one can make the test $\underline{C}_{j \sim I} N_I \geq \underline{e}$ of step 6 an immediate byproduct of step 2.

IV. THE SECOND PROBLEM - Let \underline{Y} and \underline{X} be two vectors, possibly equal. Assume \underline{X} to be n -dimensional and for $k \leq n$ denote by \mathcal{V}_k the set of all k -dimensional subvectors of \underline{X} . If I is a set of k distinct indices, we shall denote by \underline{X}_I the subvector of \underline{X} retaining those elements of \underline{X} with indices in I and by $\widehat{\underline{Y}}(\underline{X}_I)$ the vector $(\underline{\Delta} \underline{X}_I)$ solution of the first problem studied in Section III. Thus this second problem can be stated as :

- Determine the vector $\underline{X}_I \in \mathcal{V}_k$ which maximizes
- $$RV [\underline{Y}, \widehat{\underline{Y}}(\underline{X}_I)] \text{ over } \mathcal{V}_k.$$

When n and k are large, it becomes unpractical to try all k -dimensional subvectors of \underline{X} and for this reason we propose a sequential algorithm analogous to the one used in step-wise regression.

From the analysis carried in Section III, it is easily seen that for a given set I the vector \underline{X}_I is feasible for the present problems if and only if the vector $N_I = \underline{C}_I^{-1} \underline{e}$ is non negative and that the coefficient $RV [\underline{Y}, \widehat{\underline{Y}}(\underline{X}_I)]^2$ is then proportional to $\alpha_I = \underline{e}' \underline{C}_I^{-1} \underline{e}$. Thus our problem becomes that of maximizing.

$\alpha_I = \underline{e}' \underline{C}_I^{-1} \underline{e}$ subject to $\underline{C}_I^{-1} \underline{e} \geq 0$,
over all subsets I of k distinct indices.

Suppose now that I is a subset of only $(k-1)$ indices which is known to be feasible (i.e. $\underline{N}_I = \underline{C}_I^{-1} \underline{e} \geq 0$). The step-wise technique consists in the determination of that index $j \notin I$ such that, with $(I, j) = I \cup \{j\}$, $\underline{N}_{(I, j)}$ will be feasible and $\alpha_{(I, j)}$ will be maximized (over all choices of j).

The technicalities will now be described, assuming for simplicity that the set I contains the first $(k-1)$ indices :
 $I = \{1, 2, \dots, k-1\}$.

Consider the matrix

$$\underline{C}^+ = \begin{bmatrix} & & & & 1 \\ & & & & 1 \\ & & & & \vdots \\ \underline{C} & & & & \\ & & & & 1 \\ & & & & 1 \\ & & & & 1 \end{bmatrix}$$

where \underline{C} is given by (3), and let $\underline{C}^+ = \underline{R} \underline{R}'$ be the Choleski decomposition [FORSYTHE, MOLER, 1967] of \underline{C}^+ (which is positive semi-definite), i.e. \underline{R} is upper-triangular. We assume that the choice of the elements of the set I has been made sequentially by this same algorithm and that the first $(k-1)$ steps of the Choleski decomposition of \underline{C}^+ have been carried, column-wise, leading to the $(n+1) \times (k-1)$ matrix

$$\begin{bmatrix} \underline{R}_I \\ \hline \underline{r}'_{j \bullet} \\ \vdots \\ \underline{r}'_{(n+1) \bullet} \end{bmatrix} \quad (4)$$

where $\underline{r}'_{j \bullet} = (r_{j1} \ r_{j2} \ \dots \ r_{j(k-1)})$ for $j \geq k$ and

$$\underline{C}_I = \underline{R}_I \underline{R}'_I$$

$$\underline{R}_I \underline{r}'_{(n+1) \bullet} = \underline{e}$$

For $j \notin I$ ($j \neq n+1$), define

$$w_j = C_{jj} - \sum_{i=1}^{k-1} r_{ji}^2 \quad (5)$$

$$s_j = 1 - \sum_{i=1}^{k-1} r_{ji} r_{(n+1)i}$$

($s_j = 1 - \underline{r}'_{j \bullet} \underline{R}_I^{-1} \underline{e}$). If the index j is to be added to the set I

with the feasible vector $\underline{X}_{(I,j)}$, we must have the corresponding matrix $\underline{C}_{(I,j)}$ non-singular, hence $w_j > 0$. The Choleski decomposition $\underline{C}_{(I,j)} = \underline{R}_{(I,j)} \underline{R}'_{(I,j)}$ would then give:

$$\underline{R}_{(I,j)}^{-1} = \left[\begin{array}{c|c} \underline{R}_I^{-1} & 0 \\ \hline \frac{-1}{\sqrt{w_j}} \underline{r}'_{j \bullet} \underline{R}_I^{-1} & \frac{1}{\sqrt{w_j}} \end{array} \right], \quad \underline{R}_{(I,j)}^{-1} \underline{e} = \left[\begin{array}{c} \underline{R}_I^{-1} \underline{e} \\ \hline s_j / \sqrt{w_j} \end{array} \right]$$

The coefficient RV $[\underline{X}, \widehat{\underline{V}}(\underline{X}_{(I,j)})]^2$ will be proportional to

$$\alpha_{(I,j)} = \underline{e}' \underline{R}_I^{-1} \underline{R}_I^{-1} \underline{e} + (s_j^2/w_j) = \alpha_I + (s_j^2/w_j) \quad (6)$$

The feasibility condition of non-negativity of

$\underline{N}_{(I,j)} = \underline{C}_{(I,j)}^{-1} \underline{e}$ must be respected. It can be verified by substitution of the above relations that

$$\underline{N}_{(I,j)} = \underline{R}_{(I,j)}^{-1} \underline{R}_{(I,j)}^{-1} \underline{e} = \left[\begin{array}{c} \underline{N}_I - \frac{s_j}{w_j} \underline{Z}_j \\ \hline \frac{s_j}{w_j} \end{array} \right]$$

where \underline{Z}_j is the solution of the triangular system $\underline{R}'_I \underline{Z}_j = \underline{r}_{j \bullet}$.

We can now summarize one basic cycle of the algorithm assuming that k is given, that I is a set of indices with less than k elements and that the corresponding matrix (4) is available, as s_j and w_j for all $j \in I$:

- 1 - Define $J = \{j : j \notin I, w_j > 0 \text{ and } s_j > 0\}$;
- 2 - If J is empty, stop. Otherwise select $j_0 \in J$ which maximizes (s_j^2/w_j) (see (6));
- 3 - Solve for \underline{Z}_{j_0} the triangular system $\underline{R}'_I \underline{Z}_{j_0} = \underline{r}_{j_0 \bullet}$;
- 4 - If $\underline{N}_{(I,j_0)} = \underline{N}_I - (s_{j_0}/w_{j_0}) \underline{Z}_{j_0}$ is non-negative, then accept j set $I := I \cup \{j_0\}$, $\underline{N}_I := \underline{N}_{(I,j_0)}$ and go to Step 5. Otherwise set $J := J - \{j_0\}$ and repeat Step 2;
- 5 - Carry one more step of the Choleski decomposition of \underline{C}^+ , computing the column corresponding to column " j_0 " of \underline{C}^+ (to obtain $\underline{R}_{(I,j_0)}$) and, for $j \in J$, set ((5)):

$$s_j := s_j - r_{(n+1)j_0} r_{jj_0}, \quad w_j := w_j - r_{jj_0}^2;$$

6 - If $k_0 < k$, set $k_0 := k_0 + 1$ and repeat a complete cycle from Step 1. Otherwise, stop.

Note that a "stop" a Step 2 implies that the set of indices already chosen cannot be enlarged by addition of a single element so as to increase the RV coefficient.

To initiate the algorithm one may choose any one element X_i of \underline{X} and set $I = \{i\}$. The square of the RV coefficient is then proportional to C_{ii}^{-1} ; this suggests initiation with that index for which the diagonal element of \underline{C} is minimum.

It must be made clear that this algorithm does not necessarily produce the optimal solution. As in the step-wise regression technique, it produces the best solution of order k knowing which solution has been selected to the order $(k-1)$.

Subject to this conditionality restriction upon convergence to the optimal solution, a count of the number of operations for a k -th cycle shows the algorithm to be quite efficient.

We conclude the study of this second problem by pointing out that the first problem (Section III) can be interpreted as a special case of the second: the case where $k = n$. Thus the algorithm of this section could be used in an attempt to solve Problem P2. The following proposition, deduced from the Kuhn-Tucker conditions for Problem P2 and the above analysis, provides a criterion to detect the optimal solution:

Proposition 6. A solution produced by the algorithm of this Section is optimal for Problem P2 of Section III if and only if, at stop,

$$s_j < 0 \quad \text{for all } j \in J$$

or all n variables have been accepted (The proof is omitted).

CONCLUSION. - The theory summarized in Section II has shed new lights on two practical problems of interest to the applied statistician. Arguments on convexity and a factorization of the matrices involved have led to algorithms to solve those problems. A deeper numerical analysis of the problems, particularly the first one, should lead to more efficient algorithms and computer programs.

REFERENCES

- BEALE E.M.L., KENDALL M.G. and MANN D.W. (1967) - The Discarding of Variables in Multivariate Analysis - *Biometrika*, 54, 3 and 4, p. 357-365.
- CAMBON J. (1974) - Vecteurs équivalents à un autre au sens des composantes principales : Applications hydroliques - Note du laboratoire d'Hydrologie et d'Aménagement des eaux , USTL-Montpellier n°13-7.
- ESCOUFIER Y. (1970) - Echantillonnage dans une population de variables aléatoires réelles - *Publ. Inst. Stat. Univ. Paris* - XIX-4 p. 1 à 4
- ESCOUFIER Y. (1973-a) - Le traitement des variables vectorielles - *Biometrics* XXIX, p. 751-760.
- ESCOUFIER Y. (1973-b) - Vecteurs aléatoires équivalents du point de vue de l'analyse en composantes principales. *Rap. Techn. 7301* , *Cen de Recherche en Informatique et Gestion, Av. d'Occitanie* , 34000-MONTPPELLIER.
- FORSYTHE G., MOLER C.B. (1967) - *Computer Solution of Linear Algebraic Systems* - Prentice Hall.
- GOWER J.C. (1971) - A General Coefficient of Similarity and Some of its Properties - *Biometrics*, XXVII, p. 857-874.
- JOLIFFE I.T. (1973) - Discarding Variables in a principal Component Analysis I : Artificial Data. *Applied Statistics*, Vol. 21 n° 2 , p. 160-173.